

Semester I

Semester 1

Subject 1

ME010101 Abstract Algebra

1.1 Introduction to Abstract Algebra

Definitions 1.1.1. Set-theoretical foundation of Abstract Algebra

- A **cartesian product**, $A \times B = \{(a, b) : a \in A, b \in B\}$
- A **relation** on a set A is a subset of $A \times A$.
- A **function** f from A into B , $f : A \rightarrow B$ is a relation such that *every element of A is related to some unique element of B* (well defined).

Definitions 1.1.2. Functions : A **binary operation** on a set A is a function $* : A \times A \rightarrow A$.

“A binary operation on a set A gives an algebra on A .”

Abstract Algebra It is the study of algebraic structures. We are interested in a few algebraic structures : 1. Group 2. Ring 3. Integral Domain 4. Field

Definitions 1.1.3. A **binary algebraic structure** $\langle G, * \rangle$ is a set G together with a binary operation $*$.

Definitions 1.1.4 (Group). A set G , closed under a binary relation $*$ satisfying the following three axioms -G1, G2, & G3 is a group.

G1 Associativity

For any three elements $a, b, c \in G$, $a * (b * c) = (a * b) * c$.

G2 Identity element

There exists a unique element $e \in G$ such that $a * e = a = e * a$ for any element $a \in G$.

G3 Inverse elements

For any element $a \in G$ there exists a unique element a^{-1} such that $a * a^{-1} = e = a^{-1} * a$.

Definitions 1.1.5. Group terminologies :

- A group is **abelian** if the binary operation is commutative. $a * b = b * a$
- The **order** of a group G , $*$ is the number of elements in G .

Definitions 1.1.6. $\mathbb{Z}_n = \{0, 1, 2, \dots, n-1\}$

Remark. Consider $3, 4 \in \mathbb{Z}_5$, $3 * 4 = 2 = 4 * 3$ since $7 \cong 2 \pmod{5}$. $\langle \mathbb{Z}_5, +_5 \rangle$ is an abelian group of order 5.

Definitions 1.1.7. Homomorphism & Isomorphism

- A function $\phi : A \rightarrow B$ is a **homomorphism** if for any two elements $x, y \in A$, $\phi(xy) = \phi(x)\phi(y)$
- A function $\phi : A \rightarrow B$ is an **isomorphism** if ϕ is a bijective, homomorphism. If two binary structures are isomorphic, then they have the same (algebraic) structure.

Definitions 1.1.8. Subgroup

- A subset H of a group $\langle G, * \rangle$ is a **subgroup** of G if H is group with the same binary operation $*$. And is denoted by $H \leq G$.
- G is the **improper** subgroup of G and every other subgroup is **proper**.S5.5
- $\{e\}$ is the **trivial** subgroup of G and every other subgroup is **non-trivial**.
- The **subgroup generated by** $g \in G$ is the subgroup $\{g^n : n \in \mathbb{Z}\}$.
- The **order** of an element g is order of the subgroup generated by g .
- An element $g \in G$ is a **generator** of G if g generates G .
- A group is **cyclic** if it has a generator.

Remark. Cyclic Groups :

- Cyclic groups are abelian.
- Subgroups of cyclic groups are cyclic.

1.1.1 Some Proof Techniques

Equality of two Sets $A = B \iff A \subset B$ and $B \subset A$

If $x \in A \implies x \in B$, then $A \subset B$

Uniqueness Suppose there are two elements that qualify our conditions. We show (using the conditions) that they are the same, that is, unique.

For example, $3 + a = \pi$. Suppose $a = x, y$. Then $3 + x = 3 + y \implies x = y$, provided that the values of a comes from a set in which left cancelation law can be applied. Then a is unique.

Remember : We usually don't care to show what this unique element is. It may be also be the case that there is no such element, that is, proof of uniqueness doesn't imply existence.

Existence There are constructive and non-constructive proofs for existence problems. Suppose we want to prove that $a * b$ has an inverse element. We know that a, b has inverse elements a^{-1}, b^{-1} . From those elements, we construct an element $b^{-1} * a^{-1}$ which is an inverse of $a * b$ by construction.

And we may also prove existence without actually giving an object. Suppose we want to prove that $x^y \in \mathbb{Q}$ for some irrational numbers x and y . We know that $\sqrt{2} \notin \mathbb{Q}$. Then, $\sqrt{2}^{\sqrt{2}}$ is either rational or irrational. Suppose it is irrational, then $\left(\sqrt{2}^{\sqrt{2}}\right)^{\sqrt{2}} = 2$ is rational. Thus the proof is complete, but we are yet to know whether $\sqrt{2}^{\sqrt{2}}$ is an irrational or rational.

1.2 Direct Products and Finitely Generated Abelian Groups

Definitions 1.2.1 (Cartesian product of sets). Let S_1, S_2, \dots, S_n be a sets. Their cartesian product,

$$S_1 \times S_2 \times \dots \times S_n = \prod_{i=1}^n S_i = \{(a_1, a_2, \dots, a_n) : a_i \in S_i\} \quad (1.1)$$

For example, $\{A, B, C\} \times \{1, 2\} = \{(A, 1), (A, 2), (B, 1), (B, 2), (C, 1), (C, 2)\}$.

Question 1. How many elements in $\mathbb{Z}_3 \times \mathbb{Z}_{10} \times \mathbb{Z}_9$?

Theorem 1.2.1 (Direct product of Groups). Let G_1, G_2, \dots, G_n be groups. Then their cartesian product is a group with the binary operation $*$,

$$(a_1, a_2, \dots, a_n) * (b_1, b_2, \dots, b_n) = (a_1 * b_1, a_2 * b_2, \dots, a_n * b_n) \quad (1.2)$$

where the binary operation in $a_i * b_i$ is the binary operation of the group G_i .

Proof. $\prod_{i=1}^n G_i$ is a group if it satisfies the group axioms.

G1 Associativity

$$\begin{aligned} & (a_1, a_2, \dots, a_n)((b_1, b_2, \dots, b_n)(c_1, c_2, \dots, c_n)) \\ &= (a_1, a_2, \dots, a_n)(b_1 c_1, b_2 c_2, \dots, b_n c_n) \\ &= (a_1(b_1 c_1), a_2(b_2 c_2), \dots, a_n(b_n c_n)) \\ &= ((a_1 b_1) c_1, (a_2 b_2) c_2, \dots, (a_n b_n) c_n) \\ &= (a_1 b_1, a_2 b_2, \dots, a_n b_n)(c_1, c_2, \dots, c_n) \\ &= ((a_1, a_2, \dots, a_n)(b_1, b_2, \dots, b_n))(c_1, c_2, \dots, c_n) \end{aligned}$$

G2 Existence of a unique identity element in $\prod_{i=1}^n G_i$

Let e_i be the identity element in G_i . Then (e_1, e_2, \dots, e_n) is the identity

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element in $\prod_{i=1}^n G_i$.

$$\begin{aligned} (a_1, a_2, \dots, a_n)(e_1, e_2, \dots, e_n) \\ = (a_1 e_1, a_2 e_2, \dots, a_n e_n) \\ = (a_1, a_2, \dots, a_n) \end{aligned}$$

G3 Existence of unique inverse element for each element in $\prod_{i=1}^n G_i$

Let (a_1, a_2, \dots, a_n) be in $\prod_{i=1}^n G_i$. Then it has the inverse element $(a_1^{-1}, a_2^{-1}, \dots, a_n^{-1})$ in $\prod_{i=1}^n G_i$.

$$\begin{aligned} (a_1, a_2, \dots, a_n)(a_1^{-1}, a_2^{-1}, \dots, a_n^{-1}) \\ = (a_1 a_1^{-1}, a_2 a_2^{-1}, \dots, a_n a_n^{-1}) \\ = (e_1, e_2, \dots, e_n) \end{aligned}$$

□

Remark. We usually write ab instead of $a * b$ and relevant binary operations are used in different contexts. Student should be able to recognise the difference from the context.

Remark. $\mathbb{Z}_n = \{0, 1, \dots, (n-1)\}$ is a group with $+_n$. (addition modulo n)

For example, Consider $(1, 2) \in \mathbb{Z}_2 \times \mathbb{Z}_3$. We have, $(1, 2) + (1, 2) = (0, 1)$ since $1 +_2 1 = 0$ and $2 +_3 2 = 1$.

Definitions 1.2.2. Suppose all the groups G_i are abelian. Then $\prod_{i=1}^n G_i$ is the direct sum of the groups G_i . And is represented by $\oplus_{i=1}^n G_i$.

Theorem 1.2.2. $\mathbb{Z}_m \times \mathbb{Z}_n$ is cyclic and is isomorphic to \mathbb{Z}_{mn} if and only if m and n are relatively prime.

Proof. Sufficient part : Consider the cyclic subgroup¹ H generated by $(1, 1) \in \mathbb{Z}_m \times \mathbb{Z}_n$. It is enough to prove that the order of this cyclic subgroup H is mn . The order of H is the smallest power of $(1, 1)$ that gives the identity $(0, 0)$. The first component gives 0 for multiples of m . And the second component gives 0 for multiples of n . Since m, n are relatively prime, mn is the smallest power of $(1, 1)$ that will give $(0, 0)$. Thus $\mathbb{Z}_m \times \mathbb{Z}_n = H$ and is cyclic. Every cyclic group of order mn is isomorphic to \mathbb{Z}_{mn} . Therefore, $\mathbb{Z}_m \times \mathbb{Z}_n \approx \mathbb{Z}_{mn}$.

Necessary part : Suppose $\gcd(m, n) = d > 1$. Then mn/d is the smallest integer divisible by both m and n . Consider $(r, s) \in \mathbb{Z}_m \times \mathbb{Z}_n$. r gives 0 in mn/d since it is a multiple of m . Similarly, s gives 0 in mn/d since it is a multiple of n . Thus, $\frac{mn}{d}(r, s) = (0, 0)$. And the cyclic group generated by any element of $\mathbb{Z}_m \times \mathbb{Z}_n$ is a proper subgroup. Therefore $\mathbb{Z}_m \times \mathbb{Z}_n$ has no generators and it is not cyclic. □

¹ $(1, 1) \in \mathbb{Z}_m \times \mathbb{Z}_n$. The cyclic group generated by $(1, 1)$ has all its elements in $\mathbb{Z}_m \times \mathbb{Z}_n$. And therefore, it is a subgroup of $\mathbb{Z}_m \times \mathbb{Z}_n$

Corollary 1.2.2.1. $\prod_{i=1}^n \mathbb{Z}_{m_i}$ is cyclic and is isomorphic to $\mathbb{Z}_{m_1 m_2 \dots m_n}$ if and only if any two of the numbers m_i are relatively prime.

Question 2. Prove : For any non-negative integer n , there exists a cyclic group of order n , which is unique upto isomorphism.

Theorem 1.2.3. Let $(a_1, a_2, \dots, a_n) \in \prod_{i=1}^n G_i$. And a_i are of finite order r_i in G_i . Then the order of $\prod_{i=1}^n G_i$ is the least common multiple of r_i s.

Proof. Least common multiple of r_i s is the smallest positive integer d which is a multiple of all r_i s. For each i , the r_i th multiple of a_i gives 0 (identity). Thus, the order of the cyclic subgroup generated by (a_1, a_2, \dots, a_n) is the least common multiple of all the r_i s. \square

Remark. Consider $(3, 6, 12, 16) \in \mathbb{Z}_4 \times \mathbb{Z}_{12} \times \mathbb{Z}_{20} \times \mathbb{Z}_{24}$. Order of $3 \in \mathbb{Z}_4$ is $4/\gcd(3, 4) = 4$ ie, $\langle 3 \rangle = \{3, 2, 1, 0\}$
Order of $6 \in \mathbb{Z}_{12}$ is $12/\gcd(6, 12) = 2$ ie, $\langle 6 \rangle = \{6, 0\}$
Order of $12 \in \mathbb{Z}_{20}$ is $20/\gcd(12, 20) = 5$ ie, $\langle 12 \rangle = \{12, 4, 16, 8, 0\}$
Order of $16 \in \mathbb{Z}_{24}$ is $24/\gcd(16, 24) = 3$ ie, $\langle 16 \rangle = \{16, 8, 0\}$
Order of $(3, 6, 12, 16)$ is $\text{lcm}(4, 2, 5, 3) = 2^2 \cdot 3 \cdot 5 = 60$.

Remark. Define $\overline{G_i} = \{(e_1, e_2, \dots, e_{i-1}, a_i, e_{i+1}, \dots, e_n) : a_i \in G_i\}$. Then $G_i \approx \overline{G_i}$. And $\prod_{i=1}^n G_i$ is the internal direct product of $\overline{G_i}$ s.

For example, $\mathbb{Z}_2 \times \mathbb{Z}_3 \approx (\mathbb{Z}_2 \times \{0\}) \otimes (\{0\} \times \mathbb{Z}_3)$

Question 3. Internal direct product form of $\mathbb{Z}_{12} \times \mathbb{Z}_{60} \times \mathbb{Z}_{24}$?

1.3 Fundamental Theorem

Definitions 1.3.1. A group G is **finitely generated** if G has a finite subset that generates G .

Theorem 1.3.1 (fundamental theorem of finitely generated abelian groups). *Every finitely generated abelian group G is isomorphic to a direct product of cyclic groups in the form*

$$\mathbb{Z}_{p_1^{r_1}} \times \mathbb{Z}_{p_2^{r_2}} \times \dots \times \mathbb{Z}_{p_n^{r_n}} \times \mathbb{Z} \times \mathbb{Z} \times \dots \times \mathbb{Z} \quad (1.3)$$

where p_i are primes, not necessarily distinct and r_i are positive integers. The direct product is unique, except for the possible rearrangement of the factors.

Proof. —proof is omitted— \square

For example, $G = \mathbb{Z}_{20} \times \mathbb{Z} \times \mathbb{Z}_{15} \times \mathbb{Z} \approx \mathbb{Z}_{2^2} \times \mathbb{Z}_3 \times \mathbb{Z}_5 \times \mathbb{Z}_5 \times \mathbb{Z} \times \mathbb{Z}$. In the above case, Betti number of G is 2 (number of \mathbb{Z} factors). For any finite abelian group, Betti number is 0.

Remark (finite abelian groups). Every finite group is finitely generated. And thus we can enumerate finite abelian group of any order.

Remark. There are precisely 6 different abelian groups of order $360 = 2^3 3^2 5$.

1. $\mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_5$
2. $\mathbb{Z}_2 \times \mathbb{Z}_{2^2} \times \mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_5$
3. $\mathbb{Z}_{2^3} \times \mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_5$
4. $\mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_{3^2} \times \mathbb{Z}_5$
5. $\mathbb{Z}_2 \times \mathbb{Z}_{2^2} \times \mathbb{Z}_{3^2} \times \mathbb{Z}_5$
6. $\mathbb{Z}_{2^3} \times \mathbb{Z}_{3^2} \times \mathbb{Z}_5$

Question 4. Group of order 360 with at least an element of order 8

Definitions 1.3.2. A group G is **decomposable** if it is isomorphic to a direct product of two proper, non-trivial subgroups. Otherwise G is **indecomposable**.

For example, $\mathbb{Z}_{2^3} \times \mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_5 \approx \mathbb{Z}_{24} \times \mathbb{Z}_{15}$ is decomposable.

Remark. $G = \mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_3 \times \mathbb{Z}_3 \times \mathbb{Z}_5 \approx \mathbb{Z}_2 \times \mathbb{Z}_6 \times \mathbb{Z}_{30}$ is also decomposable. Since $G \approx (\mathbb{Z}_2 \times \mathbb{Z}_6) \times \mathbb{Z}_{30}$.

Theorem 1.3.2. *The finite indecomposable abelian groups are exactly the cyclic groups with order a power of a prime.*

Proof. Necessary part: Let G be a finite, indecomposable abelian group. By fundamental theorem of finitely generated abelian groups, G is isomorphic to a direct product of cyclic groups of prime power order.

$$G \approx \mathbb{Z}_{p_1^{r_1}} \times \mathbb{Z}_{p_2^{r_2}} \times \cdots \times \mathbb{Z}_{p_n^{r_n}}$$

Thus for G to be indecomposable the direct product should be a cyclic group of prime power order. $G \approx \mathbb{Z}_{p_1^{r_1}}$.

Sufficient part : Let p be a prime and r a non-negative integer. Cyclic group of order p^r is isomorphic to \mathbb{Z}_{p^r} . Since every cyclic groups are abelian, \mathbb{Z}_{p^r} is an abelian group of finite order p^r . It is enough to prove that \mathbb{Z}_{p^r} is indecomposable.

A proper, non-trivial subgroup of \mathbb{Z}_{p^r} is of the form \mathbb{Z}_{p^i} where $0 < i < r$. Suppose \mathbb{Z}_{p^r} is decomposable. Then $\exists i, j \in \mathbb{Z}^+$ such that $\mathbb{Z}_{p^r} \approx \mathbb{Z}_{p^i} \times \mathbb{Z}_{p^j}$ and $i + j = r$. Clearly, p^i and p^j are not relatively prime, thus $\mathbb{Z}_{p^r} \not\approx \mathbb{Z}_{p^i} \times \mathbb{Z}_{p^j}$. Therefore, cyclic groups of order prime power are indecomposable. \square

Theorem 1.3.3. *If m divides the order of a finite abelian group G , then G has a subgroup of order m .*

Proof. Let $G \approx \mathbb{Z}_{p_1^{r_1}} \times \mathbb{Z}_{p_2^{r_2}} \times \cdots \times \mathbb{Z}_{p_n^{r_n}}$. Then $|G| = p_1^{r_1} p_2^{r_2} \cdots p_n^{r_n}$. Suppose m divides $|G|$, then $m = p_1^{s_1} p_2^{s_2} \cdots p_n^{s_n}$ where $0 \leq s_i \leq r_i$. Define $H = \mathbb{Z}_{p_1^{s_1}} \times \mathbb{Z}_{p_2^{s_2}} \times \cdots \times \mathbb{Z}_{p_n^{s_n}}$. Then H is subgroup of order m . \square

Theorem 1.3.4. *If m is square-free integer, then every abelian group of order m is cyclic.*

Proof. Let m be a square-free integer and G be an abelian group of order m . By fundamental theorem of finitely generated abelian groups

$$G \approx \mathbb{Z}_{p_1^{r_1}} \times \mathbb{Z}_{p_2^{r_2}} \times \cdots \times \mathbb{Z}_{p_n^{r_n}}$$

We have, m is square-free. Thus $r_i = 1$ and p_i are distinct. Therefore, $G \approx \mathbb{Z}_{p_1 p_2 \cdots p_n}$ is a cyclic group of order m . \square

1.4 Exercises §11

Question 5. Enumerate subgroups of $\mathbb{Z}_2 \times \mathbb{Z}_2 \times \mathbb{Z}_4$

1.4.1 Abelian Groups

Remark. Direct product of abelian groups is abelian.

Question 6. Enumerate abelian groups of order 72 ?

Remark. Let G be an abelian group. The subset H of G with identity element and all elements of order n is subgroup of G if and only if n is a prime.

Proof. Suppose $a \in H$. Then every power of a has order n . Suppose n is not prime. Then d divides n and a^d has order n/d . \square

1.4.2 Torsion Group and Torsion Coefficients

Remark. If group G is abelian, then its elements of finite order forms a subgroup. (hint : $a, b \in G$ has finite order, then ab has finite order. And $a \in G$ has finite order, then a^{-1} has finite order)

Definitions 1.4.1. The **torsion group** of an abelian group G is the subgroup of G containing only those elements of finite order. An abelian group is **torsion free** if identity element is the only element of finite order. G is Torsion free if Torsion group of G is trivial, $\{e\}$.

Definitions 1.4.2. The integers m_1, m_2, \dots, m_n are torsion coefficients of G such that $G \approx \mathbb{Z}_{m_1} \times \mathbb{Z}_{m_2} \times \dots \times \mathbb{Z}_{m_n}$ where m_i divides m_{i+1} .

For example, $\mathbb{Z}_6 \times \mathbb{Z}_{12} \times \mathbb{Z}_{20}$ has torsion coefficients 2, 12, 60

Remark (Algorithm to find torsion coefficients of a group). Suppose G has a direct product form.

Step 1 Find power of each prime in the direct product form

Step 2 List power of each prime

Step 3 Append 1s on left to make all lists to equal length

Step 4 Product of i th number on each list gives m_i

For example, $G \approx \mathbb{Z}_6 \times \mathbb{Z}_{12} \times \mathbb{Z}_{20}$

Step 1 $\mathbb{Z}_6 \times \mathbb{Z}_{12} \times \mathbb{Z}_{20} \approx \mathbb{Z}_2 \times \mathbb{Z}_3 \times \mathbb{Z}_{2^2} \times \mathbb{Z}_3 \times \mathbb{Z}_{2^2} \times \mathbb{Z}_5$

Step 2 (2,4,4), (3,3), (5)

Step 3 (2,4,4), (1,3,3), (1,1,5)

Step 4 (2,12,60)

1.4.3 Torsion p -subgroup

“Caution : Torsion p -subgroup is a name suggested by ‘Jacob’. And is not among the standard terminology in group theory.”

Remark. Let G be a group. Let p be an integer, ($p > 1$). Then the set of all element of G of order p together with the identity element is a group, if p is a prime.

For example : Consider symmetric group, S_3 . It has three elements of order 2, namely $\mu_1 = (2, 3)$, $\mu_2 = (1, 3)$ and $\mu_3 = (1, 2)$. Clearly, the set of all element of order 2 together with identity () is not a subgroup of S_3 as $(1, 2)(1, 3) = (3, 1, 2)$ is an element of order 3. Thus by counter-example, for a prime p , the set of all elements of a non-abelian group G together with identity is not necessarily a subgroup of G .

Let G be an abelian group. Let p be a prime. Let H be the set of all element of G of order p together with the identity e . Let $g, h \in H$. Clearly, $g^p = e$ and $h^p = e$. For every $g, h \in G$, $gh \in G$, since G is abelian $(gh)^p = g^p h^p = e$. Also we have, $g^{-1} = g^{p-1}$ is a element of order p . Therefore, H is a subgroup of G .

Let g be an element of order 4. Then g^2 is a element of order 2. Thus, the subgroup generated by g or the smallest subgroup containing g has a element of order 2. Therefore, elements of order 4 together with identity cannot be a subgroup of G . Thus, ‘Torsion p -subgroup’ exists only if p is a square-free integer.

Let g be an element of order 6. Then g^2 is an element of order 3. Again, elements of order 6 together with identity cannot be a subgroup of G . Thus, ‘Torsion p -subgroup’ exists only if p is a power of a prime. Thus, ‘Torsion p -subgroup’ of G exists only if G is abelian and p is a prime.

1.4.4 Normal Factors of G

This is a warm-up exercise for §37.5, where the theory is discussed by Fraleigh. However, we are able to conclude the following :

Remark. Let $G = H \times K$. Let $g \in G$. Then $g = (h, k) \in H \times K$. Clearly, H is a subset of G , and $H \times \{e\} \leq G$. This subgroup is isomorphic to H . Thus $h \in H$ suggests the existence of $(h, e) \in G$.

Similarly $k \in K$ suggests $(e, k) \in G$. Therefore, $hk \in G$ suggests $(h, e)(e, k) = (h, k) = (e, k)(h, e)$. We know that, $kh \in G$ suggest $(e, k)(h, e) \in G$. Thus, $hk = kh$ for every $h \in H$ and every $k \in K$.

In other words, if $G = H \times K$, then H, K are isomorphic to normal subgroups H', K' of G such that $H' \cap K' = \{e\}$ and $G \simeq K' \times H'$.

1.5 Cosets and Homomorphism

Definitions 1.5.1. A **permutation group** S_n is the set of all permutations on the set $\{1, 2, \dots, n\}$.

Remark. Consider, $(1, 2, 3)(4, 5), (1, 2)(3, 4) \in S_5$. I was wrong about the order in which the permutations are carried out. It follows the same order as function composition. That is, $f \circ g(x)$ implies $f(g(x))$. Similarly, $\sigma\rho$ implies

$\sigma(\rho(1, 2, \dots, n))$.

$$(1 \ 2) (3 \ 4) * (1 \ 2 \ 3) (4 \ 5) = \begin{pmatrix} 1 & 2 & 3 & 4 & 5 \\ 2 & 3 & 1 & 5 & 4 \\ 1 & 4 & 2 & 5 & 3 \end{pmatrix} = (2 \ 4 \ 5 \ 3)$$

$$(1 \ 2 \ 3) (4 \ 5) * (1 \ 2) (3 \ 4) = \begin{pmatrix} 1 & 2 & 3 & 4 & 5 \\ 2 & 1 & 4 & 3 & 5 \\ 3 & 2 & 5 & 1 & 4 \end{pmatrix} = (1 \ 3 \ 5 \ 4)$$

Clearly, S_5 is a non-abelian group of order 120.

Definitions 1.5.2. Kernel of a function $\phi : G \rightarrow G'$ is the inverse image of the identity element in G' .

Definitions 1.5.3. Cosets and Normal Subgroup,

- A **left coset** gH is the subset $\{gh \in G : h \in H\}$ where $g \in G$ and $H \leq G$.
- A **right coset** Hg is the subset $\{hg \in G : h \in H\}$ where $g \in G$ and $H \leq G$.
- A subgroup H of group G is **normal** if $gH = Hg$, $\forall g \in G$.
- All subgroups of abelian groups are normal.

Remark. For example, $H = \{1, \rho^2, \mu, \mu\rho^2\}$ is a normal subgroup of D_4 . And $K = \{1, \mu\}$ is a subgroup of D_4 which is not normal. Note that, $\rho\mu \neq \mu\rho$. Clearly, $\rho K \neq K\rho$. However, $\rho H = \{\rho, \rho^3, \mu\rho^3, \mu\rho\} = H\rho$.

Remark (Lagrange's Theorem). Let G be a finite group. If $H \leq G$, then order of H divides order of G .

- $aH \cap bH \neq \emptyset \implies aH = bH$.
- $\forall g \in G, g \in gH$.
- $\forall g \in G, |gH| = |H|$.

Remark (Cayley's Theorem). Every group is isomorphic to a group of isomorphisms.

Definitions 1.5.4. Let G, G' be groups. A **group homomorphism** is a function $\phi : G \rightarrow G'$ such that $\phi(x)\phi(y) = \phi(xy)$. Clearly $\phi(e) = e'$. A **trivial homomorphism** is a function $\phi : G \rightarrow G'$ such that $\phi(G) = \{e'\}$.

Remark. Group homomorphism $\phi : G \rightarrow G'$ preserves identity, inverses and subgroups. And kernel of group homomorphism is a normal subgroup of G .

Remark. For example, $\phi : D_4 \rightarrow \mathbb{Z}_2$ defined by $\phi(\rho) = 1$ and $\phi(\mu) = 0$ is a group homomorphism with $\ker(\phi) = \{1, \rho^2, \mu, \mu\rho^2\}$.

1.6 Factor Groups

Remark (factor group). Let H be a normal subgroup of a group G . Then the **factor group** of G over H , G/H is the group of cosets of H in G .

Remark. For example, $H = \{1, \rho^2, \mu, \mu\rho^2\}$ is normal subgroup of D_4 . And the factor group $D_4/H = \{1H, \rho H\}$.

Theorem 1.6.1. *Let $\phi : G \rightarrow G'$ be a group homomorphism with kernel H . Then cosets of H form a factor group, G/H where $(aH)(bH) = (ab)H$. Also, $\mu : G/H \rightarrow \phi[G]$ defined by $\mu(aH) = \phi(a)$ is an isomorphism.*

Proof. Let $\phi : G \rightarrow G'$ be a group homomorphism with $\ker(\phi) = H$. We have, $\phi^{-1}(\phi(a)) = \{g \in G : \phi(g) = \phi(a)\}$.

Let $x \in aH$. Then $x = ah$ for some $h \in H$. And $\phi(x) = \phi(ah) = \phi(a)\phi(h) = \phi(a)$, since $\phi(h) = e'$. Thus, $x \in \phi^{-1}(\phi(a))$ and $aH \subset \phi^{-1}(\phi(a))$.

Let $x \in \phi^{-1}(\phi(a))$. Then $\phi(x) = \phi(a)$. And $\phi(a)^{-1}\phi(x) = e' \implies \phi(a^{-1}x) = e'$. Clearly, $a^{-1}x \in \ker(\phi)$. Thus, there exists $h \in H$ such that $a^{-1}x = h$. Therefore, $x = ah$ for some $h \in H$. Thus, $x \in aH$ and $\phi^{-1}(\phi(a)) \subset aH$. Therefore, $\phi^{-1}(\phi(a)) = aH$.

Similarly, $\phi^{-1}(\phi(a)) = Ha$. Thus $aH = Ha$ and H is a normal subgroup of G . Therefore, we have the factor group G/H .

To prove : $\mu : G/H \rightarrow \phi[G]$ is a one-one correspondence. ie, $aH \xleftrightarrow{\mu} \phi(a)$. To prove : μ is injective. Suppose $\mu(aH) = \mu(bH)$. Then $\phi(a) = \phi(b)$. And $b \in \phi^{-1}(\phi(a)) = aH$. Therefore, $bH = aH$.

To prove : μ is surjective. Let $\phi(a) \in \phi[G]$. Then, there exists aH such that $\mu(aH) = \phi(a)$.

We have, $\mu(aH) = \phi(a)$, $\mu(bH) = \phi(b)$, and $\mu((ab)H) = \phi(ab)$. Therefore, $\mu((aH)(bH)) = \mu((ab)H) = \phi(ab) = \phi(a)\phi(b) = \mu(aH)\mu(bH)$. Thus μ is a homomorphism. Therefore μ is an isomorphism. \square

Theorem 1.6.2. *Let G be a group and $H \leq G$. Then left coset multiplication is well-defined by $(aH)(bH) = (ab)H$ if and only if H is a normal subgroup of G .*

Proof. Necessary part : Suppose $(aH)(bH) = (ab)H$ is well-defined. Let $a \in G$. It is enough to prove that $aH = Ha$. Let $x \in aH$. Then $(xH)(a^{-1}H) = (xa^{-1})H$. Also $(aH)(a^{-1}H) = eH = H$. We have, coset multiplication is well-defined. Thus $xa^{-1} = h \in H \implies x = ha \in Ha$. Then, $aH \subset Ha$. Similarly, $Ha \subset aH$ and $aH = Ha$. Therefore, H is a normal subgroup of G .

Sufficient part: Suppose H is a normal subgroup of G , and let $x \in aH$ and $y \in bH$. $x \in aH \implies x = ah_1$ for some $h_1 \in H$ $y \in bH \implies y = bh_2$ for some $h_2 \in H$. Therefore $xy = (ah_1)(bh_2) = a(h_1(bh_2)) = a((h_1b)h_2) = a((bh_3)h_2) = a(b(h_3h_2)) = a(bh_4)$. Since H is a group, $h_3h_2 = h_4 \in H$. Thus, $xy = a(bh_4) = (ab)h_4 \in (ab)H$ for all $x \in aH$ and $y \in bH$. Thus $(aH)(bH) = (ab)H$. \square

Corollary 1.6.2.1. *Let H be a normal subgroup of G . Then the cosets of H form a group G/H under the binary operation $(aH)(bH) = (ab)H$.*

Proof. Let H be a normal subgroup and aH, bH, cH are cosets of H in G .

G1 Associativity

$$(aH)[(bH)(cH)] = (aH)[(bc)H] = [a(bc)]H = [(ab)c]H = [(ab)H](cH) = [(aH)(bH)](cH)$$

G2 Existence of identity, eH

$$(aH)(eH) = (ae)H = aH \text{ and } (eH)(aH) = (ea)H = aH.$$

G3 Existence of inverse $(a^{-1}H)$

$$(aH)(a^{-1}H) = (aa^{-1})H = eH \text{ and } (a^{-1}H)(aH) = (a^{-1}a)H = eH.$$

□

Remark. $n\mathbb{Z}$ is a normal subgroup of \mathbb{Z} . And $\mathbb{Z}/n\mathbb{Z} \approx \mathbb{Z}_n$. \mathbb{Z}_n is a torsion group isomorphic to a factor group of torsion free group \mathbb{Z} .

Remark. Let $c \in \mathbb{R}^*$. Then the cyclic group generated by c is a normal subgroup of \mathbb{R} and $\mathbb{R}/\langle c \rangle \approx \mathbb{R}_c$.

Question 7. Let $c = 0.31$. Find the coset $x + \langle 0.31 \rangle$ containing 2.

1.7 Fundamental Homomorphism & Automorphisms

1.7.1 Fundamental Homomorphism Theorem

Theorem 1.7.1. Let H be a normal subgroup of G . Then $\gamma : G \rightarrow G/H$ is defined by $\gamma(x) = xH$ is a homomorphism with kernel H .

Proof. $\gamma(x)\gamma(y) = (xH)(yH)$ Let $h_1, h_2 \in H$, $(xh_1)(yh_2) = xyh_3h_2 = xyh_4$ for some $h_3, h_4 \in H$. Therefore $(xH)(yH) = (xy)H$. $\gamma(xy) = (xy)H = \gamma(x)\gamma(y)$. $\gamma(x) = xH = H \iff x \in H$. Therefore, $\ker(\gamma) = H$. □

Remark. Suppose H is a normal subgroup of a group G . Then, a homomorphism $\gamma : G \rightarrow G/H$ is a natural homomorphism.

Theorem 1.7.2 (Fundamental Homomorphism). Let $\phi : G \rightarrow G'$ be a group homomorphism with kernel H . Then $\phi[G]$ is a group, and $\mu : G/H \rightarrow \phi[G]$ given by $\mu(gH) = \phi(g)$ is an isomorphism. If $\gamma : G \rightarrow G/H$ is the homomorphism given by $\gamma(g) = gH$, then $\phi(g) = \mu\gamma(g)$ for each $g \in G$.

Proof. μ is an isomorphism $G/H \xrightarrow{\mu} \phi[G]$. $\mu(\gamma(g)) = \mu(gH) = \phi(g)$. Thus $\mu\gamma = \phi$. □

“Every group homomorphism $\phi : G \rightarrow G'$ with kernel N has a unique natural group homomorphism $\gamma : G \rightarrow G/N$ and a unique isomorphism $\mu : G/N \rightarrow G'$ such that $\phi = \mu\gamma$. That is, $\phi(g) = \mu(\gamma(g)) = \mu(gN)$.”

1.7.2 Inner Automorphism

Theorem 1.7.3. Let H be a subgroup of G , then the following statements are equivalent:

1. $ghg^{-1} \in H, \forall g \in G, h \in H$
2. $gHg^{-1} = H, \forall g \in G$
3. $gH = Hg, \forall g \in G$

Proof. Let G be a group and $H \leq G$. By right multiplication, $gHg^{-1} = H \iff gH = Hg$. Trivially, $\forall h \in H, ghg^{-1} \in H \iff gHg^{-1} \subset H$. Therefore, it is enough to prove that $H \subset gHg^{-1}$. Let $h \in H$ and $x \in G$, then $xhx^{-1} = h'$ for some $h' \in H$. Then $h = x^{-1}h'x = x^{-1}h'(x^{-1})^{-1} \in gHg^{-1}$ where $x^{-1} = g \in G$. Thus $h \in gHg^{-1}$ and $H \subset gHg^{-1}$. Therefore $gHg^{-1} = H$. \square

Definitions 1.7.1. Automorphisms :

- An **automorphism** of G is an isomorphism $\phi : G \rightarrow G$
- The **inner automorphism** of G by $g \in G$ is the isomorphism $i_g : G \rightarrow G$ defined by $i_g(x) = gxg^{-1}$ for all $x \in G$
- The **conjugate** of x by g is the element $gxg^{-1} \in G$
- The **conjugate subgroup** of subgroup H , $i_g[H] = \{ghg^{-1} : h \in H\}$

Remark. For example, $i_\rho : D_4 \rightarrow D_4$ defined by $i_\rho(x) = \rho x \rho^{-1}$ is an inner automorphism. We have, $H = \{1, \mu\}$ is a subgroup of D_4 . The conjugate subgroup $i_\rho[H] = \{1, \mu\rho^2\}$.

Remark. Normal subgroups are invariant under any inner automorphism.

1.8 Exercise §14

1.8.1 Normal subgroups

Question 8. Prove that the notion of normality is stronger than abelian.

Remark. Let G be a group. Then intersection of normal subgroups of G is a normal subgroup of G .

Proof. Let \mathcal{N} be a nonempty subfamily of normal subgroups of G . Let N be the intersection of subgroups in \mathcal{N} . Clearly, intersection of subgroups of G is also subgroup of G .

Suppose $x \in N$. Then $x \in H$ for every $H \in \mathcal{N}$. Suppose $g \in G$. Then $gxg^{-1} \in H$, for every normal subgroup $H \in \mathcal{N}$. Thus $gxg^{-1} \in N$. Therefore, N is a normal subgroup of G . \square

Challenge 1. Let $\phi : G \rightarrow G'$ and $\psi : G \rightarrow G'$ be two group homomorphism with kernel H and K respectively. Show that $\phi\psi : G \rightarrow G'$ defined by $\phi\psi(g) = \phi(g)\psi(g)$ is also a group homomorphism, but $\ker(\phi\psi) \neq H \cap K$.
(hint : Construct ψ such that $\phi(g)\psi(g) = e$ for some $g \notin H$)

Remark. Let $S \subset G$, then G has a smallest, normal subgroup containing S .

Proof. Let N be the intersection of all normal subgroups of G containing S . Then $S \subset N$ and N is a normal subgroup of G . Let H be the smallest, normal subgroup of G containing S . Then $N \subset H$. Therefore, $H = N$. \square

Remark. If a finite group G has exactly one subgroup H of order m , then H is normal.

Proof. Let G be a finite group. Suppose G has only one subgroup of order m , say H . We know that, conjugate of a subgroup is a subgroup of same order. Thus, conjugates of H are H itself. Thus, $xHx^{-1} = H$ for every $x \in G$. Therefore, $xH = Hx$ for every $x \in G$ and H is a normal subgroup of G . \square

Remark. If G has a subgroup of order s , then the intersection of all subgroups of order s is a normal subgroup of G .

Proof. Let G be a group. Let H be a subgroup of G of order s . Let N be the intersection of all subgroups of G of order s . ?? \square

1.8.2 Linear Groups

Definitions 1.8.1. The set of all $n \times n$, non-singular matrices with real entries is a group under matrix multiplication. This group is the **General Linear Group** and is denoted by $GL(n, \mathbb{R})$.

Definitions 1.8.2. The set of all $n \times n$ matrices with real entries and determinant ± 1 is a group under matrix multiplication. This group is the **Special Linear Group** and is denoted by $SL(n, \mathbb{R})$.

Remark. $SL(n, \mathbb{R})$ is a normal subgroup of $GL(n, \mathbb{R})$

Proof. Let $A \in GL(n, \mathbb{R})$ and $B \in SL(n, \mathbb{R})$. Let $r = |A|$ and we have $|B| = \pm 1$. Then $|AB| = |A| |B| = \pm r$.

Let M_r be the set of all matrices with determinant $\pm r$ where $r \in \mathbb{R}$. Clearly, for every matrix A with determinant r , $A \in M_r$. And there exists a matrix $C = B^{-1}AB$. Then, $|C| = |B^{-1}| |A| |B| = \pm r$. Thus, $C \in M_r$ and $AB = BC$. Clearly, M_r is a left coset of $GL(n, \mathbb{R})$. Therefore, $SL(n, \mathbb{R})$ is a normal subgroup of $GL(n, \mathbb{R})$. \square

1.8.3 Factor Group

Remark. A_n is a normal subgroup of S_n . And $S_n/A_n \simeq \mathbb{Z}_2$.

Proof. Let function $\phi : S_n \rightarrow \mathbb{Z}_2$ be defined by $\phi(\sigma) = 0$ if σ is an even permutation and $\phi(\sigma) = 1$ if σ is an odd permutation. Then ϕ is a homomorphism with kernel A_n , the set of all even permutations in S_n . We know that, the kernel of a homomorphism is a normal subgroup of the domain. Therefore, A_n is a normal subgroup of S_n . \square

Remark. If H is normal subgroup of G and $(G : H) = m$, then $a^m \in H$ for all $a \in G$.

Proof. Let H be a normal subgroup of G such that $(G : H) = m$. Then $|G/H| = m$. Let $a \in G$, then $aH \in G/H$. Then by Lagrange's theorem, order of aH divides m . Therefore, $(aH)^m = eH$. We have, $(aH)^m = a^m H$. Therefore, $a^m \in H$. \square

Remark. Every factor group of an abelian group is also abelian.

Proof. Let G be an abelian group and H be a normal subgroup of G . Then G/H is a factor group of G . Let $aH, bH \in G/H$. Then, $(aH)(bH) = (ab)H = (ba)H = (bH)(aH)$. Clearly, factor group of an abelian group is abelian. \square

Remark. Let G be a group and T be the torsion subgroup of G , then the factor group, G/T is torsion free.

Proof. Let G be a group T its torsion subgroup. Let $g \in G$. If the order of g is finite, then $g \in T$. Suppose G/T has a element of finite order $m > 1$, say xT where $x \notin T$. Then x is an element of infinite order.

We have, order of xT is m . That is, $(xT)^m = eT$. Thus, $(xT)^m = (x^m)T = eT$. Therefore, $x^m = y \in T$. Since $y \in T$, y is an element of finite order, say r . Then $(x^m)^r = x^{mr} = e$. This is a contradiction as x is an element of infinite order. Therefore, every element of G/T except eT are of infinite order. Clearly, G/T is torsion free. \square

1.8.4 Commutator subgroup

Definitions 1.8.3. A **commutator** c in group G is an element in the form $c = aba^{-1}b^{-1}$ for some $a, b \in G$. The **commutator subgroup** is the smallest normal subgroup containing all commutators in G .

Remark. Let C be the commutator subgroup of G , then G/C is abelian.

Proof. Let C be the commutator subgroup of G . Let $x, y \in G$. Then $x^{-1}yxy^{-1} \in C$. Therefore, $x(x^{-1}yxy^{-1}) \in xC$ and $y(x^{-1}yxy^{-1}) \in yC$. But, $x(x^{-1}yxy^{-1})y(x^{-1}yxy^{-1}) = yx(x^{-1}yxy^{-1}) \in (yx)C$. Therefore, $(xC)(yC) = (yx)C = (yC)(xC)$. Clearly, G/C is abelian. \square

Remark. The factor group G/C is the abelianised version of G .

Remark. Let G be a group and C be the commutator group of G . Let H be a normal subgroup of G . If G/H is abelian, then C is a subgroup of H .

Question 9. Find commutator subgroup of the dihedral group D_4 ?

1.8.5 Automorphism

Remark. Every inner automorphism is an identity map for an abelian group.

Proof. Let G be an abelian group and $g \in G$. Then $i_g : G \rightarrow G$ is defined by $i_g(x) = gxg^{-1} = gg^{-1}x = x$. Clearly, i_g is an identity map. \square

Remark. Set of all $g \in G$ such that the inner automorphism i_g is an identity map is normal.

Proof. Let H be the set of all $g \in G$ such that i_g is an identity map.

$$\begin{aligned} H &= \{g \in G : i_g(x) = x, \forall x \in G\} \\ &= \{g \in G : gxg^{-1} = x, \forall x \in G\} \\ &= \{g \in G : gx = xg, \forall x \in G\} \end{aligned}$$

Therefore, H is a normal subgroup of G . \square

Remark. Set of automorphisms Γ of a group G is a group under composition. And the set of inner automorphisms is a normal subgroup of Γ .

Proof. Let G be a group. We know that, the composition of two automorphisms is also an automorphism. Also, the composition of functions is associative. Let i be the identity map and μ be any automorphism. Then $i\mu = \mu = \mu i$. Since automorphisms are bijective, there exists a unique inverse μ^{-1} for each automorphism μ such that $\mu\mu^{-1} = i$. Clearly, the set of all automorphisms of a group G is also a group. \square

Remark. Subgroup conjugacy is an equivalence relation on the set of subgroups.

Proof. Let H be a subgroup of a group G . Conjugation is reflexive, since $eHe^{-1} = H$ and $H \sim H$.

Let K be a conjugate subgroup of H . That is, $H \sim K$. Then $K = gHg^{-1}$. Clearly, $H = g^{-1}K(g^{-1})^{-1}$. Thus, H is a conjugate of K . ie, $K \sim H$.

Let $H \sim K$ and $K \sim L$. Then we have $x, y \in G$ such that $K = xHx^{-1}$ and $L = yKy^{-1}$. Now, $L = y(xHx^{-1})y^{-1} = (yx)H(yx)^{-1}$. Thus, $H \sim L$. \square

Question 10. Find the automorphism group $\Gamma(\mathbb{Z}_2 \times \mathbb{Z}_4)$?

1.9 Simple Groups

Remark. Factor Group Computations :

- The converse of Lagrange's theorem is false.
For example, A_4 has order 12, but doesn't have a subgroup of order 6.
- Factor group of a cyclic group is cyclic.
(hint : if g is a generator of G , then gH is a generator of G/H .)

Question 11. Show that $\mathbb{Z}_4 \times \mathbb{Z}_6 / \langle (2, 3) \rangle \approx \mathbb{Z}_4 \times \mathbb{Z}_3$.

Definitions 1.9.1. A group is simple if it is non-trivial and has no proper, non-trivial normal subgroups.

Remark. Abelian, simple groups are \mathbb{Z}_p , the cyclic groups of prime order.

Remark. Symmetric group, S_3 is not simple.

The subgroup, $\{\rho_0, \rho_1, \rho_2\}$ is a normal subgroup of S_3 .

Remark. Smallest nonabelian, simple group is A_5 .

Every nonabelian, simple group of order 60 is isomorphic to A_5 .

Remark. Simple groups :

- Alternating groups A_n are simple for $n \geq 5$.
- Every finite group can be factorised into simple groups.
- Every finite, non-abelian, simple group is of even order.
- Group homomorphism preserves normal subgroups.
- M is maximal normal subgroup of G if and only if G/M is simple.
- Center $Z(G) = \{z \in G : zg = gz, \forall g \in G\}$ is normal.
- Center of non-abelian groups of order pq are trivial if p, q are primes.
- Factor group, G/N is abelian if and only if C is a subgroup of N .

Question 12. G is simple and H is subgroup of G , then H is simple ?

1.10 Group Action on a Set

Definitions 1.10.1. An action of a group G on a set X is a map. $*$: $G \times X \rightarrow X$ such that

1. $ex = x, \forall x \in X$
2. $(g_1g_2)(x) = g_1(g_2x), \forall x \in X, \forall g_1, g_2 \in G$

Then X is a G -set.

Theorem 1.10.1. Let X be a G -set. $\forall g \in G, \sigma_g : X \rightarrow X$ defined by $\sigma_g(x) = gx$ is a permutation of X . Also, the map $\phi : G \rightarrow S_X$ defined by $\phi(g) = \sigma_g$ is a homomorphism with the property that $\phi(g)(x) = gx$.

Proof. Suppose X is a G -set. Let $g \in G$, and $x_1, x_2 \in X$.

Suppose $\sigma_g(x_1) = \sigma_g(x_2) \implies gx_1 = gx_2 \implies g^{-1}(gx_1) = g^{-1}(gx_2) \implies (g^{-1}g)x_1 = (g^{-1}g)x_2 \implies ex_1 = ex_2 \implies x_1 = x_2$. Thus, σ_g is injective.

Let $x \in X$. Then $\sigma_g(g^{-1}x) = g(g^{-1}x) = (gg^{-1})x = ex = x$. Thus, σ_g is surjective. Therefore, σ_g is a permutation of $X, \sigma_g \in S_X$.

Let $g_1, g_2 \in G$. And $\phi(g_1)(x) = \sigma_{g_1}(x) = g_1x, \phi(g_2)(x) = \sigma_{g_2}(x) = g_2x$. $\phi(g_1g_2)(x) = \sigma_{g_1g_2}(x) = (g_1g_2)x = g_1(g_2x) = \sigma_{g_1}(g_2x) = \phi(g_1)(g_2x) = \phi(g_1)(\sigma_{g_2}(x)) = \phi(g_1)(\phi(g_2)(x)) = \phi(g_1)\phi(g_2)(x)$. Therefore, $\phi(g_1g_2) = \phi(g_1)\phi(g_2)$ and ϕ is a homomorphism. \square

Definitions 1.10.2. Group Action,

- G acts **faithfully** on X , if e is the only element that leaves every $x \in X$ fixed.
- G is **transitive** on X if for every $x_1, x_2 \in X, \exists g \in G$ such that $gx_1 = x_2$.
 G is transitive on X iff the subgroup $\phi[G]$ of S_X is transitive.

1.11 Isotropy subgroups & Orbits

Definitions 1.11.1. Let X be a G -set.

- The subset fixed by $g, X_g = \{x \in X : gx = x\}$
- The isotropy subgroup of $x, G_x = \{g \in G : gx = x\}$
Let $Y \subset X$, then $G_Y = \{g \in G : gy = y, \forall y \in Y\}$ is a subgroup of G .
- The orbit of x in X under $G, Gx = \{gx \in X : g \in G\}$

Theorem 1.11.1. Let X be a G -set. Then G_x is a subgroup of $G, \forall x \in X$.

Proof. Let $x \in X$. And $g_1, g_2 \in G_x$. Then $g_1x = x$ and $g_2x = x$.

Clearly, $(g_1g_2)x = g_1(g_2x) = g_1x = x$. Therefore, $g_1g_2 \in G_x$. Also $ex = x \implies e \in G_x$. Let $g \in G_x$. Then $gx = x \implies g^{-1}(gx) = g^{-1}x \implies (g^{-1}g)x = g^{-1}x \implies x = g^{-1}x$. Thus, for any $g \in G_x, g^{-1} \in G_x$. Therefore, G_x is a subgroup of G for any $x \in X$. \square

Theorem 1.11.2. Let X be a G -set and $x_1, x_2 \in X$. Then the relation \sim defined by $x_1 \sim x_2$ iff $gx_1 = x_2$ is an equivalence relation.

Proof. Let $x \in X$. Then $ex = x \implies x \sim x$. Let $x_1, x_2 \in X$ and $x_1 \sim x_2$. Then there exists some $g \in G$ such that $gx_1 = x_2$. We have, $g^{-1}x_2 = g^{-1}(gx_1) = (g^{-1}g)x_1 = ex_1 = x_1$. Therefore, $x_2 \sim x_1$. Let $x_1, x_2, x_3 \in X$ and $x_1 \sim x_2$ and $x_2 \sim x_3$. Then there are $g_1, g_2 \in G$ such that $g_1x_1 = x_2$ and $g_2x_2 = x_3$. Clearly, $g_2g_1 \in G$ and $(g_2g_1)x_1 = g_2(g_1x_1) = g_2x_2 = x_3$. Therefore, $x_1 \sim x_3$. \square

Theorem 1.11.3. *Let X be a G -set and $x \in X$. Then $|Gx| = (G : G_x)$. If $|G|$ is finite, then $|Gx|$ is a divisor of $|G|$.*

Proof. We have Gx is the orbit of x in X under G and L_{G_x} is the left cosets of G_x in G . Let $x_1 \in Gx$. Then there exists $g_1 \in G$ such that $x_1 = g_1x$. Define $\psi : Gx \rightarrow L_{G_x}$ by $\psi(x_1) = g_1G_x$.

Step 1 : ψ is well-defined.

Let $x_1 \in Gx$. Suppose there exists $g_1, g'_1 \in G$ such that $g_1x = x_1$ and $g'_1x = x_1$. Then we have, $g_1x = g'_1x \implies x = g_1^{-1}(g'_1x) = (g_1^{-1}g'_1)x$. Thus, $g_1^{-1}g'_1 \in G_x$. Therefore, $g_1(g_1^{-1}g'_1) \in g_1G_x$. Clearly, $g_1(g_1^{-1}g'_1) = (g_1g_1^{-1})g'_1 = g'_1 \in g_1G_x$. Therefore, $g_1G_x = g'_1G_x$. And $\psi(x_1) = g_1G_x$ is well-defined.

Step 2 : ψ is one-to-one.

Suppose $\psi(x_1) = \psi(x_2)$. Let $x_1, x_2 \in Gx$ such that $x_1 = g_1x$ and $x_2 = g_2x$. Then we have $\psi(x_1) = \psi(x_2) \implies g_1G_x = g_2G_x$. Thus, $g_2 = g_1g$ for some $g \in G_x$. Clearly, $x_2 = g_2x = (g_1g)x = g_1(gx) = g_1x = x_1$.

Step 3 : ψ is onto.

Let g_1G_x be a left coset of G_x in G . Then we have, $g_1 \in G$ and $g_1x \in Gx$, say x_1 . Therefore, there exists $x_1 \in Gx$ such that $\psi(x_1) = g_1G_x$. \square

1.12 Exercise §16

Definitions 1.12.1. Let X be a G -set. And $S \subset X$. Then S is a **sub- G -set** if the orbit Gs of each $s \in S$ is contained in S .

$$S = \bigcup_{s \in S} Gs \quad (1.4)$$

Remark. • Every G -set is a union of its orbits.

• Every G -set is isomorphic to the disjoint union of left coset G -sets.

Definitions 1.12.2. Let G be a group. Let X, Y be two G -sets. Then function $\phi : X \rightarrow Y$ is a G -set isomorphism if

1. ϕ is a bijection and
2. ϕ is a G -set homomorphism
ie, $g\phi(x) = \phi(gx)$, for every $x \in X$ and every $g \in G$.

1.13 Application of G -Sets to Counting

Theorem 1.13.1 (Burnside). *Let G be a finite group and X a finite G -set. If r is the number of orbits in X under G , then*

$$r|G| = \sum_{g \in G} |X_g| \quad (1.5)$$

Proof. Let $N = \{(g, x) \in G \times X : gx = x\}$.

Step 1 : $|N| = \sum_{g \in G} |X_g|$.

Let $g \in G$. We have, X_g is the set of all $(g, x) \in N$ with g as first member. Enumerating elements of N for each $g \in G$, we get $\sum_{g \in G} |X_g| = N$.

Step 2 : $|N| = r|G|$.

Let $x \in X$. We have, G_x is the set of all $g \in G$ such that $gx = x$. In other words, G_x is the set of all $g \in G$ such that $(g, x) \in N$ with x as second member. However, $|Gx| = (G : G_x)$.

$$\implies |Gx| = \frac{|G|}{|G_x|} \implies |G_x| = \frac{|G|}{|Gx|}$$

Enumerating element of N for each $x \in X$, we get

$$|N| = \sum_{x \in X} |G_x| = \sum_{x \in X} \frac{|G|}{|Gx|} = |G| \sum_{x \in X} \frac{1}{|Gx|}$$

Let $Gx = \mathcal{O}$ be an orbit containing x with length k .

$$\implies \sum_{x \in \mathcal{O}} \frac{1}{|Gx|} = \sum_{i=1}^k \frac{1}{k} = 1$$

Let r be the number of orbits in X under the group action of G . Clearly, the orbits of X are disjoint. Thus,

$$|N| = |G| \sum_{i=1}^r \sum_{x \in \mathcal{O}_i} \frac{1}{|Gx|} = |G| \sum_{i=1}^r 1 = r|G|$$

Therefore, $\sum_{g \in G} |X_g| = |N| = r|G|$. □

Corollary 1.13.1.1. *If G is a finite group and X is a finite G -set, then*

$$\text{number of orbits in } X \text{ under } G = \frac{1}{|G|} \sum_{g \in G} |X_g| \quad (1.6)$$

Proof. From Burnside formula, we have

$$\sum_{g \in G} |X_g| = r|G| \implies r = \frac{1}{|G|} \sum_{g \in G} |X_g|$$

□

1.14 Isomorphism Theorems 1-2

1.14.1 First Isomorphism Theorem

Theorem 1.14.1 (first isomorphism). *Let $\phi : G \rightarrow G'$ be a homomorphism with kernel K , and let $\gamma_K : G \rightarrow G/K$ be the canonical homomorphism. There is a unique isomorphism $\mu : G/K \rightarrow \phi[G]$ such that $\phi(x) = \mu(\gamma_K(x))$ for each $x \in G$.*

Proof. By, fundamental homomorphism theorem.[Fraleigh, 2013, §14.1] □



Figure 1.1: First Isomorphism Theorem

1.14.2 Second Isomorphism Theorem

Definitions 1.14.1. Join $H \vee N$ is the smallest subgroup of G containing HN where $HN = \{hn : h \in H, n \in N\}$.

Lemma 1.14.2. Let N be a normal subgroup of G and let $\gamma : G \rightarrow G/N$ be the canonical homomorphism. Then the map ϕ from the set of normal subgroups of G containing N to the set of normal subgroup of G/N given by $\phi(L) = \gamma[L]$ is one-to-one and onto.

Proof. Let G be a group and N be a normal subgroup of G . Given that $\gamma : G \rightarrow G/N$ the canonical homomorphism. That is, $\gamma(g) = gN$ for every $g \in G$. Let L, M be normal subgroups of G containing N . Since homomorphism preserves normality, $\gamma(L)$ is a normal subgroup of $\gamma[G]$.

Suppose $\phi(L) = \phi(M)$. By definition of ϕ ,

$$\gamma[L] = \phi(L) = \phi(M) = \gamma[M]$$

Since N is normal, $\gamma^{-1}(\gamma(g)) = gN$ for every $g \in G$. And

$$\gamma^{-1}(\gamma[L]) = \gamma^{-1}\left(\bigcup_{g \in L} \gamma(g)\right) = \bigcup_{g \in L} \gamma^{-1}(\gamma(g)) = \bigcup_{g \in L} gN = L$$

for every point $g \in L$, the left coset gN is contained in L . ($\because N \leq L$) Therefore,

$$\gamma^{-1}(\phi(L)) = \gamma^{-1}(\gamma[L]) = L$$

$$\gamma^{-1}(\phi(M)) = \gamma^{-1}(\gamma[M]) = M$$

Thus, $L = M$. Therefore, ϕ is injective.

Let H be a normal subgroup of G/N . Then $\gamma^{-1}(H)$ is a normal subgroup of G . We have, $eN \in H$ and $\gamma^{-1}(eN) = eN = N$. Thus, $N \subset \gamma^{-1}(H)$. Thus, there exists $\gamma^{-1}(H)$, a normal subgroup of G containing N such that $\phi(\gamma^{-1}(H)) = H$. Therefore, ϕ is surjective. \square

Lemma 1.14.3. If N is a normal subgroup of G , then $H \cap N = HN = NH$. Furthermore, if H is also normal in G , then HN is normal in G .

Proof. Let G be a group and N be a normal subgroup of G . Also let H be a subgroup of G .

Claim : $HN = \{hn \in G : h \in H, n \in N\}$ is a subgroup of G .

G1 Closure : $h_1n_1h_2n_2 = h_1(h_2n_3)n_2 = h_3n_4 \in HN$, where $h_1, h_2 \in H$ and $n_1, n_2, n_4 \in N$. Since N is normal, $n_1h_2 = h_2n_3$ for some $n_3 \in N$.



Figure 1.2: Second Isomorphism Theorem

G2 $HN \subset G$, thus HN satisfies associativity.

G3 Since $H, N \leq G$, $e \in H$ and $e \in N$. Thus, $e = ee \in HN$.

G4 Let $h_1n_1 \in HN$. $H \leq G$ and $h_1 \in H \implies h_1^{-1} \in H$. Then $n_1^{-1}h_1^{-1} = h_1^{-1}n_2$ for some $n_2 \in N$, since N is normal. Thus, every element in HN , h_1n_1 has an inverse $h_1^{-1}n_2 \in HN$.

Let H be a normal subgroup of G . Let $g \in G$. Then $g(h_1n_1) = (gh_1)n_1 = (h_2g)n_1 = h_2(gn_1) = h_2(n_2g) = (h_2n_2)g$ for some $h_2 \in H$ and $n_2 \in N$ since both H and N are normal subgroups of G . Thus $gHN = HN g$ for every $g \in G$. Therefore, HN is a normal subgroup of G . \square

Theorem 1.14.4 (second isomorphism). *Let H be a subgroup of G and let N be a normal subgroup of G . Then $(HN)/N \simeq H/(H \cap N)$.*

Proof. Consider the canonical homomorphism $\gamma : G \rightarrow G/N$ with kernel N . Then γ restricted to HN is a homomorphism from HN onto $\gamma[H]$.

$$\gamma|_{HN} : HN \rightarrow \gamma[H] \text{ where } \gamma|_{HN}(hn) = \gamma(hn) = (hn)N = hN = \gamma(h) \quad (1.7)$$

Since $\ker(\gamma) = N$ and $N \subset HN$. We have, $\ker(\gamma|_{HN}) = N$. By first isomorphism theorem there exists a unique isomorphism $\mu_1 : HN/N \rightarrow \gamma[H]$ where $\mu_1(hnN) = \gamma(h) = hN$.

Similarly, γ restricted to H is also a homomorphism onto $\gamma[H]$.

$$\gamma|_H : H \rightarrow \gamma[H] \text{ where } \gamma|_H(h) = \gamma(h) = hN. \quad (1.8)$$

Since $\ker(\gamma) = N$ and $H \cap N \neq \phi$. $\ker(\gamma|_H) = H \cap N$. By first isomorphism theorem, there exists a unique isomorphism $\mu_2 : H/(H \cap N) \rightarrow \gamma[H]$

Then $HN/N \simeq H/(H \cap N)$, since the composition of two isomorphisms, $\mu_2^{-1} \circ \mu_1 : HN/N \rightarrow H/(H \cap N)$ is an isomorphism \square

1.15 Third Isomorphism Theorem

Theorem 1.15.1 (third isomorphism). *Let H and K be normal subgroup of G with $K \leq H$. Then $G/H \simeq (G/K)/(H/K)$.*



Figure 1.3: Third Isomorphism Theorem

Proof. Consider the function $\phi : G \rightarrow (G/K)/(H/K)$ defined by $\phi(g) = gK(H/K)$. Then, ϕ is a homomorphism.

$$\begin{aligned}
 \phi(ab) &= (ab)K(H/K) \\
 &= (aK)(bK)(H/K), \because (ab)K = (aK)(bK) \\
 &= aK(H/K) bK(H/K), \because (xy)H/K = xH/K yH/K \\
 &= \phi(a)\phi(b)
 \end{aligned}$$

The kernel of ϕ is the set $\{a \in G : \phi(a) = H/K\}$. Since the coset H/K was originally the points in H , we have $\ker(\phi) = H$. By first isomorphism theorem, there exists a unique isomorphism $\mu : G/H \rightarrow (G/K)/(H/K)$. Thus, $G/H \simeq (G/K)/(H/K)$. \square

1.16 Finite, non-abelian Groups

Theorem 1.16.1. Let p be a prime. Let G be a group of order p^n . Let X be a finite G -set. Then $|X| \equiv |X_G| \pmod{p}$.

Proof. Suppose there are r orbits in X . Choose an element from each orbit, say x_1, x_2, \dots, x_r . We have,

$$|X| = \sum_{i=1}^r |Gx_i| \quad (1.9)$$

Let $X_G = \{x \in X : gx = x, \forall g \in G\}$. Then each element in X_G belongs to an orbit of length 1. Let $|X_G| = s$. Then,

$$|X| = |X_G| + \sum_{i=s+1}^r |Gx_i| \quad (1.10)$$

We have, $|Gx| = (G : G_x)$. Clearly, both G and G_x are groups of order a multiple of prime p . Thus, for every $x \in X$, $|Gx|$ is multiple of prime p . Therefore, $|X| \equiv |X_G| \pmod{p}$. \square

Challenge 2. Let p be a prime and G be a finite group G . Design a mechanism to enumerate all elements of order p ?

Definitions 1.16.1. Let G be a group. Let p be a prime. G is a **p -group** if every element of G has order a power of p . A subgroup H of G is a **p -subgroup** of G if every element of H has order a power of p .

Theorem 1.16.2 (Cauchy). *Let p be a prime. Let G be a finite group and p divides $|G|$, then G has a subgroup of order p .*

Proof. Let X be the set of all n -tuples of elements of G such that the product of co-ordinates of each n -tuple is the identity element e of G .

$$X = \{(g_1, g_2, \dots, g_p) \in G^p : g_1 g_2 \cdots g_p = e\} \quad (1.11)$$

Let g_1, g_2, \dots, g_{p-1} be any elements in G . Then $g_p = (g_1 g_2 \cdots g_{p-1})^{-1}$ is uniquely determined. That is, the $(p-1)$ co-ordinates of an element in X may be chosen in $|G|$ different ways. Thus, $|X| = |G|^{p-1}$. Since p divides $|G|$, p divides $|X|$.

We have $g_1(g_2 \cdots g_p) = (g_2 g_3 \cdots g_p)g_1$, since $g_2 g_3 \cdots g_p = g_1^{-1}$. Thus, for every $(g_1, g_2, \dots, g_p) \in X$, the σ permutation of that element is in X . That is, $\sigma(g_1, g_2, \dots, g_p) = (g_2, g_3, \dots, g_p, g_1) \in X$. Similarly, $g_2(g_3 g_4 \cdots g_p g_1) = (g_3 g_4 \cdots g_p g_1)g_2$. And $\sigma^2(g_1, g_2, \dots, g_p) = (g_3, g_4, \dots, g_p, g_1, g_2)$. Clearly, the subgroup generated by σ , is a subgroup of S_p and X is a $\langle \sigma \rangle$ -set. We have, $|X| \cong |X_{\langle \sigma \rangle}| \pmod{p}$. Thus, p divides $|X_{\langle \sigma \rangle}|$.

Clearly, $(e, e, \dots, e) \in X$ and $(e, e, \dots, e) \in X_{\langle \sigma \rangle}$. Thus, $X_{\langle \sigma \rangle}$ has atleast p elements. Let $(g_1, g_2, \dots, g_p) \in X_{\langle \sigma \rangle}$, then σ fixes (g_1, g_2, \dots, g_p) .

$$\begin{aligned} \sigma(g_1, g_2, \dots, g_p) &= (g_1, g_2, \dots, g_p) \\ \implies (g_2, \dots, g_p, g_1) &= (g_1, g_2, \dots, g_p) \end{aligned}$$

Thus, $g_1 = g_2 = \cdots = g_p$, say $g \in G$. That $(g, g, \dots, g) \in X$ and $g^p = e$ by the definition of X . Therefore, G has an element g of order p . \square

Corollary 1.16.2.1. *Let G be a finite group. Then G is a p -group if and only if $|G|$ is a power of p .*

Proof. Let G be a finite group. Suppose G is a p -group. Suppose there exists another prime q , $q \neq p$ such that q divides $|G|$. Then by Cauchy's theorem, G has an element of order q . This contradicts the assumption that G is a p -group. Thus, the only prime that divides $|G|$ is p . Therefore, the order of G is a power of prime p .

Let $|G| = p^n$. Then the factors of p^n are powers of p . By Lagrange's theorem, order of subgroups of G must divide $|G|$. Thus, every subgroup of G has order a power of p . Thus, every element of G has order a power of p . Therefore, G is a p -group. \square

Definitions 1.16.2. Let G be a group and H be a subgroup of G . Consider the inner automorphisms $i_g : G \rightarrow G$ such that $i_g(x) = gxg^{-1}$. We have, $i_g(H) = gHg^{-1}$ is the conjugate of the subgroup H . The set of all elements in G which has H itself as the conjugate of H is the normaliser of H in G .

$$N[H] = \{g \in G : gHg^{-1} = H\} \quad (1.12)$$

Remark. The normaliser of H in G is a subgroup of G . And $N[H]$ is the largest subgroup of G with H as its normal subgroup.

Lemma 1.16.3. *Let H be a p -subgroup of a finite group G . Then*

$$(N[H] : H) \cong (G : H) \pmod{p} \quad (1.13)$$

Proof. Let G be a finite group. Let H be a p -subgroup of G . Let \mathcal{L} be the set of all left cosets of H in G . Then, $|\mathcal{L}| = (G : H)$.

Claim : \mathcal{L} is an H -set with group action $h(xH) = (hx)H$. We have, $e(xH) = (ex)H = xH$ and $(g_1g_2)(xH) = (g_1g_2x)H = g_1(g_2xH) = g_1(g_2(xH))$.

Let \mathcal{L}_H be the set of all left cosets that are fixed under action by all element of H .

$$\begin{aligned}\mathcal{L}_H &= \{xH \in \mathcal{L} : h(xH) = xH, \forall h \in H\} \\ &= \{xH \in \mathcal{L} : x^{-1}h(xH) = H, \forall h \in H\} \\ &= \{xH \in \mathcal{L} : (x^{-1}hx)H = H, \forall h \in H\} \\ &= \{xH \in \mathcal{L} : (x^{-1}hx) \in H, \forall h \in H\} \\ &= \{xH \in \mathcal{L} : x^{-1} \in N[H]\}\end{aligned}$$

Clearly, left cosets of \mathcal{L}_H has all its elements contained in $N[H]$. Thus,

$$|\mathcal{L}_H| = (N[H] : H) \quad (1.14)$$

We have, \mathcal{L} is an H -set. And H is a p -subgroup. Thus, H has order a power of prime p . Therefore,

$$|\mathcal{L}| \cong |\mathcal{L}_H| \pmod{p} \quad (1.15)$$

□

Corollary 1.16.3.1. *Let H be a p -subgroup of finite group G . If p divides $(G : H)$, then $N[H] \neq H$.*

Proof. We have, $(G : H) \cong (N[H] : H) \pmod{p}$. And p divides $(G : H)$. Thus, p divides $(N[H] : H)$. Therefore, $H \neq N[H]$. □

1.17 Sylow Theorems

Theorem 1.17.1 (First Sylow Theorem). *Let G be a finite group of order $p^n m$ where $n \geq 1$ and p does not divide m . Then*

1. G contains a subgroup of order p^i where $1 \leq i \leq n$.
2. Every subgroup H of G of order p^i is normal subgroup of the subgroup of order p^{i+1} , for $1 \leq i \leq n$.

Proof. We have, G is finite group and p divides $|G|$. By Cauchy's theorem, G has a subgroup of order p .

Suppose G has a subgroup H of order p^i , where $(i < n)$. Then, H is a p -subgroup. Thus, $(G : H) \cong (N[H] : H) \pmod{p}$ where $N[H]$ is the normaliser of H in G . Clearly, p divides $(G : H)$. Thus, p divides $(N[H] : H)$.

And H is a normal subgroup of $N[H]$. Thus, we have factor group $N[H]/H$ and p divides the order of $N[H]/H$. By Cauchy's theorem, $N[H]/H$ has a subgroup K of order p . Consider the canonical homomorphism $\gamma : N[H] \rightarrow N[H]/H$ defined by $\gamma(x) = xH$. Then $\gamma^{-1}(K)$ is a subgroup of $N[H]$ of order p^{i+1} and contains H . Thus, H is a normal subgroup of $\gamma^{-1}(K)$. By mathematical induction, G has subgroups of order p^i for $i = 2, 3, \dots, n$. □

Definitions 1.17.1. A Sylow p -subgroup of G is a maximal p -subgroup of G .

Theorem 1.17.2 (Second Sylow Theorem). *Let P_1 and P_2 be two Sylow p -subgroups of G . Then P_1 and P_2 are conjugate subgroups of G .*

Proof. Let P_1 and P_2 be two Sylow p -subgroups of G . Let \mathcal{L} be the set of all left cosets of P_1 . Then, group P_2 act on \mathcal{L} by $y(xP_1) = (yx)P_1$. Thus, \mathcal{L} is a P_2 -set. Therefore, $|\mathcal{L}| \cong |\mathcal{L}_{P_2}| \pmod{p}$.

Clearly $|\mathcal{L}| = (G : P_1)$. And p doesn't divide $|\mathcal{L}|$. Thus, p does not divide $|\mathcal{L}_{P_2}|$. Therefore, $|\mathcal{L}_{P_2}| > 0$.

Thus, \mathcal{L} has at least an element xP_1 which is fixed in the action of every element in P_2 . That is, $yxP_1 = xP_1$ for every $y \in P_2$. Therefore, $x^{-1}yxP_1 = P_1$ for every $y \in P_2$.

In other words, $x^{-1}yx \in P_1$ for every $y \in P_2$. Therefore, $x^{-1}P_2x \leq P_1$. But, $|P_1| = |P_2|$. Thus, $x^{-1}P_2x = P_1$. Therefore, P_1 and P_2 are conjugate subgroups of G . \square

Theorem 1.17.3 (Third Sylow Theorem). *If G is a finite group and p divides $|G|$, then the number of Sylow p -subgroups is congruent to 1 \pmod{p} and divides $|G|$.*

Proof. Let \mathcal{S} be the set of all Sylow p -subgroups of G . Let P be a Sylow p -subgroup of G . The elements of P act on \mathcal{S} by conjugation. Let $x \in P$ and $T \in \mathcal{S}$, then x carries T into xTx^{-1} . Clearly, \mathcal{S} is a P -set.

We have, $|\mathcal{S}| \cong |\mathcal{S}_P| \pmod{p}$. Suppose $T \in \mathcal{S}_P$. Then, $xTx^{-1} = T$ for every $x \in P$. Thus, T is a normal subgroup of P . But, T and P are of the same order, since both are Sylow p -subgroups of G . Therefore, $T = P$. Thus, $\mathcal{S}_P = \{P\}$. And $|\mathcal{S}_P| = 1$. Thus, $|\mathcal{S}| \cong 1 \pmod{p}$.

Let G act on \mathcal{S} by conjugation. Let $x \in G$ and $P \in \mathcal{S}$, then x carries P into xPx^{-1} . Clearly, \mathcal{S} is a G -set. However, every Sylow p -subgroup of G are conjugates. Thus, every Sylow p -subgroup of G belong to the same orbit under conjugation action. We have, $|Gx| = (G : G_x)$ and thus length of orbits divides the order of G . Since the number of Sylow p -subgroups is same as the length of the orbit of P , the number of Sylow p -subgroup of G divides the order of G . \square

Remark. Let G be a group of order 15. Let $p = 3$. By Third Sylow Theorem, the number of Sylow 3-subgroup of G is congruent to 1 $\pmod{3}$ and divides 15. We have, congruence class $\hat{1} = \{1, 4, 7, 10, 13\}$. Only 1 divides 15. Thus, there is only one Sylow 3-subgroup of G .

Let $p = 5$. By Third Sylow Theorem, the number of Sylow 5-subgroup of G is congruent to 1 $\pmod{5}$ and divides 15. We have, congruence class $\hat{1} = \{1, 6, 11\}$. Only 1 divides 15. Thus, there is only one Sylow 5-subgroup of G .

Remark. Let G be a group of order 255. And 1, 3, 5, 15, 17, 51, 85, 255 are the divisors of 255. Let $p = 3$. By third Sylow theorem, either there is one or eighty-five Sylow 3-subgroups of G . Suppose there 85 Sylow 3-subgroups. Then there are 170 elements of order 3.

Let $p = 5$. By third Sylow theorem, either there are one or fifty-one Sylow 5-subgroups of G . Suppose there 51 Sylow 5-subgroups. Then there are 204 elements of order 5.

Let $p = 17$. By third Sylow theorem, there is exactly one Sylow 17-subgroup of G . Thus, there are exactly 16 elements of order 17 in G .

1.18 Sylow Theorem : Applications

Definitions 1.18.1. A group G is solvable if there is sequence of subgroups $\{e\} = H_0 \leq H_1 \leq \cdots H_n = G$ such that for $i = 0, 1, 2, \dots$

1. H_i is a normal subgroup of H_{i+1}
2. H_{i+1}/H_i is simple and
3. H_{i+1}/H_i is abelian.

Theorem 1.18.1. Every group of prime order is solvable.

Proof.

□

Definitions 1.18.2. Let G be a finite group. Consider G acting on itself by conjugation. Then,

$$|G| = |Z(G)| + \sum_{s+1}^r |Gx_i| \quad (1.16)$$

This is the class equation of G . And each orbit of G under conjugation by itself is a conjugate class in G .

Theorem 1.18.2. The center of a finite, nontrivial p -group is nontrivial.

Proof.

□

Lemma 1.18.3. Let G be a group containing normal subgroups H and K such that $H \cap K = \{e\}$ and $H \vee K = G$. Then G is isomorphic to $H \times K$.

Proof.

□

Theorem 1.18.4. For a prime number p , every group of G of order p^2 is abelian.

Proof.

□

1.19 Sylow Theorem : Further Applications

Theorem 1.19.1. If H, K are subgroups of a group G . Then,

$$|HK| = \frac{(|H|)(|K|)}{|H \cap K|} \quad (1.17)$$

Proof.

□

1.19.1 Analysis of Finite Groups

The following are a few sample tests for finite groups. For different numbers, you may have to use a combination of these tests.

Remark. Finite, non-abelian, simple groups are of order 60, 168, 360, \dots

Type 1 : p^r

Groups of order p^r for $r > 1$, are not simple as they have normal subgroup of order p^{r-1} by first Sylow theorem.

For example, any group G of order 16 is not simple as by first Sylow theorem G has a normal subgroup of order 8.

Type 2 : pq

Suppose G is a group with order pq where p, q are primes and $q > p$. Clearly, $p \not\equiv 1 \pmod{q}$. Thus, by third Sylow theorem, the Sylow p -subgroup of G is unique, say H .

Now, by second Sylow theorem, Sylow p -subgroups are conjugate subgroups. Thus, this Sylow p -subgroup H is its only conjugate. That is, $gHg^{-1} = H$ for every $g \in G$. Thus, it is a normal subgroup of G . Thus G is not simple.

Further, if $q \not\equiv 1 \pmod{p}$ for $q > p$, then group G is cyclic.

For example, $5 \not\equiv 1 \pmod{3}$. Thus any group G of order 15 is cyclic. Clearly, G is abelian and is not simple.

However, $7 \equiv 1 \pmod{3}$. Thus, groups of order 21 are not simple, but are not necessarily cyclic.

Type 3 : Only one Sylow p -subgroup

In this case, we don't have a strict form. However, the nature of prime factor of the order suggests that G has only one Sylow p -subgroup for some prime factor p of its order.

For example, group G of order 20 has only one Sylow 5-subgroup say, H (by third Sylow theorem). Thus, this subgroup H is a normal subgroup of G (by second Sylow theorem). Therefore, groups of order 20 are not simple.

Type 4 : Enumerating elements of Sylow p -subgroups

Groups of order 30 are not simple. — The distinct Sylow p -subgroups suggests $p - 1$ elements are unique to each. By counting, we can show that not all p -subgroups have a conjugate.

Type 5 : Applying the relation for $|HK|$

Groups of order 48 are not simple. — $|HK| =$

Type 6 : Normaliser of $H \cap K$

Groups of order 36 are not simple. — Let H, K be distinct Sylow 3-subgroups of G . Then, the normaliser of $H \cap K$ suggests a normal subgroup which is either $H \cap K$ or its normaliser.

Type 7 : Commutator of G

Groups of order 255 are not simple. — Its commutator subgroup has order 1. Thus, the group is cyclic. Clearly, cyclic groups are abelian and not simple.

1.20 Rings, Fields & Integral Domains

Subject 2

ME010102 Linear Algebra

2.1 Vector Spces

2.2 Linear Transformations

2.3 Determinants

2.4 Elementary Canonical Forms

Subject 3

ME010103 Basic Topology

Subject 4

ME010103 Real Analysis

Module 1 : Bounded Variation & Rectifiable Curves

4.1 Functions of Bounded Variation & Rectifiable Curves

4.1.2 Properties of Monotone Functions

Theorem 4.1.1. *Let f be an increasing function defined on closed interval $[a, b]$. And let $x_0 = a < x_1 < x_2 < \cdots < x_{n-1} < x_n = b$. Then,*

$$\sum_{k=1}^{n-1} [f(x_k+) - f(x_k-)] \leq f(b) - f(a)$$

Proof. Let f be an increasing function on $[a, b]$. Let $\{x_0 = a, x_1, \dots, x_n = b\}$ be a partition of $[a, b]$. Let $y_k \in (x_{k-1}, x_k)$, $\forall k$. Then, $f(y_k) \leq f(x_k+)$ and $f(x_k-) \leq f(y_{k-1})$. Therefore,

$$\sum_{k=1}^{n-1} [f(x_k+) - f(x_k-)] \leq \sum_{k=1}^{n-1} [f(y_k) - f(y_{k-1})] \leq f(b) - f(a)$$

□

In other words, for monotonic functions the sum of jumps is bounded.

Theorem 4.1.2. *Let f be a monotonic function defined on closed interval $[a, b]$. Then the set of discontinuities of f is countable.*

Proof. Without loss of generality, let f be an increasing function on $[a, b]$. Let S_m be the set of all points on $[a, b]$ at which the jump exceeds $\frac{1}{m}$.

We know that, the sum of jumps of an increasing function is bounded above by $f(b) - f(a)$. Thus, cardinality of S_m given by,

$$|S_m| < m[f(b) - f(a)]$$

is finite for any positive integer m .

If f is discontinuous at a point $x \in [a, b]$, then there exists some integer m' such that $0 < \frac{1}{m'} < x$ and $x \in S_{m'}$.

$$\text{Number of discontinuities} = \left| \bigcup_{m=1}^{\infty} S_m \right| \leq \sum_{m=1}^{\infty} |S_m| \text{ is countable.}$$

since countable sum of finite values is countable. Therefore, the number of discontinuities of f is countable. \square

4.1.3 Function of Bounded Variation

Definitions 4.1.1 (partition). Let $[a, b]$ be a compact interval. Let $x_0 = a$, $x_0 < x_1 < x_2 < \dots < x_n$ and $x_n = b$. Then $P = \{x_0, x_1, \dots, x_n\}$ is a **partition** of $[a, b]$. And (x_{k-1}, x_k) is the **k th subinterval** of the partition.

Definitions 4.1.2 (bounded variation). Let f be a function defined on closed interval $[a, b]$. If there exists a positive real-number M such that

$$\sum_{k=1}^n |\Delta f_k| = \sum_{k=1}^n |f(x_k) - f(x_{k-1})| \leq M$$

for any partition P on $[a, b]$. Then f is a function of **bounded variation**, where (x_{k-1}, x_k) is the k th subinterval of the partition.

In other words, a function is of bounded variation on $[a, b]$ if the sum of variations is bounded for any (finite) partition of $[a, b]$.

Theorem 4.1.3. Let f be a monotonic function on $[a, b]$. Then f is of bounded variation on $[a, b]$.

Proof. Without loss of generality, let f be an increasing function. Then for any partition $P = \{x_0, x_1, \dots, x_n\}$ of $[a, b]$, we have

$$\sum_{k=1}^n [f(x_k) - f(x_{k-1})] \leq f(b) - f(a) = M$$

Therefore, f is of bounded variation on $[a, b]$. \square

Theorem 4.1.4. Let f be a continuous function on $[a, b]$ and its derivative f' exists and f' is bounded in (a, b) . Then f is of bounded variation.

Proof. Let f be a continuous function with derivative f' on (a, b) . Since f is continuous and f' exists, from intermediate value theorem we have

$$\Delta f_k = f(x_k) - f(x_{k-1}) = f'(t_k)[x_k - x_{k-1}] \text{ where } t_k \in (x_{k-1}, x_k)$$

Since f' is bounded, $f'(t_k) \leq M'$ for any $t_k \in (a, b)$. Thus,

$$\sum_{k=1}^n |\Delta f_k| = \sum_{k=1}^n |f'(t_k)|(x_k - x_{k-1}) \leq M' \sum_{k=1}^n (x_k - x_{k-1}) = M'(b - a) = M$$

Therefore, f is of bounded variation. \square

Theorem 4.1.5. *Let f be a function on $[a, b]$. If f is of bounded variation, then f is bounded.*

Proof. Let $x \in (a, b)$. Consider the partition $P = \{a, x, b\}$. Since f is of bounded variation, there exists a positive real-number M such that

$$\sum_{k=1}^2 |\Delta f_k| = |f(x) - f(a)| + |f(b) - f(x)| \leq M$$

Clearly, $|f(x) - f(a)| \leq M$, since $|f(b) - f(x)| > 0$. We have,

$$|f(x)| = |f(x) - f(a) + f(a)| \leq |f(x) - f(a)| + |f(a)| \leq M + |f(a)|$$

Suppose $x = a$, then $|f(x)| = |f(a)|$.

Suppose $x = b$, then $|f(x)| = |f(b)| = M' + |f(a)|$ where $M' = |f(b)| - |f(a)|$.

Therefore, the function f is bounded on $[a, b]$. \square

4.1.4 Total Variation

Definitions 4.1.3 (total variation). Let f be a function of bounded variation on $[a, b]$. Let $\Sigma(P)$ be the sum of variations with respect to the partition P of $[a, b]$. Then the **total variation** of the function f on $[a, b]$ is given by,

$$V_f(a, b) = V_f = \sup\{\Sigma(P) : P \in \mathcal{P}[a, b]\}$$

Note 1 : V_f is finite, since f is of bounded variation on $[a, b]$.

Note 2 : $V_f \geq 0$, since $|\Delta f_k| \geq 0$ for any subinterval of $[a, b]$.

Note 3 : $V_f = 0$ if only if f is a constant function on $[a, b]$. (Why ?)

Theorem 4.1.6. *Let f, g be functions of bounded variation on $[a, b]$. Then their sum $f + g$, difference $f - g$, and product fg are of bounded variation. Also,*

$$V_{f \pm g} \leq V_f + V_g \quad \text{and} \quad V_{fg} \leq AV_f + BV_g$$

where $A = \sup\{|g(x)| : x \in [a, b]\}$ and $B = \sup\{|f(x)| : x \in [a, b]\}$.

Proof. Let f, g be functions of bounded variation on $[a, b]$. Then f, g are bounded and $\sup|f(x)|$ and $\sup|g(x)|$ exists.

Step 1 : $V_{f \pm g} \leq V_f + V_g$

We have,

$$|(f + g)(x_k) - (f + g)(x_{k-1})| \leq |f(x_k) - f(x_{k-1})| + |g(x_k) - g(x_{k-1})|$$

Then

$$\begin{aligned} V_{f+g} &= \sup_{P \in \mathcal{P}} \sum_{k=1}^n |(f + g)(x_k) - (f + g)(x_{k-1})| \\ &\leq \sup_{P \in \mathcal{P}} \sum_{k=1}^n |f(x_k) - f(x_{k-1})| + \sup_{P \in \mathcal{P}} \sum_{k=1}^n |g(x_k) - g(x_{k-1})| \\ &= V_f + V_g \end{aligned}$$

Similarly, we have $V_{f-g} \leq V_f + V_g$ since

$$|(f-g)(x_k) - (f-g)(x_{k-1})| \leq |f(x_k) - f(x_{k-1})| + |g(x_k) - g(x_{k-1})|$$

Step 2 : $V_{fg} \leq AV_f + BV_g$

We have,

$$\begin{aligned} |fg(x_k) - fg(x_{k-1})| &= |f(x_k)g(x_k) - f(x_{k-1})g(x_{k-1})| \\ &= |f(x_k)g(x_k) - f(x_{k-1})g(x_k) + f(x_{k-1})g(x_k) - f(x_{k-1})g(x_{k-1})| \\ &\leq |g(x_k)| |f(x_k) - f(x_{k-1})| + |f(x_{k-1})| |g(x_k) - g(x_{k-1})| \\ &\leq A|f(x_k) - f(x_{k-1})| + B|g(x_k) - g(x_{k-1})| \end{aligned}$$

where $A = \sup\{|g(x)| : x \in [a, b]\}$ and $B = \sup\{|f(x)| : x \in [a, b]\}$.

Therefore,

$$\begin{aligned} V_{fg} &\leq A \sup_{P \in \mathcal{P}} \left\{ \sum_{k=1}^n |f(x_k) - f(x_{k-1})| \right\} + B \sup_{P \in \mathcal{P}} \left\{ \sum_{k=1}^n |g(x_k) - g(x_{k-1})| \right\} \\ &= AV_f + BV_g \end{aligned}$$

□

Definitions 4.1.4 (bounded away from zero). A function f is bounded away from zero on $[a, b]$ if there exists a real-number m such that $0 < m \leq f(x)$, $\forall x \in [a, b]$.

Let f be a function of bounded variation. Then $\frac{1}{f}$ is of bounded variation if and only if f is bounded away from zero. (Why ?)

Theorem 4.1.7. Let f be a function of bounded variation on $[a, b]$ and f is bounded away from zero. Then, $g = \frac{1}{f}$ is a function of bounded variation on $[a, b]$ and $V_g \leq \frac{V_f}{m^2}$ whenever $0 < m \leq |f(x)|$.

Proof. Suppose f is of bounded variation on $[a, b]$ and f is bounded away from zero. That is, there exists positive real-number m such that $0 < m \leq |f(x)|$, $\forall x \in [a, b]$. Then, $\frac{1}{|f(x)|} \leq \frac{1}{m}$, $\forall x \in [a, b]$.

Define $g = \frac{1}{f}$. Then, we have

$$|\Delta g_k| = \left| \frac{1}{f(x_k)} - \frac{1}{f(x_{k-1})} \right| = \frac{|f(x_k) - f(x_{k-1})|}{|f(x_k)| |f(x_{k-1})|} \leq \frac{|f(x_k) - f(x_{k-1})|}{m^2}$$

The total variation of g on $[a, b]$ is given by,

$$\begin{aligned} V_g &= \sup_{P \in \mathcal{P}} \left\{ \sum_{k=1}^n |g(x_k) - g(x_{k-1})| \right\} \\ &\leq \frac{1}{m^2} \sup_{P \in \mathcal{P}} \left\{ \sum_{k=1}^n |f(x_k) - f(x_{k-1})| \right\} \\ &\leq \frac{V_f}{m^2} \text{ where } 0 < m \leq |f(x)|, \forall x \in [a, b] \end{aligned}$$

□

In other words, if h, f are functions of bounded variation of $[a, b]$. And f is bounded away from zero, then $g = \frac{1}{f}$ and $\frac{h}{f} = hg$ is a function of bounded variation and $V_{\frac{h}{f}} = V_{hg} \leq AV_h + BV_f$.

Now, we have analyzed sum, difference, product and quotient of functions of bounded variation. And we are not surprised about preservation of bounded variation under function composition. (Why ?)

4.1.5 Additive Property of Total Variation

Theorem 4.1.8 (additive property). *Let f be a function of bounded variation on $[a, b]$. Let $c \in (a, b)$. Then f is of bounded variation on both $[a, c]$ and $[c, b]$. And, $V_f(a, b) = V_f(a, c) + V_f(c, b)$.*

Proof. Let f be a function of bounded variation on $[a, b]$. Let P_1, P_2 be partitions of $[a, c]$ and $[c, b]$ respectively. Then $P_0 = P_1 \cup P_2$ is a partition of $[a, b]$. Thus,

$$\sum(P_1) + \sum(P_2) = \sum(P_0) \leq V_f(a, b)$$

Taking supremums on the left side, we get

$$V_f(a, c) + V_f(c, b) \leq V_f(a, b)$$

Clearly, $V_f(a, c) \leq V_f(a, b)$ and $V_f(c, b) \leq V_f(a, b)$. Therefore, function f is of bounded variation on both $[a, c]$ and $[c, b]$.

Let P be a partition of $[a, b]$. Then $P_0 = P \cup \{c\}$ is refinement of P . Now, we have two partitions P_1, P_2 of $[a, c]$ and $[c, b]$ such that $P_0 = P_1 \cup P_2$. Thus,

$$\sum(P) \leq \sum(P_0) = \sum(P_1) + \sum(P_2) \leq V_f(a, c) + V_f(c, b)$$

Taking supremum on the left side, we get

$$V_f(a, b) \leq V_f(a, c) + V_f(c, b)$$

Therefore,

$$V_f(a, b) = V_f(a, c) + V_f(c, b), \quad \forall c \in (a, b)$$

□

4.1.6 Total Variation on $[a, x]$ as a function of x

By additive property, we have existence of $V_f(a, x)$ for every $x \in (a, b]$. Assigning $V(x) = V_f(a, x)$, we have a well-defined function on $(a, b]$.

Theorem 4.1.9. *Let f be a function of bounded variation on $[a, b]$. Define $V : [a, b] \rightarrow \mathbb{R}$ given by,*

$$V(x) = \begin{cases} V_f(a, x) & x \in (a, b] \\ 0 & x = a \end{cases}$$

Then,

1. V is an increasing function on $[a, b]$.
2. $V - f$ is an increasing function of $[a, b]$.

Proof. Suppose $x = a$. Then $V(x) = 0$ and $V(y) = V_f(a, y) \geq 0 = V(x)$. Suppose $x \neq a$. Then, $a < x < y \leq b$. By additive property of total variation, we have $V(y) = V_f(a, y) = V_f(a, x) + V_f(x, y) = V(x) + V_f(x, y)$. Since $V_f(x, y) \geq 0$, we have $V(x) \leq V(y)$. Therefore, V is an increasing function on $[a, b]$.

Define $D : [a, b] \rightarrow \mathbb{R}$ given by $D(x) = (V - f)(x) = V(x) - f(x)$. Suppose $a \leq x < y \leq b$. Then, $D(y) - D(x) = V(y) - f(y) - V(x) + f(x) = [V(y) - V(x)] - [f(y) - f(x)]$. We have, $V(y) = V_f(a, y) = V_f(a, x) + V_f(x, y)$. Thus, $V(y) - V(x) = V_f(x, y)$. Also we have, f is of bounded variation on $[x, y]$. Consider the trivial partition $P = \{x, y\}$ of $[x, y]$. Then, we have $f(y) - f(x) = \sum(P) \leq V_f(x, y)$. Thus, $D(y) - D(x) = V_f(x, y) - [f(y) - f(x)] \geq 0$. Therefore, $D = V - f$ is an increasing function on $[a, b]$. \square

4.1.7 Function of bounded variation expressed as the difference of increasing functions

Theorem 4.1.10. *Let f be a function on $[a, b]$. Function f is of bounded variation on $[a, b]$ if and only if f can be expressed as difference of two increasing functions.*

Proof. Let f be a function of bounded variation, then $f = V - D$ where total variation V and $D = V - f$ are both increasing.

Let f be function on $[a, b]$. Let $f = V - D$ where V, D are increasing functions. Then V, D are of bounded variation, since monotonic functions on $[a, b]$ are of bounded variation. Also, we have $V - D$ is of bounded variation, since for any two functions of bounded variation their difference is also of bounded variation. \square

4.1.8 Continuous functions of bounded variation

Theorem 4.1.11. *Let f be a function of bounded variation on $[a, b]$. Let V be the total variation function of f defined on $[a, b]$. f is continuous at a point if and only if V is continuous at that point.*

Proof. Let f be a function of bounded variation on $[a, b]$. Let V be the total variation of f . Let $x, y \in [a, b]$, such that $x < y$. We have V is an increasing function on $[a, b]$. And f is difference of two increasing functions on $[a, b]$. Thus, $f(x+)$, $f(x-)$, $V(x+)$, $V(x-)$ exists for any $x \in (a, b)$. It remains to prove that V, f are continuous at $x \in (a, b)$.

Part 1 : V continuous $\implies f$ continuous

Suppose $x \neq a$, then $a < x < y \leq b$. Let P be any partition on $[a, x]$. Then there exists a partition P' on $[a, y]$ such that $P \subset P'$. Consider, $P' = P \cup \{y\}$. Then, $V(y) > V(x)$ and $V(y) - V(x) \geq |f(y) - f(x)|$. Thus,

$$0 \leq |f(y) - f(x)| \leq V(y) - V(x)$$

The inequality is true for any $y > x$. Therefore,

$$0 \leq |f(x+) - f(x)| \leq V(x+) - V(x) \text{ as } y \rightarrow x+$$

Similarly, let $a < z < x$. Then,

$$0 \leq |f(x) - f(x-)| \leq V(x) - V(x-) \text{ as } z \rightarrow x-$$

Clearly, if V is continuous at x then f is continuous at x .

Part 2 : f continuous $\implies V$ continuous

Suppose f is continuous at $c \in (a, b)$. Let $\varepsilon > 0$. We have,

$$V_f(c, b) = \sup_{P \in \mathcal{P}} \sum(P)$$

Thus, there exists a partition $P_1 \in \mathcal{P}[c, b]$ such that $V_f(c, b) - \frac{\varepsilon}{2} < \sum(P_1)$. Since f is continuous at c , there exists $x_1 \in (c, b)$ such that $|f(x_1) - f(c)| < \frac{\varepsilon}{2}$. Then,

$$V_f(c, b) - \frac{\varepsilon}{2} < \frac{\varepsilon}{2} + V_f(x_1, b)$$

We have,

$$\begin{aligned} V(x_1) - V(c) &= V_f(a, x_1) - V_f(a, c) = V_f(c, x_1) \\ &= V_f(c, b) - V_f(x_1, b) < \varepsilon \end{aligned}$$

Clearly, $V(c+h) \rightarrow V(c)$ as $h \rightarrow 0$. That is, $V(c+) = V(c)$, $\forall c \in [a, b]$. Similarly we have,

$$V_f(a, c) = \sup_{P \in \mathcal{P}} \sum(P)$$

And there exists $P_2 \in \mathcal{P}[a, c]$ such that

$$V_f(a, c) - \frac{\varepsilon}{2} < \sum(P_2)$$

And there exists $x_2 \in (a, c)$ such that $|f(c) - f(x_2)| < \frac{\varepsilon}{2}$, since f is continuous at c . Therefore,

$$V(c) - V(x_2) = V_f(a, c) - V_f(a, x_2) = V(x_2, c) < \varepsilon$$

Thus, $V(c-) = V(c)$, $\forall c \in (a, b]$. Therefore, V is continuous at c . \square

Theorem 4.1.12. *Let f be a continuous function on $[a, b]$. Function f is of bounded variation on $[a, b]$ if and only if f can be expressed as difference of two increasing continuous functions.*

Proof. Let f be a continuous function on $[a, b]$. Then total variation V is a continuous increasing function on $[a, b]$. Clearly, $D = V - f$ is also a continuous, increasing function on $[a, b]$. Therefore, $f = V - D$ where V, D are continuous, increasing functions.

Let V, D be continuous, increasing functions on $[a, b]$. Then $f = V - D$ is also a continuous function on $[a, b]$. We have, V, D are increasing functions, therefore both V, D are of bounded variation and their difference f is also of bounded variation. \square

Suppose f is of bounded variation on $[a, b]$. Let $id : [a, b] \rightarrow [a, b]$ where $id(x) = x$. Then $V + id$ is a strictly increasing function and $D = V + id - f$ is also strictly increasing. Thus, any function of bounded variation on $[a, b]$ can be characterised as difference of two strictly increasing continuous functions on $[a, b]$.

4.1.9 Curves and Paths

Definitions 4.1.5 (path). Let $f : [a, b] \rightarrow \mathbb{R}^n$ be a continuous, vector-valued function. Then f is a path in \mathbb{R}^n . And f is a motion if $[a, b]$ is a time interval.

4.1.10 Rectifiable paths and Arc length

Definitions 4.1.6 (rectifiable path). Let $f : [a, b] \rightarrow \mathbb{R}^n$ be a path in \mathbb{R}^n . Let $P = \{t_0, t_1, \dots, t_m\}$ be a partition of $[a, b]$.

$$\Lambda_f(P) = \sum_{k=1}^m \|f(t_k) - f(t_{k-1})\| = \sum_{k=1}^m \|\Delta f_k\|$$

If $\Lambda_f(P)$ is bounded for any partition $P \in \mathcal{P}[a, b]$, then path f is **rectifiable**. If $\Lambda_f(P)$ is unbounded, then f is **nonrectifiable**.

Definitions 4.1.7 (arc length). Let $f : [a, b] \rightarrow \mathbb{R}^n$ be a rectifiable path. Then **arc length** of path f is given by,

$$\Lambda_f(a, b) = \sup_{P \in \mathcal{P}} \Lambda_f(P)$$

Theorem 4.1.13. A path $f : [a, b] \rightarrow \mathbb{R}^n$ is rectifiable if and only if each component f_k of f is of bounded variation on $[a, b]$. Let $V_k(a, b)$ be the total variation of f_k on $[a, b]$. Then,

$$V_k(a, b) \leq \Lambda_f(a, b) \leq V_1(a, b) + V_2(a, b) + \dots + V_n(a, b)$$

Proof. Let $x_j \in \mathbb{R}^n$ for $j = 1, 2, \dots, m$. Then,

$$|x_r| \leq \sqrt{\sum_{j=1}^n |x_j|^2} \leq \sum_{j=1}^n |x_j| \text{ since } \sum_{j=1}^n x_j^2 \leq \left(\sum_{j=1}^n x_j \right)^2$$

Let $f : [a, b] \rightarrow \mathbb{R}^n$ be a path in \mathbb{R}^n . Then $f = (f_1, f_2, \dots, f_n)$ where f_k 's are components of the path f . Let $P = \{t_0, t_1, \dots, t_m\}$ be a partition of $[a, b]$. Now $f_r(t_j) - f_r(t_{j-1}) \in \mathbb{R}^n$ for each subinterval of $[a, b]$ and each component of f . Thus for each subinterval (t_j, t_{j-1}) we have,

$$|f_r(t_j) - f_r(t_{j-1})| \leq \|f(t_j) - f(t_{j-1})\| \leq \sum_{k=1}^n |f_k(t_j) - f_k(t_{j-1})|$$

Adding inequalities for every subinterval of the partition, we get

$$\sum_{k=1}^m |f_k(t_j) - f_k(t_{j-1})| \leq \sum_{k=1}^m \|f(t_j) - f(t_{j-1})\| \leq \sum_{k=1}^m \sum_{j=1}^n |f_k(t_j) - f_k(t_{j-1})|$$

Rearranging summation, we get

$$\sum(P) \leq \Lambda_f(P) \leq \sum_{j=1}^n \left(\sum(P)_{f_j} \right)$$

Suppose f is a rectifiable path. Then f has finite arc length $\Lambda_f(a, b)$. Let f_k be a component function of f . Let P be a partition of $[a, b]$. Then,

$$\sum(P) \leq \Lambda_f(P) \leq \Lambda_f(a, b)$$

Thus, $\sum(P)$ is bounded for any partition P . Therefore, total variation $V_k(a, b)$ exists for each component function f_k . And $V_k(a, b) \leq \Lambda_f(a, b)$.

Suppose f_k 's are of bounded variation. Then

$$\Lambda_f(P) \leq \sum_{j=1}^n \sum(P) \leq \sum_{j=1}^n V_j(a, b)$$

is bounded for any partition P . Thus, f is rectifiable. And, arc length $\Lambda_f(a, b) \leq \sum_{k=1}^n V_k(a, b)$. Combining the inequalities, we get

$$V_k(a, b) \leq \Lambda_f(a, b) \leq \sum_{j=1}^n V_j(a, b)$$

□

4.1.11 Additive and Continuity Properties of Arc length

Theorem 4.1.14 (additive). *Let $f : [a, b] \rightarrow \mathbb{R}^n$ be a rectifiable path. Let $c \in (a, b)$. Then,*

$$\Lambda_f(a, b) = \Lambda_f(a, c) + \Lambda_f(c, b)$$

Proof. Let P be a partition of $[a, b]$. Then $P' = P \cup \{c\}$ is a refinement of P such that $P' = P_1 \cup P_2$ where P_1, P_2 are partition of $[a, c]$ and $[c, b]$ respectively. We have,

$$\Lambda_f(P) \leq \Lambda_f(P') = \Lambda_f(P_1) + \Lambda_f(P_2)$$

This inequality is true for any partition of $[a, b]$. Thus,

$$\Lambda_f(a, b) \leq \Lambda_f(a, c) + \Lambda_f(c, b)$$

Let P_1, P_2 be partition of $[a, c]$ and $[c, b]$ respectively. Then,

$$\Lambda_f(P_1) + \Lambda_f(P_2) \leq \Lambda_f(P) \leq \Lambda_f(a, b)$$

This inequality is true for any partitions on $[a, c]$ and $[c, b]$. Thus,

$$\Lambda_f(a, c) + \Lambda_f(c, b) \leq \Lambda_f(a, b)$$

□

Theorem 4.1.15 (continuity). *Let $f : [a, b] \rightarrow \mathbb{R}^n$ be rectifiable path. Let function $s : [a, b] \rightarrow \mathbb{R}$ defined by*

$$s(x) = \begin{cases} 0 & x = a \\ \Lambda_f(a, x) & x \in (a, b] \end{cases}$$

Then,

1. *function s is continuous and increasing on $[a, b]$.*
2. *if there is no subinterval of $[a, b]$ in which f is constant, then s is strictly increasing.*

Proof. Let $a \leq x < y \leq b$. Then, $s(y) - s(x) = \Lambda_f(x, y) = \Lambda(a, y) - \Lambda(a, x) \geq 0$. Therefore, s is increasing.

Suppose f is not constant in any subinterval $[x, y]$ of $[a, b]$. Suppose s is not strictly increasing. Then, there exists $x, y \in (a, b)$ such that $x < y$ and $s(y) - s(x) = 0$. Thus,

$$\Lambda_f(a, y) - \Lambda(a, x) = \Lambda_f(x, y) = 0$$

Thus, $V_k(x, y) = 0$, $\forall k$ which is a contradiction since f is not constant in $[x, y]$. Therefore, s is strictly increasing. \square

4.1.12 Equivalence of path, Change of parameter

Definitions 4.1.8 (change of parameter). Let $f : [a, b] \rightarrow \mathbb{R}^n$ be a path. Let $g : [c, d] \rightarrow \mathbb{R}^n$ be another path. Then f, g are **equivalent** if there exists a continuous, real-valued function, $u : [c, d] \rightarrow [a, b]$ such that $g = f \circ u$. That is, $g(t) = f(u(t))$, $\forall t \in [c, d]$. In other words, f, g are different parametric representations of a common graph.

Function u defines a change of parameter. If u is strictly increasing, then f, g are in the same direction. And u is an orientation preserving change of parameter. If u is strictly decreasing, then f, g are in opposite directions. And u is an orientation reversing change of parameter.

Theorem 4.1.16 (change of parameter). *Let $f : [a, b] \rightarrow \mathbb{R}^n$ and $g : [c, d] \rightarrow \mathbb{R}^n$ be two paths. Let f, g be both injective functions. Then f and g are equivalent if they have the same graph.*

Proof. Let $f : [a, b] \rightarrow \mathbb{R}^n$ and $g : [c, d] \rightarrow \mathbb{R}^n$ be continuous, injective, vector-valued functions. Suppose f, g are equivalent paths, then f, g have the same graph.

Suppose f, g have the same graph. Since f is injective and continuous on its domain $[a, b]$, function f^{-1} exists and is continuous on its graph.

$$\text{Define } u : [c, d] \rightarrow [a, b], \quad u(t) = f^{-1}(g(t))$$

Then u is continuous and $g(t) = f(u(t))$. Suppose u is not a strictly monotonic function. Since u is continuous, there exists $t_1, t_2 \in [c, d]$ such that $u(t_1) = u(t_2)$. Then $f(u(t_1)) = f(u(t_2)) \implies g(t_1) = g(t_2)$ which is a contradiction since g is injective on $[c, d]$. \square

Module 2 : Riemann-Stieltjes Integral

4.2 The Riemann-Stieltjes Integral

Definitions 4.2.1 (unit step). The unit step function $I : \mathbb{R} \rightarrow \mathbb{R}$ is defined by

$$I(x) = \begin{cases} 0 & x \leq 0 \\ 1 & x > 0 \end{cases}$$

Definitions 4.2.2 (Riemann Integral). Let f be a bounded real function defined on $[a, b]$. Let $P = \{x_0, x_1, \dots, x_n\}$ be a partition of $[a, b]$. Let $M_k = \sup\{f(x) : x \in [x_{k-1}, x_k]\}$ and $m_k = \inf\{f(x) : x \in [x_{k-1}, x_k]\}$.

Then Riemann upper sum of function f with respect to partition P ,

$$U(P, f) = \sum_P M_k \Delta x_k = \sum_{k=1}^n M_k (x_k - x_{k-1})$$

And Riemann lower sum of f with respect to P ,

$$L(P, f) = \sum_{k=1}^n m_k (x_k - x_{k-1})$$

Now, Riemann upper integral of f over $[a, b]$,

$$\int_a^b f \, dx = \inf\{U(P, f) : P \in \mathcal{P}[a, b]\}$$

And, Riemann lower integral of f over $[a, b]$,

$$\int_a^b f \, dx = \sup\{L(P, f) : P \in \mathcal{P}[a, b]\}$$

A function f is Riemann integrable over $[a, b]$ if Riemann lower and upper integrals of f over $[a, b]$ are the same. Then Riemann integral of f over $[a, b]$,

$$\int_a^b f \, dx = \int_a^b f \, dx = \int_a^b f \, dx$$

Definitions 4.2.3 (Riemann-Stieltjes Integral). Let f be a bounded function on $[a, b]$. Let α be an increasing function on $[a, b]$. Let $P = \{x_0, x_1, \dots, x_n\}$ be a partition of $[a, b]$. Then, the Riemann-Stieltjes upper sum of f with respect to partition P and increasing function α ,

$$U(P, f, \alpha) = \sum_{k=1}^n M_k \Delta \alpha_k$$

where $M_k = \sup\{f(x) : x \in [x_{k-1}, x_k]\}$ and $\Delta \alpha_k = \alpha(x_k) - \alpha(x_{k-1})$. Similarly, Riemann-Stieltjes lower sum,

$$L(P, f, \alpha) = \sum_{k=1}^n m_k \Delta \alpha_k$$

where $m_k = \inf\{f(x) : x \in [x_{k-1}, x_k]\}$. And function f is Riemann-Stieltjes integrable if Riemann Stieltjes upper and lower integrals are the same.

$$\int_a^b f d\alpha = \int_a^{\bar{b}} f d\alpha = \int_{\underline{a}}^b f d\alpha$$

where $\int_a^{\bar{b}} f d\alpha = \int_a^{\bar{b}} f d\alpha(x) = \inf\{U(P, f, \alpha) : P \in \mathcal{P}[a, b]\}$ and $\int_{\underline{a}}^b f d\alpha = \sup\{L(P, f, \alpha) : P \in \mathcal{P}[a, b]\}$.

We write $f \in \mathcal{R}(\alpha)$ on $[a, b]$, which means that a bounded, real function f is Riemann-Stieltjes integrable on $[a, b]$ with respect to the increasing function α .

Note : function $\alpha : [a, b] \rightarrow \mathbb{R}$ is monotonic (increasing), but not necessarily continuous.

Remark : With $\alpha = id$ identity function, Riemann-Stieltjes integral is Riemann integral itself. That is, Riemann integral is a special case of Riemann-Stieltjes integral.

Theorem 4.2.1. Let P^* be a refinement of a partition P of $[a, b]$. Then, $L(P, f, \alpha) \leq L(P^*, f, \alpha)$ and $U(P^*, f, \alpha) \leq U(P, f, \alpha)$.

Proof. Let $P^* = P \cup \{x\}$ where x belongs to the i th subinterval (x_{j-1}, x_j) . Define $w_1 = \min\{f(x) : x \in (x_{j-1}, x)\}$ and $w_2 = \min\{f(x) : x \in (x, x_j)\}$. Clearly, $w_1, w_2 \geq \min\{f(x) : x \in (x_{j-1}, x_j)\} = m_i$.

$$\begin{aligned} L(P, f, \alpha) - L(P^*, f, \alpha) &= m_i \Delta\alpha_j - w_1(\alpha(x) - \alpha(x_{j-1})) - w_2(\alpha(x_j) - \alpha(x)) \\ &= (m_i - w_1)(\alpha(x) - \alpha(x_{j-1})) + (m_i - w_2)(\alpha(x_j) - \alpha(x)) \\ &\geq 0 \end{aligned}$$

By mathematical induction the inequality is true for any refinement P^* of P . Therefore, $L(P, f, \alpha) \geq L(P^*, f, \alpha)$.

Similarly, define $W_1 = \max\{f(x) : x \in (x_{j-1}, x)\}$ and $W_2 = \max\{f(x) : x \in (x, x_j)\}$ where $W_1, W_2 \leq \max\{f(x) : x \in (x_{j-1}, x_j)\} = M_i$.

$$\begin{aligned} U(P, f, \alpha) - U(P^*, f, \alpha) &= M_i \Delta\alpha_j - W_1(\alpha(x) - \alpha(x_{j-1})) - W_2(\alpha(x_j) - \alpha(x)) \\ &= (M_i - W_1)(\alpha(x) - \alpha(x_{j-1})) + (M_i - W_2)(\alpha(x_j) - \alpha(x)) \\ &\leq 0 \end{aligned}$$

Again, the result is true for any refinement P and we have

$$U(P, f, \alpha) \leq U(P^*, f, \alpha)$$

□

Theorem 4.2.2. Let f be a bounded, real function on $[a, b]$ and α increasing function on $[a, b]$. Then,

$$\int_a^b f d\alpha \leq \int_a^{\bar{b}} f d\alpha$$

Proof. Let P_1, P_2 be any two partition of $[a, b]$. Let $P^* = P_1 \cup P_2$ be a refinement of both partitions. Then, we have $L(P^*, f, \alpha) \leq U(P^*, f, \alpha)$. And,

$$\int_a^b f \, d\alpha \leq L(P_1, f, \alpha) \text{ and } U(P_2, f, \alpha) \leq \int_a^b f \, d\alpha$$

Therefore,

$$L(P_1, f, \alpha) \leq L(P^*, f, \alpha) \leq U(P^*, f, \alpha) \leq U(P_2, f, \alpha)$$

Clearly, the inequality holds independent of the choice of the partition. Taking supremum on right and infimum on left, we get

$$\sup_{P \in \mathcal{P}} L(P, f, \alpha) \leq \inf_{P \in \mathcal{P}} U(P, f, \alpha)$$

Therefore,

$$\int_a^b f \, d\alpha \leq \int_a^b f \, d\alpha$$

□

Theorem 4.2.3 (criterion for integrability). *Let f be a bounded, real function on $[a, b]$. Let α be an increasing function on $[a, b]$. Then, f is Riemann-Stieltjes integrable on $[a, b]$ with respect to α if and only if for every $\varepsilon > 0$ there exists a partition P of $[a, b]$ such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$.*

In other words, $f \in \mathcal{R}(\alpha)$ on $[a, b]$ if and only if

$$\forall \varepsilon > 0, \exists P \in \mathcal{P}[a, b] \text{ such that } U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$$

Proof. Let $\varepsilon > 0$. Suppose there exists a partition P of $[a, b]$ such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$. We have,

$$L(P, f, \alpha) \leq \int_a^b f \, d\alpha \leq \int_a^b f \, d\alpha \leq U(P, f, \alpha)$$

$$\int_a^b f \, d\alpha - \int_a^b f \, d\alpha \leq U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$$

Clearly, $\int_a^b f \, d\alpha = \int_a^b f \, d\alpha$. Therefore, $f \in \mathcal{R}(\alpha)$.

Suppose $f \in \mathcal{R}(\alpha)$ on $[a, b]$. Then by the definition of infimum there exists a partition P_1 of $[a, b]$ such that

$$U(P_1, f, \alpha) - \int_a^b f \, d\alpha < \frac{\varepsilon}{2}$$

Similarly, there exists a partition P_2 of $[a, b]$ such that

$$\int_a^b f \, d\alpha - L(P_2, f, \alpha) < \frac{\varepsilon}{2}$$

Consider $P^* = P_1 \cup P_2$. Clearly, P^* is a refinement for both P_1 and P_2 . Thus,

$$U(P^*, f, \alpha) - \int_a^{\bar{b}} f \, d\alpha < \frac{\varepsilon}{2}$$

And,

$$\int_a^b f \, d\alpha - L(P^*, f, \alpha) < \frac{\varepsilon}{2}$$

Thus,

$$U(P^*, f, \alpha) - \int_a^{\bar{b}} f \, d\alpha + \int_a^b f \, d\alpha - L(P^*, f, \alpha) < \varepsilon$$

Given, $f \in \mathcal{R}(\alpha)$ on $[a, b]$. Therefore, $U(P^*, f, \alpha) - L(P^*, f, \alpha) < \varepsilon$. \square

Theorem 4.2.4. Suppose $\varepsilon > 0$ and $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$ for some partition P of $[a, b]$.

1. The inequality is true for any refinement of P .
2. Let $s_i, t_i \in [x_{i-1}, x_i]$ for each subinterval of the partition P .

$$\sum_{i=1}^n |f(s_i) - f(t_i)| \Delta\alpha_i < \varepsilon$$

3. If $f \in \mathcal{R}(\alpha)$ and $t_i \in [x_{i-1}, x_i]$ for each subinterval of the partition P , then

$$\left| \sum_{i=1}^n f(t_i) \Delta\alpha_i - \int_a^b f \, d\alpha \right| < \varepsilon$$

Proof. 1. Let $\varepsilon > 0$. Suppose $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$. Let P^* be a refinement of P . We have $U(P, f, \alpha) \geq U(P^*, f, \alpha)$ and $L(P, f, \alpha) \leq L(P^*, f, \alpha)$. Thus,

$$U(P^*, f, \alpha) - L(P^*, f, \alpha) \leq U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$$

2. Let $s_i, t_i \in (x_{i-1}, x_i)$. Clearly, for each subinterval (x_{i-1}, x_i) , we have $m_i \leq f(s_i) \leq M_i$ and $m_i \leq f(t_i) \leq M_i$. Thus,

$$|f(t_i) - f(s_i)| \leq M_i - m_i$$

$$\sum_{i=1}^n |f(t_i) - f(s_i)| \Delta\alpha_i \leq \sum_{i=1}^n (M_i - m_i) \Delta\alpha_i \leq U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$$

3. Let $\varepsilon > 0$. Let P be a partition of $[a, b]$ such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$. We have,

$$L(P, f, \alpha) \leq \int_a^b f \, d\alpha \leq \int_a^{\bar{b}} f \, d\alpha \leq U(P, f, \alpha)$$

Suppose $f \in \mathcal{R}(\alpha)$ over $[a, b]$. Then,

$$L(P, f, \alpha) \leq \int_a^b f \, d\alpha \leq U(P, f, \alpha)$$

Let $t_i \in (x_{i-1}, x_i)$ for each subinterval of the partition P . Clearly, $m_i \leq f(t_i) \leq M_i$. Thus, we also have

$$L(P, f, \alpha) \leq \sum_{i=1}^n f(t_i) \Delta \alpha_i \leq U(P, f, \alpha)$$

Therefore,

$$\left| \sum_{i=1}^n f(t_i) \Delta \alpha_i - \int_a^b f d\alpha \right| \leq U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$$

□

Theorem 4.2.5. *If f is continuous on $[a, b]$, then $f \in \mathcal{R}(\alpha)$ on $[a, b]$.*

Proof. Let f be a continuous function on $[a, b]$. Since continuous functions defined on compact subsets metric spaces are uniformly continuous, we have f is uniformly continuous.

Let $\varepsilon > 0$. Choose $\eta > 0$ such that $[\alpha(b) - \alpha(a)]\eta < \varepsilon$. Since f is uniformly continuous, there exists $\delta > 0$ such that $|f(x) - f(t)| < \eta$ whenever $|x - t| < \delta$. Consider the partition P such that each subinterval is of length less than δ . For any $s_i, t_i \in (x_{i-1}, x_i)$, we have

$$\begin{aligned} \sum_{i=1}^n [f(t_i) - f(s_i)] [\alpha(x_i) - \alpha(x_{i-1})] &\leq \sum_{i=1}^n \eta [\alpha(x_i) - \alpha(x_{i-1})] \\ &\leq \eta \sum_{i=1}^n [\alpha(x_i) - \alpha(x_{i-1})] \\ &\leq \eta [\alpha(b) - \alpha(a)] \\ &< \varepsilon \end{aligned}$$

Clearly, the result is true for minimum and maximum values of f in each subinterval of P . Therefore, we have

$$U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$$

Thus, $f \in \mathcal{R}(\alpha)$ over $[a, b]$. □

Theorem 4.2.6. *If f is monotonic on $[a, b]$ and α is continuous on $[a, b]$, then $f \in \mathcal{R}(\alpha)$ on $[a, b]$.*

Proof. Let f be increasing on $[a, b]$. Let α is continuous on $[a, b]$. Let n be any integer. We can construct a partition P of $[a, b]$ such that each the variation of α in each subinterval is fixed. That is, $\Delta \alpha_i = \frac{\alpha(b) - \alpha(a)}{n}$. Since f is increasing, in each subinterval $[x_{i-1}, x_i]$, we have $M_i = f(x_i)$ and $m_i = f(x_{i-1})$. Now we

have,

$$\begin{aligned}
 U(P, f, \alpha) - L(P, f, \alpha) &= \sum_{i=1}^n M_i \Delta \alpha_i - \sum_{i=1}^n m_i \Delta \alpha_i \\
 &= \sum_{i=1}^n (M_i - m_i) \Delta \alpha_i \\
 &= \frac{\alpha(b) - \alpha(a)}{n} \sum_{i=1}^n (f(x_i) - f(x_{i-1})) \\
 &= \frac{\alpha(b) - \alpha(a)}{n} (f(b) - f(a))
 \end{aligned}$$

Thus, given $\varepsilon > 0$, there exists an integer n such that $(\alpha(b) - \alpha(a))(f(b) - f(a)) < n\varepsilon$. And, we get a partition P by fixing $\Delta \alpha_i = \frac{\alpha(b) - \alpha(a)}{n}$ such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$.

In other words, $U(P, f, \alpha) - L(P, f, \alpha)$ depends on n and can be reduced to a value less than ε by increasing the value of n . Therefore, $f \in \mathcal{R}(\alpha)$ over $[a, b]$. \square

Theorem 4.2.7. *If f bounded on $[a, b]$, f has only finitely many points of discontinuities on $[a, b]$ and α is continuous at every point at which f is discontinuous. Then $f \in \mathcal{R}(\alpha)$.*

Proof. Suppose f is bounded and has only finitely many discontinuities on $[a, b]$. Since f is bounded, we have $-M < f(x) < M$ for some real number M . Also, suppose that α is continuous at those points where f is discontinuous on $[a, b]$.

Let E be the set of all discontinuities of f on $[a, b]$. Clearly, $|E|$ is finite.

Let $\varepsilon > 0$. Since, α is continuous at each point e_j of E , there exists open intervals (u_j, v_j) such that α is continuous on those intervals, e_j belongs to the interior of these intervals and the sum of variation of α in those intervals is less than ε .

Removing these open intervals from $[a, b]$, we get a compact subset K in which f is continuous. Since K is compact and f is continuous on K , we have f is uniformly continuous on K .

Given $\varepsilon > 0$, there exists $\delta > 0$ such that $|f(x) - f(t)| < \varepsilon$ whenever $|x - t| < \delta$. Define a partition P such that $u_j, v_j \in P$, P doesn't have any point in the interior of any (u_j, v_j) . And if $x_{i-1} \neq u_j$, then x_i is so chosen that $x_i - x_{i-1} < \delta$, dividing K into subintervals of length less than δ . Now, we have

$$\begin{aligned}
 U(P, f, \alpha) - L(P, f, \alpha) &= \sum_{i=1}^n (M_i - m_i) \Delta \alpha_i \\
 &\leq \varepsilon \sum_{x_{i-1} \neq u_j} \Delta \alpha_i + 2M \sum_{x_{i-1} = u_j} \Delta \alpha_i
 \end{aligned}$$

$$\begin{aligned}
&= \varepsilon \sum_{x_{i-1} \neq u_j} (\alpha(x_i) - \alpha(x_{i-1})) + 2M\varepsilon \\
&\leq \varepsilon(\alpha(b) - \alpha(a)) + 2M\varepsilon
\end{aligned}$$

Clearly, $U(P, f, \alpha) - L(P, f, \alpha)$ is a function of ε which can be reduced below any real number greater than zero by reducing the value of ε . Therefore, $f \in \mathcal{R}(\alpha)$ over $[a, b]$. \square

Theorem 4.2.8. Suppose $f \in \mathcal{R}(\alpha)$, $m \leq f \leq M$ on $[a, b]$, ϕ is continuous on $[m, M]$ and $h(x) = \phi(f(x))$. Then $h \in \mathcal{R}(\alpha)$ on $[a, b]$.

Proof. Let f be a function on $[a, b]$ such that $m \leq f(x) \leq M$ and $f \in \mathcal{R}(\alpha)$ over $[a, b]$. Let ϕ be a continuous function on $[m, M]$. Then ϕ is uniformly continuous on $[m, M]$, since $[m, M]$ is compact. Thus, given $\varepsilon > 0$ there exists $\delta > 0$ such that $|\phi(x) - \phi(t)| < \varepsilon$ whenever $|x - t| < \delta$. Without loss of generality, we may assume that $\delta < \varepsilon$. Otherwise choose a value less than ε as δ .

Since $f \in \mathcal{R}(\alpha)$ over $[a, b]$, there exists a partition P of $[a, b]$ such that

$$U(P, f, \alpha) - L(P, f, \alpha) < \delta^2$$

Consider the i th subinterval $[x_{i-1}, x_i]$ of the partition P . Let M_i, m_i be the maximum and minimum values of the function f in i th subinterval of P . Now, we divided the collection of subintervals into two sets depending on the value of $M_i - m_i$ in comparison with δ . Define A as the set of all subintervals of P such that $M_i - m_i < \delta$. And B as the set of all subintervals of P such that $M_i - m_i \geq \delta$. We know that,

$$\delta \sum_B \Delta\alpha_i \leq \sum_B (M_i - m_i) \Delta\alpha_i < U(P, f, \alpha) - L(P, f, \alpha) < \delta^2$$

Thus, $\sum_B \Delta\alpha_i < \delta$.

Define M_i^*, m_i^* as the maximum and minimum of the function $\phi \circ f$ each subinterval $[x_{i-1}, x_i]$ of the partition P . Now, we have

$$\begin{aligned}
U(P, \phi \circ f, \alpha) - L(P, \phi \circ f, \alpha) &= \sum_{i=1}^n (M_i^* - m_i^*) \Delta\alpha_i \\
&= \sum_A (M_i^* - m_i^*) \Delta\alpha_i + \sum_B (M_i^* - m_i^*) \Delta\alpha_i
\end{aligned}$$

Since ϕ is uniformly continuous,

$$\leq \varepsilon \sum_A \Delta\alpha_i + \sum_B (M_i^* - m_i^*) \Delta\alpha_i$$

Since, continuous functions defined on compact subset attains extrema. We have, $K = \sup\{|\phi(x)| : x \in [m, M]\}$. And $M_i^* - m_i^* < 2K$.

$$\begin{aligned}
&\leq \varepsilon(\alpha(b) - \alpha(a)) + 2K \sum_B \Delta\alpha_i \\
&\leq \varepsilon(\alpha(b) - \alpha(a)) + 2K\delta \\
&\leq \varepsilon(\alpha(b) - \alpha(a)) + 2K\varepsilon
\end{aligned}$$

Since $U(P, \phi \circ f, \alpha) - L(P, \phi \circ f, \alpha)$ is function of ε , it can be reduced to any sufficiently small real number of our choice. Therefore, $\phi \circ f \in \mathcal{R}(\alpha)$ over $[a, b]$. \square

4.2.1 Properties of the Riemann-Stieltjes Integral

Theorem 4.2.9. *Let f, f_1, f_2 be bounded real functions on $[a, b]$. Let $\alpha, \alpha_1, \alpha_2$ be increasing functions on $[a, b]$.*

1. *If $f_1, f_2, f \in \mathcal{R}(\alpha)$ on $[a, b]$, then $f_1 + f_2 \in \mathcal{R}(\alpha)$ on $[a, b]$. And,*

$$\int_a^b (f_1 + f_2) d\alpha = \int_a^b f_1 d\alpha + \int_a^b f_2 d\alpha$$

If $c \in \mathbb{R}$, then $cf \in \mathcal{R}(\alpha)$ on $[a, b]$. And,

$$\int_a^b cf d\alpha = c \int_a^b f d\alpha$$

2. *If $f_1(x) \leq f_2(x)$ on $[a, b]$, then*

$$\int_a^b f_1 d\alpha \leq \int_a^b f_2 d\alpha$$

3. *If $c \in (a, b)$, then $f \in \mathcal{R}(\alpha)$ on $[a, c]$ and $[c, b]$, then*

$$\int_a^c f d\alpha + \int_c^b f d\alpha = \int_a^b f d\alpha$$

4. *If $|f(x)| \leq M$ on $[a, b]$, then*

$$\left| \int_a^b f d\alpha \right| \leq M[\alpha(b) - \alpha(a)]$$

5. *If $f \in \mathcal{R}(\alpha_1)$ and $f \in \mathcal{R}(\alpha_2)$ on $[a, b]$, then $f \in \mathcal{R}(\alpha_1 + \alpha_2)$. And,*

$$\int_a^b f d(\alpha_1 + \alpha_2) = \int_a^b f d\alpha_1 + \int_a^b f d\alpha_2$$

If $f \in \mathcal{R}(\alpha)$ on $[a, b]$, and $c \in \mathbb{R}$, then $f \in \mathcal{R}(c\alpha)$ on $[a, b]$. And,

$$\int_a^b f d(c\alpha) = c \int_a^b f d\alpha$$

Proof. 1. Let $\varepsilon > 0$. Let $f_1, f_2 \in \mathcal{R}(\alpha)$ over $[a, b]$.

Let P_1 be a partition of $[a, b]$ such that $U(P_1, f_1, \alpha) - L(P_1, f_1, \alpha) < \frac{\varepsilon}{2}$.

Let P_2 be a partition of $[a, b]$ such that $U(P_2, f_2, \alpha) - L(P_2, f_2, \alpha) < \frac{\varepsilon}{2}$.

Let $P = P_1 \cup P_2$ be the refinement of both the partitions. Then the above inequalities are true of the partition P as well.

We have,

$$\begin{aligned} L(P, f_1, \alpha) + L(P, f_2, \alpha) &\leq L(P, f_1 + f_2, \alpha) \\ &\leq U(P, f_1 + f_2, \alpha) \\ &\leq U(P, f_1, \alpha) + U(P, f_2, \alpha) \end{aligned}$$

Thus,

$$\begin{aligned} U(P, f_1 + f_2, \alpha) - L(P, f_1 + f_2, \alpha) \\ \leq U(P, f_1, \alpha) - L(P, f_1, \alpha) + U(P, f_2, \alpha) - L(P, f_2, \alpha) \\ \leq \varepsilon \end{aligned}$$

Therefore, $f_1 + f_2 \in \mathcal{R}(\alpha)$ over $[a, b]$.

2. Let c be any real number greater than zero. We have,

$$M'_i = \max\{cf(x) : x \in [x_{i-1}, x_i]\} = c \max\{f(x) : x \in [x_{i-1}, x_i]\} = cM_i$$

Similarly, $m'_i = cm_i$. Thus,

$$\sum_{i=1}^n M'_i \Delta\alpha_i = c \sum_{i=1}^n M_i \Delta\alpha_i \text{ and } \sum_{i=1}^n m'_i \Delta\alpha_i = c \sum_{i=1}^n m_i \Delta\alpha_i$$

Therefore, given $\varepsilon > 0$ we have,

$$\begin{aligned} L(P, cf, \alpha) &= cL(P, f, \alpha) \\ U(P, cf, \alpha) &= cU(P, f, \alpha) \end{aligned}$$

Since, $f \in \mathcal{R}(\alpha)$, there exists partition P such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$. Thus,

$$U(P, cf, \alpha) - L(P, cf, \alpha) = cU(P, f, \alpha) - cL(P, f, \alpha) < c\varepsilon$$

Therefore, $cf \in \mathcal{R}(\alpha)$ over $[a, b]$.

$$\begin{aligned} \int_a^b cf \, d\alpha &\leq U(P, cf, \alpha) \\ &\leq cU(P, f, \alpha) \\ &\leq c \int_a^b f \, d\alpha + c\varepsilon \\ &\leq c \int_a^b f \, d\alpha \end{aligned}$$

Similarly, considering $-f$ and $-cf$ we get

$$\int_a^b cf \, d\alpha \geq c \int_a^b f \, d\alpha$$

Therefore, without loss of generality for any real number c ,

$$\int_a^b cf \, d\alpha = c \int_a^b f \, d\alpha$$

3. Let $\varepsilon > 0$. Suppose $f_1, f_2 \in \mathcal{R}(\alpha)$ and $f_1 \leq f_2$. Let P_1, P_2 be the partitions such that $U(P_1, f_1, \alpha) - L(P_1, f_1, \alpha) < \varepsilon$ and $U(P_2, f_2, \alpha) - L(P_2, f_2, \alpha) < \varepsilon$. Consider the common refinement $P = P_1 \cup P_2$. Then, for each subinterval of the partition P , we have

$$\min_{x \in [x_{i-1}, x_i]} f_1(x) \leq \min_{x \in [x_{i-1}, x_i]} f_2(x) \text{ and } \max_{x \in [x_{i-1}, x_i]} f_1(x) \leq \max_{x \in [x_{i-1}, x_i]} f_2(x)$$

Thus, $L(P, f_1, \alpha) \leq L(P, f_2, \alpha)$ and $U(P, f_1, \alpha) \leq U(P, f_2, \alpha)$.

$$\begin{aligned} \int_a^b f_1 \, d\alpha &\leq U(P, f_1, \alpha) \\ &\leq U(P, f_2, \alpha) \\ &\leq \int_a^b f_2 \, d\alpha + \varepsilon \end{aligned}$$

Therefore,

$$\int_a^b f_1 \, d\alpha \leq \int_a^b f_2 \, d\alpha$$

4. Let $f \in \mathcal{R}(\alpha)$ on $[a, b]$. Let $\varepsilon > 0$. Then, there exists a partition P of $[a, b]$ such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$. Let $c \in (a, b)$. Then $P^* = P \cup \{c\} = P_1 \cup P_2$ is refinement of P such that P_1, P_2 are partition of $[a, c]$ and $[c, b]$ respectively. Clearly, $U(P_1, f, \alpha) - L(P_1, f, \alpha) < \varepsilon$ and $U(P_2, f, \alpha) - L(P_2, f, \alpha) < \varepsilon$. Therefore, $f \in \mathcal{R}(\alpha)$ on both $[a, c]$ and $[c, b]$.

For any two partitions P_1, P_2 of $[a, c]$ and $[c, b]$, there exists partition $P^* = P_1 \cup P_2$ of $[a, b]$. Thus,

$$\begin{aligned} \int_a^b f \, d\alpha &\leq U(P^*, f, \alpha) \\ &\leq U(P_1, f, \alpha) + U(P_2, f, \alpha) \\ &\leq \int_a^c f \, d\alpha + \int_c^b f \, d\alpha + 2\varepsilon \end{aligned}$$

Since ε is arbitrary, we have

$$\int_a^b f \, d\alpha \leq \int_a^c f \, d\alpha + \int_c^b f \, d\alpha$$

And considering $-f$ instead of f , we get

$$\int_a^b -f \, d\alpha \leq \int_a^c -f \, d\alpha + \int_c^b -f \, d\alpha$$

Thus,

$$\int_a^b f \, d\alpha \geq \int_a^c f \, d\alpha + \int_c^b f \, d\alpha$$

Therefore,

$$\int_a^b f \, d\alpha = \int_a^c f \, d\alpha + \int_c^b f \, d\alpha$$

5. Let function $f \in \mathcal{R}(\alpha)$ on $[a, b]$ and $|f| \leq M$. Let P be any partition of $[a, b]$. We have,

$$L(P, f, \alpha) \leq \int_a^b f \, d\alpha \leq U(P, f, \alpha)$$

And,

$$L(P, f, \alpha) = \sum_{i=1}^n m_i \Delta\alpha_i \geq -M \Delta\alpha_i = -M \sum_{i=1}^n \Delta\alpha_i = -M[\alpha(b) - \alpha(a)]$$

Similarly,

$$U(P, f, \alpha) = \sum_{i=1}^n M_i \Delta\alpha_i \leq M \Delta\alpha_i = M \sum_{i=1}^n \Delta\alpha_i = M[\alpha(b) - \alpha(a)]$$

Therefore,

$$\left| \int_a^b f \, d\alpha \right| \leq M[\alpha(b) - \alpha(a)]$$

6. Let α_1, α_2 be monotonic functions on $[a, b]$. Then $|\alpha_1 + \alpha_2| \leq |\alpha_1| + |\alpha_2|$. Let $f \in \mathcal{R}(\alpha_1)$ on $[a, b]$ and $f \in \mathcal{R}(\alpha_2)$ on $[a, b]$. Let $\varepsilon > 0$. Then there exists partitions P_1, P_2 of $[a, b]$ such that

$$U(P_1, f, \alpha_1) - L(P_1, f, \alpha_1) < \varepsilon$$

$$U(P_2, f, \alpha_2) - L(P_2, f, \alpha_2) < \varepsilon$$

Consider the common refinement $P = P_1 \cup P_2$. Then, the inequalities are true for P as well. And for each subinterval $[x_{i-1}, x_i]$ of P , we have $\Delta(\alpha_1 + \alpha_2)_i \leq \Delta(\alpha_1)_i + \Delta(\alpha_2)_i$. Thus,

$$\begin{aligned} U(P, f, \alpha_1 + \alpha_2) - L(P, f, \alpha_1 + \alpha_2) &= \sum_{i=1}^n (M_i - m_i) \Delta(\alpha_1 + \alpha_2)_i \\ &\leq \sum_{i=1}^n (M_i - m_i) \Delta\alpha_{1,i} + \sum_{i=1}^n (M_i - m_i) \Delta\alpha_{2,i} \\ &\leq 2\varepsilon \end{aligned}$$

Therefore, $f \in \mathcal{R}(\alpha_1 + \alpha_2)$ on $[a, b]$.

$$\begin{aligned} \int_a^b f \, d(\alpha_1 + \alpha_2) &\leq U(P, f, \alpha_1 + \alpha_2) \\ &\leq U(P, f, \alpha_1) + U(P, f, \alpha_2) \\ &\leq \int_a^b f \, d\alpha_1 + \int_a^b f \, d\alpha_2 + 2\varepsilon \end{aligned}$$

Thus,

$$\int_a^b f d(\alpha_1 + \alpha_2) \leq \int_a^b f d\alpha_1 + \int_a^b f d\alpha_2$$

Considering $-f$ instead of f , we get

$$\int_a^b f d(\alpha_1 + \alpha_2) \geq \int_a^b f d\alpha_1 + \int_a^b f d\alpha_2$$

Therefore,

$$\int_a^b f d(\alpha_1 + \alpha_2) = \int_a^b f d\alpha_1 + \int_a^b f d\alpha_2$$

Let $c > 0$ and $f \in \mathcal{R}(\alpha)$ on $[a, b]$. Let $\varepsilon > 0$. Then there exists a partition P of $[a, b]$ such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$. Clearly, $U(P, f, c\alpha) - L(P, f, c\alpha) < c\varepsilon$. Therefore, $f \in \mathcal{R}(c\alpha)$ on $[a, b]$.

$$\begin{aligned} \int f d(c\alpha) &\leq U(P, f, c\alpha) \\ &\leq cU(P, f, \alpha) \\ &\leq c \int f d\alpha + c\varepsilon \end{aligned}$$

Thus,

$$\int f d(c\alpha) \leq c \int f d\alpha$$

Taking $-f$ instead of f , we get

$$\int f d(c\alpha) \geq c \int f d\alpha$$

Therefore,

$$\int f d(c\alpha) = c \int f d\alpha$$

□

Theorem 4.2.10. If $f, g \in \mathcal{R}(\alpha)$ on $[a, b]$, then

1. $fg \in \mathcal{R}(\alpha)$ on $[a, b]$
2. $|f| \in \mathcal{R}(\alpha)$ on $[a, b]$. And,

$$\left| \int_a^b f d\alpha \right| \leq \int_a^b |f| d\alpha$$

Proof. 1. Let $f, g \in \mathcal{R}(\alpha)$ on $[a, b]$. By linearity of the integral we have $f + g \in \mathcal{R}(\alpha)$ on $[a, b]$. Let $\phi(t) = t^2$. Then ϕ is continuous. Thus, $\phi \circ f$ is continuous on $[a, b]$. Therefore, $\phi \circ f \in \mathcal{R}(\alpha)$ on $[a, b]$. Also we have, $\phi \circ (f + g) = (f + g)^2$ and $\phi \circ (f - g) = (f - g)^2$. Thus, $(f + g)^2, (f - g)^2 \in \mathcal{R}(\alpha)$ on $[a, b]$. Now, $4fg = (f + g)^2 - (f - g)^2$. Therefore, $4fg \in \mathcal{R}(\alpha)$ on $[a, b]$, from linearity of the integral. Take $c = \frac{1}{4}$, we get $fg \in \mathcal{R}(\alpha)$ on $[a, b]$ from linearity of the integral.

2. Let $f \in \mathcal{R}(\alpha)$ on $[a, b]$. Let $\phi(t) = |t|$. Then ϕ is continuous. Therefore, $\phi \circ f = |f| \in \mathcal{R}(\alpha)$ on $[a, b]$.

Let $c = \pm 1$ such that

$$c \int f d\alpha \geq 0$$

Then,

$$\left| \int f d\alpha \right| = c \int f d\alpha = \int c f d\alpha \leq \int |f| d\alpha$$

since $c f \leq |f|$.

□

Definitions 4.2.4 (step). The unit step function, $I : \mathbb{R} \rightarrow [0, 1]$ is defined by

$$I(x) = \begin{cases} 0 & x \leq 0 \\ 1 & x > 0 \end{cases}$$

Theorem 4.2.11. Let f be bounded on $[a, b]$ and continuous at $s \in (a, b)$. Let $\alpha = I(t - s)$. Then,

$$\int_a^b f d\alpha = f(s)$$

Proof. Let $P = \{a, s, x_2, b\}$ be a partition of $[a, b]$. Since f is bounded,

$$U(P, f, \alpha) = M_1 \Delta\alpha_1 + M_2 \Delta\alpha_2 + M_3 \Delta\alpha_3 = M_2$$

since $\Delta\alpha_1 = 0$, we have $\Delta\alpha_2 = \alpha(x_2) - \alpha(s) = I(x_2 - s) - I(0) = 1 - 0$ and $\Delta\alpha_3 = 0$. Similarly, $L(P, f, \alpha) = m_1 \Delta\alpha_1 + m_2 \Delta\alpha_2 + m_3 \Delta\alpha_3 = m_2$. We have,

$$m_2 = L(P, f, \alpha) \leq \int_a^b f d\alpha \leq \int_a^b f d\alpha \leq U(P, f, \alpha) = M_2$$

We know that, M_2 is the maximum value of f in the subinterval $[s, x_2]$. The lower sum and upper sum remains the same as we reduce the length of the second subinterval. We also know that, f is continuous at s . As $x_2 \rightarrow s$, the maximum value of f tends to the value of f at s , say $f(s)$. Similarly, $m_2 \rightarrow f(s)$ as $x_2 \rightarrow s$. Thus,

$$f(s) \leq \int_a^b f d\alpha \leq \int_a^b f d\alpha \leq f(s)$$

Therefore, $f \in \mathcal{R}(\alpha)$ on $[a, b]$ and

$$\int_a^b f d\alpha = f(s)$$

□

Theorem 4.2.12. Suppose $\sum_{i=1}^{\infty} c_n$ converges and $c_n \geq 0$. Let f be continuous on $[a, b]$. Let sequence $\{s_n\}$ be a strictly increasing sequence in (a, b) . Let $\alpha = \sum_{i=1}^{\infty} c_n I(t - s_n)$. Then,

$$\int_a^b f d\alpha = \sum_{i=1}^{\infty} c_n f(s_n)$$

Proof. Let $\sum_{n=1}^{\infty} c_n$ be a convergent series. Given $\varepsilon > 0$, there exists a natural number N such that

$$\sum_{n=N+1}^{\infty} c_n < \varepsilon$$

Let s_1, s_2, s_3, \dots be distinct points in (a, b) . Without loss of generality, sequence $\{s_n\}$ is a strictly increasing sequence in $[a, b]$. Let $\alpha = \sum_{n=1}^{\infty} c_n I(t - s_n)$. Then, $\alpha(a) = 0$ and $\alpha(b) = \sum_{n=1}^{\infty} c_n$.

$$0 \leq \alpha(x) \leq \sum_{n=1}^{\infty} c_n, \quad \forall x \in [a, b]$$

By comparison test, α is convergent in $[a, b]$. We have,

$$\int_a^b f(t) d(c_1 I(t - s_1)) = c_1 \int_a^b f(t) d(I(t - s_1)) = c_1 f(s_1)$$

Let $\alpha = \alpha_1 + \alpha_2$ where $\alpha_1 = \sum_{n=1}^N c_n I(t - s_n)$ and $\alpha_2 = \sum_{n=N+1}^{\infty} c_n I(t - s_n)$. And from mathematical induction, we have

$$\begin{aligned} \int_a^b f(t) d(\alpha_1) &= \int_a^b f(t) d\left(\sum_{n=1}^N c_n I(t - s_n)\right) \\ &= \sum_{n=1}^N \int_a^b f(t) d(c_n I(t - s_n)) \\ &= \sum_{n=1}^N c_n \int_a^b f dI(t - s_n) \\ &= \sum_{n=1}^N c_n f(s_n) \end{aligned}$$

We have,

$$\left| \int_a^b f(t) d\alpha_2 \right| \leq M(\alpha_2(b) - \alpha_2(a)) = M \sum_{n=N+1}^{\infty} c_n < M\varepsilon$$

where $M = \sup |f(x)|$. Thus,

$$\begin{aligned} \left| \int_a^b f d\alpha_2 \right| &= \left| \int_a^b f d(\alpha_1 + \alpha_2) - \int_a^b f d\alpha_1 \right| \\ &= \left| \int_a^b f d\alpha - \sum_{n=1}^N c_n f(s_n) \right| \\ &< M\varepsilon \end{aligned}$$

The inequality is true as $N \rightarrow \infty$. In other words, the sequence of partial sums converges to the value of integral of f . Therefore,

$$\sum_{n=1}^{\infty} c_n f(s_n) = \int_a^b f d\alpha$$

□

Theorem 4.2.13. *Let α be increasing function on $[a, b]$ and $\alpha' \in \mathcal{R}$. Let f be bounded real-valued function on $[a, b]$. Then $f \in \mathcal{R}(\alpha)$ if and only if $f\alpha' \in \mathcal{R}$. And*

$$\int_a^b f d\alpha = \int_a^b f(x)\alpha'(x)dx$$

Proof. Since $\alpha' \in \mathcal{R}$ on $[a, b]$, for any $\varepsilon > 0$, there exists a partition P of $[a, b]$ such that $U(P, \alpha') - L(P, \alpha') < \varepsilon$. We have, α is continuous on $[a, b]$, as it is differentiable on $[a, b]$. Then by intermediate value theorem, we have

$$\Delta\alpha_i = \alpha(x_i) - \alpha(x_{i-1}) = \alpha'(t_i)(x_i - x_{i-1})$$

where $t_i \in [x_{i-1}, x_i]$. Let $s_i \in [x_{i-1}, x_i]$. Then,

$$\sum_{i=1}^n |\alpha'(s_i) - \alpha'(t_i)| \Delta x_i \leq U(P, \alpha') - L(P, \alpha') < \varepsilon$$

Let $M = \sup |f|$. Let $u_i \in [x_{i-1}, x_i]$.¹ Consider,

$$\begin{aligned} \left| \sum_{i=1}^n f(u_i) \Delta\alpha_i - \sum_{i=1}^n f(s_i) \alpha'(s_i) \Delta x_i \right| &\leq M \left| \sum_{i=1}^n \Delta\alpha_i - \alpha'(s_i) \Delta x_i \right| \\ &\leq M \left| \sum_{i=1}^n \alpha'(t_i) \Delta x_i - \alpha'(s_i) \Delta x_i \right| \\ &\leq M \sum_{i=1}^n |\alpha'(t_i) - \alpha'(s_i)| \Delta x_i \\ &< M\varepsilon \end{aligned}$$

Clearly, the inequality is true independent of the choice of $u_i, s_i \in [x_{i-1}, x_i]$. Thus, selecting u_i such that $f(u_i) = M_i$, we get

$$\left| U(P, f, \alpha) - \sum_{i=1}^n \alpha'(s_i) \Delta x_i \right| < M\varepsilon$$

Therefore,

$$\sum_{i=1}^n f(s_i) \alpha'(s_i) \Delta x_i < U(P, f, \alpha) + M\varepsilon$$

Selecting s_i such that $f\alpha'$ attains maximum at s_i , we get

$$U(P, f\alpha') < U(P, f, \alpha) + M\varepsilon$$

Now, first we select s_i such that $f\alpha(s_i)$ is maximum in i th subinterval of P . Then,

$$\left| \sum_{i=1}^n f(u_i) \Delta\alpha_i - U(P, f\alpha') \right| < M\varepsilon$$

¹In my opinion, using u_i instead of s_i in the first sum makes it simpler.

Therefore,

$$U(P, f\alpha') < \sum_{i=1}^n f(u_i)\Delta\alpha_i + M\varepsilon$$

And, now selecting u_i such that $f(u_i) = M_i$.

$$U(P, f\alpha') < U(P, f, \alpha) + M\varepsilon$$

Therefore, we have $U(P, f\alpha') = U(P, f, \alpha)$. Clearly, it is true for any refinement of P . Therefore,

$$\int_a^b f\alpha' dx = \int_a^b f d\alpha$$

Similary, by selecting u_i and s_i such that $f(u_i)$, $f\alpha'(s_i)$ is minimum in each subinterval of P , we get

$$\int_a^b f\alpha' dx = \int_a^b f d\alpha$$

Clearly, $f \in \mathcal{R}(\alpha) \iff f\alpha' \in \mathcal{R}$. And,

$$\int_a^b f\alpha' dx = \int_a^b f d\alpha$$

□

Theorem 4.2.14 (change of variable). *Let $\varphi : [A, B] \rightarrow [a, b]$ be a strictly increasing, continuous function onto $[a, b]$. Let $\alpha : [a, b] \rightarrow \mathbb{R}$ be an increasing function and $f \in \mathcal{R}(\alpha)$ on $[a, b]$. Define $\beta = \alpha \circ \varphi$ and $g = f \circ \varphi$. Then $g \in \mathcal{R}(\beta)$ on $[A, B]$ and*

$$\int_A^B g d\beta = \int_a^b f d\alpha$$

Proof. Let $P = \{x_0, x_1, \dots, x_n\}$ be any partition of $[a, b]$. Then there exists a partition $Q = \{y_0, y_1, \dots, y_n\}$ of $[A, B]$ such that $x_j = \varphi(y_j)$ since φ is a continuous, bijection. ^{†2} Similarly, for any partition $Q = \{y_0, y_1, \dots, y_n\}$ of $[A, B]$, there exists a partition $P = \{x_0, x_1, \dots, x_n\}$ of $[a, b]$ where $x_j = \varphi(y_j)$. Clearly, minimum/maximum of g in i th subinterval of Q is same as the minimum/maximum of f in i th subinterval of P . And

$$\Delta\beta_i = \beta(y_i) - \beta(y_{i-1}) = \alpha(\varphi(y_i)) - \alpha(\varphi(y_{i-1})) = \alpha(x_i) - \alpha(x_{i-1}) = \Delta\alpha_i$$

$$\begin{aligned} \min\{g(y) : y \in [y_{i-1}, y_i]\} &= \min\{f(\varphi(y)) : y \in [y_{i-1}, y_i]\} \\ &= \min\{f(x) : x \in [x_{i-1}, x_i]\} \\ \implies L(Q, g, \beta) &= L(P, f, \alpha) \\ \max\{g(y) : y \in [y_{i-1}, y_i]\} &= \max\{f(\varphi(y)) : y \in [y_{i-1}, y_i]\} \\ &= \max\{f(x) : x \in [x_{i-1}, x_i]\} \\ \implies U(Q, g, \beta) &= U(P, f, \alpha) \end{aligned}$$

²Strictly increasing functions are always injective.

Since $f \in \mathcal{R}(\alpha)$ on $[a, b]$, there exists a partition P of $[a, b]$ such that $U(P, f, \alpha) - L(P, f, \alpha) < \varepsilon$. Therefore, there exists a partition Q of $[A, B]$ such that $U(Q, g, \beta) - L(Q, g, \beta) < \varepsilon$. Thus, $g \in \mathcal{R}(\beta)$. And

$$\int_A^B g d\beta = \int_a^b f d\alpha$$

□

4.2.2 Integration and Differentiation

Theorem 4.2.15. *Let $f \in \mathcal{R}$ on $[a, b]$. Let $a \leq x \leq b$. Define*

$$F(x) = \int_a^x f(t) dt$$

Then F is continuous on $[a, b]$. Furthermore, if f is continuous at x_0 , then F is differentiable at x_0 . And $F'(x_0) = f(x_0)$.

Proof. Let $a \leq x < y \leq b$. Let $M = \sup |f|$. We have,

$$\begin{aligned} |F(y) - F(x)| &= \left| \int_a^y f(t) dt - \int_a^x f(t) dt \right| \\ &= \left| \int_a^x f(t) dt + \int_x^y f(t) dt - \int_a^x f(t) dt \right| \\ &= \left| \int_x^y f(t) dt \right| \\ &\leq M(y - x) \end{aligned}$$

Thus, given $\varepsilon > 0$, there exists $\delta > 0$ such that $|F(y) - F(x)| < \varepsilon$ whenever $|y - x| < \delta \leq \frac{\varepsilon}{M}$. Therefore, F is continuous on $[a, b]$.

Let f be continuous at $x_0 \in [a, b]$. Given $\varepsilon > 0$, there exists $\delta > 0$ such that $|f(t) - f(x_0)| < \varepsilon$ whenever $|t - x_0| < \delta$ and $t \in [a, b]$.

Let $x - \delta < s \leq x_0 \leq t < x + \delta$. Clearly, $|f(u) - f(x_0)| < \varepsilon$ whenever $u \in [s, t]$. Consider,

$$\begin{aligned} \left| \frac{F(t) - F(s)}{t - s} - f(x_0) \right| &= \left| \frac{1}{t - s} \left(\int_a^t f(u) du - \int_a^s f(u) du \right) - \frac{1}{t - s} f(x_0)(t - s) \right| \\ &= \left| \frac{1}{t - s} \int_s^t f(u) du - \frac{1}{t - s} \int_s^t f(x_0) du \right| \\ &= \left| \frac{1}{t - s} \left(\int_s^t (f(u) - f(x_0)) du \right) \right| \\ &< \frac{\varepsilon}{t - s} \int_s^t du \quad \text{since } |f(u) - f(x_0)| < \varepsilon \\ &< \varepsilon \end{aligned}$$

Clearly, as $\varepsilon \rightarrow 0$, $\delta \rightarrow 0$. Then, $s, t \rightarrow x_0$. Therefore, F differentiable at x_0 and

$$F'(x_0) = \lim_{s, t \rightarrow x_0} \frac{F(t) - F(s)}{t - s} = f(x_0)$$

□

Theorem 4.2.16 (fundamental theorem of calculus). *Let $f \in \mathcal{R}$ on $[a, b]$. Let F be a differentiable function on $[a, b]$ such that $F' = f$. Then,*

$$\int_a^b f(x)dx = F(b) - F(a)$$

Proof. Let $\varepsilon > 0$. Let $f \in \mathcal{R}$ on $[a, b]$. Then, there exists a partition $P = \{x_0, x_1, \dots, x_n\}$ of $[a, b]$ such that $U(P, f) - L(P, f) < \varepsilon$.

Let F be differentiable function such that $F' = f$. Then, F is continuous. By intermediate value theorem, there exists $t_i \in [x_{i-1}, x_i]$ such that

$$F(x_i) - F(x_{i-1}) = F'(t_i)(x_i - x_{i-1}) = f(t_i)\Delta x_i$$

Clearly,

$$F(b) - F(a) = \sum_{i=1}^n (F(x_i) - F(x_{i-1})) = \sum_{i=1}^n f(t_i)\Delta x_i$$

Since $t_i \in [x_{i-1}, x_i]$, we have $m_i \leq f(t_i) \leq M_i$. Thus,

$$\left| \int_a^b f(x)dx - (F(b) - F(a)) \right| = \left| \int_a^b f(x)dx - \sum_{i=1}^n f(t_i)\Delta x_i \right| < \varepsilon$$

Therefore,

$$\int_a^b f(x)dx = F(b) - F(a)$$

□

Theorem 4.2.17 (integration by parts). *Let F, G be differentiable $[a, b]$. Let $F' = f \in \mathcal{R}$ and $G' = g \in \mathcal{R}$. Then,*

$$\int_a^b F(x)g(x)dx = F(b)G(b) - F(a)G(a) - \int_a^b f(x)G(x)dx$$

Proof. Let F, G be differentiable functions on $[a, b]$ and $F' = f \in \mathcal{R}$ and $G' = g \in \mathcal{R}$. Let $H = FG$. Then, $H' = FG' + F'G = Fg + fG$.

$$\int_a^b H'(x)dx = \int_a^b F(x)g(x)dx + \int_a^b f(x)G(x)dx$$

By fundamental theorem of calculus, we also have

$$\int_a^b H'(x)dx = H(b) - H(a) = FG(b) - FG(a) = F(b)G(b) - F(a)G(a)$$

Rearranging the terms, we get

$$\int_a^b F(x)g(x)dx = F(b)G(b) - F(a)G(a) - \int_a^b f(x)G(x)dx$$

□

4.2.3 Integration of Vector-valued Functions

Definitions 4.2.5 (integrable). Let $\bar{f} : [a, b] \rightarrow \mathbb{R}^k$ be a vector-valued function. Let α be a monotonic function on $[a, b]$. Let f_1, f_2, \dots, f_k be the component functions of \bar{f} . That is, $\bar{f}(x) = (f_1(x), f_2(x), \dots, f_k(x))$. Then $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$ if and only if every component function $f_j \in \mathcal{R}(\alpha)$.

In other words, vector-valued function \bar{f} is integrable if and only if every component function of \bar{f} is integrable.

Theorem 4.2.18 (properties). Suppose \bar{f}, \bar{g} be vector-valued functions from $[a, b]$ into \mathbb{R}^k .

1. Let $\bar{f}, \bar{g} \in \mathcal{R}(\alpha)$ on $[a, b]$. Then $\bar{f} + \bar{g} \in \mathcal{R}(\alpha)$. And,

$$\int_a^b \bar{f} + \bar{g} d\alpha = \int_a^b \bar{f} d\alpha + \int_a^b \bar{g} d\alpha$$

If $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$ and $c \in \mathbb{R}$, then $c\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$. And,

$$\int_a^b c\bar{f} d\alpha = c \int_a^b \bar{f} d\alpha$$

2. Let $c \in (a, b)$. Let $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$. Then $\bar{f} \in \mathcal{R}(\alpha)$ on both $[a, c]$ and $[c, b]$. And,

$$\int_a^b \bar{f} d\alpha = \int_a^c \bar{f} d\alpha + \int_c^b \bar{f} d\alpha$$

3. Let α_1, α_2 be monotonic functions on $[a, b]$. Let $\bar{f} \in \mathcal{R}(\alpha_1)$ on $[a, b]$ and $\bar{f} \in \mathcal{R}(\alpha_2)$ on $[a, b]$. Then, $\bar{f} \in \mathcal{R}(\alpha_1 + \alpha_2)$ on $[a, b]$. And,

$$\int_a^b \bar{f} d(\alpha_1 + \alpha_2) = \int_a^b \bar{f} d\alpha_1 + \int_a^b \bar{f} d\alpha_2$$

Proof. 1. Let $\bar{f}, \bar{g} \in \mathcal{R}(\alpha)$ on $[a, b]$ where $\bar{f} = (f_1, f_2, \dots, f_k)$ and $\bar{g} = (g_1, g_2, \dots, g_k)$. By definition of integrability of vector-valued functions, the component functions $f_j, g_j \in \mathcal{R}(\alpha)$ on $[a, b]$ for $1 \leq j \leq k$.

We also know that, if $f_j, g_j \in \mathcal{R}(\alpha)$ on $[a, b]$, then $f_j + g_j \in \mathcal{R}(\alpha)$ on $[a, b]$. Thus, $f_1 + g_1, f_2 + g_2, \dots, f_k + g_k \in \mathcal{R}(\alpha)$ on $[a, b]$ for $1 \leq j \leq k$. Therefore, $\bar{f} + \bar{g} = (f_1 + g_1, f_2 + g_2, \dots, f_k + g_k) \in \mathcal{R}(\alpha)$ on $[a, b]$.

Let $c \in \mathbb{R}$. Let $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$. Then, $f_1, f_2, \dots, f_k \in \mathcal{R}(\alpha)$ on $[a, b]$. We know that, $cf_1 \in \mathcal{R}(\alpha)$ on $[a, b]$, since $f_1 \in \mathcal{R}(\alpha)$ on $[a, b]$. Thus, $cf_1, cf_2, \dots, cf_k \in \mathcal{R}(\alpha)$ on $[a, b]$. Therefore, $c\bar{f} = (cf_1, cf_2, \dots, cf_k) \in \mathcal{R}(\alpha)$ on $[a, b]$.

2. Let $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$. Then, $f_1, f_2, \dots, f_k \in \mathcal{R}(\alpha)$ on $[a, b]$. Let $c \in (a, b)$. We know that, if $f_j \in \mathcal{R}(\alpha)$ on $[a, b]$, then $f_j \in \mathcal{R}(\alpha)$ on both $[a, c]$ and $[c, b]$. Thus, $f_1, f_2, \dots, f_k \in \mathcal{R}(\alpha)$ on both $[a, c]$ and $[c, b]$. Therefore, $\bar{f} \in \mathcal{R}(\alpha)$ on both $[a, c]$ and $[c, b]$.

3. Let $\bar{f} \in \mathcal{R}(\alpha_1)$ and $\bar{f} \in \mathcal{R}(\alpha_2)$ on $[a, b]$. Then $f_1, f_2, \dots, f_k \in \mathcal{R}(\alpha_1)$ and $f_1, f_2, \dots, f_k \in \mathcal{R}(\alpha_2)$ on $[a, b]$. We know that, if $f_j \in \mathcal{R}(\alpha_1)$ and $f_j \in \mathcal{R}(\alpha_2)$ on $[a, b]$, then $f_j \in \mathcal{R}(\alpha_1 + \alpha_2)$ on $[a, b]$. □

Challenge 3. Let $\bar{f}, \bar{g} \in \mathcal{R}(\alpha)$ on $[a, b]$ where $\bar{f}, \bar{g} : [a, b] \rightarrow \mathbb{R}^k$. Define $\bar{f} \cdot \bar{g} : [a, b] \rightarrow \mathbb{R}$ by $(\bar{f} \cdot \bar{g})(x) = f_1(x)g_1(x) + f_2(x)g_2(x) + \dots + f_k(x)g_k(x)$. Then, $\bar{f} \cdot \bar{g} \in \mathcal{R}(\alpha)$ on $[a, b]$. And,

$$\int_a^b \bar{f} \cdot \bar{g} d\alpha = \left(\int_a^b \bar{f} d\alpha \right) \cdot \left(\int_a^b \bar{g} d\alpha \right)$$

Theorem 4.2.19. Let α be a monotonic function such that $\alpha' \in \mathcal{R}$ on $[a, b]$. Let \bar{f} be a bounded function on $[a, b]$. Then $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$ if and only if $\bar{f}\alpha' \in \mathcal{R}$ on $[a, b]$. And,

$$\int_a^b \bar{f} d\alpha = \int_a^b \bar{f}\alpha' dx$$

Proof. Let $\bar{f} = (f_1, f_2, \dots, f_k)$. Then, $\bar{f}\alpha' = (f_1\alpha', f_2\alpha', \dots, f_k\alpha')$. We already know that, $f_j \in \mathcal{R}(\alpha) \iff f_j\alpha' \in \mathcal{R}$. Therefore,

$$\begin{aligned} \bar{f} \in \mathcal{R}(\alpha) &\iff f_1, f_2, \dots, f_k \in \mathcal{R}(\alpha) \\ &\iff f_1\alpha', f_2\alpha', \dots, f_k\alpha' \in \mathcal{R} \\ &\iff \bar{f}\alpha' \in \mathcal{R} \end{aligned}$$

Thus, $\bar{f} \in \mathcal{R}(\alpha) \iff \bar{f}\alpha' \in \mathcal{R}$.

We also know that,

$$\int_a^b f_j d\alpha = \int_a^b f_j\alpha' dx$$

Thus,

$$\begin{aligned} \int_a^b \bar{f} d\alpha &= \left(\int_a^b f_1 d\alpha, \int_a^b f_2 d\alpha, \dots, \int_a^b f_k d\alpha \right) \\ &= \left(\int_a^b f_1\alpha' dx, \int_a^b f_2\alpha' dx, \dots, \int_a^b f_k\alpha' dx \right) \\ &= \int_a^b \bar{f}\alpha' dx \end{aligned}$$

□

Theorem 4.2.20. Let $\bar{f} \in \mathcal{R}$ on $[a, b]$ where $\bar{f} : [a, b] \rightarrow \mathbb{R}^k$. Define $\bar{F} : [a, b] \rightarrow \mathbb{R}^k$ defined by

$$\bar{F}(x) = \int_a^x \bar{f}(t) dt$$

Then \bar{F} is continuous on $[a, b]$. Furthermore, if \bar{f} is continuous at $x_0 \in [a, b]$, then \bar{F} is differentiable at x_0 and $\bar{F}'(x_0) = \bar{f}(x_0)$.

Proof. We know that, $\bar{f} = (f_1, f_2, \dots, f_k)$. And from the definition of integral, we have

$$\int_a^b \bar{f}(t) dt = \left(\int_a^b f_1(t) dt, \int_a^b f_2(t) dt, \dots, \int_a^b f_k(t) dt \right)$$

We know that, for $1 \leq j \leq k$, the function $F_j : [a, b] \rightarrow \mathbb{R}$ defined by

$$F_j(x) = \int_a^x f_j(t) dt$$

is continuous. And if f_j is continuous at x_0 , then F_j is differentiable at x_0 and $F_j'(x_0) = f_j(x_0)$. Clearly, $\bar{F} = (F_1, F_2, \dots, F_k)$ is continuous, since each component function is continuous. And, \bar{F} is differentiable at x_0 and

$$\bar{F}'(x_0) = (F_1'(x_0), F_2'(x_0), \dots, F_k'(x_0)) = (f_1(x_0), f_2(x_0), \dots, f_k(x_0)) = \bar{f}(x_0)$$

□

Theorem 4.2.21 (fundamental theorem of calculus for vector-valued functions). Let $\bar{f} : [a, b] \rightarrow \mathbb{R}^k$. Let $\bar{F} : [a, b] \rightarrow \mathbb{R}^k$. If $\bar{f} \in \mathcal{R}$ on $[a, b]$ and $\bar{F}' = \bar{f}$, then

$$\int_a^b \bar{f}(t) dt = \bar{F}(b) - \bar{F}(a)$$

Proof. By fundamental theorem of calculus, we have

$$\int_a^b f_j(t) dt = F_j(b) - F_j(a)$$

for $1 \leq j \leq k$. Therefore,

$$\begin{aligned} \int_a^b \bar{f}(t) dt &= \left(\int_a^b f_1(t) dt, \int_a^b f_2(t) dt, \dots, \int_a^b f_k(t) dt \right) \\ &= (F_1(b) - F_1(a), F_2(b) - F_2(a), \dots, F_k(b) - F_k(a)) \\ &= (F_1(b), F_2(b), \dots, F_k(b)) - (F_1(a), F_2(a), \dots, F_k(a)) \\ &= \bar{F}(b) - \bar{F}(a) \end{aligned}$$

□

Theorem 4.2.22. Let $\bar{f} : [a, b] \rightarrow \mathbb{R}^k$. If $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$, then $|\bar{f}| \in \mathcal{R}(\alpha)$ on $[a, b]$. And

$$\left| \int_a^b \bar{f} d\alpha \right| \leq \int_a^b |\bar{f}| d\alpha$$

Proof. Let $\bar{f} : [a, b] \rightarrow \mathbb{R}^k$. Then, we have $\bar{f} = (f_1, f_2, \dots, f_k)$. Suppose $\bar{f} \in \mathcal{R}(\alpha)$ on $[a, b]$. Then, $f_j \in \mathcal{R}(\alpha)$ on $[a, b]$ for $1 \leq j \leq k$. We have, $|\bar{f}| = (f_1^2 + f_2^2 + \dots + f_k^2)^{\frac{1}{2}}$. We know that if $f_j \in \mathcal{R}(\alpha)$, then $f_j^2 \in \mathcal{R}(\alpha)$. Again, $\sum f_j^2 \in \mathcal{R}(\alpha)$.

Consider $g : [a, b] \rightarrow \mathbb{R}$ given by $g(x) = \sum_{j=1}^k f_j^2(x)$. We have $g \in \mathcal{R}(\alpha)$ on $[a, b]$. Thus g is bounded and there exists m, M such that $m \leq g \leq M$. Clearly, $g \geq 0$.

Consider, the function $\phi : [m, M] \rightarrow \mathbb{R}$ given by $\phi(x) = \sqrt{x}$. Clearly, ϕ is well-defined on $[m, M]$ since $0 \leq m$. And, ϕ is continuous on $[m, M]$. Thus, $|\bar{f}| = \phi \circ g = \sqrt{\sum f_j^2} \in \mathcal{R}(\alpha)$ on $[a, b]$.

Let $\bar{y} = \int_a^b \bar{f} d\alpha$. Then, $\bar{y} = (y_1, y_2, \dots, y_k)$ and $y_j = \int_a^b f_j d\alpha$.

$$\begin{aligned} |\bar{y}|^2 &= y_1^2 + y_2^2 + \dots + y_k^2 \\ &= y_1 \int_a^b f_1 d\alpha + y_2 \int_a^b f_2 d\alpha + \dots + y_k \int_a^b f_k d\alpha \\ &= \int_a^b (y_1 f_1 + y_2 f_2 + \dots + y_k f_k) d\alpha \end{aligned}$$

By Schwarz inequality,

$$\sum_{j=1}^n y_j f_j \leq \left(\sum_{j=1}^n y_j^2 \right)^{\frac{1}{2}} \left(\sum_{j=1}^n f_j^2 \right)^{\frac{1}{2}} = |\bar{y}| |\bar{f}|$$

Thus,

$$|\bar{y}|^2 = \int_a^b \left(\sum_{j=1}^k y_j f_j \right) d\alpha \leq \int_a^b |\bar{y}| |\bar{f}| d\alpha = |\bar{y}| \int_a^b |\bar{f}| d\alpha$$

Therefore,

$$\left| \int_a^b \bar{f} d\alpha \right| = |\bar{y}| \leq \int_a^b |\bar{f}| d\alpha$$

□

Module 3

4.3 Sequence and Series of functions

Definitions 4.3.1. Let sequence $\{f_n\}$ be a sequence of functions defined on E . Suppose sequence $\{f_n(x)\}$ converges for every $x \in E$. Then, sequence $\{f_n\}$ converges. And **limit function** $f : E \rightarrow \mathbb{R}$ is defined by

$$f(x) = \lim_{n \rightarrow \infty} f_n(x)$$

Definitions 4.3.2. Let $f_n : E \rightarrow \mathbb{R}, \forall n \in \mathbb{N}$. Suppose $\sum f_n(x)$ converges for every $x \in E$. Then, series $\sum f_n$ converges. And **sum** $f : E \rightarrow \mathbb{R}$ is defined by

$$f(x) = \sum_{n=1}^{\infty} f_n(x)$$

4.3.1 Counter Examples - Illustrating Main problem

Main Problem : Can we obtain the sufficient conditions for preserving desirable properties under convergence ?

Interchanging limits

$$\text{Generally, } \lim_{n \rightarrow \infty} \lim_{m \rightarrow \infty} S_{n,m} \neq \lim_{m \rightarrow \infty} \lim_{n \rightarrow \infty} S_{n,m}$$

Proof. Consider, $S_{n,m} = \frac{m}{m+n}$.

$$\begin{aligned} \lim_{m \rightarrow \infty} \lim_{n \rightarrow \infty} S_{n,m} &= \lim_{m \rightarrow \infty} 0 = 0 \\ \lim_{n \rightarrow \infty} \lim_{m \rightarrow \infty} S_{n,m} &= \lim_{n \rightarrow \infty} 1 = 1 \end{aligned}$$

□

Continuity

$$\{f_n\} \rightarrow f, f_n \text{ continuous} \not\Rightarrow f \text{ continuous}$$

Proof. Consider $f_n(x) = \frac{x^2}{(1+x^2)^n}$. Clearly, $f_n : \mathbb{R} \rightarrow \mathbb{R}$ is continuous for every natural number n .

Case 1 : $x = 0$

Suppose $x = 0$. Then, $f_n(0) = 0$ and therefore $\lim_{n \rightarrow \infty} f_n(0) = 0$.

Case 2 : $x \neq 0$

Suppose $x \neq 0$. Then $\frac{1}{1+x^2} < 1, \forall x \in \mathbb{R}, (x \neq 0)$.
Therefore,

$$f(x) = \lim_{n \rightarrow \infty} f_n(x) = x^2 \lim_{n \rightarrow \infty} \left(\frac{1}{1+x^2} \right)^n = x^2 \frac{1}{1 - \frac{1}{1+x^2}} = x^2 \frac{1+x^2}{x^2} = 1+x^2$$

Clearly from cases 1 and 2,

$$f : \mathbb{R} \rightarrow \mathbb{R}, f(x) = \begin{cases} 0 & x = 0 \\ 1 + x^2 & x \neq 0 \end{cases} \quad \text{is not continuous at 0}$$

□

integrability

$$\{f_n\} \rightarrow f, f_n \in \mathcal{R} \not\Rightarrow f \in \mathcal{R}$$

Proof. Consider, $f_m(x) = \lim_{n \rightarrow \infty} (\cos m! \pi x)^{2n}$.

Suppose $m!x$ is not an integer. Then $f_m(x) = \lim_{n \rightarrow \infty} (y^2)^n = 0$ since $-1 < y = \cos m! \pi x < 1$. Suppose $m!x$ is an integer. Then $\cos m! \pi x = \pm 1$ and $f_m(x) = \lim_{n \rightarrow \infty} ((\pm 1)^2)^n = 1$.

We know that, if $x \in \mathbb{Q}$, then $x = \pm \frac{p}{q}$ where $p, q \in \mathbb{N}$. Clearly, for any $m > q$, $m!x$ is an integer and $f_m(x) = 1$ for any $m > q$. Therefore, $\lim_{m \rightarrow \infty} f_m(x) = 1$ if $x \in \mathbb{Q}$. And $\lim_{m \rightarrow \infty} f_m(x) = 0$ if $x \notin \mathbb{Q}$ since $m!x$ is not an integer for any $m \in \mathbb{N}$ and $f_m(x) = 0, \forall m \in \mathbb{N}$.

Now, $f : \mathbb{R} \rightarrow \mathbb{R}$ defined by $f(x) = \lim_{m \rightarrow \infty} f_m(x)$ is given by,

$$f(x) = \begin{cases} 1 & x \in \mathbb{Q} \\ 0 & x \notin \mathbb{Q} \end{cases} \quad \text{which is **everywhere discontinuous**.}$$

Clearly, f is not Riemann integrable since $\mathbb{Q}, \mathbb{R} - \mathbb{Q}$ are dense in any subinterval. Thus, $m_i = 0$ and $M_i = 1$ in any subinterval of any partition P .

$$\int_0^1 f(x) dx = 0 \neq 1 = \int_0^1 \bar{f}(x) dx$$

Let $m \in \mathbb{N}$. Consider closed interval $E = [-1, 1]$. Then there exists finitely many rational numbers x on E such that $m!x$ is an integer. Thus, $f_m(x)$ has at most finitely many discontinuities in any bounded interval. $f_m(x) = \cos m! \pi x$ is discontinuous at $S = \{\frac{k}{m!} \in E : k \in \mathbb{Z}\}$. For example, suppose $m = 3$. Then $S = \{0, \pm \frac{1}{6}, \pm \frac{2}{6}, \pm \frac{3}{6}, \pm \frac{4}{6}, \pm \frac{5}{6}, \pm 1\}$. Let α be the identity function. Then α is continuous at those finite points where f_m is discontinuous. And f_m are bounded functions, since $|f_m| \leq 1$. Therefore, for any bounded interval E , f_m is Riemann integrable on E for each m . However, f is not Riemann integrable. □

Derivative

$$\{f_n\} \rightarrow f \not\Rightarrow \{f'_n\} \rightarrow f'$$

Proof. Consider $f_n(x) = \frac{\sin nx}{\sqrt{n}}$.

$$f(x) = \lim_{n \rightarrow \infty} f_n(x) = 0, \forall x$$

However, $f'_n(x) = \sqrt{n} \cos nx$. And $f'(x) = 0$.

$$\lim_{n \rightarrow \infty} \frac{d}{dx} f_n(x) = \infty \neq f'(x) = \frac{d}{dx} \lim_{n \rightarrow \infty} f_n(x), \quad \forall x$$

Clearly, convergence doesn't preserve derivatives. \square

Integral

$$\{f_n\} \rightarrow f \not\Rightarrow \left\{ \int f_n \right\} \rightarrow \int f$$

Proof. Consider $f_n : [0, 1] \rightarrow \mathbb{R}$ defined by $f_n(x) = n^2 x(1 - x^2)^n$.

We have,

$$\lim_{n \rightarrow \infty} \frac{n^\alpha}{(1+p)^n} = 0$$

Let $\frac{1}{1+p} = 1 - x^2$, then $1 + p = \frac{1}{1-x^2}$ and $p = \frac{x^2}{1-x^2} > 0$.

Now, we have

$$f(x) = \lim_{n \rightarrow \infty} f_n(x) = \lim_{n \rightarrow \infty} n^2 x(1 - x^2)^n = x \lim_{n \rightarrow \infty} \frac{n^2}{\left(1 + \frac{x^2}{1-x^2}\right)^n} = 0$$

And, we have

$$\int_0^1 f(x) dx = 0$$

However,

$$\int_0^1 f_n(x) dx = n^2 \int_0^1 x(1 - x^2)^n dx = n^2 \int_0^1 \frac{u^n}{2} du = n^2 \left[\frac{u^{n+1}}{2(n+1)} \right]_0^1 = \frac{n^2}{2n+2}$$

Clearly,

$$\lim_{n \rightarrow \infty} \int_0^1 f_n(x) dx = \lim_{n \rightarrow \infty} \frac{n^2}{2n+2} = \infty \neq 0 = \int_0^1 f(x) dx = \int_0^1 \lim_{n \rightarrow \infty} f_n(x) dx$$

\square

4.3.2 Uniform Convergence

Uniform convergence is a partial solution to our main problem.

Definitions 4.3.3. Let sequence $\{f_n\}$ be a sequence of functions on E . Then $\{f_n\}$ converges uniformly to limit function f if for any $\varepsilon > 0$, there exists a natural number N such that for any $n > N$ and $x \in E$, $|f_n(x) - f(x)| \leq \varepsilon$.

$$\forall \varepsilon > 0, \exists N \in \mathbb{N}, \forall n > N, \forall x \in E, |f_n(x) - f(x)| \leq \varepsilon$$

The difference between pointwise convergence and uniform convergence is that the choice of natural number N is independent of the choice of x in the case of uniform convergence.

Criterion for Uniform Convergence

Theorem 4.3.1 (Cauchy criterion). *The sequence of functions $\{f_n\}$ defined on E converges uniformly on E if and only if for any $\varepsilon > 0$, there exists a natural number N such that for any $m, n \geq N$ and $x \in E$, $|f_n(x) - f_m(x)| \leq \varepsilon$.*

Proof. Suppose $\{f_n\} \rightarrow f$ uniformly on E . Let $\varepsilon > 0$. Then there exists $N \in \mathbb{N}$ such that $\forall n, m > N$ and $\forall x \in E$, $|f_n(x) - f(x)| \leq \frac{\varepsilon}{2}$ and $|f_m(x) - f(x)| \leq \frac{\varepsilon}{2}$. Therefore, $\forall \varepsilon > 0$ we have, $N \in \mathbb{N}$ such that $\forall n, m > N$ and $\forall x \in E$,

$$|f_n(x) - f_m(x)| \leq |f_n(x) - f(x)| + |f_m(x) - f(x)| \leq \varepsilon$$

Let $\varepsilon > 0$. Suppose there exists $N \in \mathbb{N}$ such that $\forall n, m > N$ and $\forall x \in E$,

$$|f_n(x) - f_m(x)| \leq |f_n(x) - f(x)| + |f_m(x) - f(x)| \leq \varepsilon$$

By Cauchy criterion, $\{f_n\} \rightarrow f$ pointwise on E . It remains to prove that this convergence is uniform on E . Clearly, above inequality is true for any value of m greater than N . As $m \rightarrow \infty$ we have,

$$\begin{aligned} \lim_{m \rightarrow \infty} |f_n(x) - f_m(x)| &\leq \varepsilon \\ \implies \left| f_n(x) - \lim_{m \rightarrow \infty} f_m(x) \right| &= |f_n(x) - f(x)| \leq \varepsilon \end{aligned}$$

Clearly, the convergence is uniform. That is, $\{f_n\} \rightarrow f$ uniformly on E . \square

Theorem 4.3.2 (Supremum Test). *Suppose sequence of function $\{f_n\} \rightarrow f$ pointwise on E . Suppose $M_n = \sup_{x \in E} |f_n(x) - f(x)|$. Then $\{f_n\} \rightarrow f$ uniformly on E , if and only if sequence $\{M_n\} \rightarrow 0$.*

Proof. Define $M_n = \sup_{x \in X} |f_n(x) - f(x)|$. Clearly, $M_n \geq 0$. Suppose $\{M_n\} \rightarrow 0$ on E . Let $\varepsilon > 0$. Then there exists $N \in \mathbb{N}$ such that for every $n > N$, we have $|M_n - 0| = M_n \leq \varepsilon$. Thus for any $n \geq N$ and $x \in E$,

$$|f_n(x) - f(x)| \leq \sup |f_n(x) - f(x)| = M_n \leq \varepsilon$$

Therefore, $\{f_n\} \rightarrow f$ uniformly on E .

Suppose $\{f_n\} \rightarrow f$ uniformly on E . Let $\varepsilon > 0$. Then, there exists $N \in \mathbb{N}$ such that for every $n \geq N$ and $x \in E$, $|f_n(x) - f(x)| \leq \varepsilon$. This inequality is true for every $x \in E$. Thus,

$$\sup_{x \in E} |f_n(x) - f(x)| = M_n \leq \varepsilon, \quad \forall n \geq N$$

Therefore, $\{M_n\} \rightarrow 0$ on E . \square

Theorem 4.3.3 (Weierstrass M-test). *Suppose $\{f_n\}$ is a sequence of functions on E . Suppose $|f_n(x)| \leq M_n$. Then $\sum f_n$ converges on E if $\sum M_n$ converges.*

Proof. Suppose $|f_n| \leq M_n$. Then $M_n \geq 0$. Suppose $\sum M_n$ converges. Let $\{s_n\}$ be the sequence of partial sums. Let $\varepsilon > 0$. By Cauchy criterion, there exists $N \in \mathbb{N}$ such that $\forall n, m \geq N$, $|s_n - s_m| \leq \varepsilon$. Thus,

$$|s_n - s_m| = \left| \sum_{j=1}^n M_j - \sum_{j=1}^m M_j \right| = \sum_{j=n+1}^m M_j \leq \varepsilon, \quad (m > n)$$

Let $\{S_n\}$ be the sequence of partial sums for $\sum f_n$. Clearly, $\forall n, m \geq N$ and $\forall x \in E$, we have $|S_n - S_m| \leq \varepsilon$ since

$$|S_n - S_m| = \left| \sum_{j=1}^n f_j(x) - \sum_{j=1}^m f_j(x) \right| = \left| \sum_{j=n+1}^m f_j(x) \right| \leq \sum_{j=n+1}^m M_j \leq \varepsilon$$

By Cauchy criterion for uniform convergence, the sequence of partial sums, $\{S_n\}$ converges uniformly on E . Therefore, the series of functions, $\sum f_n$ converges uniformly on E . \square

4.3.3 Uniform Convergence and Continuity

Theorem 4.3.4. Suppose $\{f_n\} \rightarrow f$ uniformly on E . Let x be a limit point of E . Suppose $\lim_{t \rightarrow x} f_n(t) = A_n$. Then $\{A_n\}$ converges, and $\lim_{t \rightarrow x} f(t) = \lim_{n \rightarrow \infty} A_n$. In other words,

$$\lim_{t \rightarrow x} \lim_{n \rightarrow \infty} f_n(t) = \lim_{n \rightarrow \infty} \lim_{t \rightarrow x} f_n(t)$$

Proof. Suppose $\{f_n\} \rightarrow f$ uniformly on E . Suppose $\lim_{t \rightarrow x} f_n(t) = A_n$ on E . Let $\varepsilon > 0$. By Cauchy criterion for uniform convergence, there exists $N \in \mathbb{N}$ such that for every $n \geq N$ and every $t \in E$, $|f_n(t) - f_m(t)| \leq \varepsilon/3$. As $t \rightarrow x$, we have $|A_n - A_m| \leq \varepsilon/3$. Clearly, $\{A_n\}$ is Cauchy and $\{A_n\} \rightarrow A$.

Now fix a natural number N such that for every $n \geq N$ and $t \in E$, $|f_n(t) - f(t)| < \varepsilon/3$ and $|A_n - A| \leq \varepsilon/3$. Also, choose a neighbourhood V of x such that $|f_n(t) - A_n| \leq \varepsilon/3$. Then, $\forall x \in V \cap E$ we have

$$|f(t) - A| \leq |f(t) - f_n(t)| + |f_n(t) - A_n| + |A_n - A| \leq \varepsilon$$

Therefore, $\lim_{t \rightarrow x} f(t) = A$.

$$\lim_{t \rightarrow x} \lim_{n \rightarrow \infty} f_n(t) = \lim_{t \rightarrow x} f(t) = \lim_{n \rightarrow \infty} A_n = \lim_{n \rightarrow \infty} \lim_{t \rightarrow x} f_n(t)$$

\square

Theorem 4.3.5. If $\{f_n\}$ is a sequence of continuous functions on E , and if $f_n \rightarrow f$ uniformly on E , then f is continuous on E .

Uniform convergence preserves continuity.

Proof. Suppose the sequence of continuous functions, $\{f_n\} \rightarrow f$ uniformly on E . Since f_n is continuous $\lim_{t \rightarrow x} f_n(t) = f_n(x)$, $\forall x \in E$. Since the convergence is uniform, the order of limits can be interchanged.

$$f(x) = \lim_{n \rightarrow \infty} f_n(x) = \lim_{n \rightarrow \infty} \lim_{t \rightarrow x} f_n(t) = \lim_{t \rightarrow x} \lim_{n \rightarrow \infty} f_n(t) = \lim_{t \rightarrow x} f(t)$$

Therefore, the limit function f is continuous. \square

Remark. Suppose a sequence of continuous functions $\{f_n\}$ converges to a function f . Function f being continuous doesn't imply that the convergence is uniform.

Proof. Consider $f_n : (0, 1) \rightarrow \mathbb{R}$ defined by $f_n(x) = \frac{1}{nx+1}$. Then, $\{f_n\} \rightarrow 0$. Clearly, $f_n, 0$ are continuous functions on $(0, 1)$. However, the convergence is not uniform. \square

Theorem 4.3.6. Suppose K is compact, and

1. $\{f_n\}$ is sequence of continuous functions on K
2. $\{f_n\}$ converges pointwise to a continuous function f on K .
3. $f_n(x) \geq f_{n+1}(x), \forall x \in K$

Then, $\{f_n\} \rightarrow f$ uniformly on K .

Proof. Suppose $\{f_n\} \rightarrow f$ pointwise on K . Let function $g_n = f_n - f$. Then, g_n is continuous, $\{g_n\} \rightarrow 0$ pointwise and $g_n \geq g_{n+1}$. If sequence $\{g_n\} \rightarrow 0$ uniformly, then sequence $\{f_n\} \rightarrow f$ uniformly.

Let $\varepsilon > 0$. Let K_n be the set of all points $x \in K$ such that $g_n(x) \geq \varepsilon$. Then K_n is closed, since g_n is continuous. And K_n is compact since K_n is a closed subset of a compact set K .

Suppose $x \in K_{n+1}$. Then $g_{n+1}(x) \geq \varepsilon$. We have, $g(x) \geq g_{n+1}(x)$. Thus, $g(x) \geq \varepsilon$. Therefore, $K_n \supset K_{n+1} \supset \dots$.

Let $x \in K$. We have $\{g_n(x)\} \rightarrow 0$ pointwise. Then there exists $N \in \mathbb{N}$ such that $\forall n > N, g_n(x) < \varepsilon$. Thus, $x \notin K_n, \forall n \geq N$. Therefore $\cap K_n$ is empty.

Clearly, K_N is empty for some $N \in \mathbb{N}$. Thus, $\forall n \geq N$ and $\forall x \in K$ we have, $0 \leq g_n(x) < \varepsilon$. In other words, $\{g_n\}$ converges to 0 uniformly on K . Therefore, $\{f_n\}$ converges to f uniformly on K . \square

Definitions 4.3.4. Let X be a metric space. Let $\mathcal{C}(X)$ be the set of all complex valued, continuous, bounded functions on X . Let $f \in \mathcal{C}(X)$. Then **supremum norm** on $\mathcal{C}(X)$ is defined by

$$\|f\| = \sup_{x \in X} |f(x)|$$

And, $\mathcal{C}(X)$ together with **distance function** $d : \mathcal{C}(X) \times \mathcal{C}(X) \rightarrow \mathbb{R}$ defined by $d(f, g) = \|f - g\|$ is a metric space.

Let $\mathcal{C}(X)$ be the set of all complex valued, continuous, bounded functions on metric space X . Let $f \in \mathcal{C}(X)$. Then f is bounded, $|f| < \infty$. Thus, $\sup |f(x)|$ exists. Therefore, $\|f\|$ is well-defined.

And, $\|f\| = \sup |f(x)| \geq 0$ since $|f(x)| \geq 0$. Let $f, g \in \mathcal{C}(X)$. Then $h = f + g \in \mathcal{C}(X)$ since sum of continuous functions is functions and sum of bounded functions is bounded. Also, we have

$$\|h\| = \sup_{x \in X} |(f + g)(x)| \leq \sup_{x \in X} |f(x)| + \sup_{x \in X} |g(x)| = \|f\| + \|g\|$$

Therefore, $\|\cdot\| : \mathcal{C}(X) \rightarrow \mathbb{R}$ is a norm on $\mathcal{C}(X)$.

Consider the function $d : \mathcal{C}(X) \times \mathcal{C}(X) \rightarrow \mathbb{R}$ defined by $d(f, g) = \|f - g\|$. Clearly, $d(f, g) \geq 0$ since $|(f - g)(x)| \geq 0$. And,

$$d(f, g) = 0 \iff |(f - g)(x)| = 0, \forall x \in X \iff f = g$$

Also we have,

$$d(f, g) \leq \|(f - h) + (h - g)\| \leq \|f - h\| + \|h - g\| = d(f, h) + d(h, g)$$

Therefore, the function d is a distance function/metric on $\mathcal{C}(X)$.

Remark. A sequence $\{f_n\}$ converges to f in metric space $\mathcal{C}(X)$ if and only if $\{f_n\}$ converges to f uniformly on X .

Proof.

$$\begin{aligned} \{f_n\} \rightarrow f \text{ in } \mathcal{C}(X) &\iff \forall \varepsilon > 0, \exists N \in \mathbb{N}, \forall n > N, d(f_n, f) \leq \varepsilon \\ &\iff \forall n > N, \|f_n - f\| \leq \varepsilon \\ &\iff \forall n > N, \sup |f_n(x) - f(x)| \leq \varepsilon \\ &\iff \forall n > N, \forall x \in X, |f_n(x) - f(x)| \leq \varepsilon \\ &\iff \{f_n\} \rightarrow f \text{ uniformly on } X \end{aligned}$$

□

Definitions 4.3.5 (uniformly closed). Closed subset of $\mathcal{C}(X)$ is **uniformly closed** since every convergent sequence in it corresponds to a uniform convergent sequence in X .

Definitions 4.3.6 (uniform closure). Let $\mathcal{A} \subset \mathcal{C}(X)$. Then its closure, $\bar{\mathcal{A}}$ is the **uniform closure** of \mathcal{A} .

Definitions 4.3.7 (complete metric space). A metric space is **complete** if every Cauchy sequence in it converges.

For example, \mathbb{R} is complete. But \mathbb{Q} is not complete since a sequence of rational numbers converging to π in \mathbb{R} is Cauchy sequence in \mathbb{Q} which doesn't converge to any point in \mathbb{Q} .

Theorem 4.3.7. Let X be a metric space. Let $\mathcal{C}(X)$ be the set of all complex valued, continuous, bounded functions on X . Let $f, g \in \mathcal{C}(X)$. Define norm $\|f\| = \sup_{x \in X} |f(x)|$ and metric $d(f, g) = \|f - g\|$. Then metric space $\langle \mathcal{C}(X), d \rangle$ is a complete metric space.

Proof. Let sequence of functions $\{f_n\}$ be a Cauchy sequence in $\mathcal{C}(X)$. Let $\varepsilon > 0$. Then there exists a natural number N such that for any $n, m > N$, $d(f_n, f_m) < \varepsilon$. That is,

$$d(f_n, f_m) = \sup_{x \in X} |f_n(x) - f_m(x)| < \varepsilon$$

Define $M_n = \sup_{x \in X} |f_n(x)|$. Then,

$$|M_n - M_m| = \left| \sup_{x \in X} |f_n(x)| - \sup_{x \in X} |f_m(x)| \right| \leq \sup_{x \in X} |f_n(x) - f_m(x)| < \varepsilon$$

We have, $\{M_n\}$ is a Cauchy sequence in \mathbb{R} . We know that, every Cauchy sequence in \mathbb{R} is convergent in \mathbb{R} since \mathbb{R} is complete. Thus, $\{M_n\}$ converges.

We have, corresponding sequence $\{f_n\}$ in X such that sequence $\{M_n\}$ defined by $M_n = \sup_{x \in X} |f_n(x)|$ converges. Therefore, $\{f_n\}$ converges to f uniformly on X by criterion of uniform convergence.

We have, f_n are continuous, bounded functions on X . And $\{f_n\} \rightarrow f$ uniformly on X .

$$\forall \varepsilon > 0, \exists N \in \mathbb{N}, \forall n > N, \forall x \in X, |f_n(x) - f_m(x)| < \varepsilon$$

Since, the convergence is uniform, the limit function f is continuous and **bounded**. Therefore, $\{f_n\}$ in $\mathcal{C}(X)$ converges to $f \in \mathcal{C}(X)$.

The Cauchy sequence $\{f_n\}$ is chosen arbitrarily. Thus, every Cauchy sequence in $\mathcal{C}(X)$ is convergent. Therefore, $\mathcal{C}(X)$ is complete. \square

Remark. [Rudin, 1976, Exercise 7.1] Let $\{f_n\}$ be a sequence of bounded functions on a metric space X . Suppose $\{f_n\} \rightarrow f$ uniformly on X . Then f is bounded.

Proof. Suppose the sequence of bounded functions, $\{f_n\}$ converges to f uniformly on X . Consider $\{g_n\}$ where $g_n = |f_n - f|$. Then, $\{g_n\}$ converges to 0 uniformly. Suppose f is unbounded. Then g_n is an unbounded sequence which doesn't converge. This is a contradiction. Therefore, f is bounded. \square

4.3.4 Uniform Convergence and Integration

Theorem 4.3.8. Let α be monotonically increasing on $[a, b]$. Suppose $f_n \in \mathcal{R}(\alpha)$ on $[a, b]$. Suppose $\{f_n\} \rightarrow f$ uniformly on $[a, b]$. Then $f \in \mathcal{R}(\alpha)$ on $[a, b]$. And,

$$\int_a^b f \, d\alpha = \lim_{n \rightarrow \infty} \int_a^b f_n \, d\alpha$$

Proof. Let $f_n \in \mathcal{R}(\alpha)$ on $[a, b]$. Suppose sequence $\{f_n\} \rightarrow f$ uniformly on $[a, b]$. Define

$$\varepsilon_n = \sup_{x \in [a, b]} |f_n(x) - f(x)|$$

Then, $-\varepsilon_n \leq f - f_n \leq \varepsilon_n$, $\forall x \in [a, b]$. And,

$$f_n - \varepsilon_n \leq f \leq f_n + \varepsilon_n$$

Since $f_n \in \mathcal{R}(\alpha)$, we have $f_n + \varepsilon, f_n - \varepsilon \in \mathcal{R}(\alpha)$. From the definition of lower and upper integrals,

$$\int_a^b (f_n - \varepsilon_n) \, d\alpha \leq \int_a^b f \, d\alpha \leq \int_a^b (f_n + \varepsilon_n) \, d\alpha$$

Thus,

$$0 \leq \int_a^b f \, d\alpha - \int_a^b f \, d\alpha \leq \int_a^b 2\varepsilon_n \, d\alpha = 2\varepsilon_n [\alpha(b) - \alpha(a)]$$

Clearly, $\varepsilon_n \rightarrow 0$ as $n \rightarrow \infty$. Then, $f \in \mathcal{R}(\alpha)$ on $[a, b]$ since the lower and upper integrals of f with respect to α are equal.

Also we have, $0 \leq |f - f_n| \leq \varepsilon_n$ for every $x \in [a, b]$. Therefore,

$$0 \leq \left| \int_a^b f \, d\alpha - \int_a^b f_n \, d\alpha \right| \leq \int_a^b |f - f_n| \, d\alpha \leq \int_a^b \varepsilon_n \, d\alpha = \varepsilon_n [\alpha(b) - \alpha(a)]$$

As $n \rightarrow \infty$, we have $\varepsilon_n \rightarrow 0$. Thus,

$$\int_a^b f \, d\alpha = \lim_{n \rightarrow \infty} \int_a^b f_n \, d\alpha$$

□

Corollary 4.3.8.1. *If $f_n \in \mathcal{R}(\alpha)$ on $[a, b]$ and if*

$$f(x) = \sum_{n=1}^{\infty} f_n(x)$$

the series converging uniformly on $[a, b]$, then

$$\int_a^b f \, d\alpha = \sum_{n=1}^{\infty} \int_a^b f_n \, d\alpha$$

In other words, the series may be integrated term by term.

Proof. Suppose $f_n \in \mathcal{R}(\alpha)$ and $\sum f_n$ converges to f uniformly on $[a, b]$. Let $\{S_n\}$ be the sequence of partial sums converging to f uniformly on $[a, b]$. We have, $S_n \in \mathcal{R}(\alpha)$ on $[a, b]$ by linearity of the integral and mathematical induction.

$$\sum_{j=1}^n \int_a^b f_j \, d\alpha = \int_a^b \sum_{j=1}^n f_j \, d\alpha = \int_a^b S_n \, d\alpha$$

Since the uniform limit function of integrable functions is integrable,

$$\sum_{n=1}^{\infty} f_n = \lim_{n \rightarrow \infty} S_n = f \in \mathcal{R}(\alpha) \text{ and } \lim_{n \rightarrow \infty} \int_a^b S_n \, d\alpha = \int_a^b f \, d\alpha$$

By the definition of sum of series,

$$\sum_{n=1}^{\infty} \int_a^b f_n \, d\alpha = \lim_{n \rightarrow \infty} \sum_{j=1}^n \int_a^b f_j \, d\alpha = \lim_{n \rightarrow \infty} \int_a^b S_n \, d\alpha = \int_a^b f \, d\alpha = \int_a^b \sum_{n=1}^{\infty} f_n \, d\alpha$$

Thus, the series may be integrated term by term. □

4.3.5 Uniform Convergence and Differentiation

Theorem 4.3.9. *Suppose $\{f_n\}$ is a sequence of functions, differentiable on $[a, b]$ and $\{f_n(x_0)\}$ converges for some $x_0 \in [a, b]$. If $\{f'_n\}$ converges uniformly on $[a, b]$, then $\{f_n\}$ converges uniformly on $[a, b]$ to a function f . And,*

$$f'(x) = \lim_{n \rightarrow \infty} f'_n(x)$$

Proof. Suppose $\{f_n(x_0)\} \rightarrow f(x_0)$. Suppose $\{f'_n\}$ converges uniformly on $[a, b]$. Let $\varepsilon > 0$. Then there exists a natural number N such that

$$\forall n, m > N, |f_n(x_0) - f_m(x_0)| < \frac{\varepsilon}{2} \text{ and}$$

$$\forall n, m > N, |f'_n(t) - f'_m(t)| < \frac{\varepsilon}{2(b-a)}$$

Let $x, t \in [a, b]$.

$$|f_n(x) - f_m(x) - f_n(t) + f_m(t)| = |(f_n - f_m)(x) - (f_n - f_m)(t)|$$

By mean value theorem, there exists $y \in (x, t)$ such that

$$\begin{aligned} |f_n(x) - f_m(x) - f_n(t) + f_m(t)| &= |(x - t)(f_n - f_m)'(y)| \\ &\leq \frac{|x - t|\varepsilon}{2(b-a)} \leq \frac{\varepsilon}{2} \end{aligned}$$

since $|(f_n - f_m)'(x)| = |f'_n(x) - f'_m(x)| \leq \frac{\varepsilon}{2(b-a)}$.

And the inequality is true of $t = x_0$. Thus,

$$\begin{aligned} |f_n(x) - f_m(x)| &= |f_n(x) - f_m(x) - f_n(x_0) + f_m(x_0) + f_n(x_0) - f_m(x_0)| \\ &\leq |f_n(x) - f_m(x) - f_n(x_0) + f_m(x_0)| + |f_n(x_0) - f_m(x_0)| \leq \varepsilon \end{aligned}$$

Thus, $\forall \varepsilon > 0$, we have a natural number N such that $\forall n, m > N, \forall x \in [a, b], |f_n(x) - f_m(x)| \leq \varepsilon$. In other words, $\{f_n\} \rightarrow f$ uniformly on $[a, b]$.

Fix $x \in [a, b]$. Define functions ϕ_n, ϕ on $[a, b]$ except at $t = x$,

$$\phi_n(t) = \frac{f_n(t) - f_n(x)}{t - x} \quad \text{and} \quad \phi(t) = \frac{f(t) - f(x)}{t - x}$$

Then,

$$\lim_{n \rightarrow \infty} \phi_n(t) = \lim_{n \rightarrow \infty} \frac{f_n(t) - f_n(x)}{t - x} = \frac{\lim_{n \rightarrow \infty} f_n(t) - \lim_{n \rightarrow \infty} f_n(x)}{t - x} = \frac{f(t) - f(x)}{t - x} = \phi(t)$$

since sequence $\{f_n\} \rightarrow f$ uniformly on $[a, b]$. Clearly, $\{\phi_n\}$ converges to ϕ pointwise on $[a, b] - \{x\}$.

We have,

$$\begin{aligned} \lim_{t \rightarrow x} \phi_n(t) &= \lim_{t \rightarrow x} \frac{f_n(t) - f_n(x)}{t - x} = f'_n(x) \\ \lim_{t \rightarrow x} \phi(t) &= \lim_{t \rightarrow x} \frac{f(t) - f(x)}{t - x} = f'(x) \end{aligned}$$

Also we have,

$$|\phi_n(t) - \phi_m(t)| = \frac{|f_n(t) - f_n(x) - f_m(t) + f_m(x)|}{t - x} \leq \frac{\varepsilon}{2(b-a)}$$

That is $\forall n, m > N, |\phi_n(t) - \phi_m(t)| < \varepsilon$. Now, by Cauchy criterion for uniform convergence, sequence $\{\phi_n\}$ converges to ϕ uniformly on $[a, b] - \{x\}$.

We know that continuity is preserved under uniform convergence. Therefore,

$$f'(t) = \lim_{t \rightarrow x} \phi(t) = \lim_{t \rightarrow x} \lim_{n \rightarrow \infty} \phi_n(t) = \lim_{n \rightarrow \infty} \lim_{t \rightarrow x} \phi_n(t) = \lim_{n \rightarrow \infty} f'_n(x)$$

□

Theorem 4.3.10. *There exists a real continuous function on the real line which is nowhere differentiable.*

Proof. Define $\phi : [-1, 1] \rightarrow \mathbb{R}$ by $\phi(x) = |x|$. Extend ϕ from $[-1, 1]$ to \mathbb{R} such that $\phi(x+2) = \phi(x)$. Then $|\phi(s) - \phi(t)| \leq |s - t|$, $\forall s, t \in \mathbb{R}$.

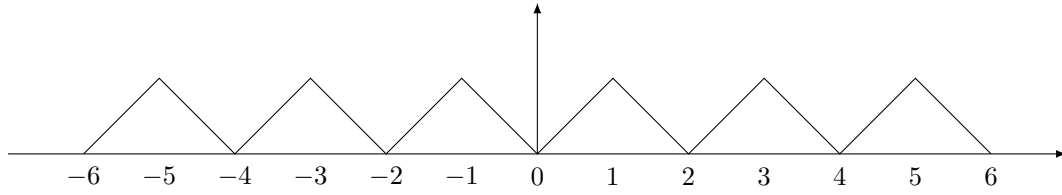


Figure 4.1: Graph of ϕ

Define $f_n(x) = \left(\frac{3}{4}\right)^n \phi(4^n x)$. Then $|f_n| \leq \left(\frac{3}{4}\right)^n$. By Weierstrass M test, $\sum_{n=0}^{\infty} f_n$ converges uniformly on \mathbb{R} . Consider the function $f : \mathbb{R} \rightarrow \mathbb{R}$ defined by $f(x) = \sum_{n=0}^{\infty} \left(\frac{3}{4}\right)^n \phi(4^n x)$. This sum function f is continuous since the convergence is uniform and f_n are all continuous.

Fix a real number x and a positive integer m . Define $\delta_m = \pm \frac{1}{2} 4^{-m}$ such that no integer lies between $4^m x$ and $4^m(x + \delta_m)$. This is possible since $|4^m(x + \delta_m) - 4^m x| = |4^m \delta_m| = \frac{1}{2}$.

The choice of sign of δ_m is made depending on the value of $4^m x$. Suppose $x = 1.1$ and $m = 3$. Then $4^m x = 70.4$. Now $\delta = \frac{1}{2} 4^{-m} = \frac{1}{128}$ so that $4^m(x + \delta_m) = 70.9$. Suppose $x = 1.2$ and $m = 3$. Then $4^m x = 76.8$. Now $\delta = -\frac{1}{2} 4^{-m} = -\frac{1}{128}$ so that $4^m(x + \delta) = 76.3$.

Define

$$\gamma_n = \frac{\phi(4^n(x + \delta_m)) - \phi(4^n x)}{\delta_m}$$

If $n > m$, then $4^n \delta_m$ is an even integer and $\gamma_n = 0$.
If $n \leq m$, then

$$|\gamma_n| = \left| \frac{\phi(4^n(x + \delta_m)) - \phi(4^n x)}{\delta_m} \right| \leq \frac{4^n \delta_m}{\delta_m} \leq 4^n$$

since $|\phi(s) - \phi(t)| \leq |s - t|$.

In particular, if $n = m$, then

$$|\gamma_m| = \left| \frac{\phi(4^m(x + \delta_m)) - \phi(4^m x)}{\delta_m} \right| = 4^m$$

since $|\phi(4^m(x + \delta_m)) - \phi(4^m x)| = |4^m \delta| = \frac{1}{2}$ as there are no integers between $4^m(x + \delta_m)$ and $4^m x$.

From the definition of γ_n we have,

$$\begin{aligned} \left| \frac{f(x + \delta_m) - f(x)}{\delta_m} \right| &= \left| \sum_{n=0}^m \left(\frac{3}{4} \right)^n \gamma_n \right| \\ &= \left| 3^m + \sum_{n=0}^{m-1} \left(\frac{3}{4} \right)^n \gamma_n \right| \\ &\geq 3^m - \sum_{n=0}^{m-1} 3^n = \frac{3^m + 1}{2} \end{aligned}$$

Therefore, the function f is not differentiable at x since the following limit does not exist as $m \rightarrow \infty$.

$$\lim_{\delta_m \rightarrow 0} \left| \frac{f(x + \delta_m) - f(x)}{\delta_m} \right| \geq \lim_{m \rightarrow \infty} \frac{3^m + 1}{2}$$

Since the choice of x is arbitrary, the function f is nowhere differentiable. \square

Module 4 : Weierstrass Approximation & Some Special Functions

4.4 Equicontinuous Family of Functions

Definitions 4.4.1. Let $\{f_n\}$ be a sequence of functions defined on E .

Sequence $\{f_n\}$ is a **bounded** sequence if every functions in the sequence is bounded.

$$\forall n \in \mathbb{N}, \exists M \in \mathbb{R} \text{ such that } \forall x \in E, |f_n(x)| < M$$

Sequence $\{f_n\}$ is a **pointwise bounded** sequence if there exists a function $\phi(x)$ such that $|f_n(x)| < \phi(x)$. In other words, sequence $\{f_n\}$ is a pointwise bounded sequence if $\{f_n(x)\}$ is bounded for every $x \in E$

$$\forall x \in E, \exists M \in \mathbb{R} \text{ such that } \forall n \in \mathbb{N}, |f_n(x)| < M$$

Sequence $\{f_n\}$ is a **uniformly bounded** sequence if there exists a real number M such that $|f_n(x)| < M$ for every $x \in E$ and $n \in \mathbb{N}$.

$$\exists M \in \mathbb{R} \text{ such that } \forall x \in E, \forall n \in \mathbb{N}, |f_n(x)| < M$$

4.4.1 Two Problems

1. Whether uniform bounded sequence of uniformly bounded functions have a convergent subsequence ? **NO**.
Sequence $\{f_n\}$ where $f_n(x) = \sin nx$ is uniformly bounded. But it doesn't have a convergent subsequence.

Proof. Suppose $\{\sin nx\}$ has a convergent subsequence, say $\{\sin n_k x\}$. Then by Cauchy criterion, $(\sin n_k x - \sin n_{k+1} x) \rightarrow 0$ as $n_k \rightarrow \infty$.

$$\lim_{n \rightarrow \infty} \sin n_k x - \sin n_{k+1} x = 0$$

$$\lim_{n \rightarrow \infty} (\sin n_k x - \sin n_{k+1} x)^2 = 0$$

By Lebesgue dominated convergence theorem,

$$\lim_{n \rightarrow \infty} \int_0^{2\pi} (\sin n_k x - \sin n_{k+1} x)^2 dx = \int_0^{2\pi} \lim_{n \rightarrow \infty} (\sin n_k x - \sin n_{k+1} x)^2 dx = 0$$

However, the integral on the left evaluates 2π which is a contradiction. Clearly, the uniformly bounded sequence $\{\sin nx\}$ of continuous functions on compact interval $[0, 2\pi]$ does not even imply existence of a subsequence which converges pointwise. \square

2. Whether every uniformly bounded, convergent sequence has a uniformly convergent subsequence ? **NO**.
Consider,

$$f_n(x) = \frac{x^2}{x^2 + (1 - nx)^2}$$

Sequence $\{f_n\}$ is a uniformly bounded sequence of functions on compact interval $[0, 1]$. doesn't have a convergent subsequence.

Proof. Suppose $\{f_n\}$ has a convergent subsequence $\{f_{n_k}\} \rightarrow f$. Then $\lim_{n \rightarrow \infty} f_{n_k}(x) = f(x)$. We have,

$$\lim_{n \rightarrow \infty} f_n(x) = \lim_{n \rightarrow \infty} \frac{x^2}{x^2 + (1 - nx)^2} = 0$$

However,

$$\lim_{n \rightarrow \infty} f_n\left(\frac{1}{n}\right) = \lim_{n \rightarrow \infty} \frac{\left(\frac{1}{n}\right)^2}{\left(\frac{1}{n}\right)^2 + 0} = 1 \neq 0 = \left(\lim_{n \rightarrow \infty} f_n\right)(x)$$

Clearly, the sequence of functions $\{f_n\}$ is a uniformly bounded sequence of continuous functions on a compact interval $[0, 1]$. Therefore, uniformly bounded, convergent sequence on compact set doesn't necessarily have a uniformly convergent subsequence. \square

Definitions 4.4.2. Let E be a subset of a metric space X . A family \mathcal{F} of complex functions on E is **equicontinuous** if for any $\varepsilon > 0$, there exists $\delta > 0$ such that $|f(x) - f(y)| < \varepsilon$ whenever $d(x, y) < \delta$ for any $x, y \in E$ and $f \in \mathcal{F}$.

$$\forall \varepsilon > 0, \exists \delta > 0 \text{ such that } \forall f \in \mathcal{F}, |f(x) - f(y)| < \varepsilon \text{ whenever } d(x, y) < \delta$$

In equicontinuity, the choice of δ is independent of the choice of f .

Theorem 4.4.1. If $\{f_n\}$ is a sequence of pointwise bounded, complex functions on countable set E , then it has a subsequence $\{f_{n_k}\}$ such that $\{f_{n_k}(x)\}$ converges for every $x \in E$.

Proof. Let E be a countable set. Let $\{f_n\}$ be a sequence of pointwise bounded, complex functions. Let $\{x_i\}$ be a sequence in E . Then $\{f_n(x_1)\}$ is a bounded sequence of complex numbers. By Bolzano-Weierstrass theorem, $\{f_n(x_1)\}$ has a subsequence S_1 , say $\{f_{1,k}(x_1)\}$ which converges at $x_1 \in E$.

$$S_1 : f_{1,1} \ f_{1,2} \ f_{1,3} \ \cdots$$

Consider, the subsequence $\{f_{2,k}(x_2)\}$ of the sequence $\{f_n(x_2)\}$ with the same values for k as in subsequence $\{f_{1,k}(x_2)\}$. Again, $\{f_{2,k}(x_2)\}$ is a bounded complex sequence. And by Bolzano-Weierstrass theorem, $\{f_{2,k}(x_2)\}$ has a convergent subsequence S_2 , say $\{f_{2,k'}(x_2)\}$ which is a subsequence of $\{f_2(x_2)\}$. And more importantly, S_2 is a subsequence of S_1 such that S_2 converges for both $x_1, x_2 \in E$.

$$S_2 : f_{2,1} \ f_{2,2} \ f_{2,3} \ \cdots$$

Continuing like this, we get a sequence of subsequences

$$\begin{array}{cccc} S_1 & : & f_{n,1} & f_{n,2} & f_{n,3} & \cdots \\ S_2 & : & f_{n,1} & f_{n,2} & f_{n,3} & \cdots \\ & & \vdots & & & \\ S_n & : & f_{n,1} & f_{n,2} & f_{n,3} & \cdots \end{array}$$

Consider the diagonal sequence, $S : f_{1,1} \ f_{2,2} \ f_{3,3} \ \cdots$. We know that discarding finitely many first terms won't affect convergence of sequences. And for

any natural number n , we have can obtain a subsequence of S_n by discarding a finite number of first terms from S . Thus, we have sequence S which converges for $x_1, x_2, \dots, x_n \in E$ since S_n converges. Therefore, as $n \rightarrow \infty$ we have a convergent subsequence S of $\{f_n\}$ which converges for every $x \in E$. \square

Theorem 4.4.2. *Let K be a compact metric space. If $\{f_n\}$ is a sequence of pointwise bounded, continuous, complex valued functions on K converges uniformly on K , then $\{f_n\}$ is equicontinuous on K .*

Proof. Let K be a compact metric space. Let $\varepsilon > 0$. Let $\{f_n\}$ converges uniformly on K . Then,

$$\forall \varepsilon > 0, \exists N \in \mathbb{N} \text{ such that } \forall n > N, \forall x \in K, \|f_n(x) - f_N(x)\| < \varepsilon$$

We have continuous functions on compact sets are uniformly continuous. Thus for any $\varepsilon > 0$ there exist $\delta > 0$ such that $|f_j(x) - f_j(y)| < \varepsilon$ whenever $d(x, y) < \delta$. Thus, $|f_N(x) - f_N(y)| \leq \varepsilon$.

Let $1 \leq j \leq N$. Since the continuity is uniform, there exists $\delta > 0$ such that $|f_j(x) - f_j(y)| < \varepsilon$.³

Let $n > N$ and $d(x, y) < \delta$. Then,

$$|f_n(x) - f_n(y)| \leq |f_n(x) - f_N(x)| + |f_N(x) - f_N(y)| + |f_N(y) - f_n(y)| < 3\varepsilon$$

Therefore, for any $\varepsilon > 0$ there exists $\delta > 0$ such that $\forall n \in \mathbb{N}$, $|f_n(x) - f_n(y)| < \varepsilon$ whenever $d(x, y) < \delta$.

$$\forall \varepsilon > 0, \exists \delta > 0, \forall n \in \mathbb{N}, \forall_{d(x,y) < \delta} y \in K, |f_n(x) - f_n(y)| < \varepsilon$$

That is, the sequence $\{f_n\}$ is equicontinuous on K . \square

Theorem 4.4.3. *Let K a compact. If $\{f_n\}$ be a sequence of pointwise bounded, equicontinuous, complex functions on K , then*

1. $\{f_n\}$ is uniformly bounded on K .
2. $\{f_n\}$ has a uniformly convergent subsequence.

Proof. Let $\{f_n\}$ be sequence of point-wise bounded, equicontinuous, complex functions on compact set K . Since f_n are equicontinuous, given $\varepsilon > 0$, there exists $\delta > 0$ such that $\forall n \in \mathbb{N}$, $|f_n(x) - f_n(y)| < \varepsilon$ whenever $d(x, y) < \delta$.

Let \mathcal{C} be a cover of K of open balls of radius δ . Since, K is compact this cover has a finite subcover, say open balls with center p_j , $j = 1, 2, \dots, r$. Thus there exists finitely many points, p_j such that every point in K is sufficiently close one of them. Then for any $x \in K$, there exists some p_j such that $d(x, p_j) < \delta$.

³For each functions f_j , given $\varepsilon > 0$, there exists $\delta_j > 0$ satisfying the ε - δ condition for uniform continuity. Define $\delta = \min\{\delta_j : j = 1, 2, \dots, N\}$, then for this δ the condition is satified by functions f_1, f_2, \dots, f_N .

Since $\{f_n\}$ is point-wise bounded, there exists M_j such that $|f_n(p_j)| < M_j$. Define $M = \max\{M_1, M_2, \dots, M_r\}$. Then

$$\begin{aligned} |f_n(x)| &= |f_n(x) - f_n(p_j) + f_n(p_j)| \\ &\leq |f_n(x) - f_n(p_j)| + |f_n(p_j)| \\ &\leq \varepsilon + M \end{aligned}$$

Therefore, $\{f_n\}$ is uniformly bounded.

Let $\varepsilon > 0$. Choose $\delta > 0$ such that $|f_n(x) - f_n(y)| < \varepsilon$ whenever $d(x, y) < \delta$. Since K is compact, K has a countable dense subset, say $E = \{x_1, x_2, \dots\}$. That is, given $\delta > 0$, for any $x \in K$, there exists $x_j \in E$ such that $d(x, x_j) < \delta$. Clearly, K has a cover of open balls with center x_j s and radius δ . Since K is compact, there exists finitely many points $x_1, x_2, \dots, x_m \in E$ such that $d(x, x_m) < \delta$.

Since E is countable and $\{f_n\}$ is point-wise bounded, $\{f_n\}$ on E has a subsequence, say diagonal sequence $f_{n_i} = g_i$ which converges for any $x_j \in E$. Thus, by Cauchy criterion there exists integer N such that

$$\forall i, j \geq N, |g_i(x_s) - g_j(x_s)| < \varepsilon, \quad s = 1, 2, \dots, m$$

Let $x \in K$. Let $x_s \in E$ such that $d(x, x_s) < \delta$. Then,

$$|g_i(x) - g_j(x)| \leq |g_i(x) - g_i(x_s)| + |g_i(x_s) - g_j(x_s)| + |g_j(x_s) - g_j(x)| \leq 3\varepsilon$$

That is, $\{g_i\}$ is uniformly convergent on K . Therefore, $\{f_n\}$ has a uniformly convergent subsequence $\{f_{n_k}\}$. \square

Theorem 4.4.4 (Weierstrass' Theorem). *If f is a continuous, complex function on $[a, b]$, there exists a sequence of polynomials P_n such that*

$$\lim_{n \rightarrow \infty} P_n(x) = f(x)$$

uniformly on $[a, b]$. If f is real, the P_n may be taken real.

In other words, any continuous function has a polynomial approximation.

Proof. Without loss of generality, suppose $[a, b] = [0, 1]$ and $f(0) = f(1) = 0$.

Step 1 : WLoG $f(0) = f(1) = 0$ If theorem is true for continuous functions satisfying $f(0) = f(1) = 0$, then it is true for any continuous function.

Let f be any continuous function on $[0, 1]$. Then there exists a function $g : [0, 1] \rightarrow \mathbb{R}$ be defined by,

$$g(x) = f(x) - f(0) - x[f(1) - f(0)]$$

such that $g(0) = g(1) = 0$. Suppose g can be expressed as limit function of a uniformly convergent sequence of polynomials, say P_n . We have,

$$(f - g)(x) = x[f(1) - f(0)] + f(0)$$

is a polynomial, say $P(x)$. Then $P_n + P \rightarrow f$ uniformly as $n \rightarrow \infty$, since $P_n \rightarrow g$ uniformly and P is a constant polynomial independent of n .

$$\lim_{n \rightarrow \infty} (P_n + P)(x) = \lim_{n \rightarrow \infty} P_n(x) + P(x) = g(x) + (f - g)(x) = f(x)$$

Therefore, it is enough to prove the theorem is true for any continuous function f satisfying $f(0) = f(1) = 0$.

Step 2 : Construction of $P_n(x)$

Since f is continuous in $[0, 1]$, f is uniformly continuous in $[0, 1]$. Extend f such that $f(x) = 0$, $\forall x \notin [0, 1]$. Then, f is uniformly continuous on \mathbb{R} . Define $Q_n(x) = c_n(1 - x^2)^n$ such that

$$\int_{-1}^1 Q_n(x) dx = 1$$

Then we have,

$$\int_{-1}^1 (1 - x^2)^n dx = 2 \int_0^1 (1 - x^2)^n dx \quad \text{since } (1 - x^2)^n \text{ is even}$$

We have, Bernoulli's inequality. $(1 + x)^r \geq (1 + rx)$, $\forall x \geq -1$, $\forall r \geq 0$

$$\begin{aligned} \int_{-1}^1 (1 - x^2)^n dx &\geq 2 \int_0^{1/\sqrt{n}} (1 - nx^2) dx \\ &\geq 2 \left(\frac{1}{\sqrt{n}} - \frac{n}{3n\sqrt{n}} \right) = \frac{4}{3\sqrt{n}} > \frac{1}{\sqrt{n}} \end{aligned}$$

Clearly, $c_n < \sqrt{n}$.

For $\delta > 0$, $Q_n(x) \leq \sqrt{n}(1 - \delta^2)^n$ for $\delta \leq |x| \leq 1$. Then $Q_n \rightarrow Q$ uniformly for all such that $\delta \leq |x| \leq 1$. Define $P_n : [0, 1] \rightarrow \mathbb{R}$ defined by

$$P_n(x) = \int_{-1}^1 f(x+t) Q_n(t) dt$$

Then,

$$\begin{aligned} P_n(x) &= \int_{-1}^1 f(x+t) Q_n(t) dt \\ &= \int_{-x}^{1-x} f(x+t) Q_n(t) dt \quad \text{since } f \text{ vanishes outside } [0, 1] \\ &= \int_0^1 f(t) Q_n(t-x) dt \end{aligned}$$

Continuous function f is uniformly continuous on compact interval $[0, 1]$. Thus f is Riemann integrable on $[0, 1]$. And $Q_n(t-x) = c_n[1 - (t-x)^2]^n$. From

integration by parts, we know that $\{P_n\}$ is a sequence of polynomials.

$$\begin{aligned}
 P_n(x) &= \int_0^1 f(t)Q_n(t+x)dt \\
 &= [f(t)Q'_n(t+x)]_0^1 - \int_0^1 Q'_n(t+x) \int_0^1 f(t)dt \\
 &= f(1)Q'_n(1+x) - f(0)Q'_n(x) - [F(1) - F(0)] \int_0^1 Q'_n(t+x)dt \\
 &= f(1)Q'_n(1+x) - f(0)Q'_n(x) - [F(1) - F(0)][Q_n(1+x) - Q_n(x)]
 \end{aligned}$$

And for each natural number n , we have $P_n(x)$ is real, if f is real.

Step 3 : $P_n \rightarrow f$ uniformly

Let $\varepsilon > 0$. Since extended f is uniformly continuous on real line, there exists $\delta > 0$ such that $|f(y) - f(x)| < \frac{\varepsilon}{2}$ whenever $|y - x| < \delta$.

$$\begin{aligned}
 |P_n(x) - f(x)| &= \left| \int_{-1}^1 f(x+t)Q_n(t) dt - f(x) \int_{-1}^1 Q_n(t) dt \right| \\
 &\leq \int_{-1}^1 |f(x+t) - f(x)|Q_n(t) dt
 \end{aligned}$$

Let $M = \sup f(x)$. Then $|f(x+t) - f(x)| \leq 2M$. And we have an upper bound for the value $Q_n(x)$ for $\delta \leq |x| \leq 1$. Therefore, we split the domain of integral into three parts so that we may apply uniform continuity on the middle part and bound of Q_n on other two parts.

$$|P_n(x) - f(x)| \leq 2M \int_{-1}^{-\delta} Q_n(t)dt + \frac{\varepsilon}{2} \int_{-\delta}^{\delta} Q_n(t)dt + 2M \int_{\delta}^1 Q_n(t)dt$$

We have an upper bound for Q_n , say $Q_n(x) \leq \sqrt{n}(1 - \delta^2)^n$ for $\delta \leq |x| \leq 1$.

$$\begin{aligned}
 |P_n(x) - f(x)| &\leq 2M\sqrt{n}(1 - \delta^2)^n \int_{-1}^{-\delta} dt + \frac{\varepsilon}{2} \int_{-\delta}^{\delta} Q_n(t)dt + 2M\sqrt{n}(1 - \delta^2)^n \int_{\delta}^1 dt \\
 &\leq 4M\sqrt{n}(1 - \delta^2)^n + \frac{\varepsilon}{2} \quad \text{since } 2(1 - \delta) < 2 \\
 &\leq \varepsilon \text{ for sufficiently large } n
 \end{aligned}$$

Therefore, there exists $N \in \mathbb{N}$ such that $\forall n > N$, $|P_n(x) - f(x)| < \varepsilon$. In other words, $P_n \rightarrow f$ uniformly on $[0, 1]$. \square

Corollary 4.4.4.1. *For every interval $[-a, a]$ there is a sequence of real polynomials P_n such that $P_n(0) = 0$ and*

$$\lim_{n \rightarrow \infty} P_n(x) = |x|$$

uniformly on $[-a, a]$.

Proof. By Weierstrass theorem, there exists a sequence $\{P_n^*(x)\}$ of real polynomials which converges to $|x|$ uniformly on $[-a, a]$. Thus, $P_n^*(0) \rightarrow 0$ as $n \rightarrow \infty$.

Define $P_n(x) = P_n^*(x) - P_n^*(0)$. Clearly, $P_n(0) = 0$ and the sequence $\{P_n(x)\}$ converges uniformly on $[-a, a]$. And,

$$\lim_{n \rightarrow \infty} P_n(x) = \lim_{n \rightarrow \infty} P_n^*(x) - \lim_{n \rightarrow \infty} P_n^*(0) = |x|$$

Therefore, $P_n(0) = 0$ and $P_n(x) \rightarrow |x|$ as $n \rightarrow \infty$. \square

Remark (Out of Syllabus). On the other side of this corollary, we have Stone-Weierstrass theorem which study functions that doesn't vanish anywhere.

Let \mathcal{A} be an algebra of functions defined on compact set K that separates points and vanishes nowhere. By Stone-Weierstrass theorem,^{†4} any continuous function f on K has a sequence of functions in \mathcal{A} which converges to f uniformly on K . The algebra of even polynomials doesn't separate points since $P_n(x) = P_n(-x)$.

4.4.2 Some special functions

Definitions 4.4.3 (analytic function). A function f is (real) analytic if it can be represented by a power series.

$$f(x) = \sum_{j=1}^{\infty} c_n x^n$$

Remark. The open interval in which a power series $\sum c_n x^n$ converges is the **interval of convergence**.

Theorem 4.4.5. Suppose series $\sum c_n x^n$ converges for $|x| < R$. Suppose function $f : (-R, R)$ is defined by

$$f(x) = \sum_{n=0}^{\infty} c_n x^n, \quad |x| < R$$

Then, for any $\varepsilon > 0$, the series $\sum c_n x^n$ converges uniformly on $[-R + \varepsilon, R - \varepsilon]$. Also the function f is continuous and differentiable in $(-R, R)$ and

$$f'(x) = \sum_{n=1}^{\infty} n c_n x^{n-1}, \quad |x| < R$$

Proof. Let $\varepsilon > 0$. Then for $|x| \leq R - \varepsilon$ we have $|c_n x^n| \leq |c_n (R - \varepsilon)^n|$.

We have,

$$\limsup_{n \rightarrow \infty} |R - \varepsilon| \sqrt[n]{|c_n|} = (R - \varepsilon) \limsup_{n \rightarrow \infty} \sqrt[n]{|c_n|} = 0 < 1$$

By ratio test, the series $\sum c_n (R - \varepsilon)^n$ converges absolutely. By Weierstrass M test, series $\sum c_n x^n$ converges uniformly on $[-R + \varepsilon, R - \varepsilon]$. In the same fashion, the series $\sum n c_n x^{n-1}$ converges uniformly on $[-R + \varepsilon, R - \varepsilon]$ since

⁴[Rudin, 1976, §7.32] Let \mathcal{A} be a family of functions defined on compact set K .
 \mathcal{A} separates points if $\forall x \in K, \exists f, g \in \mathcal{A}$ such that $f(x) \neq g(x)$.
 \mathcal{A} vanishes at a point $x \in K$ if $\forall x \in K, \exists f \in \mathcal{A}$ such that $f(x) \neq 0$

$$\limsup_{n \rightarrow \infty} \sqrt[n]{n|c_n|} = 0.$$

Let $x_0 \in (-R, R)$. Then there exists $\varepsilon > 0$ such that $x_0 \in [-R + \varepsilon, R - \varepsilon]$. The series $\sum n c_n x^{n-1}$ converges to $f'(x)$ uniformly on $[-R + \varepsilon, R - \varepsilon]$ and $f(x_0) = \sum c_n x_0^n$. Thus, f is differentiable on $[-R + \varepsilon, R - \varepsilon]$ and

$$f'(x) = \lim_{n \rightarrow \infty} \sum_{k=1}^n k c_k x^{k-1} = \sum_{n=1}^{\infty} n c_n x^{n-1}$$

We know that f is continuous at a point if it is differentiable at that point. Thus, f is continuous on $[-R + \varepsilon, R - \varepsilon]$. \square

Corollary 4.4.5.1. Suppose $\sum_{n=0}^{\infty} c_n x^n$ converges for $|x| < R$ and function f is defined by $f(x) = \sum_{n=0}^{\infty} c_n x^n$. Then f has derivatives of all ordres, say $f^{(k)}(x)$ given by

$$f^{(k)}(x) = \sum_{n=k}^{\infty} n(n-1) \cdots (n-k+1) c_n x^{n-k}$$

In particular,

$$f^{(k)}(0) = k! c_k$$

Proof. Let $f(x) = \sum c_n x^n$. Then, we have

$$f'(x) = \sum_{n=1}^{\infty} n c_n x^{n-1}$$

By mathematical induction, we have

$$f^{(k)}(x) = \sum_{n=k}^{\infty} n(n-1) \cdots (n-k+1) c_n x^{n-k}$$

When $x = 0$, we get

$$f^{(k)}(0) = \sum_{n=k}^{\infty} n(n-1) \cdots (n-k+1) c_n 0^{n-k} = k! c_k$$

\square

Theorem 4.4.6. Suppose $\sum_{n=0}^{\infty} c_n$ converges. Define

$$f(x) = \sum c_n x^n, \quad -1 < x < 1$$

Then,

$$\lim_{x \rightarrow 1} f(x) = \sum_{n=0}^{\infty} c_n$$

Proof. Let $s_n = c_0 + c_1 + \cdots + c_n$ and $s_{-1} = 0$. Suppose $\sum c_n$ converges, then s_n converges, say $s_n \rightarrow s$. Let $\varepsilon > 0$. Then there exists $N \in \mathbb{N}$ such that $\forall n > N$, $|s - s_n| < \frac{\varepsilon}{2}$.

Also we have,

$$\begin{aligned}
 \sum_{n=0}^m c_n x^n &= \sum_{n=0}^m (s_n - s_{n-1}) x^n \\
 &= \sum_{n=0}^m s_n x^n - x s_{n-1} x^{n-1} \\
 &= \sum_{n=0}^m s_n x^n - x \sum_{n=0}^m s_{n-1} x^{n-1} \\
 &= \sum_{n=0}^m s_n x^n - x \sum_{n=-1}^{m-1} s_n x^n \\
 &= \sum_{n=0}^{m-1} s_n x^n + s_m x^m - x \sum_{n=0}^{m-1} s_n x^n \text{ since } s_{-1} = 0 \\
 &= (1-x) \sum_{n=0}^{m-1} s_n x^n + s_m x^m \\
 \lim_{m \rightarrow \infty} \sum_{n=0}^m c_n x^n &= (1-x) \lim_{m \rightarrow \infty} \sum_{n=0}^{m-1} s_n x^n + \lim_{m \rightarrow \infty} s_m x^m \\
 f(x) &= (1-x) \sum_{n=0}^{\infty} s_n x^n \text{ since } s_m \rightarrow s, |x| < 1 \text{ and } x^m \rightarrow 0
 \end{aligned}$$

We know that,

$$(1-x) \sum_{n=0}^m x^n = (1-x)(1+x+x^2+\cdots+x^m) = 1-x^{m+1}$$

Thus,

$$\lim_{m \rightarrow \infty} (1-x) \sum_{n=0}^m x^n = \lim_{m \rightarrow \infty} 1 - x^{m+1} = 1$$

Thus,

$$\begin{aligned}
 |f(x) - s| &= \left| (1-x) \sum_{n=0}^{\infty} s_n x^n - s(1-x) \sum_{n=0}^{\infty} x^n \right| \\
 &= (1-x) \left| \sum_{n=0}^{\infty} (s_n - s) x^n \right| \\
 &\leq (1-x) \sum_{n=0}^{\infty} |(s_n - s) x^n| \\
 &\leq (1-x) \sum_{n=0}^{\infty} |s_n - s| |x|^n
 \end{aligned}$$

$$\leq (1-x) \sum_{n=0}^N |s_n - s| |x|^n + \frac{\varepsilon}{2} (1-x) \sum_{n=N+1}^{\infty} x^n$$

Let $1-x < \delta < 1$.^{†5}

$$\lim_{x \rightarrow 1} |f(x) - s| \leq \lim_{\delta \rightarrow 0} \delta \sum_{n=0}^N |s_n - s| (1-\delta)^n + \frac{\varepsilon}{2} = 0$$

Therefore,

$$f(1) = \lim_{x \rightarrow 1} f(x) = s = \sum_{n=0}^{\infty} c_n x^n$$

□

Theorem 4.4.7. Given a double sequence $\{a_{ij}\}$. Suppose that,

$$\sum_{j=1}^{\infty} |a_{ij}| = b_i$$

and $\sum b_i$ converges. Then,

$$\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} a_{ij} = \sum_{j=1}^{\infty} \sum_{i=1}^{\infty} a_{ij}$$

Proof. Step 1 : Construction of f_i

Given series $\sum_{j=1}^{\infty} |a_{ij}|$ converges to b_i . Let $E = \{x_0, x_1, x_2, \dots\}$ be a countable set such that $x_n \rightarrow x_0$ as $n \rightarrow \infty$. Define sequence of functions $\{f_i\}$ on E such that

$$f_i(x_n) = \sum_{j=1}^n a_{ij}, \quad \forall n \in \mathbb{N} \quad \text{and} \quad f_i(x_0) = \sum_{j=1}^{\infty} a_{ij}$$

Clearly, $f_i(x_0) = b_i$ and $f_i(x_n) \rightarrow f_i(x_0)$ as $n \rightarrow \infty$.

Step 2 : f_i is continuous at x_0

We have, $f_i(x_n) \rightarrow f_i(x_0)$. Then ^{†6},

$$\lim_{x_n \rightarrow x_0} f_i(x_n) = \lim_{n \rightarrow \infty} \sum_{j=1}^n a_{ij} = \sum_{j=1}^{\infty} a_{ij} = f_i(x_0)$$

Therefore, function f_i is continuous at x_0 .

⁵In earlier version, the term $|s_n - s|$ was neglected. Corrected as per the seminar by Haripriya

⁶The method of contradiction was an unnecessary complication. Corrected as per the seminar by Mekha

Step 3 : Construction of g

Given $|f_i(x)| < b_i$ and $\sum b_i$ converges. Thus, $\sum f_i(x)$ converges. Define $g : E \rightarrow \mathbb{R}$ such that

$$g(x) = \sum_{i=1}^{\infty} f_i(x)$$

Since f_i are continuous functions defined on a countable set E , the convergence is uniform and the sum g is continuous. And we have,

$$\lim_{n \rightarrow \infty} \lim_{m \rightarrow \infty} \sum_{i=1}^m f_i(x_n) = \lim_{x_n \rightarrow x_0} g(x_n) = g(x_0) = \lim_{m \rightarrow \infty} \sum_{i=1}^m f_i(x_0) = \lim_{m \rightarrow \infty} \lim_{n \rightarrow \infty} \sum_{i=1}^m f_i(x_n)$$

Therefore,

$$\sum_{i=1}^{\infty} \sum_{j=1}^{\infty} a_{ij} = \sum_{j=1}^{\infty} \sum_{i=1}^{\infty} a_{ij}$$

□

Theorem 4.4.8 (Taylor). *Suppose $f(x) = \sum c_n x^n$, the series converging in $|x| < R$. If $-R < a < R$, then f can be expanded in a power series about the point $x = a$ which converges in $|x - a| < R - |a|$. And,*

$$f(x) = \sum_{n=0}^{\infty} \frac{f^{(n)}(a)}{n!} (x - a)^n, \quad |x - a| < R - |a|$$

Proof. Suppose

$$\begin{aligned} f(x) &= \sum_{n=0}^{\infty} c_n [(x - a) + a]^n \\ &= \sum_{n=0}^{\infty} c_n \sum_{m=0}^n \binom{n}{m} a^{n-m} (x - a)^m \end{aligned}$$

Changing the order of summation, we may combine coefficients of $(x - a)^m$.

$$= \sum_{m=0}^{\infty} \sum_{n=m}^{\infty} \binom{n}{m} c_n a^{n-m} (x - a)^m$$

Therefore, it is enough to prove that the order of summation can be changed. We know that the order of summation can be changed if,

$$\sum_{n=0}^{\infty} \sum_{m=0}^n \left| c_n \binom{n}{m} a^{n-m} (x - a)^m \right| \text{ converges}$$

We know that,

$$\sum_{n=0}^{\infty} \sum_{m=0}^n \left| c_n \binom{n}{m} a^{n-m} (x - a)^m \right| = \sum_{n=0}^{\infty} |c_n| (|x - a| + |a|)^n$$

and it converges if $|x - a| + |a| < R - |a|$.

Also we know that, if $f(x) = \sum c_n x^n$ converges in $|x| < R$, then the convergence is uniform in $[-R + \varepsilon, R - \varepsilon]$ and it is differentiable in $(-R, R)$. And the derivatives are given by,

$$f^{(m)}(a) = \sum_{n=m}^{\infty} \frac{n!}{(n-m)!} c_n a^{n-m} = m! \sum_{n=m}^{\infty} \binom{n}{m} c_n a^{n-m}$$

Thus,

$$f(x) = \sum_{m=0}^{\infty} \sum_{n=m}^{\infty} c_n \binom{n}{m} a^{n-m} (x-a)^m = \sum_{m=0}^{\infty} \frac{f^{(m)}(a)}{m!} (x-a)^m$$

□

Theorem 4.4.9. Suppose the series $\sum a_n x^n$ and $\sum b_n x^n$ converge in the segment $S = (-R, R)$. Let E be the set of all $x \in S$ at which

$$\sum_{n=0}^{\infty} a_n x^n = \sum_{n=0}^{\infty} b_n x^n$$

If E has a limit point on S , then $a_n = b_n$.

In other words, If two power series covers to the same function in $(-R, R)$, then the series are identical.

Proof. —continue—

□

Whether the power series representation is unique ?

Theorem 4.4.10. Let e^x be defined on \mathbb{R} . Then

1. e^x is continuous and differentiable for all x .
2. $(e^x)' = e^x$
3. e^x is a strictly increasing function
4. $e^{x+y} = e^x e^y$
5. $e^x \rightarrow +\infty$ as $x \rightarrow \infty$ and $e^x \rightarrow 0$ as $x \rightarrow -\infty$
6. $\lim_{x \rightarrow \infty} x^n e^{-x} = 0$ for every n .

Proof. Let $E(z) = \sum_{k=0}^{\infty} \frac{z^k}{k!}$.

We know that,

$$E(1) = \sum_{k=0}^{\infty} \frac{1^k}{k!} = e \quad \text{and} \quad E(0) = 1$$

And,

$$\begin{aligned}
 E(z)E(w) &= \sum_{n=0}^{\infty} \frac{z^n}{n!} \sum_{m=0}^{\infty} \frac{w^m}{m!} \\
 &= \sum_{n=0}^{\infty} \sum_{k=0}^{\infty} \frac{z^k w^{n-k}}{k!(n-k)!} \\
 &= \sum_{n=0}^{\infty} \frac{1}{n!} \sum_{k=0}^{\infty} \binom{n}{k} z^k w^{n-k} \\
 &= \sum_{n=0}^{\infty} \frac{(z+w)^n}{n!} \\
 &= E(z+w)
 \end{aligned}$$

Then we have,

$$E(z)E(-z) = E(0) = 1$$

We $E(1) = e > 1$. Let M be any integer. Then $E(M) = e^M > M$. Thus there exists $x \in \mathbb{R}$ such that $E(x) > M$. Therefore, $E(x) \rightarrow +\infty$ as $x \rightarrow +\infty$.

From definition, $E(x) > 0$ for any $x > 0$. And for any $x < 0$, $-x > 0$ and $E(-x) > 0$. Therefore, $E(x) = 1/E(-x) > 0$. Thus, $E(x) > 0$ for any $x \in \mathbb{R}$. Let $\varepsilon > 0$. Then there exists an integer M such that $\varepsilon > 1/M$. We know that there exists $x \in \mathbb{R}$ such that $E(x) > M$. Then, $0 < E(-x) = 1/E(x) < 1/M < \varepsilon$. Thus, $E(x) \rightarrow 0$ as $x \rightarrow -\infty$.

$$\begin{aligned}
 E'(z) &= \lim_{h \rightarrow 0} \frac{E(z+h) - E(z)}{h} \\
 &= \lim_{h \rightarrow 0} \frac{E(z)E(h) - E(z)}{h} \\
 &= E(z) \lim_{h \rightarrow 0} \frac{E(h) - 1}{h} \\
 &= E(z) \lim_{h \rightarrow 0} \left(h + \frac{h^2}{2!} + \frac{h^3}{3!} + \dots \right) / h \\
 &= E(z)
 \end{aligned}$$

Thus, $E(x)$ is differentiable at every $x \in \mathbb{R}$. Therefore, $E(x)$ is continuous at every $x \in \mathbb{R}$. Thus, the function $E : \mathbb{R} \rightarrow (0, \infty)$ is strictly increasing.

By mathematical induction

$$E(z_1 + z_2 + \dots + z_n) = E(z_1)E(z_2) \dots E(z_n)$$

Thus,

$$E(n) = E(1)^n = e^n, \quad \forall n \in \mathbb{N}$$

Also we have,

$$E(-1) = \frac{1}{E(1)} = E(1)^{-1} = e^{-1}$$

$$E(-n) = E(-1)^n = \frac{1}{E(n)^n} = E(1)^{-n} = e^{-n}, \quad \forall n \in \mathbb{N}$$

Thus,

$$E(n) = e^n, \quad \forall n \in \mathbb{Z}$$

Let $p \in \mathbb{Q}$. Then $p = n/m$. And,

$$E(p)^m = E(mp) = E(n) = e^n \implies E(p) = e^p, \quad \forall p \in \mathbb{Q}$$

Let $x \in \mathbb{R}$. Define

$$E(x) = \sup_{p < x, p \in \mathbb{Q}} E(p)$$

By continuity and monotonicity of E , we have $E(x) = e^x$, $\forall x \in \mathbb{R}$. Now rewriting the results, we get

$$e^{x+y} = e^x e^y \quad (4.1)$$

$$\frac{d}{dx} e^x = e^x \quad (4.2)$$

We have,

$$\begin{aligned} e^x &= \sum_{k=0}^{\infty} \frac{x^k}{k!} \\ &> \frac{x^{n+1}}{(n+1)!} \\ x^n e^{-x} &< \frac{(n+1)!}{x} \end{aligned}$$

$$\lim_{x \rightarrow \infty} x^n e^{-x} < (n+1)! \lim_{x \rightarrow \infty} \frac{1}{x} = 0, \quad \forall n \in \mathbb{N}$$

Thus, $x^n e^{-x} \rightarrow 0$ as $x \rightarrow \infty$. As $x \rightarrow +\infty$, $e^x \rightarrow +\infty$ faster than any positive power x . **This is no surprise since $E(1) = e > 1 = E(0)$.** \square

Remark. The function $E : \mathbb{R} \rightarrow (0, \infty)$ is a bijection. Thus, there exists an inverse function $L : (0, \infty) \rightarrow \mathbb{R}$ such that $E \circ L = I_{(0, \infty)}$ and $L \circ E = I_{\mathbb{R}}$ such that

$$L(y) = \int_1^y \frac{1}{x} dx$$

Furthermore, $L(x) \rightarrow +\infty$ as $x \rightarrow +\infty$ and $L(x) \rightarrow 0$ as $x \rightarrow -\infty$.

Proof. Let $u, v \in (0, \infty)$. Then there exists $x, y \in \mathbb{R}$ such that $E(x) = u$ and $E(y) = v$. Thus,

$$L(uv) = L(E(x)E(y)) = L(E(x+y)) = x+y$$

Let $y = E(x)$. We have, $L(E(x)) = x$. And $L'(E(x))E'(x) = yL'(y) = 1$ since $E'(x) = E(x) = y$. Thus, $L'(y) = 1/y$.

From fundamental theorem for calculus, we have

$$\int_a^b f(t) dt = F(b) - F(a)$$

where $f = F'$ on (a, b) . We have $E(x)$ is strictly monotonic and continuous, thus $L(x)$ is monotonic and continuous on $(0, \infty)$. And $L(x)$ is Riemann integrable on any interval $[a, b]$ subset of $(0, \infty)$. Put $F = L$. Then,

$$\int_a^b L'(x)dx = \int_a^b \frac{dx}{x} = L(b) - L(a)$$

We know that $L(1) = 0$. Put $a = 1$ and $b = y$. Then,

$$L(y) = \int_1^y \frac{dx}{x}$$

By mathematical induction,

$$L(x^m) = mL(x) \quad \forall m \in \mathbb{N}$$

$$L(x) = L(x^{\frac{m}{m}}) = L(x^{\frac{1}{m}})^m = mL(x^{\frac{1}{m}}) \implies L(x^{\frac{1}{m}}) = \frac{1}{m}L(x)$$

Thus, $x^n = E(nL(x))$ and $x^{\frac{1}{m}} = E(\frac{1}{m}L(x))$. And,

$$x = E(L(x)) = E(L(\frac{1}{m}x)^m) = E(mL(\frac{1}{m}x))$$

Therefore,

$$x^\alpha = E(\alpha L(x)), \quad \forall x \in \mathbb{R}, x > 0, \forall \alpha \in \mathbb{Q}$$

Rewriting the relation with $\log x$ instead of $L(x)$ and e^x instead of $E(x)$, we get $x^\alpha = e^{\alpha \log x}$ for any rational α .

We have,

$$\frac{d}{dx}x^\alpha = \frac{d}{dx}E(\alpha L(x)) = E'(\alpha L(x))\alpha L'(x) = E(\alpha L(x))\frac{\alpha}{x} = \alpha x^{\alpha-1}$$

Suppose $\alpha \neq -1$. For irrational values of x , $L(x)$ can be obtained from,

$$E(\alpha L(x)) = \int_1^x y^\alpha dy$$

We have,

$$\begin{aligned} x^{-\alpha} \log x &= x^{-\alpha} \int_1^x t^{-1} dt \\ &< x^{-\alpha} \int_t^x t^{\varepsilon-1} dt \\ &< x^{-\alpha} \frac{x^\varepsilon - 1}{\varepsilon} \end{aligned}$$

Clearly,

$$\lim_{x \rightarrow +\infty} x^{-\alpha} \log x = 0$$

As $x \rightarrow +\infty$, we have $\log x \rightarrow +\infty$ slower than any positive power of x . \square

Theorem 4.4.11. 1. The function E is periodic with period $2\pi i$.

2. The functions C and S are periodic with period $2\pi i$.
3. If $0 < t < 2\pi$, then $E(it) \neq 1$.
4. If z is a complex number with $|z| = 1$, there is a unique $t \in [0, 2\pi]$ such that $E(it) = z$.

Proof. Let $E(z) = \sum_{n=0}^{\infty} \frac{z^n}{n!}$. □

Theorem 4.4.12. Suppose a_0, a_1, \dots, a_n are complex numbers.

$$P(z) = \sum_{n=0}^{\infty} a_n z^n$$

Then $P(z) = 0$ for some complex number z .

Proof. Without loss of generality, suppose $a_n = 1$. Put $\mu = \inf |P(z)|$, infimum exists since $|P(z)| > 0$. It is enough to prove that $\mu = 0$.

Suppose $|z| = R$. Then

$$\begin{aligned} |a_n z^n| &= |P(z) - a_0 - a_1 z - a_2 R^2 - \dots - a_{n-1} R^{n-1}| \\ R^n &\leq |P(z)| + |a_0| + |a_1| R + \dots + |a_{n-1}| R^{n-1} \end{aligned}$$

Rearranging the terms, we get

$$\begin{aligned} |P(z)| &\geq R^n - |a_0| - |a_1| R - |a_2| R^2 - \dots - |a_{n-1}| R^{n-1} \\ &\geq R^n (1 - |a_{n-1}| R^{-1} - |a_{n-2}| R^{-2} - \dots - |a_0| R^{-n}) \end{aligned}$$

Thus $|P(z)| \rightarrow \infty$ as $R \rightarrow \infty$. Hence, there exists R_0 such that $|P(z)| > \mu$ if $|z| > R_0$. Since $P(z)$ is continuous, $|P(z)|$ is also continuous. And the continuous function $|P(z)|$ on closed disc centered at origin with radius R_0 attains minima, say μ , since closed disc is compact. Thus, $|P(z_0)| = \mu$ for some z_0 such that $|z_0| \leq R_0$.

Suppose $\mu > 0$. Then $P(z_0) \neq 0$. Define $Q(z) = P(z + z_0)/P(z_0)$. Then, Q is a nonconstant polynomial with $Q(0) = 1$ and $|Q(z)| \geq 1$, $\forall z$. Clearly, $Q(z)$ is a polynomial of at most n degrees.

$$Q(z) = b_0 + b_1 z + b_2 z^2 + \dots + b_n z^n$$

where $b_0 = 1$ since $Q(0) = 1$. Let k be the smallest positive integer such that $b_k \neq 0$. That is, $1 \leq k \leq n$. We have $b_k = |b_k| e^{i\phi}$. Choose θ such that $\phi = \pi - k\theta$. Then $b_k = |b_k| e^{i(\pi - k\theta)} = -|b_k| e^{-ik\theta}$. Then $e^{ik\theta} b_k = -|b_k|$.

Let $z = r e^{i\theta}$ such that $r > 0$ and $r^k |b_k| < 1$. Then,

$$|1 + b_k r^k e^{ik\theta}| = 1 - r^k |b_k|$$

Thus,

$$|Q(z)| = |Q(r e^{i\theta})| \leq 1 - r^k (|b_k| - r |b_{k+1}| - \dots - r^{n-k} |b_n|)$$

For sufficiently small r , $|Q(r e^{i\theta})| < 1$ which is a contradiction. Thus $\mu = 0$. Therefore, there exists $|z_0| \leq R_0$ such that $P(z_0) = 0$. □

In other words, complex field \mathbb{C} is algebraically complete.

Proof. Since every polynomial with complex coefficients has a zero, there is no irreducible polynomial of degree greater than or equal to two. Thus, \mathbb{C} doesn't have an algebraic field extension properly containing \mathbb{C} . Therefore, \mathbb{C} is complete. \square

Subject 5

Graph Theory

5.1 Basic Results

5.1.1 Introduction

[R. Balakrishnan,]

5.1.2 Basic Concepts

Definitions 5.1.1. A **graph** is an ordered triple $G = (V, E, I)$ where V is a nonempty set of vertices, E is a set of edges and I is an incidence map which associates each edge with an unordered pair of vertices.

end vertices Let e be an edge with $I_G(e) = \{u, v\}$. Then u, v are the end vertices of e . We may write $e = uv$.

incident An edge $e = uv$ said to be incident on both vertices u and v . Then vertices u and v are incident on edge e as well.

parallel edges are those edges which have same pair of end vertices.

loop is an edge whose both end vertices are the same.

neighbour Vertex u is neighbour of vertex v if uv is an edge of the graph.

Open neighbourhood $N(u)$ is the set of all neighbours of the vertex u .

Closed neighbourhood $N[u] = N(u) \cup \{u\}$.

simple graph does not have any parallel edges or loops.

adjacent Two vertices u, v are adjacent if both u, v are incident on an edge, say uv . Two edges are adjacent if they have a common end vertex.

Definitions 5.1.2. A graph is **finite** if both vertex set V and edge set E are finite. A finite graph, G order, $n(G) = |V(G)|$ and size, $m(G) = |E(G)|$. Or simply $n = |V(G)|$ and $m = |E(G)|$.

Definitions 5.1.3. A graph G is **labeled** if its vertices are distinguished from one another by means of distinct labels.

Definitions 5.1.4. Two graphs G and H are **isomorphic** if there exists a pair (ϕ, θ) where $\phi : V(G) \rightarrow V(H)$ and $\theta : E(G) \rightarrow E(H)$ are bijections such that $I_G(e) = \{u, v\} \iff I_H(\theta(e)) = \{\phi(u), \phi(v)\}$.

Two simple graphs G and H are **isomorphic** if there exists a bijection $\phi : V(G) \rightarrow V(H)$ which induced another bijection $\theta : E(G) \rightarrow E(H)$ such that $I_G(e) = \{u, v\} \iff I_H(\theta(e)) = \{\phi(u), \phi(v)\}$.

Exercise 5.1.1. Let G, H be simple graph and let $\phi : V(G) \rightarrow V(H)$ be a bijection such that $uv \in E(G) \implies \phi(u)\phi(v) \in E(H)$. Show that ϕ is not an isomorphism.

Let G, H be simple graph with bijection $\phi : V(G) \rightarrow V(H)$. If there exists two vertices u, v which are non-adjacent in G , but $\phi(u), \phi(v)$ are adjacent in H . Then ϕ is not an isomorphism.

Definitions 5.1.5. A simple graph is **complete** if every pair of distinct vertices are adjacent. A complete graph with n vertices is denoted by K_n . Then $m(K_n) = \binom{n}{2} = \frac{n(n-1)}{2}$. A **totally** disconnected graph has no edges. Thus, $0 \leq m \leq \binom{n}{2}$. A graph G is trivial if it has only one vertex and no edges.

Definitions 5.1.6. A graph G is **bipartite** if its vertex set can be partitioned into two nonempty sets X and Y such that every end of G has one end in X and the other end in Y . We write $G(X, Y)$ is bipartite with partition X, Y . A simple, bipartite graph $G(X, Y)$ is **complete bipartite** if every vertex in X is adjacent to every vertex in Y . Let $G(X, Y)$ be a complete bipartite graph with $|X| = p$ and $|Y| = q$, then we write $K_{p,q}$. A graph of the form $K_{1,q}$ is called a **star**.

Remark. Let G be a complete bipartite graph, $K_{p,q}$. Then it has order $n = p + q$ and size $m = pq$.

Definitions 5.1.7. Let G be a graph. The complement G^c of graph G is the graph with same vertex set. Two vertices u, v of G^c are adjacent if and only if u, v are non-adjacent in G .

$$V(G^c) = V(G) \quad \& \quad uv \in E(G^c) \iff uv \notin E(G)$$

Remark. We have, $(G^c)^c = G$, since $uv \in G \iff uv \notin G^c \iff uv \in (G^c)^c$ and $V(G) = V(G^c) = V((G^c)^c)$. Let G be a graph of order n , then $E(G) + E(G^c) = E(K_n)$ as each edge of K_n is either an edge of G or an edge of G^c .

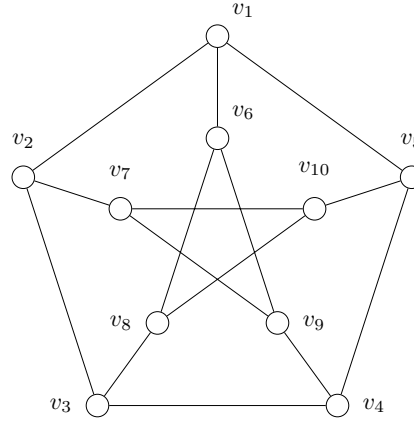
Definitions 5.1.8. A simple graph is self-complementary if $G \cong G^c$.

Remark. The order of a self complementary graph G of order n is $n(n-1)/4$ since $m(G) = m(G^c)$ and $m(G) + m(G^c) = m(K_n)$.

5.1.3 Subgraphs

subgraph A graph H is a subgraph of G if $V(H) \subset V(G)$, $E(H) \subset E(G)$ and I_H is I_G restricted to $E(H)$. Then G is a **supergraph** of H .

induced subgraph Let G be a graph and S be a subset of the vertex set of G . The subgraph induced by S , $G[S]$ has vertex set S and two vertices are adjacent only if they are adjacent in G .



edge induced subgraph Let G be a graph and S be a subset of the edge set of G . The subgraph induced by the edge set S is denoted by $G[S]$. Vertex u is a vertex of $G[S]$ only if S has an edge incident on it.

spanning subgraph Let H be a subgraph of graph G . If $V(H) = V(G)$, then H is a spanning subgraph.

clique is a subgraph which is complete. A clique is maximal if it is not contained in another clique.

Remark. If an edge e is deleted from a graph G , then vertex set remains the same. The graph $G - \{e\}$ is a spanning subgraph of G . If a vertex u is deleted from a graph G , then all the edge incident on u are also deleted. The graph $G - \{u\}$ is an induced subgraph of G .

5.1.4 Degree of Vertices

Definitions 5.1.9. Let G be a graph and u be a vertex of G . The **degree** of u is the number of edges incident on it with multiplicities. That is, every loop incident on u is counted twice.

Remark. In a simple graph, the degree of a vertex is the cardinality of its open neighbourhood. $\deg_G(u) = |N_G(u)|$.

Definitions 5.1.10. A graph G is **k -regular** if every vertex is of degree k . Graph G is **regular** if it is k -regular for some k . **Cubic** graphs are the 3-regular graphs.

Remark. Complete graph K_{n+1} are n -regular. And complete graphs are the smallest regular graphs. K_4 is cubic. Petersen graph is cubic.

Remark. The complement of a regular graph is also regular. If G is k -regular, then G^c is r -regular where $k + r = n - 1$.

1-factor is a spanning, 1-regular subgraph.

isolated vertex is a vertex with degree 0.

pendent vertex is a vertex with degree 1.

pendent edge is the only edge incident on a pendent vertex.

Theorem 5.1.1 (Euler). *The degree sum of a graph is twice its size.*

Proof. Every edge $e = uv$ contributes 1 to the degree of both the end vertices u and v . Thus every edge contributes 2 to the degree sum. There are m edges, thus degree sum is $2m$. \square

Corollary 5.1.1.1. *In a graph G , the number of vertices of odd degree is even.*

Proof. Let G be a graph of order n . Let V_1, V_2 be the set of vertices of even, odd degree respectively. Then, $\sum d_i = \sum_{v \in V_1} \deg(v) + \sum_{v \in V_2} \deg(v)$. The degree sum is even and the first sum on RHS is also even. Thus the second sum should be even. Since, each term in the second sum is an odd integer, there are even number of terms in the second sum. In other words, there are even number of vertices with an odd degree. \square

Exercise 5.1.2. *If $G \stackrel{\phi}{\cong} H$, then each pair $u, \phi(u)$ have the same degree.*

Let G, H be isomorphic graphs. Let u be a vertex of G . A vertex v is adjacent of u in G if and only if $\phi(v)$ is adjacent to $\phi(u)$ in H . Thus, $\deg_G(u) = \deg_H(\phi(u))$.

Remark. Clearly, a graph isomorphism preserves adjacency, degree of vertices and neighbourhoods, etc. $N_H(\phi(u)) = \{\phi(v) : v \in N_G(u)\}$.

Exercise 5.1.3. *Let $d : d_1, d_2, \dots, d_n$ be the degree sequence of a graph G . Let r be a positive integer, then $\sum d_i^r$ is even.*

Let G be a graph of order n . Let V_1, V_2 be the set of vertices of even and odd degree respectively. Clearly, $\sum d_i^r = \sum_{v \in V_1} d_i^r + \sum_{v \in V_2} d_i^r$. Once again, the first sum on RHS is even as d_i^r is even when d_i is even. And the second sum is odd as d_i^r is odd when d_i is odd. By Euler's theorem, there are an even number of such odd terms. Thus, the second sum on RHS is also even. Therefore, $\sum d_i^r$ is even.

Definitions 5.1.11. A sequence of nonnegative integers $d : d_1, d_2, \dots, d_n$ is **graphical** if there exists a *simple* graph with degree sequence d .

Example 5.1.1. *The sequence $d : 7, 6, 3, 3, 2, 1, 1, 1$ is nongraphical. Let v_0, v_1 be the vertices with degree 7, 6 respectively. Then for d to be graphical, there should be at least another 4 vertices with degree greater than or equal to 2. This is not the case, therefore d is not graphical.*

Exercise 5.1.4. *Let $d : d_1, d_2, \dots, d_n$ is a sequence of nonnegative integers with $\sum d_i$ even. Show that there exists a non-simple graph with d as its degree sequence.*

Let $d : d_1, d_2, \dots, d_n$ be a sequence of nonnegative integers with $\sum d_i$ even. Then there are even number of odd integers (if any).

Step 1 : Draw a vertex u_i for each term d_i . If a pair (d_i, d_j) is odd¹, then draw an edge $u_i u_j$ and subtract 1 from both d_i and d_j .

Step 2 : If $d_i = 2k$. Draw k loops on u_i .

Challenge 4. Draw maximal connected graph for a degree sequence ?

Application 5.1.1. In any group of n persons ($n \geq 2$), there are at least two with same number of friends.

Proof. Suppose group of n persons is modelled by a graph with n vertices and two vertices are adjacent if the respective persons are friends. It is assumed that friendship is mutual, otherwise it is not considered. Suppose all of them have different number of friends. Then every vertex in the graph has different degree.

The possible degree of a vertex in a graph of order n is $0, 1, \dots, (n-1)$. Since all n vertices have different degree and we have only n options. There exists a vertex of degree j for each $0 \leq j < n$. This leads to a contradiction.

Let v_1, v_n be the vertices with degree $0, (n-1)$. Then v_n is adjacent to every other vertex and v_1 is non-adjacent to every other vertex which is not possible. Thus, at least two vertices should have same degree. Therefore, in a group of n persons, at least two of them have same number of friends. \square

Exercise 5.1.5. Let G be a graph in which every vertex is of degree k or $k+1$. Then the number of vertices with degree k is $(k+1)n - 2m$.

Proof. Let G be a graph of order n and size m . Let x be the number of vertices with degree k . Then there are $n-x$ vertices with degree $k+1$. By Euler's theorem, $xk + (n-x)(k+1) = 2m \implies x = (k+1)n - 2m$. \square

5.1.5 Paths and Connectedness

Definitions 5.1.12. Let G be a graph. A **walk** on G is an alternating finite sequence of vertices and edges which starts and ends at some vertices, say $W : v_0 e_0 v_1 e_1 \dots e_k v_k$. The vertex v_0 is the **origin** and v_k is the **terminus** of the walk W .

closed walk A walk is **closed** if its origin and terminus are identical. Otherwise the walk is **open**.

trail is a walk in which edges are distinct.

path is a trail in which vertices are distinct.

cycle is a closed trail in which vertices are distinct.

Remark. A walk of length zero is a single vertex. And this walk is called a **trivial path**. Let $P : v_0, e_1, v_1, e_2, v_2, \dots, v_{k-1}, e_k, v_k$ be a path in G . Then the **inverse of the path** is given by, $P^{-1} : v_k, e_k, v_{k-1}, \dots, v_1, e_1, v_0$. And $v_j, e_j, v_{j+1}, \dots, v_{k-1}, e_{k-1}, v_k$ is the $v_j - v_k$ **section** of the path P .

¹If you want maximal simple graph, you may add an edge $u_i u_j$ (if it not already there) and subtract 1 from both d_i and d_j , provided both d_i and d_j are nonzero.

Definitions 5.1.13. Let G be a graph. Then connectedness is an equivalence relation on $V(G)$. Let $V_1, V_2, \dots, V_\omega$ be the equivalence classes. Then the induced subgraphs $G[V_1], G[V_2], \dots, G[V_\omega]$ are the **components** of G .

Remark. For a connected graph, $\omega = 1$. And for disconnected graphs $\omega \geq 2$. And components are maximal connected subgraphs of G . The number of components of G is denoted by $\omega(G)$.

Definitions 5.1.14. Let $d : V(G) \rightarrow V(G)$, $d(u, v)$ is the length of the shortest $u - v$ path. If there is no $u - v$ path, then $d(u, v) = \infty$.

Exercise 5.1.6. Let G be a simple graph. The vertex set $V(G)$ together with distance function $d(u, v)$, length of shortest $u - v$ path is a metric space.

Proof. Let G be a simple graph with vertex set $V(G)$ and $d : V(G) \rightarrow V(G)$, $d(u, v)$ is the length of a shortest $u - v$ path.

1. Every path has non-negative length. Thus, $\forall u, v \in V(G)$, $d(u, v) \geq 0$.
And $d(u, u) = 0$ since trivial path has zero length.
2. Suppose u, v are connected in G . Otherwise $d(u, v) = d(v, u) = \infty$.
Let P be a shortest $u - v$ path. Then P^{-1} is a shortest $v - u$ path.
Suppose there exists a $v - u$ path, Q which is shorter than P^{-1} . Then Q^{-1} is shorter than P which is a contradiction. Thus, $d(u, v) = d(v, u)$.
3. Suppose $d(u, v) < \infty$. Otherwise the result is trivial.
Let P, Q be shortest $u - w$ path, $w - v$ path in G . Then $P + Q$ is a $u - v$ walk. Thus, there exists a $u - v$ path² of length less than or equal to the length of $P + Q$. Therefore, $d(u, v) \leq d(u, w) + d(w, v)$.

□

Proposition 5.1.1. If G is a simple graph with $\delta(G) \geq \frac{n-1}{2}$, then G is connected.

Proof. Let G be a graph of order n . Suppose G is not connected. Then G has at least two components, say G_1 and G_2 . Let $v \in V(G_1)$. Since $\deg(v) \geq \frac{n-1}{2}$, there are at least $\frac{n-1}{2}$ other vertices in G_1 . Thus, each component of G has at least $\frac{n+1}{2}$ vertices. And G has at least $\frac{n+1}{2} + \frac{n+1}{2}$ vertices which is a contradiction. Therefore, G is connected. □

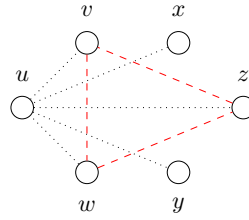
Exercise 5.1.7. A simple graph with $\delta(G) \geq \frac{n-2}{2}$ is not necessarily connected.

Proof. Let G be a simple graph with $\delta(G) = \frac{n-2}{2}$ where n is even. Then G is not necessarily connected. For example, $2K_2$ has $\delta = 1$ and is disconnected. □

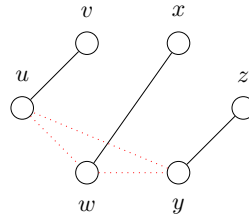
Exercise 5.1.8. In a group of six people, there must be three people who are mutually acquainted or three people who are mutually nonacquainted.

Proof. Suppose the result is not true. Then every person u has at least one friend. Suppose u has no friends. Then (v, w, x) are mutual friends. Otherwise, either (u, v, w) , (u, v, x) or (u, w, x) are mutually non-acquainted.

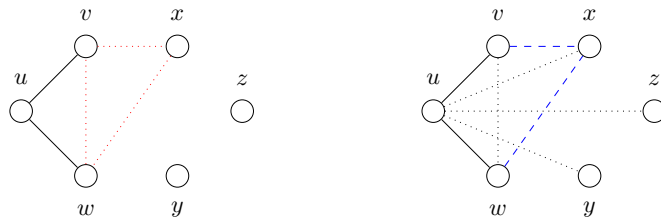
²Every $u - v$ walk, contains a $u - v$ path which can be obtained by subsequently replacing each $w - w$ section of the walk with w .

Figure 5.1: Minimum Degree $\delta(G) > 0$

There exists a person u with at least two friends. Suppose every person has exactly one friend, say $(u, v), (w, x), (y, z)$ are friends. Then (u, w, y) are mutually non-acquainted.

Figure 5.2: Maximum degree $\Delta(G) > 1$

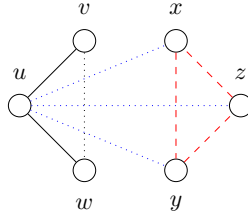
Suppose u has at least two friends, say v, w . Then v, w are not friends, otherwise (u, v, w) are mutual friends. Every other person x, y and z is friend of either v or w . Suppose x is not a friend of both v and w . Then (v, w, x) are mutually non-acquainted. Suppose x is a friend of v or w or both. Then u is not a friend of x . Otherwise either (u, v, x) or (u, w, x) are mutual friends.

Figure 5.3: $\deg_G(u) = 2$ or $\Delta(G) = 2$

Similarly, u is not a friend of y and z . Thus, $\deg(u) = 2$. And $\Delta(G) = 2$.

Suppose (x, y, z) are not mutual friends. Then either $(u, x, y), (u, y, z)$, or (u, z, x) are mutually non-acquainted which is a contradiction. \square

Theorem 5.1.2. *If a simple graph G is not connected, then G^c is connected.*



Proof. Let G be a disconnected graph with at least two components G_1, G_2 . Let $u, v \in V(G)$. It is enough to prove that there exists a $u - v$ path in G^c .

Case 1 : Two vertices u, v belong to different components of G . Then u, v are non-adjacent in G . Thus, u, v are adjacent in G^c . Therefore, there exists a $u - v$ path in G^c .

Case 2 : Suppose u, v belongs to the same component of G . Let w be a vertex from another component of G . Then u, w, v is a $u - v$ path in G^c .

Therefore, there exists a $u - v$ path between any pair of vertices in G^c . \square

Exercise 5.1.9. Let G be a self-complementary graph. Then $n(G) \cong 0$ or $1 \pmod{4}$.

Proof. Let G be a self-complementary graph of order n . Then $m(G) = m(G^c)$ since $G \cong G^c$. And $m(G) + m(G^c) = m(K_n)$ since every edge in K_n is either an edge of G or an edge of G^c . We know that, $m(K_n) = n(n-1)/2$. Therefore, $m(G) = n(n-1)/4$. Since $n, n-1$ are consecutive, at most one of them is even. Therefore, either $n \cong 0 \pmod{4}$ or $n-1 \cong 0 \pmod{4}$. \square

Exercise 5.1.10. Let G be a self-complementary graph. If G has one pendent vertex, then it has another pendent vertex.

Proof. Let G be a self-complementary graph with isomorphism ϕ from $V(G)$ to $V(G^c)$. Let u be a pendent vertex of G . Then $\phi(u)$ is also a pendent vertex of G . Suppose $u = \phi(u)$. Then, $n-1 = \deg(u) + \deg(\phi(u)) = 2$. However, no graph of order 3 is self-complementary. Therefore, $u \neq \phi(u)$. \square

Exercise 5.1.11. Let G be a simple graph. Then $m < \frac{(n-\omega)(n-\omega+1)}{2}$.

Proof. Let $G_1, G_2, \dots, G_\omega$ be the components of G . Let n_i, m_i be the order and size of G_i . Since every component is nonempty, $n - \omega + 1$ is an upperbound for

order of any component of G .

$$\begin{aligned}
 m &\leq \sum_{i=1}^{\omega} m_i \\
 &\leq \sum_{i=1}^{\omega} n_i(n_i - 1)/2, \quad \text{by Euler's theorem} \\
 &\leq \frac{n - \omega + 1}{2} \sum_{i=1}^{\omega} (n_i - 1), \quad \text{applying upperbound for } n_i \\
 &\leq \frac{n - \omega + 1}{2} (n - \omega)
 \end{aligned}$$

□

Definitions 5.1.15. A graph G is **locally connected** at a vertex v if the subgraph induced by open neighbourhood of v in G is connected. And G is locally connected if it is locally connected at every vertex.

Theorem 5.1.3 (Odd cycle characterisation of Bipartite graphs). .
A graph is bipartite if and only if it has no odd cycles.

Proof. Let G be a bipartite graph with partite sets X, Y . Let $C : v_0, v_1, \dots, v_k, v_0$ be a cycle in G . WLOG suppose $v_0 \in X$. Then $v_1 \in Y, v_2 \in X, \dots$. Clearly, if j is even, then $v_j \in X$. Since v_k, v_0 are adjacent k is odd and the cycle is of even length.

Suppose G has no odd cycles. Suppose G is connected. Let u be vertex of G . Define $X = \{v \in V(G) : d(u, v) \text{ is even}\}$ and $Y = \{v \in V(G) : d(u, v) \text{ is odd}\}$. Suppose vertices $v, w \in X$ are adjacent. Since $v, w \in X$, there exists $u - v$ path P and $u - w$ path Q both are of even length. Then $P + Q$ is not necessarily a $v - w$ path, let w' be the last vertex they have in common. Then the $w' - v$ section of P , say P' and $w' - w$ section of Q , say Q' are disjoint. If $u - w'$ is of even(odd) length, then both P', Q' are even(odd). And $P' + Q' + vw$ is an odd cycle which is a contradiction.

Similary, if $v, w \in Y$ are adjacent then respective disjoint path P', Q' are both of odd length. And $P' + Q' + vw$ is again an odd cycle which is a contradiction. Thus, $G(X, Y)$ is bipartite.

Suppose G is disconnected. Then each component $G_i(X_i, Y_i)$ is bipartite. And $G(X, Y)$ is bipartite where $X = \cup X_i$ and $Y = \cup Y_i$. □

Exercise 5.1.12. *Let G be a simple, nontrivial graph. Graph G is connected if and only if there exists an edge between any two nonempty partitions of $V(G)$.*

Proof. Suppose G is connected. Let V_1, V_2 be two nonempty partitions of $V(G)$. Let $v_1 \in V_1$ and $v_2 \in V_2$. Then there exists a $v_1 - v_2$ path, P in G . Let u be the first vertex in P which is not from V_1 and w be its preceeding vertex in P . Then wu is an edge between V_1 and V_2 .

Suppose every nonempty partition V_1, V_2 has an edge between them. Let $u, v \in V(G)$ and $V_1 = \{u\}$ and $V_2 = V(G) - V_1$. Then there exists an edge uw between V_1 and V_2 . If $w = v$, then we have a $u - v$ path. Suppose $w \neq v$. Consider $V_1 = \{u, w\}$ and $V_2 = V(G) - V_1$. Again there exists an edge between V_1 and V_2 . Continuing like this, we get a $u - v$ path as the vertex set of G is finite. Since every pair of vertices (u, v) are connected by a $u - v$ path, the graph G is connected. \square

Exercise 5.1.13. *Let G be a connected graph of order at least 3. Then any two longest paths in G has a common vertex.*

Proof. Let G be a connected graph of order at least 3. Let $P : u_1, u_2, \dots, u_k$, $Q : v_1, v_2, \dots, v_k$ be two longest paths in G which are disjoint. Since G is connected, there exists a $u_1 - v_1$ path, P' in G . Let u_r, v_s be vertices in P' such that $u_r - u_k$ section of P and $u_r - v_1$ section of P' are disjoint and $v_s - v_k$ section of Q and $u_1 - v_s$ section of Q are disjoint. Let P'' be the $u_r - v_s$ section of P' of length at least 1.

WLOG $u_1 - u_r$ section of P , say P_1 and $v_s - v_1$ section of Q , say Q_1 are of length at least $k/2$. Then $P_1 + P'' + Q_1$ is path of length at least $k/2 + 1 + k/2$. This is a contradiction as this path is longer than the longest paths in G . Therefore, two longest paths in G must have a common vertex. \square

Exercise 5.1.14. *The union of two distinct paths joining two vertices contains a cycle.*

Proof. Let P, Q be two distinct paths between two distinct vertices u, v in G . The vertex u belongs to both P and Q . Let w be the first vertex such that the vertex after w is different in P and Q . Let x be a vertex after w which is in both P and Q . There exists such a vertex since v belongs to both P and Q . Then $w - x$ section of P, Q are disjoint $w - x$ paths. Therefore, they form a cycle. \square

Exercise 5.1.15. *If a simple, connected graph G of order $n \geq 3$ is not complete. Then there exists three vertices u, v, w such that uv, vw are edges of G , but uw is not an edge of G .*

Proof. Let $n \geq 3$. Since G is not complete, there exists two vertices u, v which are non-adjacent in G . Since G is connected, there exists a $u - v$ path P in G , let w be the vertex preceeding v in P . If u, w are adjacent, then the proof is complete.

Suppose u, w are not adjacent. Rename w as v . Again, we have a pair of vertices u, v which are adjacent and a vertex w preceeding v on the $u - v$ path P' . The length of P' is one less than the length of P . Suppose there is no such vertex for every path of length ≥ 3 . Then a $u - v$ path of length two, say u, w, v has a such a vertex. \square

Definitions 5.1.16. A simple graph G is **highly irregular** at a vertex v if all its neighbours have different degrees. A simple graph G is highly irregular if it is highly irregular at every vertex of G .

Exercise 5.1.16. *There does not exist highly irregular, simple connected graphs of order 3 and 5.*

Proof. Let G be a simple connected graph of order 3. Then G has a vertex v with degree 2. Let u, w be the neighbours of v . Since G is connected, degree of these vertices are 1 or 2. By Euler's theorem, the sum of degrees should be even. Therefore, the degree of u, w are the same. Therefore simple connected graphs of order 3 are not highly irregular.

Let G be a simple connected graph of order 5. Suppose G is highly irregular. If G has a vertex v with degree 4, then the degree of its neighbours are 1, 2, 3 and 4. Then the degree sequence of G is 4, 4, 3, 2, 1. This is a contradiction as G has two vertices of degree 4 and one vertex of degree 1. Therefore, G does not have a vertex of degree 4.

If G has a vertex v with degree 3, then degree of its neighbours are 1, 2, 3 as G does not have a vertex of degree 0 or 4. By Euler's theorem, the vertex which is not adjacent to v must have an odd degree. Therefore the degree sequence of G is either 3, 3, 3, 2, 1 or 3, 3, 2, 1, 1. Let u be the vertex with degree 2. Then the neighbours u of must have degree different degree say, 1 and 3 which is not possible as the remaining vertex needs three neighbours.

Then G has vertices of degree 2 or 1. By Euler's theorem, the degree sequences are (2, 1, 1, 1, 1), (2, 2, 2, 1, 1) or (2, 2, 2, 2, 2). Clearly, not every vertex of degree 2 have neighbours with different degree. Therefore, no graph of order 5 is highly irregular. \square

Challenge 5. Prove that : highly irregular graphs of all orders exists, except 3, 5 and 7.

Remark. P_4 is highly irregular.³ And there exists a unique connected simple graph of order 6 with degree sequence 3, 3, 2, 2, 1, 1 which is highly irregular.

Definitions 5.1.17. The **generalised Petersen graph** $P(n, k)$ is given by

$$V(P(n, k)) = \{a_i, b_i : 0 \leq i \leq n-1\}$$

$$E(P(n, k)) = \{a_i a_{i+1}, a_i b_i, b_i b_{i+k} : 0 \leq i \leq (n-1)\}$$

where the additions $i+1, i+k$ are modulo n .

Exercise 5.1.17. *The generalised Petersen graph $P(n, k)$ is bipartite if n is even and k is odd.*

Proof. Consider the sets $X = \{a_{2j}, b_{2j-1} : 1 \leq j \leq n/2\}$ and $Y = \{a_{2j-1}, b_{2j} : 1 \leq j \leq n/2\}$. Clearly, edges of the form $a_i b_i$ has one end vertex in X and other in Y .

Suppose n is even. Then $a_1 \in Y$ and $a_n \in X$. Therefore, edge of the form a_i, a_{i+1} has one end vertex in X and other in Y .

³I thought, we should call it P_3 .

Suppose k is odd. If i is odd, then $i + k \pmod{n}$ is even. And $b_i \in Y$ and $b_{i+k} \in X$. Similarly, if i is even, then $b_i \in X$ and $b_{i+k} \in Y$. Therefore, edges of the form $b_i b_{i+k}$ has one end vertex in X and other in Y . Therefore, $P(n, k)$ is bipartite if n is odd and k is even. \square

Exercise 5.1.18. If G is simple and $\delta(G) \geq k$, then G contains a path of length at least k .

Proof. Let $P : v_0, v_1, \dots, v_r$ be a longest path in G . Then v_r is at most adjacent to v_1, v_2, \dots, v_{r-1} . Otherwise, there exists a longer path in G . Therefore, $\deg(v_r) < r$. And G has a vertex with minimum degree less than the length of its longest path. In other words, if G has a vertex with minimum degree k then it has a path which is at least k long. \square

5.1.6 Automorphisms of a simple graph

Definitions 5.1.18. An **automorphism** on a graph G is an isomorphism of G onto itself. The set $\text{aut}(G)$ is the set of all automorphisms on G .

Theorem 5.1.4. The set $\text{aut}(G)$ of all automorphisms on a simple graph G together with function composition is a group.

Proof. Let G be a simple graph. Then the identity map $\phi : G \rightarrow G$ defined by $\phi(v) = v$ is a trivial automorphism on G . Therefore the $\text{aut}(G)$ is nonempty.

We need to prove that the proposed group operation is closed. Let ϕ_1, ϕ_2 be two automorphisms on G . Let u, v be two vertices of G .

$$\begin{aligned} u, v \text{ are adjacent} &\iff \phi_2(u), \phi_2(v) \text{ are adjacent} \\ &\iff \phi_1(\phi_2(u)), \phi_1(\phi_2(v)) \text{ are adjacent} \end{aligned}$$

Clearly, $\phi_1 \circ \phi_2$ are bijective and preserves adjacency and non-adjacency. Therefore, $\phi_1 \circ \phi_2$ is an automorphism. Thus, the group operation is closed.

We know that function composition is associative. The trivial automorphism is the identity element of the group. And the inverse of each automorphism ϕ is the inverse map ϕ^{-1} . Therefore, $\langle \text{aut}(G), \circ \rangle$ is a group. \square

Theorem 5.1.5. For any simple graph $\text{aut}(G) = \text{aut}(G^c)$.

Proof. Let G be a simple graph and ϕ be an automorphism on G . Let u, v be two vertices of G .

$$\begin{aligned} u, v \text{ are non-adjacent in } G^c &\iff u, v \text{ are adjacent in } G \\ &\iff \phi(u), \phi(v) \text{ are adjacent in } G \\ &\iff \phi(u), \phi(v) \text{ are nonadjacent in } G^c \end{aligned}$$

Clearly, every automorphism ϕ on G is essentially a bijection on the vertex set of both G and G^c . The ϕ preserves adjacency/non-adjacency of G^c as well.

Every automorphism ϕ on G is an automorphism on G^c . If $\phi \in \text{aut}(G)$ then $\phi \in \text{aut}(G^c)$. Thus, $\text{aut}(G^c) \subset \text{aut}(G)$. Similarly, every automorphism on G^c is an automorphism on G as well. If $\phi \in \text{aut}(G^c)$ then $\phi \in \text{aut}(G)$. Thus, $\text{aut}(G) \subset \text{aut}(G^c)$. Therefore, $\text{aut}(G) \cong \text{aut}(G^c)$. \square

Exercise 5.1.19. $\text{aut}(K_n) \cong S_n$.

Proof. Let K_n be a complete graph with vertices v_1, v_2, \dots, v_n . Any automorphism ϕ on K_n is a bijection/permutation of n vertices. Thus, $\phi \in S_n$. \square

5.1.7 Line Graph

Definitions 5.1.19. Let G be a graph. The **line graph** of G is graph $L(G)$ with vertex set in 1 – 1 correspondence with edge set of G and two vertices of $L(G)$ are adjacent if the corresponding edges are adjacent in G .

Let G be a graph with vertex set $V(G) = \{u, v, w, x\}$ and edge set $E(G) = \{uv, vw, wx, uw\}$. Then the line graph $L(G)$ is a graph with vertex set $V(L(G)) = \{a, b, c, d\}$ with 1 – 1 correspondence $a \rightarrow uv, b \rightarrow vw, c \rightarrow wx, d \rightarrow uw$ and edge set $E(L(G)) = \{ab, ad, bc, bd, cd\}$ with adjacency at vertices v, u, w, w, w respectively.

Properties of Line graph

Suppose G has no isolated vertices.

1. G is connected if and only if $L(G)$ is connected.

Proof. Let G be a connected graph. Let e, f be two vertices of $L(G)$. Then e, f corresponds to edges of G say $e = uv$ and $f = wx$. Since G is connected, there exists a $u - v$ path in G . The edges of $u - v$ path corresponds to the vertices of an $e - f$ path in $L(G)$. Therefore, two arbitrary vertices e, f in $L(G)$ have a path between them. Thus, $L(G)$ is a connected graph.

Let G be a graph and $L(G)$ is the line graph of G . Suppose $L(G)$ is connected. Let u, v be two vertices of G . Since G has no isolated vertices. There exists some edges e, f incident of u, v . Suppose $e \neq f$. Otherwise edge uv is a $u - v$ path of length 1. Since $L(G)$ is connected, there exists a $e - f$ path in $L(G)$. The vertices of $e - f$ path corresponds to edges in a $u - v$ path in G . Therefore, two arbitrary vertices of G have a path between them. Therefore, G is connected. \square

2. If H is a subgraph of G , then $L(H)$ is a subgraph of $L(G)$.

Proof. Let H be a subgraph of G . The vertices of $L(H)$ corresponds to edges of H . Since H is a subgraph of G , these edges also belong to G . Thus, $L(G)$ have vertices corresponding to these edges as well. In other words, $V(L(H)) \subset V(L(G))$.

Two vertices in $L(H)$ are adjacent if the corresponding edges are adjacent in H . Clearly, these edges are adjacent in G as well. Therefore, the vertices are adjacent in $L(G)$ as well. That is, $E(L(H)) \subset E(L(G))$. Therefore $L(H)$ is a subgraph of $L(G)$. \square

3. The edges incident at a vertex of G give rise to a clique of $L(G)$.

Proof. Suppose k edges are incident on a vertex of G . Let e_1, e_2, \dots, e_k be the vertices of $L(G)$ corresponding to those edges in G . Every pair of vertices e_i, e_j among them are adjacent in $L(G)$ as the respective edges are adjacent in G . Thus, we have a complete subgraph of $L(G)$. Since G is simple, no other edge in G is simultaneously adjacent to all those edges. Therefore, the complete subgraph induced by $\{e_1, e_2, \dots, e_k\}$ is maximal in $L(G)$. \square

4. Let G be a simple graph. Then the degree of the vertex e of $L(G)$ corresponding to an edge $e = uv$ of G is given by

$$d_{L(G)}(e) = d_G(u) + d_G(v) - 2$$

Proof. Let $e = uv$ be an edge in G . The number of edges adjacent to e at u is $d(u) - 1$. Similarly, the number of edges adjacent to e at v is $d(v) - 1$. Therefore, there are $d(u) + d(v) - 2$ edges adjacent to e in G . Clearly, the corresponding vertex e is adjacent to $d(u) + d(v) - 2$ vertices in $L(G)$. In other words, $d_{L(G)}(e) = d_G(u) + d_G(v) - 2$. \square

5. Let G be a simple graph with degree sequence $\mathcal{D}(G) = d_1, d_2, \dots, d_n$. Then size of the line graph $L(G)$ is given by $m(L(G)) = \frac{1}{2} \sum d_i^2 - m(G)$.

Proof. By Euler's theorem,

$$\begin{aligned} m(L(G)) &= \frac{1}{2} \sum_{e \in V(L(G))} d_{L(G)}(e) \\ &= \frac{1}{2} \sum_{e \in E(G)} [d_G(u) + d_G(v) - 2] \\ &= \left[\sum_{v \in V(G)} d_G(v)^2 \right] - m(G) \end{aligned}$$

\square

Exercise 5.1.20. Line graph of the star $K_{1,n}$ is the complete graph K_n .

$$L(K_{1,n}) \cong K_n$$

Proof. Let e_1, e_2, \dots, e_n be the edges of $K_{1,n}$. Clearly, $L(K_{1,n})$ has n vertices. Since every pair of edges in $K_{1,n}$ are adjacent, the respective vertices are adjacent in $L(K_{1,n})$. Therefore, line graph $L(K_{1,n})$ is a complete graph with n vertices, say K_n . \square

Exercise 5.1.21. Line graph of the cycle C_n ($n \geq 3$) is C_n itself.

$$L(C_n) = C_n$$

Proof. Let C_n be a cycle of length $n \geq 3$. Let e_1, e_2, \dots, e_n be the edges of C_n where e_j, e_{j+1} are adjacent. Let G be the line graph of C_n . Then G has n vertices corresponding to n edges of C_n . We know that $d_{L(G)}(e) = d(u) + d(v) - 2$.

Since C_n are 2-regular, $d(u) = d(v) = 2$. Then the line graph G is 2-regular. We know that, line graph of a connected graph is connected. Since C_n is connected, its line graph G is also connected. Since G is a connected 2-regular graph, G is isomorphic to C_n itself. \square

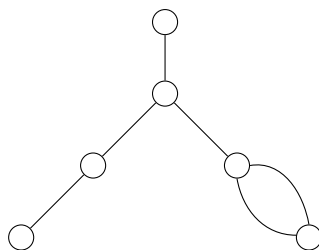
Theorem 5.1.6. *Line graph of a simple graph is a path if and only if G is a path.*

$$G \cong P_n \iff L(G) \cong P_{n-1}$$

Proof. Let G be a path of length n , say P_n . Let e_1, e_2, \dots, e_n be the edges of P_n such that each pair of edges e_j, e_{j+1} are adjacent except the pair e_n, e_1 . (We usually consider addition modulo n .) Then line graph $L(G)$ has n vertices. Since G is connected, $L(G)$ is also connected. Clearly, every vertex e_j is adjacent to e_{j+1} except e_n, e_1 . Therefore, $L(P_n)$ is a path of length $n-1$, say P_{n-1} .

Let G be a simple graph with no isolated vertices. Suppose $L(G)$ is a path of length m , say P_m . By the definition of line graph the size of G is m . Since $L(G)$ is connected, G is also connected. Since $L(G)$ has two pendent vertices, G has two pendent edges. Since the order of maximal subgraphs are at most 2, the degree of vertices of G is at most 2. Therefore, G is a path of length $m+1$, say P_{m+1} . \square

Exercise 5.1.22. Let $H = L^2(G)$ be defined as $L(L(G))$. Find $m(H)$ if G is the graph in figure ??.



Proof. (do it yourself) \square

Exercise 5.1.23. Show that $d_{L(G)}(uv) = d_G(u) + d_G(v) - 2$ is not valid if G has a loop.

Proof. (do it yourself) \square

Exercise 5.1.24. Prove that a simple connected graph G is isomorphic to its line graph if and only if it is a cycle.

Proof. We know that the line graph of a cycle is always a cycle of same length. Thus, if G is a cycle then it is isomorphic to its line graph.

Let G be a connected graph. Suppose G is isomorphic to its line graph. Then order and size of G are the same. (yet to complete) \square

Exercise 5.1.25. Disprove : If the graph H is a spanning subgraph of G , then $L(H)$ is a spanning subgraph of G .

Proof. (do it yourself) \square

Theorem 5.1.7. *If the simple graphs G_1, G_2 are isomorphic, then $L(G_1), L(G_2)$ are isomorphic.*

Proof. Let two simple graphs G_1, G_2 are isomorphic. Then there exists an isomorphic $\phi : V(G_1) \rightarrow V(G_2)$ which gives rise to an edge set bijection $\theta : E(G_1) \rightarrow E(G_2)$ such that $\phi(u)\phi(v) = \theta(uv)$.

Let e be an edge of G_1 . Then e corresponds to a vertex e of $L(G_1)$. By graph isomorphism, $\theta(e)$ is an edge of G_2 and $\theta(e)$ corresponds to a vertex $\theta(e)$ of $L(G_2)$. Consider the map $\phi' : V(L(G_1)) \rightarrow V(L(G_2))$ defined by $\phi'(e) = \theta(e)$. Clearly, ϕ' is a bijection.

Suppose two edges e, e' are adjacent in G_1 . Then by isomorphism $\theta(e), \theta(e')$ are adjacent in G_2 . Therefore, if two vertices e, e' are adjacent in $L(G_1)$, then the respective vertices $\phi'(e), \phi'(e')$ are adjacent in $L(G_2)$. Therefore ϕ' is a graph isomorphism between $L(G_1)$ and $L(G_2)$. Clearly, $L(G_1) \cong L(G_2)$. \square

Remark. Suppose G_1, G_2 be simple, connected graphs. Then $L(G_1) \cong L(G_2)$ does not imply that $G_1 \cong G_2$.

Proof. Suppose $G_1 = K_{1,3}$ and $G_2 = K_3$. Then $L(G_1) \cong L(G_2) \cong K_3$ and $G_1 \not\cong G_2$. \square

5.1.8 Graph Operations

Definitions 5.1.20. Let G_1, G_2 be two simple graph. Their **union** $G_1 \cup G_2$ is given by

$$V(G_1 \cup G_2) = V(G_1) \cup V(G_2) \quad E(G_1 \cup G_2) = E(G_1) \cup E(G_2)$$

If $V(G_1) \cap V(G_2) = \phi$, then the graph $G_1 \cup G_2$ is their **sum** $G_1 + G_2$.

Remark. The finite union of graph is associative.

Definitions 5.1.21. Let G_1, G_2 be two simple graph with $V(G_1) \cap V(G_2) \neq \phi$. Their **intersection** $G_1 \cap G_2$ is given by

$$V(G_1 \cap G_2) = V(G_1) \cap V(G_2) \quad E(G_1 \cap G_2) = E(G_1) \cap E(G_2)$$

Definitions 5.1.22. Let G_1, G_2 be two vertex-disjoint graphs. Then the **join** $G_1 \vee G_2$ is the supergraph of $G_1 + G_2$ in which each vertex of G_1 is adjacent to every vertex of G_2 .

If $G_1 = K_1$ and $G_2 = C_n$, then $G_1 \vee G_2 \cong W_n$ a **wheel** of order $n + 1$.

Properties of Graph Operations

1. $n(G_1 \cup G_2) = n(G_1) + n(G_2) - n(G_1 \cap G_2)$ and
 $m(G_1 \cup G_2) = m(G_1) + m(G_2) - m(G_1 \cap G_2)$.

Proof. do it yourself \square

2. $n(G_1 + G_2) = n(G_1) + n(G_2)$ and
 $m(G_1 + G_2) = m(G_1) + m(G_2)$.

Proof. do it yourself □

3. $n(G_1 \vee G_2) = n(G_1) + n(G_2)$ and
 $m(G_1 \vee G_2) = m(G_1) + m(G_2) + n(G_1)n(G_2)$.

Proof. do it yourself □

5.1.9 Graph Products

Definitions 5.1.23. The **general graph product** $G_1 * G_2$ is defined as a graph with vertex set $V(G_1 * G_2) = V(G_1) \times V(G_2)$ and the adjacency/non-adjacency among a pair of vertices is given by the structure table S ,

$$S = \begin{bmatrix} \circ & \circ & \circ \\ \circ & = & \circ \\ \circ & \circ & \circ \end{bmatrix}$$

where \circ 's are either N or E . The letter 'E' means Edge and 'N' means No edge.

The first, second and third rows of S corresponds to adjacency, equality and non-adjacency of respective vertices in G_1 . Similarly, first, second and third columns of S corresponds to adjacency, equality and non-adjacency of respective vertices in G_2 .

For example, G_1, G_2 be two simple graph given in diagram 5.4. Let $G_1 * G_2$ be a graph product with the following strcture table S given in table 5.1.

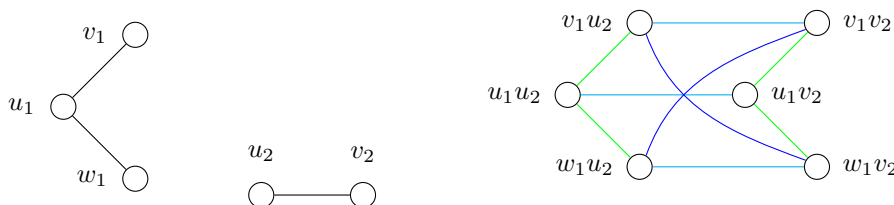


Figure 5.4: A graph product $G_1 * G_2$

For example, v_1u_2, w_1v_2 are two vertices of $G_1 * G_2$. We know that v_1, w_1 are non-adjacent in G_1 . Thus, we need to look at the third row of the structure table S . Again, u_2, v_2 are adjacent in G_2 . Thus, we need to look at the first column of the structure table S . Clearly, $S_{3,1}$ which is E . Therefore, they are adjacent in $G_1 * G_2$. Refer table 5.1 to find element of S corresponding to each pair of vertices in $G_1 * G_2$.

Definitions 5.1.24. The **cartesian product** $G_1 \square G_2$ is the graph product with structure table S given by

$$S = \begin{bmatrix} N & E & N \\ E & = & N \\ N & N & N \end{bmatrix}$$

$$S = \begin{bmatrix} N & E & N \\ E & = & N \\ E & N & N \end{bmatrix}$$

	u_1u_2	v_1u_2	w_1u_2	u_1v_2	v_1v_2	w_1v_2
u_1u_2	$S_{2,2}$	$S_{1,2}$	$S_{1,2}$	$S_{2,1}$	$S_{1,1}$	$S_{1,1}$
v_1u_2	$S_{1,2}$	$S_{2,2}$	$S_{3,2}$	$S_{1,1}$	$S_{2,1}$	$S_{3,1}$
w_1u_2	$S_{1,2}$	$S_{3,2}$	$S_{2,2}$	$S_{1,1}$	$S_{3,1}$	$S_{2,1}$
u_1v_2	$S_{2,1}$	$S_{1,1}$	$S_{1,1}$	$S_{2,2}$	$S_{1,2}$	$S_{1,2}$
v_1v_2	$S_{1,1}$	$S_{2,1}$	$S_{3,1}$	$S_{1,2}$	$S_{2,2}$	$S_{3,2}$
w_1v_2	$S_{1,1}$	$S_{3,1}$	$S_{2,1}$	$S_{1,2}$	$S_{3,2}$	$S_{2,2}$

Table 5.1: Adjacency/Non-adjacency Table for $G_1 \star G_2$

Definitions 5.1.25. The **direct product** (tensor/Kronecker product) $G_1 \times G_2$ is the graph product with structure table S given by

$$S = \begin{bmatrix} E & N & N \\ N & = & N \\ N & N & N \end{bmatrix}$$

Definitions 5.1.26. The **composition product** (wreath/lexicographic product) $G_1[G_2]$ is the graph product with structure table S given by

$$S = \begin{bmatrix} E & E & E \\ E & = & N \\ N & N & N \end{bmatrix}$$

Definitions 5.1.27. The **strong product** (normal product) $G_1 \boxtimes G_2$ is the graph product with structure table S given by

$$S = \begin{bmatrix} E & E & N \\ E & = & N \\ N & N & N \end{bmatrix}$$

Remark. From structure table S , it is evident that all the above graph products are associative and all except the composition product are commutative.

Challenge 6. Can you find a non-associative graph product ?

Exercise 5.1.26. Show that $G_1[G_2] \not\cong G_2[G_1]$

Proof. $K_2[K_{1,3}] \not\cong K_{1,3}[K_2]$ □

Exercise 5.1.27. Prove that $G_1 \square G_2 \cong G_2 \cong G_1$.

Proof. Since the structure table S of the cartesian graph product is symmetric the graph product is commutative. □

Exercise 5.1.28. Prove that $G_1 \boxtimes G_2 \cong G_2 \boxtimes G_1$

Proof. Since the structure table S of the strong product is symmetric the graph product is commutative. □

Definitions 5.1.28. Let (x, y) be a vertex of $G_1 \square G_2$. The **G_2 fiber at \mathbf{x}** in $G_1 \square G_2$ denoted by $(G_2)_x$ is the subgraph of $G_1 \square G_2$ induced by the set of all vertices $\{(x, z) : z \in V(G_2)\}$. Similarly, the **G_1 -fiber at \mathbf{y}** , $(G_1)_y$ is the subgraph of $G_1 \square G_2$ induced by $\{(z, y) : z \in V(G_1)\}$.

Exercise 5.1.29. *Properties of graph products*

1. The order of a general graph product is the product of their orders.

$$\begin{aligned} n(G_1 \square G_2) &= n(G_1[G_2]) = n(G_2[G_1]) = \\ n(G_1 \times G_2) &= n(G_1 \boxtimes G_2) = n(G_1)n(G_2) \end{aligned}$$

Proof. The vertex set of a general graph product is the cartesian product of the vertex sets, $G_1 * G_2$ is $V(G_1) \times V(G_2)$. Therefore, order of a general graph product is $n(G_1)n(G_2)$. \square

2. $m(G_1 \square G_2) = n(G_1)m(G_2) + m(G_1)n(G_2)$.

Proof. The structure table of cartesian graph product has entry E at $S_{1,2}$ and $S_{2,1}$. Thus the edge conditions for cartesian graph product $G_1 \square G_2$ are (a) $u_1 = v_1$ and u_2, v_2 are adjacent or (b) $u_2 = v_2$ and u_1, v_1 are adjacent.

Case 1 : Let xy be an edge of G_2 and v be a vertex of G_1 . Then $(v, x), (v, y)$ are adjacent in $G_1 \square G_2$. Every edge of type 1 is of this form. Therefore, there are $n(G_1)m(G_2)$ edges of type 1 in $G_1 \square G_2$.

Case 2 : Let xy be an edge of G_1 and v be a vertex of G_2 . Then $(x, v), (y, v)$ are adjacent in $G_1 \square G_2$. Every edge of type 2 is of this form. Therefore, there are $m(G_1)n(G_2)$ edges of type 2 in $G_1 \square G_2$. Therefore, $m(G_1 \square G_2) = n(G_1)m(G_2) + m(G_1)n(G_2)$. \square

3. $m(G_1[G_2]) = n(G_1)m(G_2) + n(G_2)^2m(G_1)$.

Proof. The structure table for composition product has entry E at first row of S and $S_{2,1}$. Thus the condition for adjacency of two vertices $(u_1, u_2), (v_1, v_2)$ in $G_1[G_2]$ are (a) u_1, v_1 are adjacent or (b) $u_1 = v_1$ and u_2, v_2 are adjacent.

Case 1 : Let xy be an edge of G_1 and u, v be two vertices of G_2 . Then an edge of type 1 is of the form $(x, u)(y, v)$. Thus, there are $m(G_1)n(G_2)^2$ edges of type 1 in $G_1[G_2]$.

Case 2 : Let xy be an edge in G_2 and v be a vertex in G_1 . Then an edge of type 2 is of the form $(v, x)(v, y)$. Thus, there are $m(G_2)n(G_1)$ edges of type 2 in $G_1[G_2]$.

Therefore $m(G_1[G_2]) = m(G_1)n(G_2)^2 + m(G_2)n(G_1)$. \square

4. $m(G_1 \times G_2) = 2m(G_1)m(G_2)$.

Proof. The only E entry in the structure table of direct product is $S_{1,1}$. Thus the condition for adjacency between $(u_1, u_2), (v_1, v_2)$ in $G_1 \times G_2$ is u_1, v_1 are adjacent and u_2, v_2 are adjacent.

Let uv, xy be edges in G_1, G_2 respectively. Then an edge of direct product is of the form $(u, x)(v, y)$ or $(u, y)(v, x)$. There are $m(G_1)m(G_2)$ edges of both kind. Therefore, $m(G_1 \times G_2) = 2m(G_1)m(G_2)$. \square

$$5. \quad m(G_1 \boxtimes G_2) = n(G_1)m(G_2) + m(G_1)n(G_2) + 2m(G_1)m(G_2).$$

Proof. We know that the strong product is the disjoint union of the cartesian and direct graph products, $G_1 \boxtimes G_2 = G_1 \square G_2 \cup G_1 \times G_2$. By the properties of sum of two graph, $E(G_1 \boxtimes G_2) = E(G_1 \square G_2) \cup E(G_1 \times G_2)$. Therefore, $m(G_1 \boxtimes G_2) = m(G_1 \square G_2) + m(G_1 \times G_2)$. \square

Exercise 5.1.30. Prove that $L(K_{m,n}) \cong K_m \square K_n$.

Proof. Let u_j, v_k be labels for the vertices of K_m, K_n respectively. Abusing the language we may write that a vertex of $K_m \square K_n$ is of the form $u_j v_k$. By the definition of cartesian graph product, $u_j v_s, u_k v_t$ are adjacent if and only if u_j, u_k are adjacent and v_s, v_t are adjacent.

Consider the line graph $L(K_{m,n})$ of the bipartite graph $K_{m,n}$ with partite set X, Y . Let x_j, y_k be labels for vertices in the partite sets X, Y respectively. The edge of $K_{m,n}$ are of the form $x_j y_k$. Two vertices $x_j y_s, x_k y_t$ are adjacent in $L(K_{m,n})$ if the respective edges have a common vertex. In other words, either $x_j = x_k$ or $y_s = y_t$.

Define $\psi : X \rightarrow V(K_m)$ where $\psi(x_j) = u_j$ and define $\xi : Y \rightarrow V(K_n)$ where $\xi(y_k) = v_k$. Define $\phi : V(L(K_{m,n})) \rightarrow V(K_m \square K_n)$ by $\phi(x_j y_k) = u_j v_k$. Clearly, $x_j y_s, x_k y_t$ are adjacent if and only if $\phi(x_j y_s) = u_j v_s$ and $\phi(x_k y_t) = u_k v_t$ are adjacent. Therefore, $L(K_{m,n}) \cong K_m \square K_n$. \square

Definitions 5.1.29. The **kth power** G^k of G has $V(G^k) = V(G)$ where two vertices u, v are adjacent in G^k whenever $d_G(u, v) \leq k$.

Exercise 5.1.31. Show that if G is a connected graph with at least two edges, then each edge of G^2 belongs to a triangle.

Proof. Let $e = uv$ be an edge of G^2 . Then there are two possibilities, either $d_G(u, v) = 1$ or $d_G(u, v) = 2$.

Case 1 : Suppose $d_G(u, v) = 1$. Since G is connected, without loss of generality there exists an edge adjacent to uv at v , say vw . Then $d_G(u, w) \leq 2$ and $uw \in E(G^2)$. Clearly, the edge uv belongs to the triangle $\triangle uvw$ in G^2 .

Case 2 : Suppose $d_G(u, v) = 2$. Then there exists $w \in V(G)$ such that $uw, vw \in E(G)$. Clearly, the edge uv belongs to the triangle $\triangle u w v$ in G^2 . \square

Exercise 5.1.32. If $d_G(u, v) = p$, determine $d_{G^k}(u, v)$.

Proof. Suppose p is finite. Otherwise $d_G(u, v) = \infty$ and $d_{G^k}(u, v) = \infty$. Since p, k are natural numbers, there exists unique integers q, r such that $p = qk + r$, $q \geq 0$ and $0 \leq r < k$. Then the shortest $u-v$ path contains vertices v_1, v_2, \dots, v_q such that $d_G(u, v_1) = k$ and $d_G(v_j, v_{j+1}) = k$ and $d_G(v_{q+1}, v) = r$. Then u, v_1 are adjacent in G^k . Similarly, v_j, v_{j+1} are adjacent in G^k .

Suppose $r = 0$. Then $v_q = v$ and $u, v_1, v_2, \dots, v_{q-1}, v$ is the shortest $u-v$ path in G^k . Thus, $d_{G^k}(u, v) = q$. Suppose $r > 0$. Then $d_G(v_q, v) = r < k$. Thus, v_q, v are adjacent in G^k and $u, v_1, v_2, \dots, v_q, v$ is the shortest $u-v$ path in G^k . Therefore, $d_{G^k}(u, v) = q + 1$. \square

5.1.12 Miscellaneous Exercises

1. Show that the sequences (a) $(7, 6, 5, 4, 3, 3, 2)$ and (b) $(6, 6, 5, 4, 3, 3, 1)$ are not graphical
2. Give an example of a degree sequence that is realizable as the degree sequence by only a disconnected graph.

Solution. $(1, 1, 1, 1)$ or $(1, 1, 0)$. □

3. Show that for a simple bipartite graph, $m \leq n^2/4$.

Solution. Let G be a bipartite graph. The size of a bipartite graph is maximum when it is a complete bipartite graph with cardinality of partite set X, Y almost equal.

Suppose order of G is even, $n = 2k$. Then the size of the bipartite graph is maximum when $p = q = k$. And size

$$m \leq pq = k^2 = \frac{n^2}{4}$$

Suppose order of G is odd, $n = 2k + 1$. The size of the bipartite graph is maximum when $p = k$ and $q = k + 1$. And size

$$m \leq pq = k(k + 1) \leq \left(k + \frac{1}{2}\right)^2 = \frac{n^2}{4}$$

□

4. Show that $\delta \leq \frac{2m}{n} \leq \Delta$.

Solution. By Euler's theorem, the sum of degrees of a graph is twice its size. Thus $2m/n$ is the average of the degrees of vertices. We know that, *minimum* \leq *average* \leq *maximum* is true for any finite sequence. □

5. For every $n \geq 3$, construct a 3-regular simple graph on $2n$ vertices containing no triangles.

Solution. The generalised Petersen graphs $P(n, k)$ where $0 \leq k \leq n$ and $\gcd(n, k) = 1$ are 3-regular and does not contain any cycles. □

6. If a bipartite graph $G(X, Y)$ is regular, show that $|X| = |Y|$.

Solution. Since G is bipartite the sum of degree of vertices both partite sets are equal, as every edge contributes one to both the sum. If G is regular, then □

7. *Solution.* □

8. *Solution.* □

- 9. *Solution.* ☐
- 10. *Solution.* ☐
- 11. *Solution.* ☐
- 12. *Solution.* ☐
- 13. *Solution.* ☐
- 14. *Solution.* ☐
- 15. *Solution.* ☐
- 16. *Solution.* ☐
- 17. *Solution.* ☐
- 18. *Solution.* ☐
- 19. *Solution.* ☐
- 20. *Solution.* ☐
- 21. *Solution.* ☐

Semester II

Semester 2

Subject 6

ME010201 Advanced Abstract Algebra

Module 1 Field Extensions

6.1 Extension Fields §29

Context

- Let R be a commutative ring with unity. If M is a maximal ideal in R , then R/M is a field. [Fraleigh, 2013, §27.9]
- Let F be a field. Every polynomial in $F[x]$ has a unique factorisation into irreducible polynomials except for order and unit. cite[§27.27]fraleigh
- If α is a zero of $f(x) \in F[x]$, then $f(\alpha) = 0$. cite[§22.10]fraleigh
- If $p(x)$ is irreducible over field F , then the principal ideal generated by $p(x)$, denoted by $\langle p(x) \rangle$ is a maximal ideal in $F[x]$. [Fraleigh, 2013, §27.25]
- Let R be a ring with unity. And N be an ideal of R containing a unit. Then $N = R$. [Fraleigh, 2013, §27.5]

Basic Goal Let F be a field and $f(x) \in F[x]$. Find a field E such that F is a subfield of E and there exists a zero of $f(x)$ in E ?

Extension Field Let F be a field. Field E is an extension field of F if F is a subfield of E .

Example : $\mathbb{Q} \leq \mathbb{R} \leq \mathbb{C}$

Tower of Fields A diagrammatic representation emphasising the hierarchy of field extensions in which extension fields appears above their subfields.

Theorem 6.1.1 (Kronecker). *Let F be field. And $f(x)$ be a non-constant polynomial in $F[x]$. Then there exists an extension field E of F and an $\alpha \in E$ such that $f(\alpha) = 0$.*

Proof. Let $f(x) \in F[x]$. Then $f(x)$ has a unique factorisation into irreducible polynomials in $F[x]$ (except for order and unit). Let $p(x)$ be an irreducible factor of $f(x)$. If $f(x)$ is irreducible over F , then $f(x) = cp(x)$. It is enough to construct an extension field E containing both F and α such that $p(\alpha) = 0$.

If $p(x)$ is irreducible over F , then $\langle p(x) \rangle$ is maximal ideal in $F[x]$. Therefore, $F[x]/\langle p(x) \rangle$ is a field, say E .

Consider the function $\psi : F \rightarrow F[x]/\langle p(x) \rangle$ defined by $\psi(a) = a + \langle p(x) \rangle$. We claim that $\psi : F \rightarrow \psi[F]$ is a field isomorphism. ψ is a canonical homomorphism with trivial kernel. Thus, ψ is one-to-one.

Let $a, b \in F$. And suppose $\psi(a) = \psi(b)$. It is enough to prove that $a = b$. By the definition of ψ , we have $a + \langle p(x) \rangle = b + \langle p(x) \rangle \implies a - b \in \langle p(x) \rangle$. Suppose $a \neq b$. Then $a - b \neq 0$ and $\deg(a - b) = 0$. Then, $\langle p(x) \rangle = F[x]$ which is a contradiction since $\langle p(x) \rangle$ is maximal ideal. Therefore, ψ is one-to-one.

We have $p(x)$ is a factor of $f(x)$. Thus $p(\alpha) = 0 \implies f(\alpha) = 0$. Thus, it remains to prove that there exists $\alpha \in F[x]/\langle p(x) \rangle$ such that $p(\alpha) = 0$.

Let $p(x) = a_0 + a_1x + a_2x^2 + \cdots + a_nx^n$. Consider $\alpha = x + \langle p(x) \rangle$. Then $p(\alpha) = \phi_\alpha(p)$. Thus, $p(\alpha) = a_0 + a_1(x + \langle p(x) \rangle) + \cdots + a_n(x + \langle p(x) \rangle)^n$. Thus, $p(\alpha) = (a_0 + a_1x + a_nx^n) + \langle p(x) \rangle = p(x) + \langle p(x) \rangle = \langle p(x) \rangle = 0$. Therefore, $p(\alpha) = 0$ and $f(\alpha) = 0$. \square

algebraic over F Let $F \leq E$. An element $\alpha \in E$ is algebraic over a field F , if there exists $f(x) \in F[x]$ such that $f(\alpha) = 0$.

transcendental over F Let $F \leq E$. An element $\alpha \in E$ is transcendental over the field F , if it is not algebraic over F .

algebraic number We have, $\mathbb{Q} \leq \mathbb{C}$. A complex number $\alpha \in \mathbb{C}$ is algebraic if it is algebraic over \mathbb{Q} . Example : $2, \sqrt{2}, i$

transcendental number A complex number $\alpha \in \mathbb{C}$ is transcendental if it is not an algebraic number. Example : π, e (proof excluded)

Note 1 : A polynomial $f(x) \in F[x]$ is reducible/irreducible depending upon the choice of the field F . For example : $x^2 - 2$ is irreducible over \mathbb{Q} , but is reducible over \mathbb{R} .

Note 2 : An element $\alpha \in E$ is algebraic/transcendental depending on the choice of the field F . For example : $\sqrt{2} \in \mathbb{C}$ is algebraic over \mathbb{Q} since $x^2 - 2 \in \mathbb{Q}[x]$.

Theorem 6.1.2. Let E be an extension field of F , and $\alpha \in E$. Let function $\phi_\alpha : F[x] \rightarrow E$ be an evaluation homomorphism. Then α is transcendental if and only if ϕ_α is one-to-one.

Proof. An element $\alpha \in E$ is transcendental if and only if $f(\alpha) \neq 0$ for any nonzero $f(x) \in F[x]$ where $f(\alpha) = \phi_\alpha(f)$. Thus, kernel of ϕ_α is trivial. That is, $\ker(\phi_\alpha) = \{0\}$. Therefore, ϕ_α is one-to-one. \square

Theorem 6.1.3. *Let E be an extension field of F and $\alpha \in E$ be algebraic over F . Then there exists a unique irreducible polynomial $p(x)$ with minimum degree in $F[x]$ and $p(\alpha) = 0$. If there exists a nonzero polynomial $f(x) \in F[x]$ with $f(\alpha) = 0$, then $p(x)$ divides $f(x)$.*

Proof. Consider evaluation homomorphism $\phi_\alpha : F[x] \rightarrow E$ defined by $\phi_\alpha(f) = f(\alpha)$. Then $\ker(\phi)$ is an ideal in $F[x]$. Since every ideal in $F[x]$ is principal, there exists $p(x) \in F[x]$ such that $\langle p(x) \rangle = \ker(\phi)$. And $f(\alpha) = 0 \implies f \in \ker(\phi) = \langle p(x) \rangle$. Thus, $p(x)$ divides $f(x)$.

Suppose $p(x) = r(x)s(x)$. Then $p(\alpha) = r(\alpha)s(\alpha) = 0$. However, E is a field and has no zero divisors. Thus, there exists a polynomial of lesser degree in $\langle p(x) \rangle$ which is a contradiction. \square

monic polynomial A polynomial which has 1 as the coefficient of highest power of x . For example : $x^3 - 3x \in \mathbb{Q}[x]$

$\text{irr}(\alpha, F)$ The unique monic, irreducible polynomial $p(x) \in F[x]$ such that $p(\alpha) = 0$. For example, $\text{irr}(\sqrt{3}, \mathbb{Q}) = x^2 - 3$. And $\sqrt{3}$ is the **minimal** polynomial for $\sqrt{3}$ over \mathbb{Q}

$\text{deg}(\alpha, F)$ The degree of the unique monic, irreducible polynomial $p(x) \in F[x]$ such that $p(\alpha) = 0$. For example, $\text{deg}(\sqrt{3}, \mathbb{Q}) = 2$.

simple extension An extension field E of field F is a simple extension if $E = F(\alpha)$ for some $\alpha \in E$.

Theorem 6.1.4. *Let E be a simple extension $F(\alpha)$ of field F where $\alpha \in E$ is algebraic over F . Let $\text{deg}(\alpha, F) = n \geq 1$. Then any element $\beta \in E$ can be uniquely expressed in the form $\beta = b_0 + b_1\alpha + \cdots + b_{n-1}\alpha^{n-1}$ where $b_k \in F$.*

Proof. Consider the evaluation homomorphism $\phi_\alpha : F[x] \rightarrow E$ defined by $\phi_\alpha(f) = f(\alpha)$. Then $\phi_\alpha[F[x]] = F(\alpha)$.

Let $\text{irr}(\alpha, F) = p(x) = x^n + a_{n-1}x^{n-1} + \cdots + a_1x + a_0$. Then, $p(\alpha) = 0$. Thus,

$$\alpha^n = -a_{n-1}\alpha^{n-1} - a_{n-2}\alpha^{n-2} - \cdots - a_1\alpha - a_0 \quad (6.1)$$

Clearly, any higher power of α can be eliminated from $f(\alpha)$ as shown below,

$$\begin{aligned} \alpha^{n+1} &= \alpha\alpha^n = -a_{n-1}\alpha^n - a_{n-2}\alpha^{n-1} - \cdots - a_1\alpha^2 - a_0\alpha \\ &= a_{n-1}(a_{n-1}\alpha^{n-1} + a_{n-2}\alpha^{n-2} + \cdots + a_1\alpha + a_0) \\ &\quad - a_{n-2}\alpha^{n-1} - a_{n-3}\alpha^{n-2} - a_1\alpha^2 - a_0\alpha \end{aligned}$$

Thus, in the representation of the element in $F(\alpha)$, the maximum degree of α is $\text{deg}(\alpha, F) - 1$. Therefore, $\beta \in F(\alpha) \implies \beta = b_0 + b_1\alpha + b_2\alpha^2 + \cdots + b_{n-1}\alpha^{n-1}$.

We can also show that this representation is unique for any $\beta \in F(\alpha)$. Suppose $\beta = b'_0 + b'_1\alpha + b'_2\alpha^2 + \cdots + b'_{n-1}\alpha^{n-1}$. Then $0 = (b_0 - b'_0) + (b_1 - b'_1)\alpha + \cdots + (b_{n-1} - b'_{n-1})\alpha^{n-1} = \phi_\alpha(g)$ where $g(x) = (b_0 - b'_0) + (b_1 - b'_1)x + \cdots + (b_{n-1} - b'_{n-1})x^{n-1}$. Clearly, degree of $g(x)$ is $\text{deg}(\alpha, F) - 1$ which is less than the minimum degree for a non-zero irreducible polynomial for α over F . Thus $g(x) = 0$. In other words, $b_j = b'_j, \forall j$ and representation for $\beta \in F(\alpha)$ is unique. \square

6.1.1 Exercises §29

Irreducibility Conditions for Polynomials

1. Irreducibility over a finite field
 $x^2 + 1$ is irreducible in \mathbb{Z}_3 since for every $x \in \{0, 1, 2\}$, $x^2 + 1 \neq 0$.
2. Irreducibility in rational field : Eisentein's Criteria (§23.15)
 Consider $f(x) = x^3 + 60x^2 + 30x + 12$ since for $p = 3$, $f(x)$ satisfies Eisentein's Criteria. And thus is $f(x)$ irreducible over \mathbb{Q} .
 Note that for $p = 2, 5, 7, \dots$ the Eisentein's Criteria is not satisfied.

Algebraic over a Field

1. $\sqrt{2} + \sqrt{3}$ is algebraic over \mathbb{Q} [Fraleigh, 2013, Exercise 29.2]

$$\begin{aligned}
 \alpha &= \sqrt{2} + \sqrt{3} \\
 \implies \alpha^2 &= 2 + 2\sqrt{6} + 3 = 5 + 2\sqrt{6} \\
 \implies \alpha^2 - 5 &= 2\sqrt{6} \\
 \implies (\alpha^2 - 5)^2 &= 24 \\
 \implies \alpha^4 - 10\alpha^2 + 1 &= 0 \\
 \implies \phi_\alpha(x^4 - 10x^2 + 1) &= 0
 \end{aligned}$$

We have, $\text{irr}(\sqrt{2} + \sqrt{3}, \mathbb{Q}) = x^4 - 10x^2 + 1$ and $\deg(\sqrt{2} + \sqrt{3}, \mathbb{Q}) = 4$.

2. π, e are transcendental numbers. (proof excluded)
 However, π is algebraic over $\mathbb{Q}(\pi)$ since $x - \pi \in \mathbb{Q}(\pi)[x]$.
3. Consider $\alpha = \pi^2$ and $F = \mathbb{Q}(\pi^3)$. [Fraleigh, 2013, Exercise 29.16]

$$\begin{aligned}
 \alpha^3 &= \pi^6 = (\pi^3)^2 \\
 \alpha^3 - (\pi^3)^2 &= 0 \\
 \implies \phi_\alpha(x^3 - (\pi^3)^2) &= 0
 \end{aligned}$$

Thus, $\text{irr}(\pi^2, \mathbb{Q}(\pi^3)) = x^3 - (\pi^3)^2$ and $\deg(\pi^2, \mathbb{Q}(\pi^3)) = 3$.
 Note that $x^3 - (\pi^3)^2 \in \mathbb{Q}(\pi^3)[x]$ since $\pi^3 \in \mathbb{Q}(\pi^3) \implies -(\pi^3)^2 \in \mathbb{Q}(\pi^3)$.
 However, $x^3 - (\pi^3)^2 \notin \mathbb{Q}[x]$.

Factorisation over Extended Field

1. Factorisation over Finite Extension of Finite Field [Fraleigh, 2013, Exercise 29.25]
 Let α be a zero of $f(x) = x^3 + x^2 + 1 \in \mathbb{Z}_2[x]$. Clearly, $f(x)$ is irreducible over \mathbb{Z}_2 . And $x - \alpha$ is a factor of $f(x)$ in $\mathbb{Z}_2(\alpha)$. We have, $f(x) = (x - \alpha)g(x)$

$$\text{By long division, } g(x) = \frac{x^3 + x^2 + 1}{x - \alpha} = x^2 + (1 + \alpha)x + (\alpha + \alpha^2)$$

Therefore, $x^3 + x^2 + 1 = (x - \alpha)[x^2 + (1 + \alpha)x + \alpha(1 + \alpha)]$.

The elements of $\mathbb{Z}_2(\alpha)$ are of the form $a_0 + a_1\alpha + a_2\alpha^2$ where $a_0, a_1, a_2 \in \{0, 1\}$. Also we have, α is a zero of $x^3 + x^2 + 1$. Thus, $\alpha^3 = \alpha^2 + 1$.

In order to find a zero of $g(x)$ it is sufficient to evaluate $g(x)$ for all the eight elements $0, 1, \alpha, (1 + \alpha), \alpha^2, (1 + \alpha^2), (\alpha + \alpha^2), (1 + \alpha + \alpha^2) \in \mathbb{Z}_2(\alpha)$.

Clearly, $g(1) = \alpha^2 + 2\alpha + 2 = \alpha^2$. And $g(\alpha) = \alpha^2 + 2\alpha(1 + \alpha) = \alpha^2$. However, $g(\alpha^2) = \alpha^4 + \alpha^3 + 2\alpha^2 + \alpha = \alpha(\alpha^2 + 1) + (\alpha^2 + 1) + 0\alpha^2 + \alpha = \alpha^3 + \alpha^2 + 2\alpha + 1 = \alpha^3 + \alpha^2 + 1 = 0$. Thus, α^2 is a zero of $g(x)$ and $g(x) = (x - \alpha^2)h(x)$.

$$\text{By long division, } h(x) = \frac{x^2 + (1 + \alpha)x + (\alpha + \alpha^2)}{x - \alpha^2} = x + (1 + \alpha + \alpha^2)$$

Therefore, we have the following linear factorisation for $f(x)$,
 $f(x) = (x - \alpha)(x - \alpha^2)(x - 1 - \alpha - \alpha^2) = (x + \alpha)(x + \alpha^2)(x + 1 + \alpha + \alpha^2)$
 since $-\alpha = 0 - \alpha = 2\alpha - \alpha = \alpha$ in \mathbb{Z}_2 .

Note : Students should be able to perform long division of polynomials over extended fields.

6.2 Algebraic Extensions §31

$(G : H)$ is the number of H -left cosets in G .

algebraic extension A extension field E of a field F is algebraic if every element in E is algebraic over F .

For example, \mathbb{C} is algebraic over \mathbb{R} . But, \mathbb{R} is not algebraic over \mathbb{Q} .

finite extension A extension field E of field F is a finite extension if E is a finite dimensional vector space over F . And $[E : F]$ is the **dimension of the vector space** E over F . Again, $[E : F]$ is the **degree of the finite extension** E over F .

For example, \mathbb{C} is a finite extension of degree 2 over \mathbb{R} , $[\mathbb{C} : \mathbb{R}] = 2$. But, \mathbb{R} is not a finite extension of \mathbb{Q} and $[\mathbb{R} : \mathbb{Q}]$ is infinite.

Theorem 6.2.1. *Every finite field extensions is an algebraic extension.*

Proof. Let E be a finite extension of degree n over F . Then $[E : F] = n$. Suppose $\alpha \in E$. Clearly, $\{1, \alpha, \alpha^2, \dots, \alpha^n\}$ is set of $n + 1$ vectors from the vector space E over F . We know that in a vector space of dimension n , any set having $n + 1$ vector is linearly dependant. In other words, there exists scalars $c_0, c_1, \dots, c_n \in F$ (not all zero) such that $c_0 + c_1\alpha + c_2\alpha^2 + \dots + c_n\alpha^n = 0$.

Clearly, $f(x) = c_0 + c_1x + c_2x^2 + \dots + c_nx^n$ is a polynomial in $F[x]$ such that $\phi_\alpha(f) = f(\alpha) = 0$. Since $\alpha \in E$ is arbitrary, every element in E is algebraic over F . \square

Theorem 6.2.2. *If E is a finite extension of F and K is a finite extension of E . Then K is a finite extension of F . And $[K : F] = [K : E][E : F]$.*

Proof. Let $[E : F] = n$ and $[K : E] = m$. Let $\{\alpha_1, \alpha_2, \dots, \alpha_n\}$ be basis for vector space $E(F)$ and let $\{\beta_1, \beta_2, \dots, \beta_m\}$ be basis for vector space $K(E)$. We claim that $\{\alpha_i \beta_j : 1 \leq i \leq n, 1 \leq j \leq m\}$ is a basis for the vector space $K(F)$.

Let $\gamma \in K$. Then we have $b_1, b_2, \dots, b_m \in E$ such that $\gamma = b_1 \beta_1 + b_2 \beta_2 + \dots + b_m \beta_m$. Again, for each $b_j \in E$, we have $a_{ij} \in F$ such that $b_j = a_{1j} \alpha_1 + a_{2j} \alpha_2 + \dots + a_{nj} \alpha_n$. Therefore,

$$\gamma = \sum_{i=1}^n \sum_{j=1}^m a_{ij} \alpha_i \beta_j$$

That is, $\{\alpha_i \beta_j : i = 1, 2, \dots, n, j = 1, 2, \dots, m\}$ spans K . It remains to prove that $\{\alpha_i \beta_j\}$ is linearly independent. Suppose it is linearly dependent. Then there exists scalars $c_{i,j} \in F$ (not all zero) such that

$$\sum_{i=1}^n \sum_{j=1}^m c_{i,j} \alpha_i \beta_j = \sum_{j=1}^m \left(\sum_{i=1}^n c_{i,j} \alpha_i \right) \beta_j = 0$$

Let $\sum_{i=1}^n c_{i,j} \alpha_i = b_j$. We know that, $\{\beta_j : j = 1, 2, \dots, m\}$ is linearly independent. Thus, $\sum_j b_j \beta_j = 0 \implies b_j = 0, \forall j$.

Again $b_j = 0 \implies \sum_{i=1}^n c_{i,j} \alpha_i = 0$. Once again, $\{\alpha_i : i = 1, 2, \dots, n\}$ is linear independent. Thus, $c_{i,j} = 0, \forall i, j$. Thus, $\{\alpha_i \beta_j\}$ is linearly independent. Therefore, $\{\alpha_i \beta_j\}$ is a basis for the vector space $K(F)$. And $[K : F] = |\{\alpha_i \beta_j : i = 1, 2, \dots, n \text{ and } j = 1, 2, \dots, m\}| = mn$. \square

Corollary 6.2.2.1. *Let F_i be fields and F_{i+1} are finite extensions of F_i for $i = 1, 2, \dots, r$. Then $[F_r : F_1] = [F_r : F_{r-1}][F_{r-1} : F_{r-2}] \cdots [F_2 : F_1]$.*

Proof. We have, F_3 is a finite extension of F_1 and

$$[F_3 : F_1] = [F_3 : F_2][F_2 : F_1] \quad (6.2)$$

Suppose F_k is a finite extension of F_1 and

$$[F_k : F_1] = [F_k : F_{k-1}][F_{k-1} : F_{k-2}] \cdots [F_2 : F_1] \quad (6.3)$$

$$\begin{aligned} [F_{k+1} : F_1] &= [F_{k+1} : F_k][F_k : F_1] \text{ since } F_k \text{ is a finite extension of } F_1 \\ &= [F_{k+1} : F_k][F_k : F_{k-1}][F_{k-1} : F_{k-2}] \cdots [F_2 : F_1] \end{aligned}$$

\square

Corollary 6.2.2.2. *If E is an extension field of F and $\alpha \in E$ is algebraic over F and $\beta \in F(\alpha)$, then $\deg(\beta, F)$ divides $\deg(\alpha, F)$.*

Proof. We have, $\deg(\alpha, F) = [F(\alpha) : F]$ and $\deg(\beta, F) = [F(\beta) : F]$. Also given that $\beta \in F(\alpha) \implies F(\beta) \leq F(\alpha)$. Clearly, $F \leq F(\beta) \leq F(\alpha)$. Therefore, $[F(\alpha) : F] = [F(\alpha) : F(\beta)][F(\beta) : F]$. And $[F(\alpha) : F(\beta)] = [F(\alpha) : F]/[F(\beta) : F]$. Clearly, $[F(\beta) : F]$ divides $[F(\alpha) : F]$. \square

Theorem 6.2.3 (algebraic closure). *Let E be an extension field of F . Then $\bar{F}_E = \{\alpha \in E : \alpha \text{ is algebraic over } E\}$ is a subfield of E .*

Proof. Let $\alpha, \beta \in \bar{F}_E$. Then $\alpha, \beta \in E$ are algebraic over F . And $F(\alpha, \beta)$ is a finite extension field of F . Thus every element in $F(\alpha, \beta)$ are algebraic over F . Thus, $\alpha + \beta, \alpha\beta, \alpha - \beta, \alpha/\beta \in \bar{F}_E$. Therefore, \bar{F}_E is a subfield of E . \square

Corollary 6.2.3.1. *The set of all algebraic numbers forms a field.*

Proof. Let α be an algebraic number. Then $\alpha \in \mathbb{C}$ and α is algebraic over \mathbb{Q} . Clearly, the set of all algebraic numbers, $\bar{\mathbb{Q}}$ is a subfield of \mathbb{C} . \square

algebraic closure Let F be a field and E be an extension field of F . Then the (smallest) field containing all elements of E which are algebraic over F is the algebraic closure \bar{F}_E of F in E .

algebraically closed Let F be a field. F is algebraically closed if every non-constant polynomial in $F[x]$ has a zero in $F[x]$.

Note : Let F be algebraically closed. Then every irreducible polynomial in $F[x]$ are linear since every non-constant polynomial has a linear factor.

Theorem 6.2.4. *A field F is algebraically closed if and only if every non-constant polynomial $f(x)$ can be factorised in $F[x]$ into linear factors.*

Proof. Let F be algebraically closed and $f(x)$ be a non-constant polynomial in $F[x]$. Then $f(x)$ has a zero $\alpha \in F$. Then $x - \alpha$ is a factor of $f(x)$. That is, $f(x) = (x - \alpha)g(x)$. If $g(x) \in F[x]$ is non-constant, then it has a zero in F . Continuing like this, $f(x)$ can be factorised in $F[x]$ into linear factors.

Suppose every non-constant polynomial in $F[x]$ can be factorised into linear factors. Let $f(x)$ be a non-constant polynomial in $F[x]$. Then $f(x)$ has a linear factor $(ax + b) \in F[x]$. Clearly, $-b/a$ is a zero of $f(x)$. \square

Theorem 6.2.5. *Algebraically closed field has no proper algebraic extensions.*

Proof. Let E be an algebraic extension field of F . Then if $\alpha \in E$, we have $\text{irr}(\alpha, F) = (x - \alpha)$ since F is algebraically closed, every irreducible polynomial in $F[x]$ are linear. Thus $\alpha \in F$. Since $\alpha \in E$ is arbitrary, $F = E$. \square

Theorem 6.2.6 (Fundamental Theorem of Algebra). *\mathbb{C} is algebraically closed.*

Proof. Let $f(z) \in \mathbb{C}[z]$. Suppose $f(z)$ has no zeroes in \mathbb{C} . Then $1/f(z)$ is an entire function and as $|z| \rightarrow \infty$, $|f(z)| \rightarrow \infty$. Thus, $\lim_{|z| \rightarrow \infty} \frac{1}{|f(z)|} = 0$. And $1/f(z)$ is bounded.

By Liouville's theorem, every bounded entire function is constant. Therefore, $1/f(z)$ is constant and $f(z)$ is also constant. Thus, every non-constant polynomial function in $\mathbb{C}[z]$ has a zero in \mathbb{C} . \square

POSET Partial Ordered Set - A set together with partial order (reflexive, antisymmetric, transitive relation).

For example $(\mathbb{R}, <)$, the set of all real numbers together with less than relation is a poset. However, in a poset it is not necessary that two arbitrary elements are comparable. (\mathbb{C}, R) defined by aRb if $\Re(a) = \Re(b)$ and $\text{Im}(a) < \text{Im}(b)$ is a poset in which $2 + 3i$ and $3 + 3i$ are not comparable.

chain A subset of a poset in which any two elements are comparable. That is,
 $x, y \in T \implies x < y \text{ OR } y < x.$

For example, For above defined poset (\mathbb{C}, R) , $T = \{2 + ib \in \mathbb{C} : b \in \mathbb{R}\}$ is a chain in \mathbb{C} .

Lemma 6.2.7 (Zorn). *If every chain in a poset S has an upper bound. Then S has at least one maximal element in it.*

Proof. Not required (I think, there is no proof. We just take it as an axiom - always true !. If it is not true for a collection then it is not a set !!) \square

Theorem 6.2.8 (Existence of Algebraic Closure). *Every field F has an algebraic closure \bar{F} .*

Proof. Not required (as per syllabus) \square

6.2.1 Exercise §31

1.

6.3 Geometric Constructions §32

6.3.1 Basic Constructions

Finding Midpoint of a line

Let OA be a line. The line passing through the intersection of circles with center O and A with diameter greater than the length of the line gives a perpendicular line through its mid point (say, perpendicular bisection).

Drawing Perpendicular line through a point

Let OA be a line and B be a point on that line. Find points P, Q on OA which are equidistant from B . Then the perpendicular bisection of PQ is a line perpendicular to OA through B .

Drawing Parallel Line

Let OA be a line. Then the perpendicular line segment of any perpendicular line segment is a line segment parallel to OA .

6.3.2 Constructible Numbers

Constructible Number A real number α is constructible if you can draw a line of length $|\alpha|$, given a line of unit length, in finite steps using straight-edge and compass.

Theorem 6.3.1. *Let α, β be constructible real numbers. Then $\alpha + \beta$, $\alpha - \beta$, $\alpha\beta$, α/β ($\beta \neq 0$) are also constructible.*

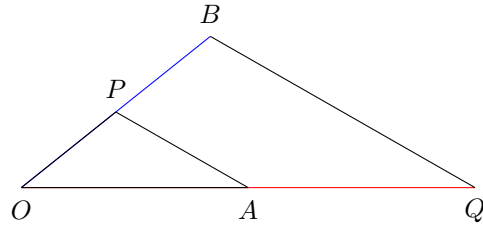
Proof. $\alpha + \beta$ Draw a line OA of length $|\alpha|$ and extend that line using straight edge OE . And draw the line AB of length $|\beta|$ on that extended line, at an point A of the former line extending it. Then OB is a line of length $|\alpha| + |\beta|$.



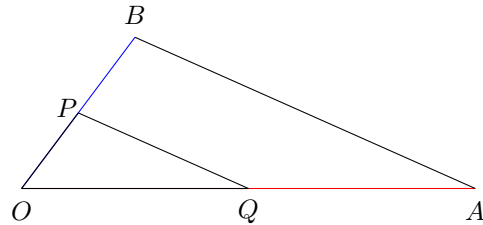
$\alpha - \beta$ Draw a line OA of length $|\alpha|$ and extend that line using straight edge OE . And draw the line AB of length $|\beta|$ on that extended line, at an end point A of former line, but in the opposite direction of extension (towards O). Then the line OB is of length $|\alpha| - |\beta|$.



$\alpha\beta$ Draw two lines OA and OB of length $|\alpha|$ and $|\beta|$ respectively (with a common end point O). Now draw the line of unit length OP along the line OB . Construct triangle OAP . Draw the line BQ parallel to the line PA through B such that OQ and OA are colinear. Now we have two similar triangles OPA and OBQ . Then, the line OQ of length $\|\alpha\beta\|$.



α/β Draw two lines OA and OB of length $\|\alpha\|$ and $\|\beta\|$, with a common end point O . Draw a line of unit length OP along the line OB . Now construct triangle OBA . Draw line PQ parallel to BA so that OQ and OA are colinear. Now the triangles OAB and OQP are similar. And the line OQ has length $\|\alpha/\beta\|$.



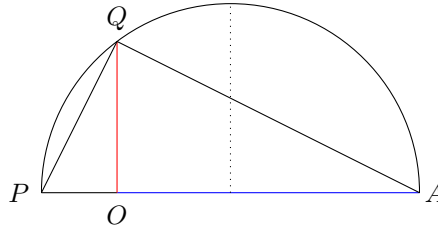
□

Corollary 6.3.1.1. *The set of all constructible real numbers forms a subfield of the field of real numbers.*

Proof. The set of all constructible real numbers say H , contains both 0 and 1. since the line of length zero is trivial and line of unit length is provided. And we have, $\forall \alpha, \beta \in H$, $\alpha + \beta, \alpha - \beta, \alpha\beta, \alpha/\beta \in H$. Since $\alpha - \beta \in H$, $0 - \beta = -\beta$ which is the additive inverse of β . And $1/\beta = \beta^{-1}$ is the multiplicative inverse of β . Thus, the set of all constructible numbers is a subfield of \mathbb{R} . \square

Theorem 6.3.2. *The field of F of constructible numbers consists **precisely** of all real numbers that we can obtain from \mathbb{Q} by taking square root of positive numbers a finite number of times and applying a finite number of field operations.*

Proof. The constructible numbers are closed under field operations and forms a subfield H of real numbers. We have, \mathbb{Q} is the prime field of \mathbb{R} . That is, every subfield of \mathbb{R} contains \mathbb{Q} . Thus, the subfield H of constructible numbers contains \mathbb{Q} . That is, all the rational numbers are constructible. Therefore, it remains to prove that if $\alpha > 0$ is constructible then $\sqrt{\alpha}$ is also constructible.



Let OA and OP be colinear lines such that OA be a line of length $\|\alpha\|$ and OP be a line of unit length. Find the mid point of PA and draw a circle of diameter PA . Then the length of the perpendicular OQ from PA to the circle is of length $\|\sqrt{\alpha}\|$ since $\triangle OAQ$ and $\triangle OQP$ are similar triangles.

Therefore, real numbers obtained from rational numbers through finite number of additions/subtractions, multiplications/divisions, and square root are constructible. \square

Note : For example $\sqrt[4]{5\sqrt[8]{3}-2}$ is constructible, but $\sqrt[6]{2}$ and π are not constructible. We skip the proof that a real number which can't be obtained from rationals by a finite number of these operations is not constructible. And assume that these three operations define the entire field of constructible numbers.

Corollary 6.3.2.1. *Let γ be constructible number which is not rational. Then there exists a sequence of real numbers $\alpha_1, \alpha_2, \dots, \alpha_n = \gamma$ such that for every $i = 2, \dots, n$, the extension field $\mathbb{Q}(\alpha_1, \alpha_2, \dots, \alpha_i)$ is an extension of $\mathbb{Q}(\alpha_1, \alpha_2, \dots, \alpha_{i-1})$ of degree two.*

In other words, for any constructible number γ , $[\mathbb{Q}(\gamma) : \mathbb{Q}] = 2^n$ for some positive integer n .

Proof. Let γ be a constructible number which can be obtained from rationals by n square root operations and a finite number of field operations. Then, we have a sequence of constructible numbers $\alpha_1, \alpha_2, \dots, \alpha_n$ such that $[\mathbb{Q}(\alpha_1, \alpha_2, \dots, \alpha_i) : \mathbb{Q}(\alpha_1, \alpha_2, \dots, \alpha_{i-1})] = 2$. Therefore, $[\mathbb{Q}(\gamma) : \mathbb{Q}] = 2^n$. \square

For example, $\gamma = \sqrt[4]{5\sqrt[8]{3}-2}$. Then $\alpha_1 = \sqrt{3}$, $\alpha_2 = \sqrt[4]{3}$, $\alpha_3 = \sqrt[8]{3}$, $\alpha_4 = \sqrt{5\sqrt[8]{3}-2}$ and $\alpha_5 = \gamma$. Clearly, the geometric construction of γ contains five instances of square root operation.

6.3.3 Impossible Problems from Ancient Times

Doubling the cube

Theorem 6.3.3. *There exists a cube such that it is impossible to construct the side of the cube with double the volume.*

Proof. Suppose the cube is of unit side. Then the side of the cube with double the volume is $\gamma = \sqrt[3]{2}$. And $\text{irr}(\gamma, \mathbb{Q}) = x^3 - 2$ and $[\mathbb{Q}(\sqrt[3]{2} : \mathbb{Q})] = 3 \neq 2^n$. Therefore, γ is not constructible. \square

Squaring the circle

Theorem 6.3.4. *There exists a circle such that it is impossible to construct the side of the square with the same area.*

Proof. Suppose the circle is of unit radius. Then area of the circle is 2π . And the side of the square with same area is $\sqrt{2\pi}$. Since π is transcendental, $\pi, \sqrt{\pi}$ and $\sqrt{2\pi}$ are not constructible. \square

Trisecting an angle

Theorem 6.3.5. *There exists an angle which can be trisected.*

Proof. We have $\cos 3\theta = 4\cos^3\theta - 3\cos\theta$. Consider $\theta = 20^\circ$. Then $\gamma = \cos\theta$ is a root of the irreducible polynomial $4x^3 - 3x - 0.5$. That is, γ is a root of the monic irreducible polynomial $x^3 - \frac{3}{4}x - \frac{1}{8}$ of degree 3. Thus, γ is not constructible, since $[\mathbb{Q}(\gamma) : \mathbb{Q}] = 3 \neq 2^n$. \square

6.3.4 Exercise §32

1.

6.4 Finite Fields §33

6.4.1 Structure of a Finite Field

Theorem 6.4.1. *Let E be a finite extension of degree n over a finite field F . If F has q elements, then E has q^n elements.*

Proof. We have, extension field E is an n -dimensional vector space over the field F . Let $\{\alpha_1, \alpha_2, \dots, \alpha_n\}$ be a basis of the vector space E over F . Then every elements of E can be uniquely written as linear combination of basis vectors.

$$\forall \beta \in E, \beta = b_1\alpha_1 + b_2\alpha_2 + \dots + b_n\alpha_n$$

Suppose $\beta \in E$ has two distinct linear combinations. Then, the vectors $\{\alpha_1, \alpha_2, \dots, \alpha_n\}$ are not linearly independent, which is a contradiction.

Since the representation is unique and F has q elements, there are q^n distinct linear combinations possible. Therefore, E has q^n elements. \square

Corollary 6.4.1.1. *If E is a finite field of characteristic p . Then F contains exactly p^n elements for some integer n .*

Proof. We have, E is a finite field of characteristic p . Thus, \mathbb{Z}_p is the prime subfield of E . And E is a finite extension of \mathbb{Z}_p . Thus, E is a finite dimensional vector space over \mathbb{Z}_p . Let n be the dimension of E over \mathbb{Z}_p . And \mathbb{Z}_p has p elements. Therefore, E has p^n elements. \square

Theorem 6.4.2. *Let E be a field of p^n elements contained in the algebraic closure $\overline{\mathbb{Z}_p}$ of \mathbb{Z}_p . Then the elements of E are precisely the zeroes of the polynomial $x^{p^n} - x \in \mathbb{Z}_p[x]$.*

Proof. We have E^* is a multiplicative group of non-zero elements in E . And E^* has $p^n - 1$ elements. Thus, order of any element $\alpha \in E^*$ should divide $p^n - 1$. In other words, if $\alpha \in E^*$, then $\alpha^{p^n-1} = 1$. Clearly, $\alpha^{p^n} - \alpha = 0, \forall \alpha \in E$. However, x^{p^n} can have at most p^n zeroes in $\overline{\mathbb{Z}_p}$. Thus, E is precisely the set of all zeroes of $x^{p^n} - x \in \overline{\mathbb{Z}_p}$. \square

n th Root of Unity α is an n th root of unity if $\alpha^n = 1$. ie, $\alpha = \sqrt[n]{1}$

Primitive n th Root of Unity α is a primitive n th root of unity if n is the smallest positive integer such that $\alpha^n = 1$.

That is, $\alpha^n = 1$ and $\forall m \in \mathbb{N}, m < n \implies \alpha^m \neq 1$

Theorem 6.4.3. *The multiplicative group of non-zero elements of a finite field F is cyclic.*

Proof. Refer : [Fraleigh, 2013, Theorem 23.6] \square

Corollary 6.4.3.1. *Finite extension of finite fields are simple extensions.*

Proof. Let E be a finite extension field of the finite field F . Then the multiplicative group of non-zero elements E^* is cyclic. Let α be a generator of the cyclic group E^* . Then, $E = F(\alpha)$. \square

6.4.2 Galois Field $GF(p^n)$

Lemma 6.4.4. *If F is a field of prime characteristic p with algebraic closure \overline{F} , then $x^{p^n} - x$ has p^n distinct zeroes in \overline{F} .*

Proof. We have, \overline{F} is algebraically closed. And $x^{p^n} - x \in \overline{F}[x]$. Thus, $x^{p^n} - x$ can be factorised into p^n linear components. It remains to prove that these factors are distinct.

Clearly, 0 is a zero of multiplicity 1, since $x^{p^n} - x = x(x^{p^n-1} - 1)$. Let $\alpha \neq 0$ be a zero of $x^{p^n} - x$. Then α is a zero of $x^{p^n-1} - 1$. ie, $\alpha^{p^n-1} = 1$.

$$\begin{aligned}(x - \alpha)g(x) &= x^{p^n-1} - 1 \\ g(x) &= \frac{x^{p^n-1} - 1}{x - \alpha}\end{aligned}$$

By long division, we get

$$\begin{aligned}g(x) &= x^{p^n-2} + \alpha x^{p^n-3} + \alpha^2 x^{p^n-4} + \cdots + \alpha^{p^n-3} x + \alpha^{p^n-2} \\ g(\alpha) &= (p^n - 1)\alpha^{p^n-2} \\ &= (p^n - 1) \frac{\alpha^{p^n-1}}{\alpha} \\ &= p^n \frac{1}{\alpha} - \frac{1}{\alpha} \\ &= -\frac{1}{\alpha} \neq 0\end{aligned}$$

Thus, every zero of $x^{p^n-1} - x$ is of multiplicity 1. \square

Lemma 6.4.5. *If F is a field of prime characteristic p , then $(\alpha + \beta)^{p^n} = \alpha^{p^n} + \beta^{p^n}$.*

Proof. Since F is a field of characteristic p , for every $\alpha \in F$, $p\alpha = 0$.

For $n = 1$, we have

$$\begin{aligned}(\alpha + \beta)^p &= \alpha^p + p\alpha^{p-1}\beta + \frac{p(p-1)}{2}\alpha^{p-2}\beta^2 + \cdots + p\alpha\beta^{p-1} + \beta^p \\ &= \alpha^p + 0\alpha^{p-1}\beta + 0\alpha^{p-2}\beta^2 + \cdots + 0\alpha\beta^{p-1} + \beta^p \\ &= \alpha^p + \beta^p\end{aligned}$$

Suppose $(\alpha + \beta)^{p^{n-1}} = \alpha^{p^{n-1}} + \beta^{p^{n-1}}$, then

$$\begin{aligned}(\alpha + \beta)^{p^n} &= \left[(\alpha + \beta)^{p^{n-1}}\right]^p \\ &= \left[\alpha^{p^{n-1}} + \beta^{p^{n-1}}\right]^p \\ &= \alpha^{p^n} + \beta^{p^n}\end{aligned}$$

Therefore, by mathematical induction the result is true. \square

Theorem 6.4.6 (Existence of Galois Field). *For every prime power p^n , a finite field of p^n elements exists.*

Hint : $\alpha^{p^n} = \alpha \iff \alpha^{p^n} - \alpha = 0 \iff \alpha \text{ is a zero of } x^{p^n} - x$

Proof. Consider the algebraic closure $\overline{\mathbb{Z}_p}$ of \mathbb{Z}_p . Let K be a subset of $\overline{\mathbb{Z}_p}$ containing all zeroes of $x^{p^n} - x \in \overline{\mathbb{Z}_p}$.

Let $\alpha, \beta \in K$. Then $\alpha^{p^n} = \alpha$ and $\beta^{p^n} = \beta$. Since $\overline{\mathbb{Z}_p}$ is a field of characteristic p , we have $(\alpha + \beta)^{p^n} = \alpha^{p^n} + \beta^{p^n} = \alpha + \beta$. Thus, $(\alpha + \beta)$ is a zero of $x^{p^n} - x$. ie, $(\alpha + \beta) \in K$. Clearly, $(\alpha\beta)^{p^n} = \alpha^{p^n}\beta^{p^n} = \alpha\beta$. Thus, $\alpha\beta \in K$.

Again $(-\alpha)^{p^n} = (-1 \cdot \alpha)^{p^n} = (-1)^{p^n} \cdot \alpha^{p^n} = -1 \cdot \alpha = -\alpha$. Thus, $-\alpha \in K$. Also $(\alpha^{-1})^{p^n} = (\frac{1}{\alpha})^{p^n} = \frac{1}{\alpha^{p^n}} = \frac{1}{\alpha} = \alpha^{-1}$. Thus, $\alpha^{-1} \in K$.

Trivially, $0, 1 \in K$. Therefore, K is a subfield of $\overline{\mathbb{Z}_p}$ with p^n elements since the p^n zeroes of $x^{p^n} - x$ are distinct. \square

Corollary 6.4.6.1. *If F is a finite field, then for every positive integer n , there exists an irreducible polynomial in $F[x]$ of degree n .*

Proof. Let F be a finite field. Then \mathbb{Z}_p is prime field of F for some prime p . And F is of characteristic p and has p^r elements for some positive integer r .

Let K be the subfield of \overline{F} containing precisely all the zeroes of the polynomial $x^{p^{rn}} - x \in \mathbb{Z}_p[x]$. Then, K has a subfield isomorphic to \mathbb{Z}_p since F is of characteristic p and every subfield of F has a subfield isomorphic to \mathbb{Z}_p .

By existence theorem of Galois Fields, every element of F is a zero of the polynomial $x^{p^r} - x \in \mathbb{Z}_p[x]$. That is, $\alpha \in F \iff \alpha^{p^r} = \alpha$.

$$\begin{aligned} \alpha^{p^{rn}} &= \left[\alpha^{p^r} \right]^{p^{r(n-1)}} &&= \alpha^{p^{r(n-1)}} \\ &= \left[\alpha^{p^r} \right]^{p^{r(n-2)}} &&= \alpha^{p^{r(n-2)}} \\ &\vdots &&\vdots \\ &= \left[\alpha^{p^r} \right]^{p^r} &&= \alpha^{p^r} = \alpha \end{aligned}$$

Thus, $\alpha \in F \implies \alpha^{p^{rn}} = \alpha \implies \alpha \in K$. Therefore F is a subfield of K . Clearly, K is a finite extension of the finite field F . And the vector space K over F is n -dimensional, since K has $p^{rn} = [p^r]^n$ elements and F has p^r elements.

Since every finite extension of finite fields are simple extensions, we have $K = F(\beta)$ and $\text{irr}(\beta, F) = n$. That is, there exists an unique monic, irreducible polynomial $p(x) \in F[x]$ of degree n such that $p(\beta) = 0$. Therefore, $\forall n \in \mathbb{Z}^+$, there exists an irreducible polynomial in $F[x]$ of degree n . \square

Theorem 6.4.7 (Uniqueness of Galois Field). *Let p be a prime and n a positive integer. If E and E' are fields of order p^n , then E and E' are isomorphic.*

*There exists a unique finite field of order p^n , say **Galois Field**, $GF(p^n)$.*

Proof. Let E, E' be fields of order p^n . Then both fields have \mathbb{Z}_p as prime field. Thus, E is a simple extension of \mathbb{Z}_p of degree n . ie, $[E : \mathbb{Z}_p] = n$. And there exists an irreducible polynomial $f(x)$ such that $E \simeq \mathbb{Z}_p[x] / \langle f(x) \rangle$.

Elements of E are zeroes of $x^{p^n} - x$, thus $f(x)$ is a factor of $x^{p^n} - x$. Clearly, elements of E' are zeroes of x^{p^n} and therefore E' has all zeroes of $f(x)$. And

$E' \simeq \mathbb{Z}_p / \langle f(x) \rangle \simeq E$. Therefore, there exists a unique field of order p^n (upto isomorphism), say $GF(p^n)$. \square

6.4.3 Exercise §33

- 1.

Module 2 : Unique Factorisation Domains

6.5 Unique Factorisation Domains §45

Definitions 6.5.1 (divides). An element $a \in R$ divides b if there exists an element $c \in R$ such that $b = ac$.

Definitions 6.5.2 (associate). Two elements a, b are associates if $a = bu$ where u is a unit.

Definitions 6.5.3 (UFD). **Unique factorisation domain**, UFD is an integral domain such that

1. Every element can be factored into a finite number of irreducibles, except 0 and units \dagger^1 .
2. The above factorisation is unique except for order and associates.

For example, In \mathbb{Z} , $24 = 2 \times 2 \times 2 \times 3 = -2 \times -3 \times 2 \times 2$. Here 2 and -2 are associates. And 2 and 3 are not units, since $2^{-1}, 3^{-1} \notin \mathbb{Z}$.

Definitions 6.5.4 (PID). An integral domain D is a **Principal Ideal Domain** if every ideal in D is a principal ideal.

There are two important results on UFDs.

1. Every PID is a UFD.
2. If D is a UFD, then $D[x]$ is a UFD.

6.5.1 Every PID is UFD

Lemma 6.5.1. Let R be a commutative ring and let $N_1 \subset N_2 \subset \dots$ be an ascending chain of ideals in R . Then $N = \cup_i N_i$ is an ideal of R .

Proof. **Step 1 : N is a subring of R**

Suppose N_i, N_j are two ideals in the chain, $N_i \subset N_j$ and $a \in N_i, b \in N_j$. Clearly, $a \in N_j$.

And $a \pm b, ab \in N_j$. And $0 \in N_j \implies 0 \in N$. We have, 0 in every ideal². Take $a = 0$, we know that for every element $b \in N_j$, its additive inverse $-b \in N_j$. Thus $b \in N \implies b \in N_j \implies -b \in N_j \implies -b \in N$. Clearly, N is a subring of R .

Step 2 : N is a ideal of R

Let $a \in N$ and $r \in R$. We have, $a \in N \implies a \in N_j$ for some ideal N_j in the chain. Since N_j is an ideal $ar = ra \in N_j$. Therefore, $ar \in N$. And N is a ideal of R . \square

¹Unit is a element which has multiplicative inverse.

²Suppose $b \in N$. We have $-b \in R$ and $-b + b = 0 \in bN \subset N$.

Lemma 6.5.2 (Ascending Chain Condition). *Let D be a PID. If $N_1 \subset N_2 \subset \dots$ is an ascending chain of ideal, then there exists a positive integer r such that $N_r = N_s$ for every $s \geq r$.*

In other words, every strictly³ ascending chain of ideals in a PID is of finite length.

Proof. Let D be an integral domain. And $N_1 \subseteq N_2 \subseteq \dots$ be an ascending chain of ideals in D . Then $N = \cup_i N_i$ is an ideal in D . Since D is a PID, by definition of **p**ri**n**cip**a**l **i**de**a**l **d**om**a**in every ideal in it is a principal ideal. Thus, N is a principal ideal in D . That is, there exists $c \in D$ such that $\langle c \rangle = N$.

Since $c \in N$, and $N = \cup_i N_i$ we have $c \in N_r$ for some $r \in \mathbb{N}$. Then $\langle c \rangle = N_r = N$. Again, $N_r \subset N_s \implies c \in N_s \implies \langle c \rangle = N_s = N$.

Suppose the chain of ideals are strictly ascending, that is every ideal in the chain is properly containing the former ideal. Then, the ideal N_r such that $c \in N_r$ is the last ideal in its chain. Thus, strictly ascending chain of ideals is finite. \square

Theorem 6.5.3. *Let D be a PID. Every element that is neither 0 nor a unit in D is a product of irreducibles.*

Proof. Suppose D is a PID. And suppose $a \in D$ is a neither a zero nor a unit. If a is an irreducible, then the result is trivial. Suppose a is not an irreducible.

Step 1 : a has an irreducible factor

By the definition of irreducibility, a has a factorisation $a = a_1 b_1$ where a_1, b_1 are non-units. And every element $ar \in \langle a \rangle$ can be expressed as $ar = a_1 b_1 r = a_1 r' \in \langle a_1 \rangle$. Thus, the ideal generated by a , $\langle a \rangle$ is contained in the ideal generated by a_1 , $\langle a_1 \rangle$. That is, $\langle a \rangle \subset \langle a_1 \rangle$.

If a_1 is an irreducible, then the proof is complete. Suppose a_1 is not an irreducible. That is, $a_1 = a_2 b_2$. Then $\langle a_1 \rangle \subset \langle a_2 \rangle$. Also we have, $\langle a_1 \rangle \neq \langle a_2 \rangle$.

Suppose $\forall d \in D, a_2 d \in \langle a_1 \rangle \implies a_2 \cdot 1 = a_2 \in \langle a_1 \rangle \implies a_2 = a_1 c_2 \implies a_1 = (a_1 c_2) b_2 \implies c_2 b_2 = 1 \implies b_2$ has multiplicative inverse c_2 . This contradicts the assumption that b_2 is a unit.

Continuing like this, we get a **strictly** ascending chain of ideals $\langle a \rangle \subsetneq \langle a_1 \rangle \subsetneq \langle a_2 \rangle, \dots$. And we know that, every strictly ascending chain of ideals is finite. Thus, a_r is an irreducible. And $a = a_1 b_1 = (a_2 b_2) b_1 = \dots = a_r (b_r b_{r-1} \dots b_1) = a_r b$. Clearly, a has an irreducible factor a_r .

Step 2 : a is a product of irreducibles

Suppose a is neither zero nor a unit. Then $a = p_1 c_1$ where p_1 is an irreducible and c_1 is not a unit. Continuing like this, we get another strictly ascending chain of ideals $\langle a \rangle \subsetneq \langle c_1 \rangle \subsetneq \langle c_2 \rangle \subsetneq \dots$ such that $a = p_1 c_1 = p_1 (p_2 c_2) = \dots$ where p_1, p_2, \dots are irreducibles and c_j are non-units. Again, by ascending

³Strictly ascending chain : $N_1 \subsetneq N_2 \subsetneq N_3 \subsetneq \dots \subsetneq N_k$.

chain condition this strictly ascending chain is finite. That is, $a = p_1 p_2 \dots p_k c_k$ for some $k \in \mathbb{N}$. Since $\langle c_k \rangle$ is a maximal ideal in D , c_k is an irreducible say, p_{k+1} . Thus, any element $a \in D$ (except zero and units) can be expressed as product of irreducibles. \square

6.5.2 Irreducible element and Maximal Ideal

We have seen earlier that a non-trivial principal ideal $\langle p(x) \rangle$ in $F[x]$ is maximal if and only if the generator $p(x)$ is irreducible over F . Now we have a generalisation, which says ideals in PIDs are maximal if and only if the generator is irreducible. Remember that for any field F , $F[x]$ is a PID.

Lemma 6.5.4. *An ideal $\langle p \rangle$ in a PID is maximal if and only if p is an irreducible.*

Proof. Part A : $\langle p \rangle$ maximal in $D \implies p$ irreducible in D

Let D be a PID. Suppose $\langle p \rangle$ is a maximal ideal in D . Suppose $p = ab$ in D . That is, $p = ab$ where $a, b \in D$. Then, $\langle p \rangle \subset \langle a \rangle$.

Suppose $\langle p \rangle = \langle a \rangle$. Then $a = pc = (ab)c \implies bc = 1$ and b is a unit.

Suppose $\langle p \rangle \neq \langle a \rangle$. Then $\langle a \rangle = D$, since $\langle p \rangle$ is a maximal ideal. That is $\langle a \rangle = \langle 1 \rangle \implies a, 1$ are associates. And a is a unit.

Thus for any factorisation $p = ab$, either a or b is a unit. Therefore, p is an irreducible.

Part B : p is irreducible in $D \implies \langle p \rangle$ is maximal in D

Suppose p is an irreducible in D . Then $p = ab$ where $a, b \in D$ implies that either a or b is a unit. Since $p = ab$, $\langle p \rangle \subset \langle a \rangle$. We know that, if a is a unit then $\langle a \rangle = \langle 1 \rangle = D$.

Suppose a is not a unit. Then b is a unit. That is, there exists $u \in D$ such that $bu = 1$. Then $pu = abu = a$. Thus, $\langle a \rangle \subset \langle p \rangle \implies \langle p \rangle = \langle a \rangle$.

Thus, for any factorisation $p = ab$, the ideals generated by a, b are either D or $\langle p \rangle$ itself. In other words, there doesn't exist a proper ideal containing $\langle p \rangle$. Therefore, $\langle p \rangle$ is a maximal ideal in D . \square

6.5.3 Every PID is UFD

We have seen earlier that $F[x]$ has unique factorisation of all its elements. Now, we generalise that result into "every PID has unique factorisation for all its elements".

Lemma 6.5.5. *In a PID, if an irreducible p divides ab , then either $p|a$ or $p|b$.*

Proof. Let D be a PID. Suppose $p \in D$ is irreducible. Suppose $p|ab$. Then $ab \in \langle p \rangle$. And we have, $\langle p \rangle$ is a maximal ideal. However, every maximal ideal in D is a prime ideal. Thus, $ab \in \langle p \rangle \implies a \in \langle p \rangle$ or $b \in \langle p \rangle$. In other words, $p|a$ or $p|b$. \square

Corollary 6.5.5.1. *If p is an irreducible in a PID D and p divides $a_1 a_2 \dots a_n$ where $a_i \in D$, then $p|a_i$ for at least one i .*

Proof. We know that, $p|a_1 a_2 \implies p|a_1$ or $p|a_2$.

Suppose $p|a_1 a_2 \dots a_k \implies p|a_1$ or $p|a_2 \dots$ or $p|a_k$.

Suppose $p|a_1 a_2 \dots a_{k+1} \implies p|a_1 a_2 \dots a_k$ or $p|a_{k+1}$.

$\implies p|a_1$ or $p|a_2$ or \dots or $p|a_k$ or $p|a_{k+1}$.

That is, if p divides a finite product, then it divides at least one of them. \square

Definitions 6.5.5. A nonzero, nonunit element p in an integral domain D is a prime if $\forall a, b \in D$, $p|ab \implies p|a$ or $p|b$.

Theorem 6.5.6. *Every PID is a UFD.*

Proof. Let D be a PID. Let $a \in D$ is neither a zero nor a unit. Then a has a factorisation into irreducibles $a = p_1 p_2 \dots p_r$. Suppose a has another factorisation $a = q_1 q_2 \dots q_s$.

We have $p_1|q_1 q_2 \dots q_s$. Thus, p_1 divides at least one of them, say q_j . Without loss of generality, we assume that $p_1|q_1$ (rearranging the terms if necessary).

Since q_1 is an irreducible, $q_1 = p_1 u_1 \implies u_1$ is a unit. Thus, $p_1 p_2 \dots p_r = p_1 u_1 q_2 q_3 \dots q_s$. That is, $p_2 p_3 \dots p_r = u_1 q_2 q_3 \dots q_s$.

Continuing like this we get, $p_r = u_1 u_2 \dots u_{r-1} p_r u_r q_{r+1} q_{r+s} \dots q_s$. Thus $u_1 u_2 \dots u_r q_{r+1} q_{r+2} \dots q_s$ is a unit. We have $s = r$, otherwise $1 = u_1 u_2 \dots q_{r+1} \dots$ is a contradiction. Therefore, any element $a \in D$ has a unique factorisation except for order, associates and units. \square

Corollary 6.5.6.1. *The integral domain \mathbb{Z} is a UFD.*

Proof. The ideals in \mathbb{Z} are of the form $n\mathbb{Z}$ where $n \in \mathbb{Z}$. Thus every ideal in \mathbb{Z} are principal ideals. Therefore, \mathbb{Z} is a PID. We know that every PID is a UFD. Thus, \mathbb{Z} is a **unique factorisation domain**. \square

6.5.4 D is UFD $\implies D[x]$ is UFD

Definitions 6.5.6 (gcd). Let D be a UFD and a_1, a_2, \dots, a_n be nonzero elements in D . An element $d \in D$ is a **greatest common divisor** of a_i if d is a common divisor of all a_i s and also divides any common divisor of a_i s.

Definitions 6.5.7 (primitive). Let D be a UFD. A nonconstant polynomial $a_0 + a_1 x + \dots + a_n x^n \in D[x]$ is a primitive if 1 is the gcd of all a_i s.

For example, $4x^3 + 3x^2 + 2 \in \mathbb{Z}[x]$ is a primitive since $\gcd(4, 3, 2) = 1$.

Lemma 6.5.7. *If D is a UFD, then for every nonconstant $f(x) \in D[x]$ we have $f(x) = cg(x)$ where $c \in D$ and $g(x) \in D[x]$ and $g(x)$ is primitive. The element $c \in D$ is unique upto a unit factor in D and is the **content** of $f(x)$. Also $g(x)$ is unique upto a unit factor in D .*

Proof. Step 1 : There exists $c \in D$ such that $f(x) = cg(x)$

Let $f(x) = a_0 + a_1x + \cdots + a_nx^n$ be a non-constant polynomial in $D[x]$. Let $c = \gcd(a_0, a_1, \dots, a_n)$. In other words, c is the greatest common divisor of the coefficients of $f(x)$. Let $q_i = ca_i$. Clearly, $\gcd(q_1, q_2, \dots, q_n) = 1$. Thus, we have $f(x) = cg(x)$ where $g(x) = q_0 + q_1x + \cdots + q_nx^n$ is a primitive.

Step 2 : There exists unique c such that $f(x) = cg(x)$

Suppose $f(x) = cg(x) = dh(x)$ where $c, d \in D$ and $g(x), h(x)$ are primitives.

Clearly, c and d are associates. Otherwise, there exists an irreducible element p such that $p|cg(x)$ but $p \nmid dh(x)$ which is a contradiction. Thus, we can cancel all irreducible factors of c to obtain, $ug(x) = vh(x)$ where u and v are units. Therefore, the content of $f(x)$, c unique upto associates.

Again, $f(x) = cg(x) = (cu)(u^{-1}g(x))$ and thus $g(x)$ is unique upto associates/units. \square

Definitions 6.5.8 (content). Let D be a UFD. Let $f(x) \in D[x]$. Then $f(x) = cg(x)$ for some primitive $g(x) \in D[x]$ and $c \in D$. The element c which is unique upto units, is the content of $f(x)$.

For example, Let $f(x) = 8x^3 + 6x^2 + 4 \in \mathbb{Z}[x]$. Then $f(x) = 2(4x^3 + 3x^2 + 2)$ where $4x^3 + 3x^2 + 2 \in \mathbb{Z}[x]$ is a primitive and $2 \in \mathbb{Z}$. Thus, $\text{content}(f) = 2$.

Lemma 6.5.8 (Gauss). *If D is a UFD, then product of two primitive polynomials in $D[x]$ is again primitive.*

Proof. Let $f(x) = a_0 + a_1x + \cdots + a_nx^n$ and $g(x) = b_0 + b_1x + \cdots + b_mx^m$. Suppose $f(x)$ and $g(x)$ are primitives. Let $h(x) = f(x)g(x) = c_0 + c_1x + \cdots + c_{n+m}x^{n+m}$. Let p be an irreducible in D . Then $f(x)$ has a coefficient which is not divisible by p . Suppose p divides every coefficient $f(x)$, then $f(x)$ is not a primitive.

Let r be the smallest integer such that $p \nmid a_r$. Similarly, let s be the smallest integer such that $p \nmid b_s$. Now consider the coefficient c_{r+s} of $h(x)$.

$$c_{r+s} = a_0b_{r+s} + \cdots + a_{r-1}b_{s+1} + a_rb_s + a_{r+1}b_{s-1} + \cdots + a_{r+s}b_0$$

By the selection of r , $p|a_j$ for every $j < r$. Thus, $p|a_0b_s + \cdots + a_{r-1}b_{s+1}$. Similarly, by the selection of s , $p|b_k$ for every $k < s$. Thus, $p|a_{r+1}b_{s-1} + \cdots + a_{r+s}b_0$. Clearly, $p|c_{r+s} \iff p|a_rb_s \implies p|a_r$ or $p|b_s$. This is not possible, thus $p \nmid c_{r+s}$. Thus, for any irreducible $p \in D$, $h(x)$ has a coefficient which is not divisible by p . Therefore, product of two primitives is always a primitive. \square

Corollary 6.5.8.1. *If D is a UFD, then a finite product of primitive polynomials in $D[x]$ is again primitive.*

Proof. Let $\{f_k\}_{k=1}^n$ be a family of primitives. Then $f_1(x)f_2(x)$ is a primitive, since product of two primitives is always a primitive.

Suppose $g(x) = f_1(x)f_2(x) \cdots f_k(x)$ is a primitive.

Consider $h(x) = f_1(x)f_2(x)\dots f_{k+1}(x) = g(x)f_{k+1}(x)$. Since both $g(x)$ and $f_{k+1}(x)$ are primitives their product $h(x)$ is also a primitive.

Thus by finite mathematical induction, every finite product of primitives is a primitive. \square

Lemma 6.5.9. *Let D be a UFD and let F be a field of quotients of D . Let $f(x) \in D[x]$ where $\deg f(x) > 0$. If $f(x)$ is an irreducible in $D[x]$, then $f(x)$ is also an irreducible in $F[x]$. Also, if $f(x)$ is primitive in $D[x]$ and irreducible in $F[x]$, then $f(x)$ is irreducible in $D[x]$.*

Proof. Let D be a UFD and F be the field of quotients of D . Then elements of F are of the form $(a, b) \in D \times D$. Also $(a, b) + (c, d) = (ad + bc, bd)$ and $(a, b) \cdot (c, d) = (ac, bd)$. And we say express (a, b) as a/b .

Part 1 : irreducible in $D[x] \implies$ irreducible in $F[x]$

Let $f(x) \in D[x]$ be an irreducible in $D[x]$. Suppose $f(x)$ is not irreducible in $F[x]$. Then $f(x) = r(x)s(x)$ where $r(x), s(x) \in F[x]$.

The coefficients of $r(x)$ are of the form a/b . By multiplying $r(x)$ with the product of all denominators, we can remove its denominators and obtain an associate $r_1(x) \in D[x]$. Similarly, we have $s(x) = vs_1(x)$ where $s_1(x) \in D[x]$. Thus, $f(x) = r(x)s(x) = ur_1(x)vs_1(x) = uvr_1(x)s_1(x)$. Therefore, $f(x) \in D[x]$ is not an irreducible, which is a contradiction. Thus, $f(x)$ is irreducible in the quotient field F as well.

Part 2 : irreducible in $F[x] \implies$ irreducible in $D[x]$

Suppose $f(x) \in D[x]$ is irreducible in $F[x]$. Clearly, F contains a UFD isomorphic to D . Thus $D[x] \leq F[x]$. Therefore, any polynomial irreducible in $F[x]$ is also irreducible in $D[x]$. \square

We just saw that if F is the quotient field of a UFD D and $f(x) \in D[x]$. The irreducibility of $f(x)$ in $F[x]$ is necessary and sufficient for the irreducibility of $f(x)$ in $D[x]$. Now we prove that the factorisation is also unique upto a morphism.

Corollary 6.5.9.1. *If D is a UFD and F is a field of quotients of D , then a nonconstant polynomial $f(x) \in D[x]$ factors into a product of two polynomials of lower degrees r and s in $F[x]$ if and only if it has a factorization into polynomials of the same degrees r and s .*

Proof. We know that if $f(x) \in D[x]$ has a factorisation $f(x) = r(x)s(x)$ in $F[x]$. Then it has a factorisation in $D[x]$ as well, $f(x) = cr_1(x)s_1(x)$ where $r_1(x), s_1(x)$ are associates of $r(x), s(x)$ in $F[x]$. Clearly, the associates are always of the same degree.

Trivially, any factorisation in $D[x]$ will also be a factorisation in $F[x]$. And the factors will have the same degree, unless $D[x]$ has an irreducible polynomial which is not irreducible in $F[x]$. This is not possible. \square

Theorem 6.5.10. *If D is a UFD, then $D[x]$ is a UFD.*

Proof. Let $f(x) \in D[x]$. Suppose $\deg(f) > 0$. Otherwise, $f(x)$ is a constant and we have trivial factorisation.

Let $f(x) = g_1(x)g_2(x)\dots g_r(x)$ be a factorisation in $D[x]$ with maximum number of factors. Then $f(x) = c_1h_1(x)c_2h_2(x)\dots c_rh_r(x)$ where $g_k(x) = c_kh_k(x)$ and $h_k(x)$ are primitives. And $h_k(x)$ are irreducibles. Otherwise, $f(x)$ has a factorisation with greater number of factors than r which is a contradiction. Thus, $f(x) = uh_1(x)h_2(x)\dots h_r(x)$ is a factorisation of $f(x)$ into a product of irreducibles.

Suppose $f(x)$ has another factorisation $f(x) = G_1(x)G_2(x)\dots G_r(x)$. Then $f(x) = vH_1(x)H_2(x)\dots H_r(x)$ with $H_k(x)$ are all irreducibles. We know that, $h_k(x)|f(x) \implies h_k(x)|H_1(x)H_2(x)\dots H_r(x) \implies h_k(x)|H_k(x)$ (WLOG). Therefore, the factorisation is unique. \square

For example, let x, y be two indeterminates, then an element $f(x) \in F[x, y]$ is of the form $\sum_{i,j=0}^{n,m} a_{ij}x^i y^j$.

Corollary 6.5.10.1. *If F is a field of quotients and x_1, x_2, \dots, x_n are indeterminates, then $F[x_1, x_2, \dots, x_n]$ is a UFD.*

Proof. Suppose D is a UFD and F is its field of quotients, then $F[x]$ is a UFD.

Suppose $K = F[x_1, x_2, \dots, x_k]$ is a UFD.

Then, $F[x_1, x_2, \dots, x_k, x_{k+1}] = F[x_1, x_2, \dots, x_k][x_{k+1}] = K[x_{k+1}] = K[x]$ is a UFD. \square

6.5.5 Exercise §45

1.

6.6 Euclidean Domain §46

Definitions 6.6.1 (Euclidean Norm). Euclidean norm on an integral domain D is a function v which maps non-zero elements of D into non-negative integers such that

1. $\forall a, b \in D, (b \neq 0)$ there exists $q, r \in D$ such that $a = qb + r$ where either $r = 0$ or $v(r) < v(b)$ and
2. $\forall a, b \in D, (a, b \neq 0) v(a) \leq v(ab)$

Definitions 6.6.2 (Euclidean Domain). An integral domain D is a Euclidean Domain if there exists a Euclidean Norm in D .

For example, $v(n) = |n|$ is a Euclidean norm on Euclidean domain \mathbb{Z} . And $v(f(x)) = \deg(f(x))$ is a Euclidean norm on Euclidean domain $F[x]$.

Theorem 6.6.1. *Every Euclidean domain is a PID.*

Proof. Let D be a Euclidean domain with Euclidean norm v . Let N be an ideal in D . If $N = \{0\}$, then $N = \langle 0 \rangle$ is a principal ideal.

Suppose $N \neq \{0\}$. Then there exists $b \in N$ ($b \neq 0$). Clearly, $b \in N \implies \langle b \rangle \subset N$. Choose a $b \in N$ such that $v(b)$ is minimal in N .

Claim : $N = \langle b \rangle$

Let $a \in N$. Then there exists $q, r \in D$ such that $a = qb + r$ where $r = 0$ or $v(r) < v(b)$. Clearly, $r = a - qb \in N$. Thus, $r = 0$ since $v(r) < v(b)$ and $v(b)$ is minimal in N . That is, $a = qb$ or $b|a$ for every $a \in N$. Thus, $N \subset \langle b \rangle$. Therefore, $N = \langle b \rangle$. In other words, every ideal in D is a principal ideal. \square

Corollary 6.6.1.1. *Every Euclidean domain is a UFD.*

Proof. We know that, every Euclidean domain is a PID. And every PID is a UFD. Thus, every Euclidean domain is a UFD. \square

6.6.1 Arithmetic in Euclidean Domain

Theorem 6.6.2. *For a Euclidean domain with Euclidean norm v , $v(1)$ is minimal and $u \in D$ is a unit if and only if $v(u) = v(1)$.*

Proof. We have, $v(a) \leq v(ab)$. Take $a = 1$, $v(1) \leq v(1b) \implies v(1) \leq v(b)$ for every $b \in D$. Thus, $v(1)$ is minimal.

Suppose u is a unit with multiplicative inverse u^{-1} . Clearly, $v(1) \leq v(u)$ since $u \in D$. Take $a = u$ and $b = u^{-1}$. Then $v(u) \leq v(uu^{-1}) = v(1)$. Therefore, for every unit $u \in D$, we have $v(u) = v(1)$. \square

Theorem 6.6.3 (division algorithm). *Let D be a Euclidean domain with Euclidean norm v . And $a, b \in D$ ($a, b \neq 0$).*

$$a = bq_1 + r_1, \text{ where } r_1 = 0 \text{ or } v(r_1) < v(b)$$

if $r_1 \neq 0$

$$b = r_1q + r_2, \text{ where } r_2 = 0 \text{ or } v(r_2) < v(r_1)$$

In general, if $r_i \neq 0$

$$r_{i-1} = r_iq + r_{i+1} \text{ where } r_{i+1} = 0 \text{ or } v(r_{i+1}) < v(r_i)$$

Then the sequence r_1, r_2, \dots must terminate with some $r_s = 0$.

If $r_1 = 0$, then $\gcd(a, b) = b$.

If $r_k = 0$, then $\gcd(a, b) = r_{k-1}$.

Furthermore, if d is a gcd of a and b , then there exists $\lambda, \mu \in D$ such that $d = \lambda a + \mu b$.

Proof. Step 1 : Finite Sequence r_1, r_2, \dots, r_s

Suppose $r_1, r_2, \dots, r_{i-1} \neq 0$. Then $v(r_1) > v(r_2) > \dots > v(r_{i-1})$ is a strictly decreasing sequence of non-negative integers. Thus, the sequence will terminate in finite numbers of steps. That is, $r_s = 0$ for some integer s .

Step 2 : $\gcd(a, b) = r_{s-1}$

Suppose $r_1 = 0$. Then $a = bq$ and $\gcd(a, b) = b$ since $b|a, b|b$ and there does not exist an integer greater than b that divides b .

Suppose $r_1 \neq 0$ and $\gcd(a, b) = d$. Then $r_1 = a - bq$ and $d|r_1$.

And $d|b, d|r_1 \implies d|a$ since $a = qb + r_1$. Thus, $\gcd(a, b) = \gcd(b, r_1)$.

Continuing like this, we get $\gcd(a, b) = \gcd(r_{s-2}, r_{s-1})$ if $r_{s-1} \neq 0$.

And we have, $r_s = 0 \implies r_{s-2} = qr_{s-1}$.

Clearly, $\gcd(r_{s-2}, r_{s-1}) = r_{s-1} = \gcd(a, b)$.

Step 3 : $\gcd(a, b) = \lambda a + \mu b$

Let b be a $\gcd(a, b)$. Then, $b = 0a + 1b$ and result is complete.

Suppose $\gcd(a, b) = r_{s-1}$. And $r_{s-1} = r_{s-3} - qr_{s-2}$. Clearly, $r_i = \lambda_i r_{i-2} + \mu_i r_{i-1}$. Keep on substituting using these equations until we reach $r_{s-1} = \lambda a + \mu b$.

Suppose d' is another $\gcd(a, b)$. Then $d' = ud$ is an associate of d . And $d' = ud = (\lambda u)a + (\mu u)b$. And the result is true for any $\gcd(a, b)$. \square

6.7 Gaussian Integers and Multiplicative Norms §47

6.7.1 Gaussian Integers and Norm

Definitions 6.7.1 (Gaussian integers). Gaussian integers are complex numbers of the form $a + ib$ where $a, b \in \mathbb{Z}$.

Note : Let $\alpha = a + ib$. Clearly, Gaussian integers are elements of $\mathbb{Z}[i]$. Here, $\mathbb{Z}[i]$ is a simple extension of the integral domain \mathbb{Z} with a zero of $x^2 + 1$.

Definitions 6.7.2 (Gaussian Norm). Let $\alpha = a + ib$ be a Gaussian integer. Then $N(\alpha) = |\alpha|^2 = a^2 + b^2$ is a Euclidean norm on $\mathbb{Z}[i]$.

Lemma 6.7.1. The Gaussian norm $N : \mathbb{Z}[i] \rightarrow \mathbb{Z}$ defined by $N(a + ib) = a^2 + b^2$ has the following properties

1. $N(\alpha) \geq 0, \forall \alpha \in \mathbb{Z}[i]$
2. $N(\alpha) = 0 \iff \alpha = 0$
3. $N(\alpha\beta) = N(\alpha)N(\beta)$

Proof. We have $\alpha \in \mathbb{Z}[i]$. Then $\alpha = a + ib$ where $a, b \in \mathbb{Z}$. Clearly, $N(a + ib) = a^2 + b^2 \geq 0$.

$$N(\alpha) = 0 \iff a^2 + b^2 = 0 \iff a = 0, b = 0 \iff \alpha = 0 + i0 = 0.$$

$$\text{Also, } N(\alpha\beta) = |\alpha\beta|^2 = |\alpha|^2|\beta|^2 = N(\alpha)N(\beta). \quad \square$$

Lemma 6.7.2. $\mathbb{Z}[i]$ is an integral domain.

Proof. Let $\alpha, \beta \in \mathbb{Z}[i]$.

Then, $\alpha + \beta = (a + ib) + (c + id) = (a + c) + i(b + d) = (c + a) + i(d + b) = \beta + \alpha$. And for any $\alpha \in \mathbb{Z}[i]$, we have $\alpha \cdot 1 = (a + ib) \cdot (1 + i0) = (a + ib)$. Thus, we have $\mathbb{Z}[i]$ is a commutative ring with unity.

It remains to prove that $\mathbb{Z}[i]$ has no zero divisors. Suppose $\alpha\beta = 0$. $\alpha\beta = 0 \implies N(\alpha\beta) = 0 \implies N(\alpha)N(\beta) = 0$. But, $N(\alpha), N(\beta) \in \mathbb{Z}$ and \mathbb{Z} has no zero divisors. Thus, $N(\alpha) = 0$ or $N(\beta) = 0 \implies \alpha = 0$ or $\beta = 0$. Therefore $\mathbb{Z}[i]$ has no zero divisors. \square

Theorem 6.7.3 (Gaussian integers is a Euclidean Domain). *The function v defined by $v(\alpha) = N(\alpha)$ for every non-zero $\alpha \in \mathbb{Z}[i]$ is a Euclidean norm.*

Proof. **Step 1 :** $v(a) \leq v(ab)$

Let $\beta = b_1 + ib_2$ and $\beta \neq 0$. Then $N(\beta) = b_1^2 + b_2^2 \geq 1$. Thus, $N(\alpha) \leq N(\alpha)N(\beta) = N(\alpha\beta)$.

Step 2 : Division algorithm

That is, $\forall \alpha, \beta \in \mathbb{Z}[i], (\beta \neq 0) \exists \sigma, \rho \in \mathbb{Z}[i]$ such that $\alpha = \beta\sigma + \rho$, and $\rho = 0$ or $v(\rho) < v(\beta)$.

Since $\beta \neq 0$, α/β exists. And we have, $\frac{\alpha}{\beta} = \frac{a_1 + ia_2}{b_1 + ib_2} = \frac{(a_1 + ia_2)(b_1 - ib_2)}{(b_1 + ib_2)(b_1 - ib_2)} = \frac{a_1b_1 + a_2b_2}{b_1^2 + b_2^2} + i \frac{a_2b_1 - a_1b_2}{b_1^2 + b_2^2} = r + is$ where $r, s \in \mathbb{Q}$. Consider integers q_1, q_2 that are nearest to the rational numbers r, s . Define $\sigma = q_1 + iq_2 \in \mathbb{Z}[i]$ and $\rho = \alpha - \beta\sigma$.

Suppose $\rho \neq 0$. If $\rho = 0$, then the proof is complete. By definition of σ , $|q_1 - r| \leq 0.5$ and $|q_2 - s| \leq 0.5$. And we have,

$$N\left(\frac{\alpha}{\beta} - \sigma\right) = N((r - q_1) + i(s - q_2)) = |q_1 - r|^2 + |q_2 - s|^2 \leq 0.5$$

Since $\alpha = \beta\sigma + \rho$, we have $\rho = \beta\left(\frac{\alpha}{\beta} - \sigma\right)$

$$N(\rho) = N\left(\beta\left(\frac{\alpha}{\beta} - \sigma\right)\right) = N(\beta)N\left(\frac{\alpha}{\beta} - \sigma\right) \leq 0.5N(\beta)$$

Thus, $N(\rho) < N(\beta)$. Therefore, Gaussian integers is a Euclidean domain. \square

6.7.2 Multiplicative Norm

Definitions 6.7.3 (multiplicative norm). Let D be an integral domain. A function $N : D \rightarrow \mathbb{Z}$ is a multiplicative norm if

1. $N(\alpha) = 0 \iff \alpha = 0$
2. $N(\alpha\beta) = N(\alpha)N(\beta), \forall \alpha, \beta \in D$

Note : Gaussian norm is a multiplicative norm.

Theorem 6.7.4. *Let D be an integral domain with multiplicative norm N . Then $N(1) = 1$ and $|N(u)| = 1$ for any unit u . Suppose $|N(\alpha)| = 1$ only for units in D . Then $N(\alpha)$ is a prime in \mathbb{Z} implies α is an irreducible in D .*

Proof. We know, N is a multiplicative norm. Thus, $N(1) = N(1 * 1) = N(1)N(1) \implies N(1) = 1$.

Suppose u is a unit in D . Then $1 = N(1) = N(uu^{-1}) = N(u)N(u^{-1})$. Since 1 has only two factors $\pm 1 \in \mathbb{Z}$, we have $|N(u)| = 1$.

Suppose $|N(\alpha)| = 1$ only if α is a unit in D . Suppose $\alpha \in D$ and $|N(\alpha)| = p$, where p is a prime in \mathbb{Z} . Suppose α is not an irreducible. Then α has a factorisation $\alpha = \beta\gamma$ where $\beta, \gamma \in D$ are non-units. Then $p = N(\alpha) = N(\beta\gamma) = N(\beta)N(\gamma)$. We have, β, γ are non-units and $|N(\beta)| \neq 1$ and $|N(\gamma)| \neq 1$. This is not possible. Therefore, α is an irreducible in D . \square

6.7.3 Applications of multiplicative norm

The integral domain $\mathbb{Z}[\sqrt{-5}]$ is not a UFD. (§47.9)

Proof. Elements of $\mathbb{Z}[\sqrt{-5}]$ are of the form $a + b\sqrt{-5}$ and $N(a + b\sqrt{-5}) = a^2 + 5b^2$. If $N(a + b\sqrt{-5}) = a^2 + 5b^2 = 1$, then, $a = \pm 1$ and $b = 0$. Clearly, $|N(\alpha)| = 1$ only if α is a unit in $\mathbb{Z}[\sqrt{-5}]$.

The element $3 \in \mathbb{Z}[\sqrt{-5}]$ is an irreducible. Suppose 3 is not an irreducible. That is, there is a factorisation $3 = \beta\gamma$ where β, γ are non-units. Then $9 = N(3) = N(\beta)N(\gamma)$ where β, γ are non-units. The integer factors of 9 are 1, 3, 9. Thus, $N(\beta) = N(\gamma) = 3$. Otherwise $N(\beta) = 1$ or $N(\gamma) = 1$. But, there doesn't exist an element $a + b\sqrt{-5}$ such that $N(a + b\sqrt{-5}) = 3$. Similarly, $7 \in \mathbb{Z}[\sqrt{-5}]$ is an irreducible. We have, $49 = N(7) = N(\beta)N(\gamma) \implies N(\beta) = N(\gamma) = 7$ is not possible since there are no elements of norm 7 in $\mathbb{Z}[\sqrt{-5}]$.

Suppose $(1 + 2\sqrt{-5})$ is not an irreducible. Then, $21 = N(1 + 2\sqrt{-5}) = N(\beta)N(\gamma)$. However, we know that the integer factors of 21 are 1, 3, 7, 21. Without loss of generality, we have $N(\beta) = 3$ which is not possible. Thus, $1 + 2\sqrt{-5}$ is an irreducible.

However neither 3 nor 7 is an associate of $1 + 2\sqrt{-5}$. Since $|N(1 + 2\sqrt{-5})| \neq N(3)$ and $N(1 + 2\sqrt{-5}) \neq N(7)$. Clearly, the factorisation of $21 \in \mathbb{Z}[\sqrt{-5}]$ is not unique. Therefore, $\mathbb{Z}[\sqrt{-5}]$ is not a UFD. \square

Fermat's $p = a^2 + b^2$ theorem

Theorem 6.7.5 (Fermat). *Let p be an odd prime integer. Then $p = a^2 + b^2$ where $a, b \in \mathbb{Z}$ if and only if $p \cong 1 \pmod{4}$*

Proof. Suppose there exists an odd prime p such that $p = a^2 + b^2$. Then a, b are of different parity. Without loss of generality, a is even and b is odd. Let $a = 2r$ and $b = 2s + 1$. Then $a^2 + b^2 = 4r^2 + 4s^2 + 1 \cong 1 \pmod{4}$.

Suppose $p \cong 1 \pmod{4}$. Then \mathbb{Z}_p is a cyclic group of order $p - 1$ which has a element n of order 4 since $4|(p - 1)$. Then $n^2 = -1 \in \mathbb{Z}_p$. And $n^2 + 1 \cong 0$

(mod p). In other words, $p|(n^2 + 1)$.

We claim that, p is not an irreducible in $\mathbb{Z}[i]$. Suppose p is an irreducible then $p|(n^2 + 1) \implies p|(n + i)$ or $p|(n - i)$. And $p|(n + i) \implies n + i = p(a + bi) \implies n = pa$ and $1 = pb$. But, $1 = pb$ is not possible. Similarly, $p|(n - i) \implies n - i = p(a + bi) \implies -1 = pb$ is also not possible. Thus if $p \cong 1 \pmod{4}$, then p is not an irreducible in $\mathbb{Z}[i]$.

Since p is not an irreducible in $\mathbb{Z}[i]$, $p = (a + bi)(c + di)$ where $a + bi, c + di$ are non-units in $\mathbb{Z}[i]$. Considering the Gaussian norm N on $\mathbb{Z}[i]$ defined by $N(a + bi) = a^2 + b^2$. We have, $p^2 = N(p) = N(a + bi)N(c + di) = (a^2 + b^2)(c^2 + d^2)$. And $N(a + bi) \neq 1$ since for Gaussian norm $N(\alpha) = 1 \iff \alpha = 1$. Similarly, $N(c + di) \neq 1$. Therefore, without loss of generality $N(a + bi) = p$ since $a^2 + b^2 > 0$. In other words, $p = a^2 + b^2$. \square

6.7.4 Exercise §47

1. We have 5 is an odd prime and $5 \cong 1 \pmod{4}$. Therefore 5 is not an irreducible Gaussian integer. Clearly, $2^2 + 1^2 = 5 \implies (2 + i)(2 - i) = 5$.

And $N(2 + i) = 5$ is a prime, thus $2 + i$ is an irreducible. Similarly, $2 - i$ is also an irreducible. Therefore $5 = (2 + i)(2 - i)$ is factorisation using irreducibles Gaussian integers. (hint : $2^2 + 1^2 = 5$. But, $1 * 1 + 2 * 2 = 5$)

Note : We have $5 = (2 + i)(2 - i) = (1 + 2i)(1 - 2i)$. However 5 has a unique factorisation since $(2 + i)$ and $(1 - 2i)$ are associates $(2 + i)(0 - i) = 1 - 2i$ where $-i \in \mathbb{Z}[i]$ is a unit with multiplicative inverse i .

2. We have 7 is an odd prime. However $7 \cong 3 \pmod{4}$. Therefore 7 is an irreducible Gaussian integer.
3. We have $N(4 + 3i) = 25$. Clearly, $(4 + 3i) = \alpha\beta \implies N(\alpha) = N(\beta) = 5$. Therefore, $4 + 3i = (2 - i)(1 + 2i)$ is a factorisation of $4 + 3i$. (Hint : $2^2 + 1^2 = 5$. But, $1 * 2 + 1 * 2 = 4$.)
4. We have $N(6 - 7i) = 85 = 5 * 17$. Clearly $6 - 7i = (2 + i)(1 - 4i)$. (Hint : $2^2 + 1^2 = 5$ and $4^2 + 1^2 = 17$. But $1 * 2 + 1 * 4 = 6$.)
5. We have $6 = 2 * 3$ in $\mathbb{Z}[\sqrt{-5}]$. Then, $N(\alpha\beta) = 4 \implies a^2 + 5b^2 = 2$ is not possible. And, $N(\alpha\beta) = 9 \implies a^2 + 5b^2 = 3$ is also not possible. Therefore, $6 = 2 * 3$ is a factorisation in $\mathbb{Z}[\sqrt{-5}]$.

However $N(\alpha\beta) = 6 \implies a^2 + 5b^2 = 6 \implies |a| = |b| = 1$. Clearly, $6 = (1 - \sqrt{-5})(1 + \sqrt{-5})$ is another factorisation.

Module 3 Field Automorphisms

6.8 Automorphism and Fields §48

Context

1. Let F be a field. Then there exists a unique field \bar{F} which is algebraically closed.
2. Let F be a field. Let α be algebraic over F . Then there exists a unique, monic, irreducible, minimal polynomial $\text{irr}(\alpha, F) = p(x) \in F[x]$ such that $p(\alpha) = 0$.
3. $F(\alpha)$ is a vector space of dimension $[F(\alpha) : F] = \deg(\alpha, F)$ over F .

6.8.1 Conjugation Isomorphisms

Definitions 6.8.1 (conjugates). Let E be an algebraic extension of F . Let $\alpha, \beta \in E$. Then α and β are **conjugates over F** if $\text{irr}(\alpha, F) = \text{irr}(\beta, F)$.

For example, $i, -i$ are conjugates over \mathbb{R} . $\text{irr}(i, \mathbb{R}) = x^2 + 1 = \text{irr}(-i, \mathbb{R})$.

Theorem 6.8.1 (conjugation isomorphism). Let F be a field. Let α, β be algebraic over F with $\deg(\alpha, F) = n$. The map $\psi_{\alpha, \beta} : F(\alpha) \rightarrow F(\beta)$ defined by

$$\psi_{\alpha, \beta}(c_0 + c_1\alpha + \cdots + c_{n-1}\alpha^{n-1}) = c_0 + c_1\beta + \cdots + c_{n-1}\beta^{n-1}$$

where $c_j \in F$ is an isomorphism from $F(\alpha)$ onto $F(\beta)$ if and only if α, β are conjugates over F .

Proof. Let $\psi_{\alpha, \beta} : F(\alpha) \rightarrow F(\beta)$ be an isomorphism. Let $\deg(\alpha, F) = n$ and $\text{irr}(\alpha, F) = q(x)$ given by

$$\text{irr}(\alpha, F) = q(x) = a_0 + a_1x + \cdots + a_nx^n$$

Then

$$a_0 + a_1\alpha + \cdots + a_n\alpha^n = 0$$

Thus,

$$\psi_{\alpha, \beta}(0) = \psi_{\alpha, \beta}(a_0 + a_1\alpha + \cdots + a_n\alpha^n) = a_0 + a_1\beta + \cdots + a_n\beta^n = 0$$

Let $\text{irr}(\beta, F) = p(x) \in F[x]$. Then, $p(\beta) = 0$. We already have, $q(\beta) = 0$. Let $\phi_\beta : F[x] \rightarrow F(\beta)$ be evaluation homomorphism. Then, $\phi_\beta(q(x)) = q(\beta)$. Therefore, $q(x) \in \ker(\phi_\beta)$. However, $\ker(\phi_\beta)$ is the ideal generated by $p(x)$. Therefore, $p(x)$ must divide $q(x)$ in $F[x]$. Thus, $\text{irr}(\beta, F)$ divides $\text{irr}(\alpha, F)$.

Similarly, $\psi_{\alpha, \beta}^{-1} : F(\beta) \rightarrow F(\alpha)$ is an isomorphism and $\text{irr}(\alpha, F)$ divides $\text{irr}(\beta, F)$. We know that, both the polynomials are monic polynomials. Therefore, $\text{irr}(\alpha, F) = \text{irr}(\beta, F)$. That is, α, β are conjugates over F .

Let $\text{irr}(\alpha, F) = \text{irr}(\beta, F) = p(x)$. Let $\phi_\alpha : F[x] \rightarrow F(\alpha)$ and $\phi_\beta : F[x] \rightarrow F(\beta)$ be evaluation homomorphisms. We know that for any polynomial $f(x) \in F[x]$ with $f(\alpha) = 0$, $p(x)$ divides $f(x)$. Thus, $\ker(\phi_\alpha) = \langle p(x) \rangle$. Similarly, $\ker(\phi_\beta) = \langle p(x) \rangle$.

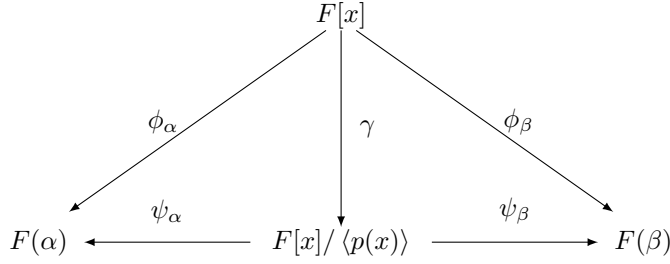


Figure 6.1: $\psi_{\alpha,\beta} = \psi_\beta \circ \psi_\alpha^{-1}$

By fundamental homomorphism theorem, there exists a canonical isomorphism

$$\psi_\alpha : F[x]/\langle p(x) \rangle \rightarrow \phi_\alpha[F[x]], \quad \psi_\alpha(r(x) + \langle p(x) \rangle) = \phi_\alpha(r(x)) = r(\alpha)$$

Similarly, there exists another canonical isomorphism

$$\psi_\beta : F[x]/\langle p(x) \rangle \rightarrow \phi_\beta[F[x]], \quad \psi_\beta(r(x) + \langle p(x) \rangle) = \phi_\beta(r(x)) = r(\beta)$$

Therefore, $\psi_{\alpha,\beta} = \psi_\beta \circ \psi_\alpha^{-1} : F(\alpha) \rightarrow F(\beta)$ is an isomorphism such that

$$\psi_\beta \circ \psi_\alpha^{-1}(c_0 + c_1\alpha + \cdots + c_{n-1}\alpha^{n-1}) = c_0 + c_1\beta + \cdots + c_{n-1}\beta^{n-1}$$

□

Challenge 7. No algebraic number over \mathbb{R} has multiple conjugates.(proved later) However, there exists fields over which there exists algebraic numbers with multiple conjugates.

Proof. Let $F = \mathbb{Z}_2$. Polynomial $x^3 + x + 1$ is an irreducible over \mathbb{Z}_2 since $\phi_0(x^3 + x + 1) = 0^3 + 0 + 1 \neq 0$ and $\phi_1(x^3 + x + 1) = 1^3 + 1 + 1 \neq 0$. Let α be a zero of $x^3 + x + 1$. Then α is algebraic over \mathbb{Z}_2 . And $\phi_\alpha(x^3 + x + 1) = \alpha^3 + \alpha + 1 = 0$. Then α has two conjugates β, γ .

We have, $[\mathbb{Z}_2(\alpha) : \mathbb{Z}_2] = 3$. Thus, $|\mathbb{Z}_2(\alpha)| = 8$. And $\mathbb{Z}_2(\alpha) = \{0, 1, \alpha, \alpha^2, \alpha + \alpha^2, 1 + \alpha, 1 + \alpha^2, 1 + \alpha + \alpha^2\}$. Clearly, $\alpha^3 = \alpha + 1$. We can test each element of $\mathbb{Z}_2(\alpha)$ to find other zeroes of the polynomial $x^3 + x + 1$. Thus, $\phi_{\alpha^2} = 0$ and $\phi_{\alpha+\alpha^2} = 0$. Let $\beta = \alpha^2$ and $\gamma = \alpha + \beta$.

Now the we have conjugate isomorphisms $\psi_{\alpha,\beta} : \mathbb{Z}_2(\alpha) \rightarrow \mathbb{Z}_2(\beta)$ and $\psi_{\alpha,\gamma} : \mathbb{Z}_2(\alpha) \rightarrow \mathbb{Z}_2(\gamma)$ defined by

Clearly, there are 6 conjugate isomorphisms on $\mathbb{Z}_2(\alpha)$. However, only two of them are distinct. □

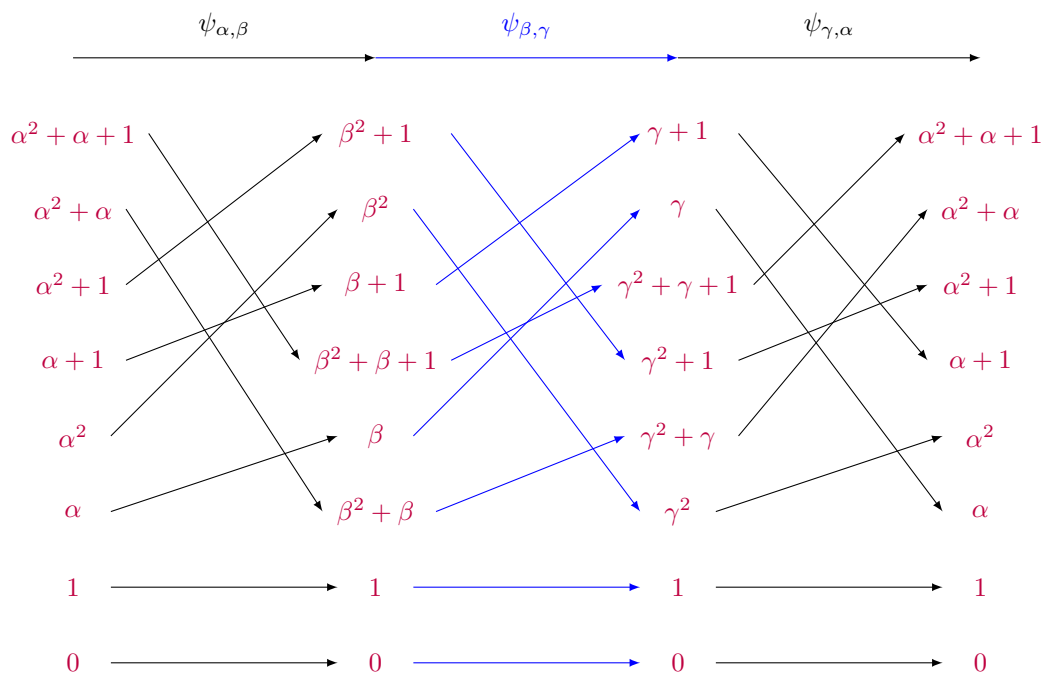


Figure 6.2: $\psi_{\alpha,\beta} : \mathbb{Z}_2(\alpha) \rightarrow \mathbb{Z}_2(\beta)$, $\psi_{\beta,\gamma} : \mathbb{Z}_2(\beta) \rightarrow \mathbb{Z}_2(\gamma)$, $\psi_{\gamma,\alpha} : \mathbb{Z}_2(\gamma) \rightarrow \mathbb{Z}_2(\alpha)$.

Corollary 6.8.1.1. *Let α be algebraic over F . Let $E \leq \bar{F}$. Then every isomorphism $\psi : F(\alpha) \rightarrow E$ with $\psi(a) = a$, $\forall a \in F$ maps α into its conjugate over F . And for any conjugate β of α over F , there exists a unique isomorphism $\psi_{\alpha,\beta} : F(\alpha) \rightarrow \bar{F}$ such that $\psi_{\alpha,\beta}(\alpha) = \beta$ and $\forall a \in F$, $\psi_{\alpha,\beta}(a) = a$.*

Proof. Let $\psi : F(\alpha) \rightarrow E$ be an isomorphism such that $\psi(a) = a$, $\forall a \in F$. Let $\text{irr}(\alpha, F) = p(x) = a_0 + a_1x + \cdots + a_nx^n$. Then $\phi_\alpha(p(x)) = p(\alpha) = a_0 + a_1\alpha + \cdots + a_n\alpha^n = 0$. And, $\psi(0) = \psi(p(\alpha)) = a_0 + a_1\psi(\alpha) + \cdots + a_n\psi(\alpha)^n = 0$. That is, $\psi(\alpha)$ is a zero of the polynomial $p(x)$. Thus, $\alpha, \psi(\alpha)$ are conjugates over F .

Let α, β be conjugates over F . Then, the conjugation isomorphism $\psi_{\alpha,\beta} : F(\alpha) \rightarrow F(\beta)$ defined by

$$\psi_{\alpha,\beta}(a_0 + a_1\alpha + \cdots + a_{n-1}\alpha^{n-1}) = a_0 + a_1\beta + \cdots + a_{n-1}\beta^{n-1}$$

is an isomorphism which maps $a \rightarrow a$, $\forall a \in F$ and $\alpha \rightarrow \beta$. Since every element in $F(\alpha)$ is of the form $a_0 + a_1\alpha + \cdots + a_{n-1}\alpha^{n-1}$. And the definition $a \rightarrow a$, $\alpha \rightarrow \beta$ completely determines the isomorphism. Therefore, there exists a unique conjugate isomorphism which maps α to its conjugate β . \square

Corollary 6.8.1.2. *Let $f(x) \in \mathbb{R}[x]$. Let $a + ib \in \mathbb{C}$. If $f(a + ib) = 0$, then $f(a - ib) = 0$. Also, complex zeroes of polynomials with real-coefficients occur in conjugate pairs.*

In other words, complex conjugates are conjugates over \mathbb{R} .

Proof. We have $\mathbb{C} = \mathbb{R}(i)$ where i is algebraic over \mathbb{R} , $\text{irr}(i, \mathbb{R}) = x^2 + 1$. Clearly, $i, -i$ are conjugates over \mathbb{R} . And $\psi_{i, -i} : \mathbb{C} \rightarrow \mathbb{C}$ is a conjugate isomorphism which maps $a + ib \rightarrow a - ib$.

Let $a + ib \in \mathbb{C}$. Let $f(x) \in \mathbb{R}[x]$ such that $\text{irr}(a + ib, \mathbb{R}) = f(x)$. Since $f(x) \in \mathbb{R}[x]$,

$$f(x) = c_0 + c_1x + \cdots + c_nx^n$$

Since $\text{irr}(a + ib, \mathbb{R}) = f(x)$,

$$\phi_{a+ib}(f(x)) = f(a + ib) = c_0 + c_1(a + ib) + \cdots + c_n(a + ib)^n = 0$$

Field \mathbb{R} is fixed under the conjugate isomorphism, $\psi_{i, -i}$. Thus, $\psi_{i, -i}(0) = 0$,

$$\psi_{i, -i}(0) = \psi_{i, -i}(f(a + ib)) = f(a - ib) = c_0 + c_1(a - ib) + \cdots + c_n(a - ib)^n = 0$$

Therefore, $(a - ib)$ is a zero of $f(x)$. Let $\text{irr}(a - ib, \mathbb{R}) = p(x)$. Then $f(x) \in \langle p(x) \rangle$. That is, $p(x)$ divides $f(x)$. Similarly, $\psi_{-i, i}(p(a - ib)) = p(a + ib) = 0$. And $f(x)$ divides $p(x)$. Therefore, $\text{irr}(a - ib, \mathbb{R}) = p(x) = f(x) = \text{irr}(a + ib, \mathbb{R})$. Clearly, $a + ib, a - ib$ are conjugates over \mathbb{R} .

Let $w = c + id \in \mathbb{C}$ be another conjugate of $z = a + ib$ where $a, b, c, d \in \mathbb{R}$. Suppose $\psi_{z, w} : \mathbb{C} \rightarrow \mathbb{C}$ is a conjugation isomorphism. If $a + ib$ is a zero of some $g(x) \in \mathbb{R}[x]$. Then $c + id$ is also a zero of $g(x)$ by conjugation isomorphism. We have, $\psi_{z, w}(a) = a, \psi_{z, w}(b) = b$ since $a, b \in \mathbb{R}$. Thus, $\psi_{z, w}(a + ib) = \psi_{z, w}(a) + \psi_{z, w}(i)\psi_{z, w}(b) = a + \psi_{z, w}(i)b = c + id$. Clearly, $\psi_{z, w}(i) = \frac{c-a}{b} + i\frac{d}{b}$. Thus, $a = c$ and $d = \pm b$ since $i, -i$ are the only zero of $x^2 + 1$. Therefore, conjugates over \mathbb{R} occur in pairs. \square

Remark. The notion of complex conjugates

1. Complex conjugates z, \bar{z} are conjugates over the real field \mathbb{R} with irreducible, monic polynomial $\text{irr}(z, \mathbb{R}) = (x - \Re(z))^2 + \Im(z)^2$.
2. Any non-real number, $z \notin \mathbb{R}$ which is algebraic over \mathbb{R} has a unique conjugate \bar{z} over \mathbb{R} .

Remark. $\sqrt{2}, -\sqrt{2}$ are conjugates over \mathbb{Q} with $\text{irr}(\sqrt{2}, \mathbb{Q}) = x^2 - 2$. Thus, $\psi_{\sqrt{2}, -\sqrt{2}} : \mathbb{Q}(\sqrt{2}) \rightarrow \mathbb{Q}(\sqrt{2})$ is a conjugation isomorphism. Elements of $\mathbb{Q}(\sqrt{2})$ are of the form $a + \sqrt{2}b$ where $a, b \in \mathbb{Q}$. Clearly, $\psi_{\sqrt{2}, -\sqrt{2}}(a + \sqrt{2}b) = a - \sqrt{2}b$.

Result : If $\sqrt{2}$ is a zero of a polynomial with rational coefficients, then $-\sqrt{2}$ is also a zero of it. Therefore, $\sqrt{2}$ is an eigen value of a rational matrix, if and only if $-\sqrt{2}$ is also an eigen value of it. Similarly, $a + ib$ is an eigen value of a real matrix, if and only if $a - ib$ is also an eigen value of it.

6.8.2 Automorphism and Fixed Fields

Definitions 6.8.2 (automorphism). Automorphism of a field is an isomorphism of a field onto itself.

Definitions 6.8.3 (fixed). Let $\sigma : F \rightarrow F$ be a field automorphism. The set $S = \{a \in F : \sigma(a) = a\}$ of elements invariant under σ is fixed under σ . Let E be a subfield of F such that $\forall a \in E, \sigma(a) = a$. Then E is left fixed by σ .

For example, the conjugation map $\psi_{\sqrt{3}, -\sqrt{3}} : \mathbb{Q}(\sqrt{2}, \sqrt{3}) \rightarrow \mathbb{Q}(\sqrt{2}, \sqrt{3})$ leaves $\mathbb{Q}(\sqrt{2})$ fixed.

We have, $\text{irr}(\sqrt{2}, \mathbb{Q}) = x^2 - 2$. Elements of $\mathbb{Q}(\sqrt{2})$ are of the form $a + b\sqrt{2}$ where $a, b \in \mathbb{Q}$.

$$\mathbb{Q}(\sqrt{2}, \sqrt{3}) = \mathbb{Q}(\sqrt{2})(\sqrt{3})$$

And $\text{irr}(\sqrt{3}, \mathbb{Q}(\sqrt{2})) = x^2 - 3$. Thus, elements of $\mathbb{Q}(\sqrt{2})(\sqrt{3})$ are of the form $x + y\sqrt{3}$ where $x, y \in \mathbb{Q}(\sqrt{2})$.

$$(a + b\sqrt{2}) + (c + d\sqrt{2})\sqrt{3} = a + b\sqrt{2} + c\sqrt{3} + d\sqrt{6}$$

Elements of $\mathbb{Q}(\sqrt{2}, \sqrt{3})$ are of the form $a + b\sqrt{2} + c\sqrt{3} + d\sqrt{6}$ where $a, b, c, d \in \mathbb{Q}$.

$$\psi_{\sqrt{3}, -\sqrt{3}}(a + b\sqrt{2} + c\sqrt{3} + d\sqrt{6}) = a + b\sqrt{2} - c\sqrt{3} - d\sqrt{2}\sqrt{3}$$

Thus, an element of $\mathbb{Q}(\sqrt{2}, \sqrt{3})$ is left fixed by $\psi_{\sqrt{3}, -\sqrt{3}}$ if and only if $c, d = 0$.

Theorem 6.8.2. Let $\{\sigma_i : i \in I\}$ be a collection of automorphisms of field E . Then, the $E_{\{\sigma_i\}} = \{a \in E : \sigma_i(a) = a, \forall i \in I\}$ is a subfield of E .

The set fixed by an automorphism of a field is a **subfield**.

Quick Revision : How to test whether F is a subfield of field E ?

1. Preservation of identities, $0, 1 \in F$ (usually trivial)
2. Closure of binary operations, $\forall a, b \in F, a \pm b, ab \in F$ (hard to prove)
3. Existence of inverses (easy to prove)

- (a) $\forall a \in F, -a \in F$
- (b) $\forall a \in F, (a \neq 0), a^{-1} \in F$.

Proof. Let $a, b \in E_{\{\sigma_i\}}$. Then $\sigma_i(a) = a, \forall i \in I$ and $\sigma_i(b) = b, \forall i \in I$. Since σ_i 's are homomorphisms on E ,

$$\sigma_i(a \pm b) = \sigma_i(a) \pm \sigma_i(b) = a \pm b$$

$$\sigma_i(ab) = \sigma_i(a)\sigma_i(b) = ab$$

Thus, $a, b \in E_{\{\sigma_i\}} \implies a \pm b, ab \in E_{\{\sigma_i\}}$.

We have, σ_i 's are automorphisms. Thus, $\sigma_i(0) = 0$ and $\sigma_i(1) = 1$ for every $i \in I$. We have, $\sigma_i(0) = 0$.

$$0 = \sigma_i(0) = \sigma_i(a + (-a)) = \sigma_i(a) + \sigma_i(-a) = a - \sigma_i(a)$$

Thus, $\sigma_i(-a) = -a, \forall i \in I$. And, $\forall a \in E_{\{\sigma_i\}}, -a \in E_{\{\sigma_i\}}$.

Similarly, $\sigma_i(1) = 1$. Let $a \neq 0$.

$$1 = \sigma_i(1) = \sigma_i(aa^{-1}) = \sigma_i(a)\sigma_i(a^{-1}) = a\sigma_i(a^{-1})$$

Thus, $\sigma_i(a^{-1}) = a^{-1}, \forall i \in I$. And, $\forall a \in E_{\{\sigma_i\}}, (a \neq 0), a^{-1} \in E_{\{\sigma_i\}}$. Therefore, $E_{\{\sigma_i\}}$ is a subfield of E . \square

Definitions 6.8.4. Let σ be an automorphism on field E . Then $E_{\{\sigma\}}$ is the subfield left fixed by σ .

For example, we have seen earlier that the subfield left fixed by the conjugation isomorphism

$$\psi_{\sqrt{3}, -\sqrt{3}} : \mathbb{Q}(\sqrt{2}, \sqrt{3}) \rightarrow \mathbb{Q}(\sqrt{2}, \sqrt{3})$$

is $\mathbb{Q}(\sqrt{2})$. Therefore,

$$\mathbb{Q}(\sqrt{2}, \sqrt{3})_{\{\psi_{\sqrt{3}, -\sqrt{3}}\}} = \mathbb{Q}(\sqrt{2})$$

Here, $E = \mathbb{Q}(\sqrt{2}, \sqrt{3})$ and $\sigma = \psi_{\sqrt{3}, -\sqrt{3}}$ and $E_{\{\sigma\}} = \mathbb{Q}(\sqrt{2})$.

Theorem 6.8.3. The set of all automorphisms of a field E is a group under function composition.

Proof. Let E be a field. Let S_E be the set of all permutations on the field E . Clearly, the field permutations are associative under function composition.

We have, identity permutation $i : E \rightarrow E$ is an automorphism of E . Let σ be any automorphisms of E . Then, we have

$$i \circ \sigma = \sigma = \sigma \circ i$$

We have σ is a permutation on E . Let σ^{-1} be the inverse permutation of σ . Clearly, $\sigma^{-1} : E \rightarrow E$ is also an automorphism. And,

$$\sigma \circ \sigma^{-1} = i = \sigma^{-1} \circ \sigma$$

Let τ be an automorphism on E . Then $\sigma \circ \tau$ is an automorphism on E . Thus, set of all automorphisms on E is a subgroup of permutation group S_E . \square

Theorem 6.8.4. Let E be a field and F be a subfield of E . Then the set of all automorphisms of E leaving F fixed, say $G(E/F)$ is a subgroup of the set of all automorphisms on E . Furthermore, $F \leq E_{G(E/F)}$.

Proof. Let $\sigma, \tau \in G(E/F)$. We have $\sigma(a) = a$ and $\tau(a) = a$ for every $a \in F$. Thus, $(\sigma \circ \tau)(a) = \sigma(\tau(a)) = \sigma(a) = a$. Therefore, $\sigma \circ \tau \in G(E/F)$.

We have, $\sigma(a) = a \implies a = \sigma^{-1}(a)$, $\forall a \in F$. Thus, $\sigma^{-1} \in G(E/F)$. Therefore, $G(E/F)$ is a subgroup of all automorphisms on E .

The subfield $E_{G(E/F)}$ is the set of all automorphisms on E leaving F fixed. Let $\sigma \in G(E/F)$. Then,

$$a \in F \implies \sigma(a) = a \implies a \in E_{\{\sigma\}} \implies a \in E_{G(E/F)}$$

Thus, subfield F is contained in $E_{G(E/F)}$. \square

Definitions 6.8.5. Let F be a subfield of a field E . The group of E over F , $G(E/F)$ is the set of all automorphisms on E leaving F fixed.

Any automorphism on field E leaves the prime field \mathbb{Q} or \mathbb{Z}_p fixed.

For example, consider the set of all automorphisms on $\mathbb{Q}(\sqrt{2}, \sqrt{3})$ (leaving \mathbb{Q} fixed.) We have,

$$i, \quad \psi_{\sqrt{2}, -\sqrt{2}}, \quad \psi_{\sqrt{3}, -\sqrt{3}}, \quad \psi_{\sqrt{2}, -\sqrt{2}} \circ \psi_{\sqrt{3}, -\sqrt{3}}$$

are the only automorphisms on $\mathbb{Q}(\sqrt{2}, \sqrt{3})$. Therefore, we have

$$G\left(\mathbb{Q}(\sqrt{2}, \sqrt{3})/\mathbb{Q}\right) \simeq V \simeq \mathbb{Z}_2 \times \mathbb{Z}_2$$

since each automorphism is an inverse of itself. Also we have, $[\mathbb{Q}(\sqrt{2}, \sqrt{3}) : \mathbb{Q}] = 4$ is the order of the group $G(\mathbb{Q}(\sqrt{2}, \sqrt{3})/\mathbb{Q})$.

Earlier we have seen, $\mathbb{Z}_2(\alpha)$ where $\text{irr}(\alpha, \mathbb{Z}_2) = x^3 + x + 1$. The conjugates of α over \mathbb{Z}_2 are β, γ . However, the distinct automorphisms are,

$$i, \quad \psi_{\alpha, \beta}, \quad \psi_{\beta, \alpha}$$

Therefore,

$$G(\mathbb{Z}_2(\alpha), \mathbb{Z}_2) \simeq \mathbb{Z}_3 \quad \text{since } \psi_{\alpha, \beta} \circ \psi_{\beta, \alpha} = i$$

Once again, $[\mathbb{Z}_2(\alpha) : \mathbb{Z}_2] = 3 = |G(\mathbb{Z}_2(\alpha)/\mathbb{Z}_2)|$ makes our task much easier.

6.8.3 Frobenius Automorphism

Theorem 6.8.5 (Frobenius Automorphism). *Let F be a finite field of characteristic p . Then the map $\sigma_p : F \rightarrow F$ defined by $\sigma_p(a) = a^p$, $\forall a \in F$ is an automorphism on F . Also $F_{\{\sigma_p\}} \simeq \mathbb{Z}_p$.*

Proof. Let $a, b \in F$. Then

$$\sigma_p(a + b) = (a + b)^p = a^p + b^p = \sigma_p(a) + \sigma_p(b)$$

$$\sigma_p(ab) = (ab)^p = a^p b^p = \sigma_p(a) \sigma_p(b)$$

Thus, σ_p is a homomorphism.

We have,

$$\sigma_p(a) = 0 \implies a^p = 0 \implies a = 0$$

Thus, $\ker(\sigma_p) = \{0\}$. Therefore, σ_p is injective.

Since F is finite and $\sigma_p : F \rightarrow F$, σ_p is surjective. Therefore, σ_p is an automorphism on F .

We have, F is of characteristic p . And $\sigma_p(1) = 1^p = 1$. Let \mathbb{Z}_p be the prime field contained in F . Then, $\forall c \in \mathbb{Z}_p$, $\sigma_p(c) = c^p = c$. Therefore, \mathbb{Z}_p is fixed by σ_p .

The polynomial $x^p - x$ has at most p zeroes. Thus, it doesn't have any zero other than elements of the prime subfield of F .

$$a \in F_{\{\sigma_p\}} \iff \sigma_p(a) = a \iff a^p = a \iff a^p - a = 0 \iff \phi_a(x^p - x) = 0$$

Therefore, $F_{\{\sigma_p\}} \simeq \mathbb{Z}_p$. □

6.8.4 Exercises §48

1.

6.9 The Isomorphism Extension Theorem §49

6.9.1 Extension Theorem

Theorem 6.9.1 (isomorphism extension). *Let E be an algebraic extension of field F . Let σ be an isomorphism on F onto F' . Let \bar{F}' be an algebraic closure of F' . Then σ can be extended to an isomorphism τ of E onto \bar{F}' such that $\tau(a) = \sigma(a)$, $\forall a \in F$.*

Proof. **Step 1 : Poset S**

Let E be an algebraic extension of F . Let $\sigma : F \rightarrow F'$ be an isomorphism. Let S be the set of all pairs (L, λ) such that $F \leq L \leq E$ and λ is an extension of σ onto a subfield of \bar{F} . S is non-empty, since $(F, \sigma) \in S$.

Consider the relation \leq on S defined by $(L_1, \lambda_1) \leq (L_2, \lambda_2)$ if $L_1 \leq L_2$ and $\lambda_1(a) = \lambda_2(a)$, $\forall a \in L_1$. Then \leq is a partial order on S since \leq is reflexive, antisymmetric and transitive.

1. Reflexive

Let $(L, \lambda) \in S$. Then $L \leq L$ and $\lambda(a) = \lambda(a)$, $\forall a \in L$.

2. Anti-symmetric

Let $(L_1, \lambda_1), (L_2, \lambda_2) \in S$. Suppose $(L_1, \lambda_1) \leq (L_2, \lambda_2)$ where $L_1 \not\leq L_2$. Then, $F \leq L_1 < L_2 \leq E$. Thus, $(L_2, \lambda_2) \not\leq (L_1, \lambda_1)$.

3. Transitive

Let $(L_1, \lambda_1) \leq (L_2, \lambda_2)$ and $(L_2, \lambda_2) \leq (L_3, \lambda_3)$. Then, $F \leq L_1 \leq L_2 \leq L_3 \leq E$. And $\lambda_3(a) = \lambda_2(a) = \lambda_1(a)$, $\forall a \in L_1 \subseteq L_2$.

Thus, $(L_1, \lambda_1) \leq (L_3, \lambda_3)$.

Step 2 : Every chain in S has an upper bound (H, λ)

Let $\{(H_i, \lambda_i) : i \in I\}$ be a chain in S . Define $H = \cup\{H_i : i \in I\}$. We know that, if $c \in H$, then $c \in H_i$ for some $i \in I$. Define λ from H onto a subfield of \bar{F}' such that $\lambda(c) = \lambda_i(c)$ where $c \in H_i$.

We need to prove that λ is an isomorphism. Let $a, b \in H$. Then $a \in H_i$ and $b \in H_j$ for some $i, j \in I$. Then $H_i \leq H_j$ or $H_j \leq H_i$ since H_i, H_j belongs to a chain. Suppose $H_j \leq H_i$, then $a, b \in H_i$. Thus, $a + b, ab \in H_i$. And, λ_i is an isomorphism on H_i . Thus,

$$\lambda(a + b) = \lambda_i(a + b) = \lambda_i(a) + \lambda_i(b) = \lambda(a) + \lambda(b)$$

Similarly,

$$\lambda(ab) = \lambda_i(ab) = \lambda_i(a)\lambda_i(b) = \lambda(a)\lambda(b)$$

If $\lambda(c) = 0$, then $\lambda_i(c) = 0$ for some $i \in I$. Since λ_i is an isomorphism on H_i , $c = 0$. Therefore, $\lambda(c) = 0 \iff c = 0$. Clearly, λ is an isomorphism.

From the definition of H and λ , (H, λ) is an upper bounded for the chain since $H_i \leq H$ and $\lambda(c) = \lambda_i(c)$, $\forall c \in H_i$.

Step 3 : (K, τ) is maximal, but $K \neq E$

We have shown that, every chain in S has an upper bounded. By Zorn's lemma, S must have a maximal element, say (K, τ) .

Suppose $K \neq E$. Let $\alpha \in E$ such that $\alpha \notin K$. Then α is algebraic over K with $p(x) = \text{irr}(\alpha, K)$. Let ψ_α be the canonical isomorphism corresponding to the evaluation homomorphism ϕ_α .

$$\psi_\alpha : K[x] / \langle p(x) \rangle \rightarrow K(\alpha), \quad \phi_\alpha : K[x] \rightarrow K(\alpha)$$

Consider the isomorphism $\tau : K \rightarrow K'$. Let $p(x) = a_0 + a_1x + \cdots + a_nx^n$. Then, $q(x) = \tau(p)(x) = \tau(a_0) + \tau(a_1)x + \cdots + \tau(a_n)x^n$ is an irreducible polynomial over K' .

Let α' be a zero of $q(x) \in K'[x]$. Consider $\psi_{\alpha'} : K'[x] / \langle q(x) \rangle \rightarrow K'(\alpha')$ which is analogous to ψ_α . Let $\bar{\tau} : K[x] / \langle p(x) \rangle \rightarrow K'[x] / \langle q(x) \rangle$ be an extension of τ such that $x + \langle p(x) \rangle \rightarrow x + \langle q(x) \rangle$. Then $\psi_{\alpha'} \circ \bar{\tau} \circ \psi_\alpha^{-1} : K(\alpha) \rightarrow K'(\alpha')$ is an isomorphism. Clearly, $(K, \tau) \leq (K(\alpha), \psi_{\alpha'} \circ \bar{\tau} \circ \psi_\alpha^{-1})$ which contradicts (K, τ) being a maximal element of S . Therefore, $K = E$ and $\tau : K \rightarrow K'$ is an extension of $\sigma : F \rightarrow F'$ where $F' \leq K' \leq \bar{F}'$. \square

Corollary 6.9.1.1. *If $E \leq \bar{F}$ is an algebraic extension of F and $\alpha, \beta \in E$ are conjugates over F . Then the conjugate isomorphism $\psi_{\alpha, \beta} : F(\alpha) \rightarrow F(\beta)$ can be extended to an isomorphism of E onto \bar{F} .*

Proof. Consider the conjugate isomorphism $\psi_{\alpha, \beta} : F(\alpha) \rightarrow F(\beta)$. Let E be an algebraic extension of $F(\alpha)$. Then, by isomorphism extension theorem $\psi_{\alpha, \beta}$ can be extended to an isomorphism from E onto a subfield of \bar{F}' . \square

Corollary 6.9.1.2. *Let \bar{F} and \bar{F}' be two algebraic closure of F . Then \bar{F} is isomorphic to \bar{F}' under an isomorphism leaving each element of F fixed.*

Proof. Let $i : F \rightarrow F$ be the identity isomorphism. Let \bar{F}, \bar{F}' , both be algebraic closures of F . By isomorphism extension theorem, i can be extended to an isomorphism τ from the algebraic extension \bar{F} onto a subfield of \bar{F}' leaving F fixed since $i(a) = a$, $\forall a \in F$.

Since, τ is an isomorphism the range of τ , $\tau[\bar{F}]$ is a subfield of \bar{F}' . We have, $\tau^{-1} : \tau[\bar{F}] \rightarrow \bar{F}$ is also an isomorphism onto \bar{F} . Again by isomorphism extension theorem, τ can be extended to an algebraic extension \bar{F}' onto \bar{F} . However, τ^{-1} is already onto \bar{F} . Thus, $\tau : \bar{F} \rightarrow \bar{F}'$ is an isomorphism leaving F fixed. \square

In other words, for any field F , its algebraic closure \bar{F} is unique upto an isomorphism leaving F fixed.

6.9.2 Index of field extension

Theorem 6.9.2. *Let E be a finite extension of F . Let σ be an isomorphism of F onto F' . Let \bar{F}' be an algebraic closure of F' . Then, the number of extensions*

of σ to an isomorphism of E onto a subfield of \bar{F}' is finite and is independent of F', \bar{F}', σ .

In other words, the number of extensions depends only on F and E .

Proof. Step 1 : Index of extension of F is independent of F_1

Let $\sigma_1 : F \rightarrow F_1$ and $\sigma_2 : F \rightarrow F_2$ be two isomorphisms. Let E be an algebraic extension of F . And $\tau_1 : E \rightarrow \tau_1[E]$ and $\tau_2 : E \rightarrow \tau_2[E]$ be the extensions of σ_1, σ_2 from E to subfields $\tau_1[E] \leq \bar{F}_1$ and $\tau_2[E] \leq \bar{F}_2$ respectively. Then $\tau_2 \circ \tau_1^{-1} : \tau_1[E] \rightarrow \tau_2[E]$ is an isomorphism.

Let λ be the extension of $\tau_2 \circ \tau_1^{-1}$ where $\lambda : \bar{F}_1 \rightarrow \bar{F}_2$. Now, we have $\lambda : \bar{F}_1 \rightarrow \bar{F}_2$ such that for any isomorphism extension τ_1 from E onto a subfield of \bar{F}_1 , we have an isomorphism extension τ_2 from E on to \bar{F}_2 is given by $\tau_2 = \lambda\tau_1$. Thus, the number of extensions of $\sigma : F \rightarrow F'$ is independent of the choice of σ and F' (and \bar{F}').

Step 2 : Number of possible extensions is finite

Let $\sigma : F \rightarrow F'$ be an isomorphism. Let E be an algebraic extension of F . Let τ be an isomorphism extension of σ to E leaving F fixed. Then, $E = F(\alpha_1, \alpha_2, \dots, \alpha_n)$ since every algebraic extensions is a finite extension.

We know that, isomorphisms preserve irreducibility. ^{†4}
Let $\text{irr}(\alpha_j, F) = p(x) \in F[x]$ with $\deg(\alpha_j, F) = k$. Then,

$$p(x) = a_0 + a_1x + \dots + a_kx^k$$

We have, $\phi_{\alpha_j}(p(x)) = p(\alpha_j) = 0$. Let $q(x) \in F'[x]$ such that

$$q(x) = \sigma_x(p)(x) = \sigma(a_0) + \sigma(a_1)x + \dots + \sigma(a_k)x^k$$

Then, τ must map α_j to a zero of $q(x)$, say α'_j such that $\sigma(p(\alpha_j)) = 0 = q(\alpha'_j)$.

That is, for any isomorphism extension $\tau : E \rightarrow \bar{F}'$, we have $\tau(\alpha_j)$ is a zero of $q(x) \in F'[x]$. Clearly, any extension of σ will have to map α_j to a zero of $q(x) \in F'[x]$. However, $q(x)$ has at most k zeroes in \bar{F}' .

Therefore, the number of extension of σ is finite as E is a finite extension (number of α_j s is finite) and for each α_j the number of possible candidates $\tau(\alpha_j)$ in \bar{F}' is finite. Thus, the number of possible extensions of an isomorphism σ to an algebraic extension E leaving F fixed, is finite. \square

Definitions 6.9.1 (index). Let E be an extension of F . The index of E over F , $\{E : F\}$ is the number of isomorphisms of E onto a subfield of \bar{F} leaving F fixed.

Corollary 6.9.2.1. If $F \leq E \leq K$, where K is a finite extension of F . Then $\{K : F\} = \{K : E\}\{E : F\}$.

^{†4}Let $p(x)$ be irreducible over F . Let $\sigma : F \rightarrow F'$ be an isomorphism. Then there exists a polynomial $q(x)$ which is irreducible over F' . And if $p(x) = a_0 + a_1x + \dots + a_nx^n$, then $q(x) = \sigma(a_0) + \sigma(a_1)x + \dots + \sigma(a_n)x^n$. We may write $q = \sigma_x(p)$ where $\sigma_x : F[x] \rightarrow F'[x]$.

Proof. Let E be an algebraic extension of F and K be an algebraic extension of E . Then K is an algebraic extension of F .

We know that, the number of isomorphisms from E onto a subfield of \bar{F} leaving F fixed, $\{E : F\}$ is finite. Now, each of those isomorphisms have $\{K : E\}$ extensions, since there are $\{K : E\}$ different isomorphisms from K onto a subfield of \bar{F} leaving E fixed. Therefore, there are $\{K : E\}\{E : F\}$ isomorphisms from K onto a subfield of \bar{F} leaving F fixed. That is,

$$\{K : F\} = \{K : E\}\{E : F\}$$

□

For example, $\mathbb{Q}(\sqrt{2}, \sqrt{3})$ has 4 different isomorphisms onto a subfield of \mathbb{C} .

$$\{\mathbb{Q}(\sqrt{2}, \sqrt{3}) : \mathbb{Q}\} = \{\mathbb{Q}(\sqrt{2})(\sqrt{3}) : \mathbb{Q}(\sqrt{2})\} \{\mathbb{Q}(\sqrt{2}) : \mathbb{Q}\}$$

We have, $\mathbb{Q}(\sqrt{2}, \sqrt{3}) = \mathbb{Q}(\sqrt{2})(\sqrt{3})$ and

$$\{\mathbb{Q}(\sqrt{2})(\sqrt{3}) : \mathbb{Q}(\sqrt{2})\} = [\mathbb{Q}(\sqrt{2})(\sqrt{3}) : \mathbb{Q}(\sqrt{2})] = \deg(\sqrt{3}, \mathbb{Q}(\sqrt{2})) = 2$$

$$\{\mathbb{Q}(\sqrt{2}) : \mathbb{Q}\} = [\mathbb{Q}(\sqrt{2}) : \mathbb{Q}] = \deg(\sqrt{2}, \mathbb{Q}) = 2$$

6.9.3 Exercise §49

1.

6.10 Splitting Fields §50

Definitions 6.10.1 (splitting field). Let F be a field with algebraic closure \bar{F} . Let $\mathcal{F} = \{f_i(x) \in F[x] : i \in I\}$ be a collection of polynomials over F . Then $E \leq \bar{F}$ is a splitting field of \mathcal{F} over F if E is the smallest subfield of \bar{F} containing F and all the zeroes of polynomials in \mathcal{F} .

For example, $\mathbb{Q}(\sqrt{2}, \sqrt{3})$ is the splitting field of $\{x^2 - 2, x^2 - 3\}$ over \mathbb{Q} .

Theorem 6.10.1. Let $F \leq E \leq \bar{F}$. Then E is a splitting field of F if and only if every automorphism of \bar{F} leaving F fixed maps E onto itself. And thus induces an automorphism of E leaving F fixed.

Proof. Let E be a splitting field of F in \bar{F} of $\{f_i(x) : i \in I\}$. Let σ be an automorphism of \bar{F} leaving F fixed.

Step 1 : $\{\alpha_j : j \in J\}$ generates E

Let $\{\alpha_j : j \in J\}$ be the set of all zeroes of $\mathcal{F} = \{f_i(x) : i \in I\}$. We know that, the elements of $F(\alpha_j)$ are of the form

$$\phi_{\alpha_j} g(x) = g(\alpha_j) = a_0 + a_1 \alpha_j + \cdots + a_{n_j-1} \alpha_j^{n_j-1}$$

where $a_k \in F$, $\deg(\alpha_j, F) = n_j$. Clearly, $g(x)$ is some polynomial in $F[x]$ of degree less than n_j .

Let S be the set of all finite sums of finite products of elements of the form $g(\alpha_j)$. Clearly, $S \subset E$ and S is closed under addition, closed under multiplication, contains identities ($0, 1 \in S$) and additive inverses of its elements.

Let $\beta \in S$. Then β is a finite sum of finite products of elements of the form $g(\alpha_{j_k})$. Then, an element $\beta \in S$ belongs to some $F(\alpha_{j_1}, \alpha_{j_2}, \dots, \alpha_{j_r}) \subset S$. Thus, S has multiplicative inverse of each of its non-zero elements. Clearly, S is a subfield of E containing F and α_j s. That is, S is the splitting field of F containing all zeroes of \mathcal{F} . Therefore, $S = E$. In other words, $\{\alpha_j : j \in J\}$ generates E over F . ^{†5}

Step 2 : σ is an automorphism of E

Since $\{\alpha_j : j \in J\}$ spans E over F , $\sigma : \bar{F} \rightarrow \bar{F}$ is completely determined on E by the images of α_j s, say $\sigma(\alpha_j)$. And $\sigma(\alpha_j)$ is a zero of $\text{irr}(\alpha_j, F)$. If $f_i(\alpha_j) = 0$, then $\text{irr}(\alpha_j, F)$ divides $f_i(x)$. Similarly $\text{irr}(\sigma(\alpha_j), F)$ divides $f_i(x)$. Thus, σ maps E onto a subfield of E isomorphically. Similarly, σ^{-1} maps E onto a subfield of E isomorphically. Therefore, σ is an automorphism on E .

Converse Part

□

Definitions 6.10.2. Let E be an extension field of F . A polynomial $f(x) \in F[x]$ splits in $E[x]$ if it factors into linear components in $E[x]$.

For example, $x^4 - 5x^2 + 6 \in \mathbb{Q}$ splits in $\mathbb{Q}(\sqrt{2}, \sqrt{3})$ since

$$x^4 - 5x^2 + 6 = (x - \sqrt{2})(x + \sqrt{2})(x - \sqrt{3})(x + \sqrt{3})$$

Corollary 6.10.1.1. If $E \leq \bar{F}$ is a splitting field over F , then every irreducible polynomial in $F[x]$ having a zero in E splits in E .

Proof.

□

Corollary 6.10.1.2. If $E \leq \bar{F}$ is a splitting field over F , then every isomorphic mapping of E onto a subfield of \bar{F} leaving F fixed is an automorphism of E . In particular, if E is splitting field of finite degree over F , then

$$\{E : F\} = |G(E/F)|$$

Proof.

□

6.10.1 Exercise §50

1.

⁵ E is a vector space over the field F spanned by $\{\alpha_j : j \in J\}$. Let $\beta \in E$. Then, $\beta = a_0\alpha_0 + a_1\alpha_1 + \dots + a_r\alpha_r$ where $a_k \in F$.

Module 4 Galois Theory

6.11 Separable Extensions §51

6.11.1 Multiplicities of zeroes of a polynomial

6.11.2 Separable extensions

6.11.3 Perfect fields

6.11.4 The primitive element theorem

6.11.5 Exercise §51

1.

6.12 Galois Theory §53

6.12.1 Context

6.12.2 Normal extensions

6.12.3 The main theorem

6.12.4 Galois groups over finite fields

6.12.5 Exercise §53

1.

6.13 Illustrations of Galois Theory §54

6.13.1 Symmetric functions

6.13.2 Exercise §54

1.

6.14 Cyclotomic Extensions §55

6.14.1 The Galois group of a cyclotomic extension

6.14.2 Constructible polygons

6.14.3 Exercise §54

1.

6.15 **Insolvability of the quintic**

There exists polynomials of degree 5 in $\mathbb{Q}[x]$ that are not solvable by radicals in \mathbb{Q} .

Subject 7

ME010202 Advanced Topology

7.1 Module I

Q : Why these axioms are called separation axioms ?

T_0, T_1, T_2 axioms separates points from points.

$T_3, T_{3\frac{1}{2}}$ axioms separates points from closed subsets.

T_4 axiom separates closed subsets from closed subsets.



Figure 7.1: Separation Axioms

7.1.1 Compactness and Separation Axioms

[Joshi, 1983, chapter 7 §2.1-10]

Proposition 7.1.1. *Let X be a T_2 space, $x \in X$ and F is a compact subset of X not containing x . Then there exist open subsets U, V such that $x \in U, F \subset V, U \cap V = \emptyset$.*

T_2 spaces separates points from compact sets.

Proof. Let x be a points in X and F be a compact subset not containing x .
 X is $T_2 \implies \forall y \in F, \exists U_y, V_y \in \mathcal{T}, x \in U_y, y \in V_y, U_y \cap V_y = \emptyset$.^{†1)}
 $\mathcal{C} = \{V_y : y \in F\}$ is an open cover of compact subset F .

^{†1}**Q :** Do we need compactness for this result ?

A : Given $x \in X$. For each point $y \in F$, we have two open sets namely U_y and V_y that separates these two points x and y in X . But, the intersection $\bigcap_{y \in F} \{U_y\}$ is not necessarily an

\implies there exists a finite subcover, $\mathcal{C}' = \{V_{y_i} : y_i \in F, i = 1, 2, \dots, n\}$.

Define $U = \cap_{i=1}^n \{U_{y_i}\}$ and $V = \cup_{i=1}^n \{V_{y_i}\}$.

$\implies x \in U, F \subset V, U \cap V = \phi$. \square

Corollary 7.1.0.1. *A compact subset in a T_2 space is closed.*

Proof. Let $x \in X - F$.

By proposition, $x \in U, F \subset V, U \cap V = \phi \implies x \in U \subset X - V \subset X - F$

$X - F$ is a nbd of each of its points. Thus $X - F$ is open. (\star^2) \square

Corollary 7.1.0.2. *Every map from a compact space into a T_2 space is closed.*

And its range is a quotient space of the domain.

Proof. Suppose X is compact, Y is T_2 and $f : X \rightarrow Y$ is continuous.

Let C be a closed subset of X .

$\implies C$ is compact, since compact is weakly hereditary. $\star^3)$

$\implies f(C)$ is compact, since compactness is preserved by continuous functions.

(\star^4)

By corollary, $f(C)$ is compact, Y is $T_2 \implies f(C)$ is a closed subset of Y .

Thus $f : X \rightarrow Y$ is closed. $\dagger^5)$

$f : X \rightarrow Y$ is closed $\implies f : X \rightarrow f(X)$ is a quotient map, since every closed, surjective map is a quotient map.. \square

Corollary 7.1.0.3. *A continuous bijection from a compact space onto a T_2 space is a homeomorphism.*

Proof. Let G be an open subset of X .

Then $X - G$ is closed.

By corollary, f is closed, since $f : \text{compact} \rightarrow T_2$.

f is closed $\implies f(X - G)$ is closed.

$f(X - G) = f(X) - f(G)$ since f is injective.

$\implies f(X - G) = Y - f(G)$ since f is surjective.

$\implies f(G)$ is open.

Thus f is an open map.

Every continuous bijective, open map is a homeomorphism. \square

Corollary 7.1.0.4. *Every continuous, one-to-one function from a compact space into a T_2 space is an embedding.*

Proof. Let $f : X \rightarrow Y$ be continuous and injective.

$f : X \rightarrow f(X)$ is surjective.

$\implies f : X \rightarrow f(X)$ is continuous, surjective.

By corollary, $f : X \rightarrow f(X)$ is a homeomorphism.

Thus f is an embedding of X onto $f(X) \subset Y$. \square

open subset of X , since F is not necessarily a finite subset of X and an **arbitrary** intersection of open subsets is not necessarily an open subset. Therefore, we have restricted this family into a finite family. We use compactness for this part.

²Neighbourhood characterisation of open subsets : Let $G \subset X$ be a nbd of each of its points. Then G is an open subset of X .

³Compactness is weakly hereditary : Suppose X is compact. Then every closed subset of X is compact.

⁴Compactness is preserved by continuous functions : Suppose G be a compact subset of a topological space X . And function $f : X \rightarrow Y$ be a continuous function. Then $f(G)$ is a compact subset of the topological space Y . [Joshi, 1983, 6.1.8]

⁵Closed function is a function which maps closed subsets into closed subsets.

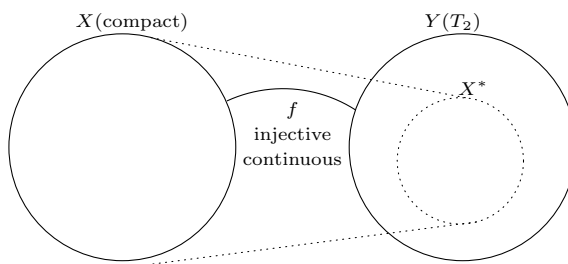


Figure 7.2: Embedding compact space into hausdorff space

Theorem 7.1.1. *Every compact T_2 space is a T_3 space.*

Proof. $T_2 \implies T_1$

It is enough to prove that compact, T_2 space is regular.

Let x be a point in X and C be a closed subset not containing x .

Then C is compact, compactness is weakly hereditary.

By proposition, T_2 space separates points from compact subsets.

Thus $T_2 + \text{compact} \implies T_3$ (\dagger^6) □

Proposition 7.1.2. *Let X be a regular space, C a closed subset of X and F a compact subset of X , such that $C \cap F = \emptyset$. Then there exist open subsets U, V such that $C \subset U, F \subset V$ and $U \cap V = \emptyset$.*

Regular spaces separates closed subsets from compact subsets.

Proof. proof technique is same as T_2 separates points from compact subsets.

Let C be a closed subset, F be a compact subset of X and $C \cap F = \emptyset$.

X is regular $\implies \forall y \in F, \exists U_y, V_y \in \mathcal{T}, C \subset U_y, y \in V_y, U_y \cap V_y = \emptyset$

Define $\mathcal{C} = \{V_y : y \in F\}$ is cover of compact subset F .

$\implies \exists \mathcal{C}'$ such that $\mathcal{C}' = \{V_{y_i} : i = 1, 2, \dots, n\}$ is a finite subcover.

Define $U = \bigcap_{i=1}^n \{U_{y_i}\}$ and $V = \bigcup_{i=1}^n \{V_{y_i}\}$.

$\implies C \subset U, F \subset V, U \cap V = \emptyset$. □

Theorem 7.1.2. *Every regular, Lindeloff space is normal.*

Proof. Let C, D be two disjoint, closed subsets of a regular, lindeloff space X .

X is regular $\implies \forall x \in C, \exists U_x, V_x$ such that $x \in U_x, D \subset V_x, U_x \cap V_x = \emptyset$

X is regular $\implies \forall y \in D, \exists U_y, V_y$ such that $C \subset U_y, y \in V_y, U_y \cap V_y = \emptyset$

Then $\{U_x : x \in C\}$ and $\{V_y : y \in D\}$ are open covers of C and D respectively.

\dagger^7)

Let $\{U_n : n = 1, 2, \dots\}$ and $\{V_n : n = 1, 2, \dots\}$ be their countable subcovers.

Define $G_n = U_n - \bigcup_{i=1}^n \overline{V_i}$ and $H_n = V_n - \bigcup_{i=1}^n \overline{U_i}$.

Define $G = \bigcup_{i=1}^\infty G_n$ and $H = \bigcup_{i=1}^\infty H_n$.

Claim : $C \subset G$ (similary $D \subset H$)

$x \in C \implies x \in U_n$ for some n , since $\{U_n : n \in \mathbb{N}\}$ is a cover of C

$\forall m, \overline{V_m} \subset X - C \implies \forall m, x \notin \overline{V_m} \implies x \in G$

Claim : $G \cap H = \emptyset$

$x \in G \cap H \implies x \in G_m \cap H_n$ for some m, n

With loss of generality, $n \geq m, x \in G_m \implies x \in U_m \implies x \notin H_n$ □

⁶ $T_3 \not\Rightarrow \text{compact}$.

⁷Here, U_x, V_y are unrelated.

Corollary 7.1.2.1. *Every regular, second countable space is normal.*

Proof. Every second countable space is lindeloff.

By theorem, every regular, lindeloff space is normal. \square

Corollary 7.1.2.2. *Every compact, T_2 space is T_4 .*

Proof. By proposition, every compact, T_2 space is regular.

Every compact space is lindeloff since finite subcovers are countable.

By theorem, every regular, lindeloff space is normal. \square

7.1.2 The Urysohn Characterisation of Normality

X is normal \iff there exist Urysohn functions

Q : Significance of Urysohn's Lemma ?

1. There is no analog of Urysohn's Lemma for Regular spaces.
2. Constructs a nice real-valued function even on non-metrisable spaces.

Q : We know that, the completely regular space separates points from closed subsets using a real-valued function. Which space separates closed subsets from closed subsets using a real-valued function ?

A : Normal space(by Urysohn's Lemma). Not completely normal(\star^8)

Proposition 7.1.3. *Let A, B be subsets of a space X and suppose there exists a continuous function $f : X \rightarrow [0, 1]$, such that $f(x) = 0, \forall x \in A$ and $f(x) = 1, \forall x \in B$. Then there exists disjoint open subsets U, V such that $A \subset U$ and $B \subset V$.*

Proof. Let f be a continuous function, $f(x) = 0, \forall x \in A$ and $f(x) = 1, \forall x \in B$. Define $G = [0, \frac{1}{2})$, $H = (\frac{1}{2}, 1]$ and $U = f^{-1}(G)$, $V = f^{-1}(H)$.

$\implies A \subset U, B \subset V$ and $U \cap V = \phi$ since $G \cap H = \phi$. \square

Corollary 7.1.2.3 (Urysohn's Lemma : Sufficient Condition). *If X has the property that for any disjoint closed subsets A, B of X , there exists a continuous function $f : X \rightarrow [0, 1]$ such that $f(x) = 0, \forall x \in A$ and $f(x) = 1, \forall x \in B$, then X is normal.*

There exists a Urysohn function $f \implies X$ is normal

Proof. Let A, B be two closed subsets of X .

By proposition, X normal. \square

Theorem 7.1.3 (Urysohn's Lemma). *A topological space X is normal iff it has the property that for every mutually disjoint, closed subsets A, B of X , there exists a continuous function $f : X \rightarrow [0, 1]$ such that $f(x) = 0$ for all $x \in A$ and $f(x) = 1$ for all $x \in B$*

⁸Completely Normal Space separates any two subsets. [Joshi, 1983, chapter 7 Exer. 1.11]
That is these subsets need not be closed. Thus every completely normal space is normal, but not the other way. Thus $T_1 + \text{Completely Normal} = T_5 \supset T_4$.

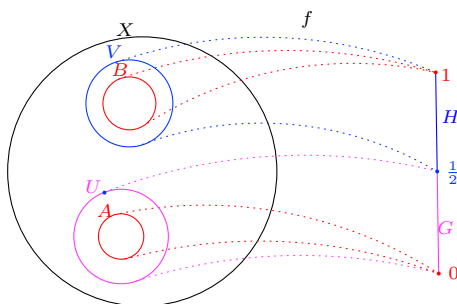


Figure 7.3: Urysohn's Lemma : Sufficient Condition

Proof. Urysohn's Lemma : necessary condition

X is normal \implies there exist Urysohn functions

Let A, B be two closed subsets in a normal space X .

Enumerate rationals in the unit interval.

$\mathbb{Q} \cap [0, 1] = \{0, 1, \dots\} = \{q_0, q_1, q_2, \dots\}$.

Define $F_1 = F_{q_1} = X - B$.

$A \subset X - B \implies \exists H$ such that $A \subset H \subset \overline{H} \subset X - B$. (\star^9)

Define $F_{q_0} = F_0 = H$.

$\overline{F_{q_0}} \subset F_{q_1} \implies$ true for $n = 1$.

Suppose : $\overline{F_{q_i}} \subset F_{q_j}$, $\forall j > i$ is true for $j = 1, 2, \dots, n-1$.

Define $q_i = \sup\{q_k : q_k < q_n, k < n\}$ and $q_j = \inf\{q_k : q_k > q_n, k < n\}$ (\dagger^{10})

$\implies q_i < q_n < q_j$ and $\overline{F_{q_i}} \subset F_{q_j}$.

$\overline{F_{q_i}} \subset F_{q_j} \implies \exists H$ such that $\overline{F_{q_i}} \subset H \subset \overline{H} \subset F_{q_j}$. Define $F_{q_n} = H$.

Therefore, $\overline{F_{q_j}} \subset F_{q_i}$, $\forall j > i$.

Now, $\{F_t : t \in \mathbb{Q}\}$ has the properties required in lemma 2.

By lemma 2, there exists a function f . This is a Urysohn function. (\dagger^{11}) \square

Lemma 7.1.4. Let $f : X \rightarrow [0, 1]$ be continuous. For each $t \in \mathbb{R}$ let $F_t = \{x \in X : f(x) < t\}$. Then the indexed family $\{F_t : t \in \mathbb{R}\}$ has the following properties

1. F_t is an open subset of X for each $t \in \mathbb{R}$
2. $F_t = \emptyset$ for $t < 0$
3. $F_t = X$ for $t > 1$
4. For any $s, t \in \mathbb{R}$, $s < t \implies \overline{F_s} \subset F_t$.

Moreover, for each $x \in X$, $f(x) = \inf\{t \in \mathbb{Q} : x \in F_t\}$.

Proof. Not needed. \square

⁹Equivalent condition for normality : Let A be closed subset and G be an open subset containing A . Then there exists open subset H such that $A \subset H \subset \overline{H} \subset G$. cite[7.1.16(3)]joshi

¹⁰Let $\mathbb{Q} \cap [0, 1] = \{0, 1, 0.3, 0.7, 0.8, \mathbf{0.5}, 0.6, \dots\}$. Consider $n = 5$. Then $q_n = q_5 = 0.5$ and in the respective induction step, $q_i = \sup\{0, 0.3\} = 0.3$ and $q_j = \inf\{1, 0.7, 0.8\} = 0.7$

¹¹You will have to replace "Urysohn function" with "There exists a continuous real-valued function, $f : X \rightarrow [0, 1]$ such that $f(x) = 0, \forall x \in A, f(x) = 1, \forall x \in B$ ".

Lemma 7.1.5. *Let X be a topological space and suppose $\{F_t : t \in \mathbb{Q}\}$ is a family of sets in X such that*

1. F_t is open in X for each $t \in \mathbb{Q}$
2. $F_t = \emptyset$ for $t \in \mathbb{Q}, t < 0$
3. $F_t = X$ for $t \in \mathbb{Q}, t > 1$
4. $\overline{F_s} \subset F_t$ for $s, t \in \mathbb{Q}, s < t$

For $x \in X$, let $f(x) = \inf\{t \in \mathbb{Q} : x \in F_t\}$. Then f is a continuous real-valued function on X and it takes values in the unit interval $[0, 1]$.

Proof. Function f is well-defined and $\text{Range}(f) = [0, 1]$.

Define $H = \{x \in X : f(x) < s\} = f^{-1}(-\infty, s)$ and

$K = \{x \in X : f(x) > s\} = f^{-1}(s, \infty)$.

$\mathcal{S} = \{(s, \infty), (-\infty, s) : s \in \mathbb{R}\}$ is a subbase for \mathbb{R} with usual topology.

Claim : $H = \cup\{F_t : t \in \mathbb{Q}, t < s\}$

$x \in H \implies f(x) < s \implies \inf G_x < s \implies \exists q \in \mathbb{Q} (q < s), x \in F_q$

$\implies x \in \cup\{F_t : t \in \mathbb{Q}, t < s\} \implies H \subset \cup\{F_t : t \in \mathbb{Q}, t < s\}$.

$x \in \cup\{F_t : t \in \mathbb{Q}, t < s\} \implies \exists t \in \mathbb{Q} (t < s), x \in F_t \implies \inf G_x < s$

$\implies f(x) < s \implies x \in H \implies \cup\{F_t : t \in \mathbb{Q}, t < s\} \subset H$.

Claim : $X - K = \cap\{\overline{F_t} : t \in \mathbb{Q}, t > s\}$.

$x \in X - K \implies x \notin K \implies f(x) \leq s \implies \inf G_x \leq s$

Let $t \in \mathbb{Q} (s < t) \implies \exists q \in G_x (s < q < t) \implies x \in \overline{F_q} \subset F_t \subset \overline{F_t}$.

$\forall t \in \mathbb{Q} (t > s), x \in \overline{F_t} \implies x \in \cap\{\overline{F_t} : t \in \mathbb{Q}, t > s\}$.

Thus, $X - K \subset \cap\{\overline{F_t} : t \in \mathbb{Q}, t > s\}$.

$x \in \cap\{\overline{F_t} : t \in \mathbb{Q}, t > s\}$

Suppose(1) $x \in K \implies s < f(x) \implies \exists q, t \in \mathbb{Q} (s < q < t), t < f(x)$

Suppose(2) $x \in \overline{F_q} \implies x \in \overline{F_q} \subset F_t \implies \inf G_x < t \implies f(x) < t$

is a contradiction(2). Thus $x \notin \overline{F_q}$.

$x \notin \overline{F_q} \implies x \notin \cap\{\overline{F_t} : t \in \mathbb{Q}, t > s\}$ is a contradiction (1). Thus $x \notin K$

Thus, $\cap\{\overline{F_t} : t \in \mathbb{Q}, s > t\} \subset X - K$.

Inverse images of subbase elements are open. Thus f is continuous. \square

Corollary 7.1.5.1. *All T_4 spaces are completely regular and hence Tychonoff.*

Proof. Let $x \in X$ and D be closed subset not containing x . We have X is a T_4 space. Therefore X is T_1 as well as normal.

Now the singleton set, $\{x\}$ is closed, since X is a T_1 space. And by **Urysohn's lemma** for disjoint, closed subsets $\{x\}, D$ there exists a Urysohn function which is a continuous, real-valued function $f : X \rightarrow [0, 1]$ such that $f(x) = 0$ and $f(y) = 1$ for all $y \in D$. Therefore the space X is completely regular and hence Tychonoff. \square

Remark (Urysohn function). The function whose existence is asserted by Urysohn's lemma is called a Urysohn function

7.1.3 Tietze Characterisation of Normality

Tietze Characterisation of normality

X is normal \iff there exist continuous extension of real-valued functions on closed subsets.

Proposition 7.1.4. *Let A be a subset of a space X and let $f : A \rightarrow \mathbb{R}$ be continuous. Then any two extensions of f to X agree on \overline{A} . In other words, if at all an extension of f exists its values on \overline{A} are uniquely determined by values of f on A .*

Proposition 7.1.5 (Tietze : necessary condition). *Suppose a topological space X has the property that for every closed subset A of X , every continuous real valued function on A has a continuous extension to X . Then X is normal.*

Definitions 7.1.1 (Pointwise Convergence). Let X be a topological space and (Y, d) a metric space. Then a sequence of functions $\{f_n\}$ from X to Y converges pointwise to f if for every $x \in X$ the sequence $\{f_n(x)\}$ converges to $f(x)$ in Y .

In other words, given a very small value, $\epsilon > 0$, there exists some $\delta > 0$ such that for every $x \in X$ there exists $N_x \in \mathbb{N}$. This N_x may be different for different values of x and for every $n > N_x$, $d(f(x), f_n(x)) < \delta$.

Definitions 7.1.2 (Uniform Convergence). Let X be a topological space and (Y, d) a metric space. Then a sequence of functions $\{f_n\}$ from X to Y converges uniformly to f if given a small $\epsilon > 0$, there exists $\delta > 0$ such that there exists $N \in \mathbb{N}$. This N is independent of the value of x and for every $n > N$, $d(f(x), f_n(x)) < \delta$.

Proposition 7.1.6. *Let X , (Y, d) , $\{f_n\}$ and f be as above and suppose $\{f_n\}$ converges to f uniformly. If each f_n is continuous, then f is continuous.*

Definitions 7.1.3 (Uniform Convergence of Series). Let X be a topological space and (Y, d) be a metric space. Then a series of function $\sum_{n=1}^{\infty} f_n$ converges uniformly to f if the sequence of partial sums converges uniformly to f .

In other words, let $g_m = \sum_{n=1}^m f_n$. Then $\sum_{n=1}^{\infty} f_n$ converges to f uniformly if the sequence of partial sums $\{g_m\}$ converges to f uniformly.

Proposition 7.1.7. *Let $\sum_{n=1}^{\infty} M_n$ be a convergent series of non-negative real numbers. Suppose $\{f_n\}$ is a sequence of real valued functions on a space X such that for each $x \in X$ and $n \in \mathbb{N}$, $|f_n(x)| \leq M_n$. Then the series $\sum_{n=1}^{\infty} f_n$ converges uniformly to a real valued function on X .*

Theorem 7.1.6. *Let A be a closed subset of a normal space X and suppose function $f : A \rightarrow [-1, 1]$ is a continuous function. Then there exists a continuous function $F : X \rightarrow [-1, 1]$ such that $F(x) = f(x)$ for all $x \in A$.*

Theorem 7.1.7. *Let A be a closed subset of a normal space X and suppose function $f : A \rightarrow (-1, 1)$ is a continuous function. Then there exists a continuous function $F : X \rightarrow (-1, 1)$ such that $F(x) = f(x)$ for all $x \in A$.*

Corollary 7.1.7.1 (Tietze : sufficient condition). *Any continuous real-valued function on a closed subset of a normal space can be extended continuously to the whole space.*

7.2 Products and Coproducts

7.2.1 Cartesian Products of Families of Sets

7.2.2 The Product Topology

7.2.3 Productive Properties

7.3 Embedding and Metrisation

7.3.1 Evaluation Functions into Products

7.3.2 Embedding Lemma and Tychonoff Embedding

7.3.3 The Urysohn Metrisation Theorem

7.4 Nets and Homotopy

7.4.1 Definition and Convergence of Nets

Definitions 7.4.1 (Directed Set). [Joshi, 1983, 10.1.1]

A directed set D is a pair (D, \geq) where D is a nonempty set and \geq is a binary relation on D such that

1. The relation ‘follows’ (\geq) is transitive. ie, $m \geq n, n \geq p \implies m \geq p$
2. The relation ‘follows’ (\geq) is reflexive. ie, For every $m \in D, m \geq m$
3. For any $m, n \in D$, there exists $p \in D$ such that $p \geq m$ and $p \geq n$.

sequence in a set X is a function f from the set of all integers into X .

Definitions 7.4.2 (Net). [Joshi, 1983, 10.1.2]

A net in a set X is a function S from a directed set D into the set X .

Remark. The set \mathbb{N} together with the relation ‘less than or equal to’ (\leq) is a directed set. Clearly, the relation ‘less than or equal to’ is reflexive and transitive. And the third condition is true iff every finite subset E of D has an element $p \in E$ such that p follows each element of E . This is a weaker notion compared to the well ordering principle¹² of the set of all integers. Thus \mathbb{N} is a directed set and every sequence in X is also a net in X .

Remark (Significance of Net). A net on a set is a generalisation of ‘a sequence on a set’ obtained by simplifying the domain of the sequence into a directed set. The notion directed set is derived by assuming a few properties of \mathbb{N} .

The convergence of sequence is not strong enough to characterise topologies as the limit of convergent sequences are unique for both Hausdorff and Co-countable spaces. The notion of Net allows us to differentiate between Hausdorff spaces from Co-countable spaces in terms of convergence of nets. The limit of a convergent net on a topological space is unique iff it is a Hausdorff space. ie, We have removed a few restrictions, so that we will have some convergent nets (which are obviously not sequences) with multiple limit points for Co-countable spaces.

Remark. Examples of Directed Sets

1. Let X be a topological space and $x \in X$. Then the neighbourhood system \mathcal{N}_x is a directed set with the binary relation \subset (subset/inclusion).
 - (a) Let U, V, W be any three neighbourhoods of $x \in X$ such that $U \subset V$ and $V \subset W$. Then, clearly $U \subset W$. Therefore, $U \geq V, V \geq W \implies U \geq W$.
 - (b) Let U be any neighbourhood of $x \in X$, then $U \subset U$. Therefore, $U \geq U$.
 - (c) Let U, V be any two neighbourhoods of $x \in X$, then there exists their intersection $W = U \cap V$, which is a neighbourhood of x . Clearly $W \subset U$ and $W \subset V$. Therefore $\forall U, V \in \mathcal{N}_x, \exists W \in \mathcal{N}_x$ such that $W \geq U$ and $W \geq V$.

¹²Well-ordering principle : Every subset of \mathbb{N} has a least element in it.

2. Let \mathcal{P} be the set of all partitions on closed unit interval $[0, 1]$. A partition $P \in \mathcal{P}$ is a refinement of $Q \in \mathcal{P}$ if every subinterval in P is contained in some subinterval of Q . Then \mathcal{P} with the binary relation refinement is a directed set.
 - (a) Suppose P, Q, R are three partitions of $[0, 1]$ such that P is a refinement of Q and Q is a refinement of R , then clearly P is a refinement of R since each subinterval of P is contained some subinterval of Q , which is contained in some subinterval of R .
Therefore, $P \geq Q, Q \geq R \implies P \geq R$
 - (b) Suppose P is a partition of $[0, 1]$. Then trivially, P is a refinement of itself since every subinterval of P is contained in the same subinterval of P .
Therefore, $\forall P \in \mathcal{P}, P \geq P$
 - (c) Suppose P, Q be any two partition of $[0, 1]$. Then $R = P \cup Q$ is a refinement of both the partitions.
Therefore $\forall P, Q \in \mathcal{P}, \exists R \in \mathcal{P}$ such that $R \geq P$ and $R \geq Q$

For example, let $P = \{0, 0.3, 0.7, 1\}$. Then the subintervals in P are $[0, 0.3]$, $[0.3, 0.7]$ and $[0.7, 1]$. Let $Q = \{0, 0.3, 0.5, 1\}$ and $R = \{0, 0.3, 0.5, 0.7, 1\}$. Then R is a refinement of P , but Q is not a refinement of P since there is a subinterval $[0.5, 1]$ in Q which is not properly contained in any subinterval of P . However, R is a refinement of Q as well.

Remark. Examples of Nets

1. Let X be a topological space and $x \in X$. Let \mathcal{N}_x be the set of all neighbourhoods of x . Let $D = (\mathcal{N}_x, X)$ be the directed set given by $(N, y) \in (\mathcal{N}_x, X)$ if $N \in \mathcal{N}_x$ and $y \in N$ and $(N, y) \geq (M, z)$ if $N \subset M$. Then the function $S : (\mathcal{N}_x, X) \rightarrow X$ given by $S(N, y) = y$ is a net on X .

For example, let $X = \{a, b, c, d\}$ and $\mathcal{T} = \{\{a\}, \{a, b\}, \{a, b, c\}, \{a, b, c, d\}\}$. Also let $S : (\mathcal{N}_b, X) \rightarrow X$ defined by $S(N, y) = y$. Suppose $C = \{a, b, c\}$. Then $C \in \mathcal{N}_b$. ie, C is a neighbourhood of b . Then $S(C, c) = c$.

2. Riemann Net - Let $D = (\mathcal{P}, \xi)$ where \mathcal{P} is the set of all partitions on $[0, 1]$ and ξ is a finite sequence in $[0, 1]$ such that consecutive terms belongs to consecutive subintervals of the partition. The set (\mathcal{P}, ξ) is directed set with \geq given by $(P, \eta) \geq (Q, \psi)$ iff P is a refinement of Q .

For example, let $P \in \mathcal{P}$ is given by $P = \{0, 0.3, 0.7, 1\}$ and $\eta = \{0.2, 0.6, 0.9\}$. Then $(P, \eta) \in (\mathcal{P}, \xi)$.

Let $f : \mathbb{R} \rightarrow \mathbb{R}$ be any function, then the function,

$$S : (\mathcal{P}, \xi) \rightarrow \mathbb{R} \text{ defined by } S(P, \eta) = \sum_{j=1}^k f(\eta_j)(a_k - a_{k-1})$$

where $P = \{a_0, a_1, \dots, a_k\}$ is the Riemann Net with respect to the real function f .

For example, let $f(x) = x^2$ and P, η are same as above example, then
 $S(P, \eta) = 0.2^2(0.3 - 0) + 0.6^2(0.7 - 0.3) + 0.9(1 - 0.7) = 3.99$

Definitions 7.4.3 (Convergence of a Net). [Joshi, 1983, 10.1.3]

A net $S : D \rightarrow X$ converges to a point $x \in X$ if for any nbd U of x , there exists $m \in D$ such that $n \in D, n \geq m \implies S(n) \in U$. And x is a limit of the net S .

Remark. The choice of m depends on the choice of neighbourhood U

$$S : D \rightarrow X, S \rightarrow x \iff (\forall U \in \mathcal{N}_x, \exists m_U \in D, \text{ such that } n \geq m_U \implies S(n) \in U)$$

Theorem 7.4.1 (Net characterisation of Hausdorff space). [Joshi, 1983, 10.1.4]
 A topological space is Hausdorff iff limits of all nets in it are unique.

Proof. Let X be a Hausdorff space. Suppose $S : D \rightarrow X$ is net on X such that S converges to two distinct points $x, y \in X$. Since X is a Hausdorff space and $x \neq y$, there exists open subsets U, V such that $x \in U, y \in V, U \cap V = \phi$.

The net S converges to $x \in X$, therefore $\exists m_x \in D$ such that $n \geq m_x \implies S(n) \in U$. And, the net S converges to $y \in X$, therefore $\exists m_y \in D$ such that $n \geq m_y \implies S(n) \in V$.

Since D is a directed set and $m_x, m_y \in D$, there exists $p \in D$ such that $p \geq m_x$ and $p \geq m_y$. Now, $n \geq p \implies n \geq m_x, n \geq m_y$, since \geq is transitive. (ie, $n \geq p, p \geq m_x \implies n \geq m_x$, and $n \geq p, p \geq m_y \implies n \geq m_y$).

We have $n \geq p \implies n \geq m_x$ and $n \geq m_x \implies S(n) \in U$. Therefore, $n \geq p \implies S(n) \in U$. Similarly, $n \geq p \implies n \geq m_y \implies S(n) \in V$. Therefore $S(n) \in U \cap V$ which is a contradiction, since $U \cap V = \phi$. Therefore, if a net S converges to two points x, y , then $x = y$. That is, if a net S in a Hausdorff space X is convergent then its limit is unique.

Conversely, suppose that X is a topological space and every convergent net in X has a unique limit. Suppose X is not a Hausdorff space. Then there exists at least two distinct points $x, y \in X$ such that every neighbourhood of x intersects with every neighbourhood of y . Now consider the set $D = \mathcal{N}_x \times \mathcal{N}_y$ and relation \geq on D such that $(U_1, V_1) \geq (U_2, V_2)$ if $U_1 \subset U_2$ and $V_1 \subset V_2$.

By the axiom of choice, a function $S : D \rightarrow X$ such that $S(U, V) \in U \cap V$ is well defined, since every nbd of x intersects every nbd of y . Thus, S is a net in X . We claim that S converges to both x and y .

Let U be a nbd of x . Then $S(U', V') \in U' \cap V'$. We have $(U, X) \in D$ such that $(U', V') \geq (U, X) \implies U' \subset U$. Then, $S(U', V') \in U' \cap V' \subset U \cap X = U$. Thus, for any nbd U containing x , we have $(U, X) \in D$ such that $(U', V') \geq (U, X) \implies S(U', V') \in U$. Therefore, S converges to $x \in X$.

Similarly, Let V be a nbd of y . Then for any nbd V containing y , we have $(X, V) \in D$ such that $(U', V') \geq (X, V) \implies S(U', V') \in V$, since

$S(U', V') \in U' \cap V' \subset X \cap V = V$. Therefore, S converges to $y \in X$ as well, where $x \neq y$. This is a contradiction to the assumption that every convergent net in X has a unique limit. Therefore, for any two points $x, y \in X$, there should be some nbd of x that doesn't intersect some nbd of y . Therefore, X is a Hausdorff space. \square

Definitions 7.4.4 (Eventual Subset). [Joshi, 1983, 10.1.5]

A subset E of a directed set D is an eventual subset of D if there exists $m \in D$ such that $n \geq m \implies n \in E$.

Remark. Let E be an eventual subset of D such that $n \geq m \implies n \in E$. Then $p \in E \not\Rightarrow p \geq m$. ie, Subset E may contain elements that doesn't follow the above m .

Remark. [Joshi, 1983, 10.1.6]

Let E be an eventual subset of D , then E is a directed set.

1. $m, n, p \in E, m \geq n, n \geq p \implies m, n, p \in D, m \geq n, n \geq p \implies m \geq p$
2. $m \in E \implies m \in D \implies m \geq m$
3. $m, n \in E \implies m, n \in D \implies \exists p \in D$ such that $p \geq m$ and $p \geq n$.

$\exists m' \in D$ such that $n' \geq m' \implies n' \in E$. (E is an eventual subset of D)

$p, m' \in D \implies \exists p' \in D$ such that $p' \geq p$ and $p' \geq m'$. (D is a directed Set)

$p' \geq m' \implies p' \in E$. (E is eventual subset of D with respect to m')

$p' \geq p, p \geq m \implies p' \geq m$ and $p' \geq p, p \geq n \implies p' \geq n$.

Therefore $\forall m, n \in E, \exists p' \in E$ such that $p' \geq m$ and $p' \geq n$.

Definitions 7.4.5 (Net eventually in A). [Joshi, 1983, 10.1.5]

Let $S : D \rightarrow X$ be a net in a topological space X . Then S is eventually in subset A of X if $S^{-1}(A)$ is an eventual subset of D .

Remark. Let $S : D \rightarrow X$ be a net in X . Then S converges to $x \in X$ if S is eventually in each nbd U of x .

Definitions 7.4.6 (Cofinal subset). [Joshi, 1983, 10.1.7]

A subset F of a directed D is a cofinal subset of D if for any $m \in D$, there exists $n \in F$ such that $n \geq m$.

Remark. Let X be a topological space and $x \in X$. Let \mathcal{N}_x be the set of all neighbourhood of x and \mathcal{L} be a local base of X at x . We have, (\mathcal{N}_x, \geq) is a directed set where $\forall U, V \in \mathcal{N}_x, U \geq V \iff U \subset V$, then \mathcal{L} is cofinal in \mathcal{N}_x .

Remark. [Joshi, 1983, 10.1.8]

Let F be a cofinal subset of D , then F is a directed set.

1. $m, n, p \in F, m \geq n, n \geq p \implies m, n, p \in D, m \geq n, n \geq p \implies m \geq p$
2. $m \in F \implies m \in D \implies m \geq m$

3. $m, n \in F \implies m, n \in D \implies \exists p \in D$ such that $p \geq m$ and $p \geq n$.

E is cofinal, $p \in D \implies \exists p' \in F$ such that $p' \geq p$.

$p' \geq p, p \geq m \implies p' \geq m$ and $p' \geq p, p \geq n \implies p' \geq n$.

Therefore $\forall m, n \in F, \exists p' \in F$ such that $p' \geq m$ and $p' \geq n$.

Definitions 7.4.7 (Net frequently in A). [Joshi, 1983, 10.1.7]

Let $S : D \rightarrow X$ be a net in a topological space X . Then S is frequently in subset B of X if $S^{-1}(B)$ is a cofinal subset of D .

Proposition 7.4.1. [Joshi, 1983, 10.1.6]

Let $S : D \rightarrow X$ be a net in a topological space X . Let E be an eventual subset of D . Then, S converges to x iff $S_{/E}$ converges to x . cite[10.1.6]joshi

Proof. Let $S : D \rightarrow X$ be a net in X , E be an eventual subset of D , and $x \in X$. Then, $S_{/E} : E \rightarrow X$ is defined by $n \in E \implies S_{/E}(n) = S(n)$

Suppose S converges to x . Let U be a nbd of x , then S is eventually in U . ie, $S^{-1}(U)$ is an eventual subset of D . Then $\exists m \in D$ such that $n \geq m \implies n \in S^{-1}(U) \implies S(n) \in U$. Since set E is eventual subset of D , $\exists m' \in D$ such that $n \geq m' \implies n \in E$.

Since E is a directed set, $S_{/E} : E \rightarrow X$ is a net in X . And $m, m' \in D \implies \exists p \in D$ such that $p \geq m$ and $p \geq m'$. We have, $p \geq m' \implies p \in E$. And $n \geq' p \implies n \geq p, p \geq m \implies n \geq m \implies S(n) \in U \implies S_{/E}(n) \in U$. Therefore, $n \geq' p \implies S_{/E}(n) \in U$. Since U is arbitrary, $S_{/E}$ converges to x .

Conversely, suppose that $S_{/E}$ converges to x . Let U be a nbd of x , then $S_{/E}$ is eventually in U . ie, $S_{/E}^{-1}(U)$ is an eventual subset of D . ie, $\exists m \in D$ such that $n \geq m \implies n \in S_{/E}^{-1}(U) \implies S_{/E}(n) \in U \implies S(n) \in U$. Therefore, $n \geq m \implies S(n) \in U$. Since, U is arbitrary, S converges to every nbd of x . ie, S converges to x . \square

Proposition 7.4.2. [Joshi, 1983, 10.1.8]

Let $S : D \rightarrow X$ be a net in a topological space X . Let F be a cofinal subset of D . If S converges to x , then $S_{/F}$ converges to x .

Proof. Let $S : D \rightarrow X$ be a net in X and S converges to $x \in X$. Also let F be a cofinal subset of D . Then $S_{/F}$ is also a net in X , since (F, \geq') is a directed set where $\forall m, n \in F, m \geq n \implies m \geq' n$.

Since S converges to x , for any nbd U of x , $\exists m \in D$, such that $n \geq m \implies S(n) \in U$. Since F is cofinal, $\exists p \in F$ such that $p \geq m$. Thus $n \geq' p \implies n \geq p, p \geq m \implies n \geq m \implies S(n) \in U \implies S_{/F}(n) \in U$. Therefore, $\exists p \in F$ such that $n \geq' p \implies S_{/F}(n) \in U$. Since U is arbitrary, $S_{/F}$ is eventually in every nbd of x . ie, $S_{/F}$ converges to x . \square

Remark. But converse of the above is not true. $S_{/F}$ converges to x does not imply that S converges to x , since cofinal subset F not necessarily contain every element following a particular m .

Definitions 7.4.8 (Cluster point). [Joshi, 1983, 10.1.9]

Let $S : D \rightarrow X$ be a net in a topological space X . Then $x \in X$ is a cluster point of S , if S is frequently in each nbd U of x in X .

Proposition 7.4.3. [Joshi, 1983, 10.1.10]

Let $S : D \rightarrow X$ be a net in a topological space X . Then $x \in X$ is a cluster point of X , if $S_{/F}$ converges to x for some cofinal subset F of D .

Proof. Let $S : D \rightarrow X$ be a net in X and (F, \geq') be a cofinal subset of (D, \geq) . Then $S_{/F}$ is also a net in X . Suppose $S_{/F}$ converges to $x \in X$. Let U be a nbd of x , then $\exists m \in F$ such that $n \geq' m \implies S_{/F}(n) \in U$.

Let $m' \in D$. Then $\exists p' \in F$ such that $p' \geq m'$, since F is a cofinal subset of D . We have, $m, p' \in F$, then $\exists p \in F$ such that $p \geq' m$ and $p \geq' p'$. Since $F \subset D$, we have $p, m \in F \implies p, m \in D$ and $p \geq' m \implies p \geq m$.

Also $p \geq' m \implies S_{/F}(p) \in U \implies S(p) \in U$. Therefore, $\forall m' \in D$, $\exists p \in D$ such that $p \geq m'$ and $S(p) \in U$. Since U, m' are arbitrary, S is frequently in every nbd of x . ie, x is a cluster point of S . \square

Definitions 7.4.9 (Subnet). [Joshi, 1983, 10.1.11]

Let $S : D \rightarrow X$ be a net in a topological space X . Then a net $T : E \rightarrow X$ in X , is a subnet of S if there exists a function $N : E \rightarrow D$ such that $S \circ N = T$ and $\forall m \in D$, $\exists p \in E$ such that $n \geq' p \implies N(n) \geq m$.

Remark. A net $T : E \rightarrow X$ is a subnet of $S : D \rightarrow X$ if $\exists N : E \rightarrow D$ such that $S \circ N = T$ and S is frequently in $T(E)$.

Let $T : E \rightarrow X$ be a subnet of $S : D \rightarrow X$ and A be a subset of X . If T eventually in A , then S is frequently in A .

Proposition 7.4.4. [Joshi, 1983, 10.1.12]

Let $S : D \rightarrow X$ be a net in a topological space X . Then $x \in X$ is a cluster point of S iff there exists a subnet of S which converges to x .

Synopsis. Let (D, \geq) , (E, \geq') be two directed sets. And $T : E \rightarrow X$ be a subnet of $S : D \rightarrow X$.

If T converges to x , then T is eventually in each nbd U of x . And since T is a subnet of S , there exists $N : E \rightarrow D$ such that $N(E)$ is a cofinal subset of D . Therefore, S is frequently in each nbd U of x . Thus, x is a cluster point of S .

If x is a cluster point of X , then S is frequently in every nbd of x . Let $N : E \rightarrow D$ be $N(n, U) = n$. Construct a directed subset E of $D \times \mathcal{N}_x$ such that $(n, U) \in E \iff S(n) \in U$. Now T is eventually in every nbd U of x , as those points with images outside U are removed by construction. Therefore, it is sufficient to show that $T : E \rightarrow X$ is a subnet of the net $S : D \rightarrow X$. Clearly, E is a directed set and $N : E \rightarrow D$ defined by $N(n, U) = n$ satisfies both $S \circ N = T$ and $\forall m \in D$, $\exists p \in E$ such that $m \geq p \implies N(m) \geq p$.

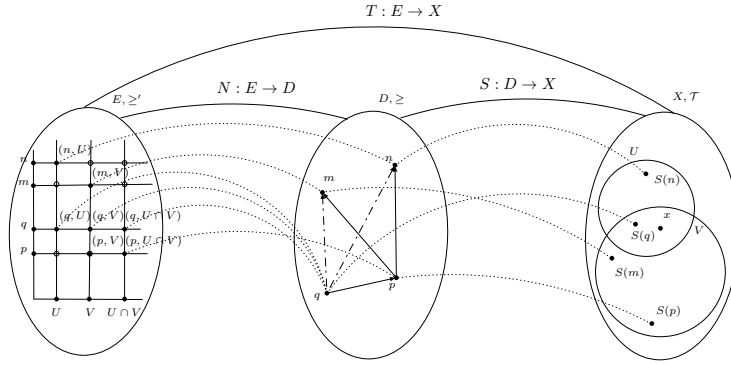


Figure 7.4: $\forall (n, U), (m, V) \in E, \exists (q, W) \in E$ such that $(q, W) \geq' (n, U)$, $(q, W) \geq' (m, V)$

Proof. Let $S : D \rightarrow X$ be a net in X . Suppose there exists a subnet $T : E \rightarrow X$ that converges to $x \in X$. By the definition of subnet, we have $\exists N : E \rightarrow D$ such that $S \circ N = T$ and S is frequently in $T(E)$.

We have, T converges to x , thus for any neighbourhood U of x , there exists $m' \in E$ such that $n' \geq' m' \implies T(n') \in U$.

Also we have, T is a subnet of S . Then $\exists N : E \rightarrow D$ such that $\forall m \in D, \exists p' \in E$ such that $n' \geq' p' \implies N(n') \geq m$.

Now, for any $m \in D$, we have $m', p' \in E$. Since E is a directed set, there exists $n' \in E$ such that $n' \geq' m'$ and $n' \geq' p'$.

Then, $n' \geq m' \implies T(n') \in U$ and $n' \geq' p' \implies N(n') \geq m$.

Thus for any $m \in D$, there exists $N(n') = n \in D$ such that $S(n) = S(N(n')) = T(n') \in U$.

Thus S is frequently in any neighbourhood U of x . Therefore, x is a cluster point of S .

Conversely, suppose that x is a cluster point of S . We have to construct a directed set (E, \geq') and a function $N : E \rightarrow D$ such that T is a subnet of S and T converges to x . Let \mathcal{N}_x be the family of all neighbourhood of x in X .

Consider $E = \{(n, U) \in D \times \mathcal{N}(x) : S(n) \in U\}$ and define \geq' by $(n, U) \geq' (m, V)$ if $n \geq m$ and $U \subset V$. Trivially, $(n, U) \geq' (m, V) \geq' (p, W) \implies (n, U) \geq' (p, W)$ and $(n, U) \geq' (n, U)$. Also, for any $(n, U), (m, V) \in E$, we have $n, m \in D$ and $U, V \in \mathcal{N}(x)$. Since D is a directed set, $\exists p \in D$ such that $p \geq n$ and $p \geq m$. Also, $U \cap V \in \mathcal{N}_x$ and there exists $q \in D$ such that $S(q) \in U \cap V$ and $q \geq' p$, since S is frequently in every nbd of x . And $U \cap V \in \mathcal{N}(x)$ such that $U \cap V \subset U$ and $U \cap V \subset V$. Thus $\exists (q, U \cap V) \in E$ such that $(q, U \cap V) \geq' (n, U)$ and $(q, U \cap V) \geq' (m, V)$. Therefore, (E, \geq') is a directed set.

Define $N : E \rightarrow D$ by $N(n, U) = n$. Again for any $(m, V) \in E$, there exists $m \in D$ such that $(n, U) \geq' (m, U)$ implies there exists $n \in D$ such that $N(n, U) = n$ and $n \geq m$.

It remains to prove that, T converges to x . Let $U \in \mathcal{N}_x$ be a nbd of x in X . We have, x is a cluster point of S . Therefore, $\forall m \in D, \exists p \in D$ such that $S(p) \in U$. By the construction of E , we have $(p, U) \in E$. Suppose $(n, V) \geq' (p, U)$, then $n \geq p$, $V \subset U$, and $S(n) \in V$. Clearly $S(n) \in U$. Therefore, $\forall (n, V) \geq' (p, U)$, $T(n, V) \in U$. That is, T is eventually in every nbd of x . ie, subnet T is convergent to x . Therefore, for each cluster point of the net S , there exists some subnet converging to it. \square

Remark. • Importance of Construction of E

If x is a cluster point of a net S in X , then S is frequently in some cofinal subset of D . Thus, if we consider any cofinal subset D' of D which is a direct set with \geq restricted to D' . Then $N : D' \rightarrow D$ defined by $N(n) = n$ gives a subnet $T : D' \rightarrow X$ of the net S . However, this subnet need not converge to x . The strongest statement, we can make on T is that ' x is a cluster point of T ', since $N : D \times \mathcal{N}(x) \rightarrow D$, $N(n) = n$ is completely independent of U . This problem is overcome by constructing E which is dependent on each nbd U of x .

- Existence of $q \in D$ such that $q \geq' p$ and $S(q) \in U$. We have, p follows both $n \& m$ and $U \cap V$ is a subset of both $U \& V$. However, since S is only frequently in U , p not necessarily be in U . But there is always someone following p which has its image in U . This q follows both $n \& m$, since \geq' is transitive.

7.5 Variations of Compactness

7.5.1 Variations of Compactness

In this chapter, we have two other notions of compactness - countable compactness and sequential compactness. (\star^{13})

Compact A topological space is compact iff every open cover of it has a finite subcover. ([Joshi, 1983, 6.1.1]) [Heine-Borel]

Countably Compact A topological space is countably compact iff every countable, open cover of it has a finite subcover. [Joshi, 1983, 11.1.1]

Sequentially Compact A topological space is sequentially compact iff every sequence in it has a convergent subsequence. [Joshi, 1983, 11.1.8] [Bolzano-Weierstrass]

Countable compactness is a weaker notion compared to compactness. (\star^{14}) However, sequentially compact and compact are not necessarily comparable. (\star^{15})

¹³For \mathbb{R} , Compactness & Sequentially compactness are equivalent to the completeness axiom.

¹⁴Every compact space is countably compact.

¹⁵ $\mathcal{T}_1, \mathcal{T}_2$ are non-comparable, if $\mathcal{T}_1 \not\subset \mathcal{T}_2$ and $\mathcal{T}_2 \not\subset \mathcal{T}_1$. [Joshi, 1983, 4.2.1]

We have seen earlier that compactness has the following properties 1. compactness is weakly hereditary. [Joshi, 1983, 6.1.10] 2. compactness is preserved under continuous functions. [Joshi, 1983, 6.1.8] 3. every continuous real functions on compact space is bounded and attains its extrema. [Joshi, 1983, 6.1.6] 4. every continuous real function on a compact, metric space is uniformly continuous by Lebesgue covering lemma. [Joshi, 1983, 6.1.7]

Countably compact spaces, Sequentially compact spaces have all the four properties listed above.

7.5.2 Countable compactness

Weakly hereditary property

A subspace (A, \mathcal{T}_A) being countably compact doesn't imply that (X, \mathcal{T}) is countably compact. However, if (X, \mathcal{T}) is a countably compact space and A is a closed subset of X , then (A, \mathcal{T}_A) is also a countably compact space. **In other words, countable compactness is weakly hereditary.**

Theorem 7.5.1. *Countable compactness is weakly hereditary. [Joshi, 1983, 11.1.3]*

Synopsis. Let A be a closed subset of countably compact space, X . If A has a countable open cover \mathcal{U} , then we can obtain a respective countable, open cover for X by attaching $X - A$ to the extensions of members of \mathcal{U} to X . This cover has a finite subcover. Then restricting them to A , we get a finite subcover of \mathcal{U} .

Proof. Suppose X is a countably compact space. And A is a closed subset of X . We need to show that A is countably compact. Without loss of generality, ^(★¹⁶) assume that A is a proper subset of X . Then $X - A$ is a non-empty, open subset of X .

Let \mathcal{U} be a countable open cover of A . Then $\mathcal{U} = \{U_1, U_2, \dots\}$ where each element $U_k \in \mathcal{U}$ is an open subset of A . Since A is a subspace of X , every open subset U_k in A is of the form $G \cap A$ for some open subset G in X . Therefore, there exists open subsets $V(U_k)$ for each U_k such that $A \cap V(U_k) = U_k$. ^{★¹⁷}

Define $\mathcal{V} = \{X - A, V(U_1), V(U_2), \dots\}$. Clearly, \mathcal{V} is a countable open cover ^(★¹⁸) of X . We have X is countably compact, thus \mathcal{V} has a finite subcover, say \mathcal{V}' . Without loss of generality assume ^(★¹⁹) that $X - A \in \mathcal{V}'$. Suppose $X - A \notin \mathcal{V}'$, then we can define another finite subcover $\mathcal{V}' \cup \{X - A\}$. Thus $\mathcal{V}' = \{X - A, V(U_{n_1}), V(U_{n_2}), \dots, V(U_{n_k})\}$.

Then the corresponding subcover $\mathcal{U}' = \{U_{n_1}, U_{n_2}, \dots, U_{n_k}\}$ is a finite subcover of \mathcal{U} . Since countable open cover \mathcal{U} and closed subset A are arbitrary, every closed subset of X with relative topology is countably compact. Therefore, countable compactness is weakly hereditary. \square

¹⁶Suppose A is not a proper subset of X . Then $X = A$ and A is countably compact.

¹⁷Relative topology, $\mathcal{T}_A = \{G \cap A : G \in \mathcal{T}\}$

¹⁸ $X - A$ is open in X . If $y \notin A$, then $y \in X - A$. If $y \in A$, then $y \in U_k$ for some k .

¹⁹Otherwise, you will have to consider two cases: $X - A \in \mathcal{V}'$ and $X - A \notin \mathcal{V}'$

Remark. Proof depends on the following,

1. There is an extension map, $\psi : P(A) \rightarrow P(X)$ that preserve open subsets (and closed subsets). This ψ is an open map which not a true inverse of the restriction, $r : P(X) \rightarrow P(A)$, defined by $r(G) = G \cap A$ for every subset G of X .
2. Also we rely on the subset A being closed. Suppose X have many countable open covers, but X has only uncountable open covers corresponding to a particular countable open cover of A . In such a case, X being countably compact is insufficient for A to be countably compact.

The behaviour of continuous functions

We will now study the nature of continuous functions defined on countably compact spaces. Suppose X, Y are topological space and function $f : X \rightarrow Y$ is continuous. If X is countably compact, then $f(X)$ is also countably compact. Continuous images of countably compact spaces are countably compact. **In other words, countable compactness is preserved under continuous functions.**(\star^{20})

Theorem 7.5.2. *Countable compactness is preserved under continuous functions. [Joshi, 1983, 11.1.2]*

Synopsis. Let X be countably compact and $f : X \rightarrow Y$ be continuous. Suppose \mathcal{U} is a countable cover of $f(X)$, then X has a countable cover \mathcal{V} obtained by taking inverse images. Since X is countably compact, \mathcal{V} has a finite subcover \mathcal{V}' . Now taking images of members of \mathcal{V}' , we get a finite subcover \mathcal{U}' of $f(X)$.

Proof. Suppose X is a countably compact space, Y is a topological space and $f : X \rightarrow Y$ is a continuous function. Let $\mathcal{U} = \{U_1, U_2, \dots\}$ be a countable cover of $f(X)$ by set open in $f(X)$. We have to show that \mathcal{U} has a finite subcover.

Define $\mathcal{V} = \{f^{-1}(U_1), f^{-1}(U_2), \dots\}$. Then \mathcal{V} is a countable open cover of X , since $f^{-1}(U_k)$ are open subsets of X and,

$$\begin{aligned} \bigcup_{k=1}^{\infty} U_k = f(X) &\implies f^{-1}\left(\bigcup_{k=1}^{\infty} U_k\right) = X \\ &\implies \bigcup_{k=1}^{\infty} f^{-1}(U_k) = X \end{aligned}$$

We have, \mathcal{V} is a countable open cover of X , which is a countably compact space. Therefore \mathcal{V} has a finite subcover $\mathcal{V}' = \{f^{-1}(U_{n_1}), f^{-1}(U_{n_2}), \dots, f^{-1}(U_{n_k})\}$.

$$\begin{aligned} \bigcup_{j=1}^k f^{-1}(U_{n_j}) = X &\implies f^{-1}\left(\bigcup_{j=1}^k U_{n_j}\right) = X \\ &\implies \bigcup_{j=1}^k U_{n_j} = f(X) \end{aligned}$$

²⁰A topological property is preserved under continuous functions if whenever a space has that property so does every continuous image of it. [Joshi, 1983, 6.1.9]

Clearly $\mathcal{U}' = \{U_{n_1}, U_{n_2}, \dots, U_{n_k}\}$ is a finite subcover of \mathcal{U} . Thus every countable open cover of $f(X)$ by sets open in $f(X)$ has a finite subcover. Therefore, continuous images of countably compact spaces are countably compact. \square

Remark. 1. For a continuous function, $f : X \rightarrow Y$ the inverse images of open subsets are open in X . The relation $f^{-1} \subset f(X) \times X$ is not a function. However, we may consider a function, $\psi : P(Y) \rightarrow P(X)$ such that $\psi(U) = f^{-1}(U)$ for any subset U of Y . This ψ is an open map which maps open subsets of Y to open subsets of X .

Theorem 7.5.3. *Every continuous, real-valued function on a countably compact, metric space is bounded and attains its extrema. [Joshi, 1983, 11.1.7]*

Synopsis. Let X be a countably compact space and function $f : X \rightarrow \mathbb{R}$ be continuous. Then $f(X) \subset \mathbb{R}$ is countably compact. Real line \mathbb{R} is metrisable(\star^{21}). Then $f(X)$ is countably compact, metric space. Therefore $f(X)$ compact(\star^{22}). The subset $f(X)$ of \mathbb{R} is bounded and closed, since every compact subset of \mathbb{R} is bounded and closed. Thus $f(X)$ contains its supremum and infimum. Therefore, f is bounded and attains its extrema.

Proof. Let X be a countably compact space and $f : X \rightarrow \mathbb{R}$ be continuous, real-valued function on the countably compact space, X . We have to show that f is bounded and attains its extrema.

Since countable compactness is preserved under continuous functions, $f(X)$ is countably compact subset of \mathbb{R} . Since, $f(X)$ is a subset of the metric space, \mathbb{R} and metrisability is hereditary, $f(X)$ is again metrisable. (suppose) We have, every countably compact, metric space is compact. Then $f(X)$ is a compact subset of \mathbb{R} .

Since every compact subset of \mathbb{R} is bounded and closed, $f(X)$ is bounded and closed. Since every closed subset of \mathbb{R} contains supremum and infimum, $f(X)$ contains its extrema. Therefore, every continuous, real-valued function on a countably compact space is bounded and attains its extrema.

We have assumed that every countably compact, metric space is compact. This result will be proved in the last section of this chapter. \square

Remark. Since countably compact, metric spaces are compact. The above theorem can be used to prove that continuous, real-valued functions on a compact, metric space attains its extrema.

Due to the Lebesgue covering lemma, next result is quite simple.*

Theorem 7.5.4. *Every continuous, real-valued function on a countably compact, metric space is uniformly continuous.*

Proposition 7.5.1. *Let X be a first countable, Hausdorff space. Then every countably compact subset A of X is closed. [Joshi, 1983, Exercises 11.1.7]*

²¹[Joshi, 1983, 4.2 Example 4], \mathbb{R} with usual metric $d : \mathbb{R} \rightarrow \mathbb{R}$, $d(x, y) = |x - y|$

²²[Joshi, 1983, 11.1.11] On metric spaces, countable compactness \implies compactness.

7.5.3 Sequential Compactness

Weakly hereditary property

Theorem 7.5.5. *Sequential compactness is weakly hereditary. [Joshi, 1983, Exercises 11.1.3]*

The behaviour of continuous functions

Theorem 7.5.6. *Sequential compactness is preserved under continuous functions. [Joshi, 1983, Exercises 11.1.4]*

Synopsis. Let X be sequentially compact and function $f : X \rightarrow Y$ be continuous. Then any sequence, $\{y_k\}$ in $f(X)$ has a sequence, $\{x_k\}$ in X such that $f(x_k) = y_k$. Sequence $\{x_k\}$ has a subsequence $\{x_{n_k}\}$ converging to x , then sequence $\{f(x_{n_k})\}$ in $f(X)$ has the subsequence $\{f(x_{n_k})\}$ converging to $f(x)$.

Proof. Let X be a sequentially compact space, function $f : X \rightarrow Y$ be continuous and $\{y_n\}$ be a sequence in $f(X)$ subset of Y . Construct a sequence $\{x_n\}$ such that $f(x_k) = y_k, \forall k$.

Every sequence in X has a convergent subsequence. Thus $\{x_n\}$ has a subsequence $\{x_{n_k}\}$ converging to $x \in X$. The image of this subsequence $\{f(x_{n_k})\}$ is a subsequence of $\{y_k\}$. We claim that, $\{f(x_{n_k})\}$ converges to $f(x) \in f(X)$.

Let U be an open subset containing $f(x)$, then $f^{-1}(U)$ is an open subset containing x . Since $\{x_{n_k}\}$ converges to x . There exists an integer n such that for every $k \geq n, x_k \in f^{-1}(U)$. Clearly, for each $k \geq n, f(x_k) \in U$. Since U is arbitrary, $\{f(x_{n_k})\}$ converges to $f(x)$. Therefore, the image of any sequentially compact space is sequentially compact. In other words, sequentially compactness is preserved under continuous functions. \square

Remark. 1. Given a sequence $\{y_n\}$ in $f(X)$, there is a sequence of subsets $\{U_n\}$ in $P(Y)$ such that $U_n = f^{-1}(y_n)$. Since each U_n is non-empty, we can construct a sequence $\{x_n\}$ in X using a choice function. The convergent subsequence of $\{y_n\}$ depends on the selection of this choice function.

Given every sequentially compact, metric space is countably compact. We may assert the properties of countably compact, metric spaces on sequentially compact, metric spaces.

Theorem 7.5.7. *Every continuous, real-valued function on a sequentially compact, metric space is bounded and attains its extrema.*

Theorem 7.5.8. *Every continuous, real-valued function on a sequentially compact, metric space is uniformly continuous. [Joshi, 1983, Exercises 11.1.6]*

7.5.4 Countable Compactness on T_1 spaces

In this section, we are going to see four different characterisations of countable compactness in T_1 spaces. The first two characterisations doesn't have anything to do with the T_1 axiom.

T_1 Space A topological space X satisfy T_1 axiom if for any two distinct points $x, y \in X$, there exists an open subset $U \subset X$ containing x but not y . [Joshi, 1983, 7.1.2]

countable compactness A topological space is countably compact if every countable open cover has a finite subcover. [Joshi, 1983, 11.1.1]

finite intersection property A family \mathcal{F} of subsets of X has finite intersection property(f.i.p.) if every finite subfamily of \mathcal{F} has a non-empty intersection. [Joshi, 1983, 10.2.6]

accumulation point A point $x \in X$ is accumulation point of a subset $A \subset X$ if every open subset containing x has atleast one point of A other than x . [Joshi, 1983, 5.2.7]

limit point A point $x \in X$ is a limit point of a sequence $\langle x_k \rangle$ in X if for every open subset U containing x , there exists an integer $N \in \mathbb{N}$ such that $x_k \in U$ for every $k \geq N$. [Joshi, 1983, 4.1.7]

cluster point A point $x \in X$ is a cluster point of a sequence $\langle x_k \rangle$ in X if for any neighbourhood V of x , the sequence $\langle x_k \rangle$ assumes a point in V infinitely many times. (\star^{23})

Countable compactness in T_1 spaces

Theorem 7.5.9. In a T_1 space X , following statements are equivalent,

1. X is countably compact
2. Every countably family of closed subsets of X with finite intersection property have non-empty intersection.
3. Every infinite subset $A \subset X$ has an accumulation point. (\star^{24})
4. Every sequence $\langle x_k \rangle$ in X has a cluster point.
5. Every infinite open cover of X has a proper subcover. [Arens-Dugundji]

Proof. 1 \implies 2

Suppose X is countably compact. Let $\mathcal{C} = \{C_1, C_2, \dots\}$ be a countable family of closed subsets of X with empty intersection. Define $\mathcal{U} = \{X - C_1, X - C_2, \dots\}$ is a family of open subsets of X . By de Morgan's law, (\star^{25})

$$\bigcap_{k=1}^{\infty} C_k = \phi, \text{ then } X = X - \left(\bigcap_{k=1}^{\infty} C_k \right) = \bigcup_{k=1}^{\infty} (X - C_k)$$

We have \mathcal{U} is a countable cover of X and X is countably compact space. Thus \mathcal{U} has a finite subcover $\mathcal{U}' = \{X - C_{n_1}, X - C_{n_2}, \dots, X - C_{n_k}\}$.

$$\mathcal{U}' \text{ is a cover of } X, \text{ then } X = \bigcup_{j=1}^k (X - C_{n_j})$$

²³ x is a cluster point of $\langle x_k \rangle$ if for every integer N , there exists $k > N$ such that $x_k \in V$. In other words, $\langle x_k \rangle$ is frequently in V . [Joshi, 1983, 10.1.9]

²⁴Every infinite subset of \mathbb{R} has a limit point is equivalent to the completeness axiom.

²⁵Complement of Intersection = Union of complements, $X - (C \cap D) = (X - C) \cup (X - D)$,

$$X - \bigcup_{j=1}^k (X - C_{n_j}) = \bigcap_{j=1}^k (X - (X - C_{n_j})) = \bigcap_{j=1}^k C_{n_j} = \phi$$

Now $\mathcal{C}' = \{C_{n_1}, C_{n_2}, \dots, C_{n_k}\}$ has empty intersection. This is a contradiction to the finite intersection property of \mathcal{C} . Thus \mathcal{C} has non-empty intersection. Therefore, every countably family of closed subsets of X have non-empty intersection.

2 \implies 1

Let $\mathcal{U} = \{U_1, U_2, \dots\}$ be a countable cover of X . Then $\mathcal{C} = \{X - U_1, X - U_2, \dots\}$ is a countable family of closed subsets of X .

Let $\mathcal{U}' = \{U_{n_1}, U_{n_2}, \dots, U_{n_k}\}$ be any finite subfamily of \mathcal{U} . Suppose X is not countably compact, then \mathcal{U} doesn't have a finite subcover. Therefore, \mathcal{U}' is not a cover of X . And \mathcal{C} is a family of closed subsets with finite intersection property.

Therefore by assumption, the countable family of closed subsets \mathcal{C} has a non-empty intersection.

$$\bigcap_{k=1}^{\infty} C_k \neq \phi, \text{ then } \bigcap_{k=1}^{\infty} C_k = \bigcap_{k=1}^{\infty} (X - U_k) = X - \left(\bigcup_{k=1}^{\infty} U_k \right) \neq \phi$$

Then \mathcal{U} is not a cover of X as well. This is a contradiction, therefore X is countably compact.

1 \implies 3

Suppose X is countably compact. Let A be an infinite subset of X . Suppose A doesn't have an accumulation point.

Let B be a countably infinite subset of A . Then B also doesn't have any accumulation point. Therefore, the derived set B' is empty. Thus B is a closed subset of X . Since countable compactness is weakly hereditary, subspace B is again countably compact.

For each point $b \in B$, there is an open subset V_b such that $V_b \cap B = \{b\}$, since $b \in B$ is not an accumulation point. Thus $\mathcal{U} = \{V_b \cap B : b \in B\}$ is a countable open cover of B . Clearly, \mathcal{U} doesn't have any finite subcover.

This is a contradiction to B being countably compact. Therefore, A has an accumulation point. \square

7.5.5 Variations of Compactness on Metric Spaces

In this document, we will see that from metric space point of view these two notions were equivalent to the compactness and were used alternatively. For example : in functional analysis (semester 3), you will find definitions like 'a normed space is compact iff every sequence in it has a convergent subsequence', which is clearly sequential compactness for a topologist.

Lindeloff A topological space is Lindeloff iff every open cover has a countable subcover.

First countable A topological space is first countable iff every point in it has a countable local base.

Second countable A topological space is second countable iff it has a countable base.

Base A family of subsets \mathcal{B} of X is a base of a topological space if every open subset can be expressed as union of some members of \mathcal{B}

Base Characterisation A family of subsets \mathcal{B} of X is a base of a topological space iff for every $x \in X$, and for every neighbourhood U of x , there is a member $B \in \mathcal{B}$ such that $x \in B \subset U$.

Local Base A family of subsets \mathcal{L} of X is a local base at point $x \in X$ if for every neighbourhood U of x , there is a member $L \in \mathcal{L}$ such that $x \in L \subset U$.

Equivalence

We are going to see when these three notions: compactness, countable compactness and sequentially compactness are equivalent.

Theorem 7.5.10. *Countably compact, metric spaces are second countable.*

Synopsis. For every positive real number r , there exists a non-empty maximal subsets A_r with every pair of points atleast r distance apart. A_r are finite. The union of maximal subsets $A_{\frac{1}{n}}$ for each natural number n is a countable, dense subset D of X . Thus countably compact, metric spaces are separable. The family \mathcal{B} of all open balls with center at $d \in D$ and rational radius is a countable, base for X . Thus countably compact, metric spaces are second countable.

Proof. Let $(X; d)$ be a countably compact, metric space. For each positive real number $r \in \mathbb{R}$, $r > 0$ construct a family of subsets $A_r \subset X$ such that it is a maximal set of points which are atleast r distances apart.

Then A_r is finite for every positive real number r . Suppose A_r is infinite for some real number $r > 0$, then A_r has a accumulation point, say x by the Characterisation of countable compactness of X .

Then every neighbourhood of x must intersect A_r at infinitely many points, since every metric space is a T_1 space. Consider $B(x, \frac{r}{2})$. Since any two points of $B(x, \frac{r}{2})$ are less than r distances apart, the intersection $B(x, \frac{r}{2}) \cap A_r$ can have atmost one point in it. Thus for every positive real number r , A_r is finite.

Define $D = \bigcup_{n=1}^{\infty} A_{\frac{1}{n}}$. We claim that D is a countable, dense subset of X .

Let $x \in X$ and $B(x, r)$ be an open ball containing x , then there exists integer $n \in \mathbb{N}$ such that $\frac{1}{n} < r$. ^{★²⁶}

²⁶By archimedean property of integers, we have $\forall r \in \mathbb{R}$, $r > 0$, $\exists n \in \mathbb{N}$ such that $nr > 1$.

Then $B(x, r) \cap D \neq \phi$, since $B(x, r) \cap A_{\frac{1}{n}} \neq \phi$. Suppose $B(x, r) \cap A_{\frac{1}{n}} = \phi$, then $A_{\frac{1}{n}}$ is not maximal. Since, x is at least $r > \frac{1}{n}$ distance apart from each point of $A_{\frac{1}{n}}$. Therefore, D intersects with every open subset and thus dense in X .

We have have a countable, dense subset D of X . Therefore, X is separable. Now define $\mathcal{B} = \{B(x, r) : r \in \mathbb{Q}, x \in D\}$. Clearly, \mathcal{B} is a countable base for X . By the construction of \mathcal{B} , X is second countable. (\star^{27}) \square

Countable Compactness, Lindeloff \iff Compactness

Theorem 7.5.11. *A topological space X is compact iff it is countably compact, Lindeloff space.*

Proof. Let X be a compact space. Let \mathcal{U} be a countable open cover of X , then \mathcal{U} has a finite subcover \mathcal{U}' . Therefore, every compact space is countably compact. (\star^{28})

Conversely, suppose X is a countably compact, Lindeloff space. Since X is Lindeloff, every open cover \mathcal{U} has a countable subcover \mathcal{U}' . Since X countably compact, every countable open cover \mathcal{U}' has a finite subcover \mathcal{U}'' . Thus every open cover \mathcal{U} has a finite subcover \mathcal{U}'' . Therefore every countably compact, Lindeloff space is compact. \square

Countable Compactness, First Countable \implies Seq. Compactness

Theorem 7.5.12. *Every countably compact, first countable space is Sequentially compact.*

Proof. Let X be a countably compact, first countable space. Let $\{x_n\}$ be a sequence in X . By, equivalent conditions (\star^{29}) of countably compact spaces, every sequence in countably compact space X has a cluster point, say x . We have, X is first countable. Therefore, X has a countable local base \mathcal{L} at $x \in X$. How to construct a subsequence of $\{x_n\}$ converging to x ? (\star^{30}) \square

Remark. Every sequentially compact space is countably compact. \star

Theorem 7.5.13. *In a second countable space, all the three forms of compactness are equivalent. [Joshi, 1983, 11.1.10]*

Proof. Every second countable space is both first countable and Lindeloff. Every countably compact, Lindeloff space is countably compact. Therefore every countably compact, second countable space compact. Again, every countably compact, first countable space is sequentially compact. Therefore every countably compact, second countable space is sequentially compact. Conversely, every compact space is countably compact and every sequentially compact space is countably compact. (\star^{31}) \square

²⁷Every separable, metric space is second countable.

²⁸Countable compactness is a weaker notion than compactness.

²⁹[Joshi, 1983, 11.1] Conditions 1, 2, and 4 are equivalent. $2 \implies 4$ without T_1 axiom is out of scope.

³⁰[Joshi, 1983, Exercises 10.1.11]

³¹Countable compactness is a weaker notion than sequential compactness as well.

Theorem 7.5.14. *In a metric space, all the three forms of compactness are equivalent. [Joshi, 1983, 11.1.11]*

Proof. In a metric space each form of compactness implies second countability. And in second countable spaces, they are all equivalent. \square

7.6 Homotopy of Paths

Definitions 7.6.1. Let X, Y be topological spaces and $f : X \rightarrow Y, f' : X \rightarrow Y$ be continuous functions. Then f, f' are homotopic if there exists a continuous function $F : X \times I \rightarrow Y$ such that for every $x \in X, F(x, 0) = f(x)$ and $F(x, 1) = f'(x)$. And we write, $f \simeq f'$.

Definitions 7.6.2. Let X be a topological space and $f : I \rightarrow X$ and $f' : I \rightarrow X$ be two paths. Then f, f' are path-homotopic if they have same initial point x_0 (ie, $x_0 = f(0) = f'(0)$), same final point x_1 (ie, $x_1 = f(1) = f'(1)$) and they are homotopic (ie, $\exists F : I \times I \rightarrow X$ such that $\forall x \in I, F(x, 0) = f(x)$ and $F(x, 1) = f'(x)$ also fixed at the end points x_0 and x_1 (ie, $\forall t \in I, F(0, t) = x_0$ and $F(1, t) = x_1$). And we write $f \simeq_p f'$.

Remark. If two paths f, f' are homotopic, then they have the same end points and there exists a (topologically) continuous deformation from one path into another.

Proposition 7.6.1. *The relations \simeq, \simeq_p are equivalence relations.*

Proof. Homotopy : Let f, f' be continuous functions from X into Y . Then f and f' are homotopic, $f \simeq f' \iff \exists F : X \times I \rightarrow Y$ such that F is continuous, $F(x, 0) = f(x)$, and $F(x, 1) = f'(x)$

1. $f \simeq f$

We have $f : X \rightarrow Y$ is continuous. Define $F : X \times I \rightarrow Y$ such that $F(x, t) = f(x)$. Clearly, F is continuous, $F(x, 0) = f(x)$ and $F(x, 1) = f(x)$. And $\exists F : X \times I \rightarrow Y \implies f \simeq f$.

2. $f \simeq f' \implies f' \simeq f$

We have, $f \simeq f'$. Thus there exists a continuous function $F : X \times I \rightarrow Y$ such that $F(x, 0) = f(x)$ and $F(x, 1) = f'(x)$.

Consider $F' : X \times I \rightarrow Y$ defined by $F'(x, t) = F(x, 1 - t)$. Clearly, F' is continuous, $F'(x, 0) = F(x, 1) = f'(x)$, and $F'(x, 1) = F(x, 0) = f(x)$. Thus, $\exists F'(x, t) : X \times I \rightarrow Y \implies f' \simeq_p f$

3. $f \simeq f', f' \simeq f'' \implies f \simeq f''$

We have, $f \simeq f' \iff \exists F : X \times I \rightarrow Y$ such that F is continuous, $F(x, 0) = f(x)$ and $F(x, 1) = f'(x)$.

Similarly, $f' \simeq f'' \iff \exists F' : X \times I \rightarrow Y$ such that F' is continuous, $F'(x, 0) = f'(x)$ and $F'(x, 1) = f''(x)$.

Consider $G : X \times I \rightarrow Y$ defined by

$$G(x, t) = \begin{cases} F(x, 2t) & , t \in [0, \frac{1}{2}] \\ F'(x, 2t - 1) & , t \in [\frac{1}{2}, 1] \end{cases}$$

We have, $G(x, \frac{1}{2}) = F(x, 1) = F'(x, 0) = f'(x)$. Thus, G is continuous by pasting lemma since $[0, \frac{1}{2}] \cap [\frac{1}{2}, 1] = \{\frac{1}{2}\}$. Also $G(x, 0) = F(x, 0) = f(x)$ and $G(x, 1) = F'(x, 1) = f''(x)$. Thus, $\exists G : X \times I \rightarrow Y \implies f \simeq f''$

Path Homotopy : Let f, f', f'' be paths in X . Then f and f' are path homotopic, $f \simeq_p f' \iff \exists F : I \times I \rightarrow X$ such that F is continuous, $\forall s \in [0, 1], F(s, 0) = f(s), F(s, 1) = f'(s)$ and $\forall t \in [0, 1], F(0, t) = f(0) = f'(0), F(1, t) = f(1) = f'(1)$

1. $f \simeq_p f$

We have $f : I \rightarrow X$ is continuous. Define $F : I \times I \rightarrow X$ such that $\forall s, t \in [0, 1], F(s, t) = f(s)$. Clearly, F is continuous, $\forall s \in [0, 1], F(s, 0) = f(s), F(s, 1) = f(s)$ and $\forall t \in [0, 1], F(0, t) = f(0), F(1, t) = f(1)$. Thus, $\exists F : I \times I \rightarrow X \implies f \simeq_p f$.

2. $f \simeq_p f' \implies f' \simeq_p f$

We have, $f \simeq_p f'$. Thus there exists a continuous function $F : I \times I \rightarrow X$ such that $\forall s \in [0, 1], F(s, 0) = f(s), F(s, 1) = f'(s)$ and $\forall t \in [0, 1], F(0, t) = f(0) = f'(0), F(1, t) = f(1) = f'(1)$.

Consider $F' : I \times I \rightarrow X$ defined by $F'(s, t) = F(s, 1 - t)$. Clearly, F' is continuous. And $F'(s, 0) = F(s, 1) = f'(s)$, and $F'(s, 1) = F(s, 0) = f(s)$. Also, $F'(0, t) = F(0, 1 - t) = f(0) = f'(0)$ and $F'(1, t) = F(1, 1 - t) = f(1) = f'(1)$. Thus, $\exists F'(s, t) : I \times I \rightarrow X \implies f' \simeq_p f$

3. $f \simeq f', f' \simeq f'' \implies f \simeq f''$

We have, $f \simeq f' \iff \exists F : I \times I \rightarrow X$ such that F is continuous, $\forall s \in [0, 1], F(s, 0) = f(s), F(s, 1) = f'(s)$ and $\forall t \in [0, 1], F(0, t) = f(0) = f'(0), F(1, t) = f(1) = f'(1)$

Similarly, $f' \simeq f'' \iff \exists F' : I \times I \rightarrow X$ such that F' is continuous, $\forall s \in [0, 1], F'(s, 0) = f'(s), F'(s, 1) = f''(s)$ and $\forall t \in [0, 1], F'(0, t) = f'(0) = f''(0), F'(1, t) = f'(1) = f''(1)$

Consider $G : I \times I \rightarrow X$ defined by

$$G(s, t) = \begin{cases} F(s, 2t) & , t \in [0, \frac{1}{2}] \\ F'(s, 2t - 1) & , t \in [\frac{1}{2}, 1] \end{cases}$$

We have, $G(s, \frac{1}{2}) = F(s, 1) = F'(s, 0) = f'(s)$. Thus, G is continuous by pasting lemma[Munkres, 2003, §18.3 pp. 106], since $[0, \frac{1}{2}] \cap [\frac{1}{2}, 1] = \{\frac{1}{2}\}$.

Also $G(s, 0) = F(s, 0) = f(s)$ and $G(s, 1) = F'(s, 1) = f''(s)$.

Again, $\forall t \in [0, \frac{1}{2}], G(0, t) = F(0, 2t) = f(0) = f'(0) = f''(0)$ and $\forall t \in [\frac{1}{2}, 1], G(0, t) = F'(0, 2t - 1) = f(0) = f'(0) = f''(0)$. Therefore, $\forall t \in [0, 1], G(0, t) = f(0) = f''(0)$.

Similarly, $\forall t \in [0, \frac{1}{2}]$, $G(1, t) = F(1, 2t) = f(1) = f'(1) = f''(1)$ and $\forall t \in [\frac{1}{2}, 1]$, $G(1, t) = F'(1, 2t - 1) = f(1) = f'(1) = f''(1)$. Therefore, $\forall t \in [0, 1]$, $G(1, t) = f(1) = f''(1)$. Thus, $\exists G : I \times I \rightarrow X \implies f \simeq_p f''$

□

Definitions 7.6.3. Let f be a path in X (ie, $f : I \rightarrow X$), then $[f]$ is the equivalence class of all paths homotopic to f in X . (ie, $g \in [f] \iff f \simeq_p g$)

Remark. The set of homotopy classes of functions from X into Y is denoted by $[X, Y]$. And, the set of all path-homotopic classes on X is denoted by $[I, X]$.

Remark (Straight-line homotopy). [Munkres, 2003, §51 Example 1 pp. 320] Let X be a topological space, and f, g be continuous functions from X into a euclidean space, say \mathbb{R}^2 . Then f, g are straight line homotopic if there exists a continuous function F from $X \times I$ such that F deforms f into g along straight line segments joining them.

For example, $F(x, t) = (1 - t)f(x) + tg(x)$.

Remark. Let A be a convex subspace of \mathbb{R}^n . Then any two paths in A from x_0 to x_1 are path homotopic in A .

Proof. —continue page 321—

□

Remark. [Munkres, 2003, §51 Example 2 pp. 321] **This demonstrates that the straight-line homotopy is very sensitive to the holes in the space.**

Definitions 7.6.4. Let f, g be two paths in X (ie, $f : I \rightarrow X$ and $g : I \rightarrow X$) such that $f(0) = x_0, f(1) = g(0) = x_1$ and $g(1) = x_2$. Then the product $h = f * g$ is given by $h : I \rightarrow X$ and

$$h(s) = \begin{cases} f(2s) & , s \in [0, \frac{1}{2}] \\ g(2s - 1) & , s \in [\frac{1}{2}, 1] \end{cases}$$

This h is well-defined, and continuous by pasting lemma.^(★³²)

Definitions 7.6.5. The product operation on path-homotopy classes is defined by $[f] * [g] = [f * g]$.

Remark. The product of path-homotopic classes is well-defined.

Proof. Let F be a path-homotopy between $f, f' \in [f]$ and G be a path-homotopy between $g, g' \in [g]$. Then $H : I \times I \rightarrow X$ defined by

$$H(s, t) = \begin{cases} F(2s, t) & s \in [0, \frac{1}{2}] \\ G(2s - 1, t) & s \in [\frac{1}{2}, 1] \end{cases}$$

Then H is well-defined, and continuous by pasting lemma.

³²Pasting Lemma : Let $X = A \cup B$, where A and B are closed in A . Let $f : A \rightarrow Y$ and $g : B \rightarrow Y$ be continuous. If $f(x) = g(x)$ for every $x \in A \cap B$, then f and g combine to give a continuous function $h : X \rightarrow Y$, defined by setting $h(x) = f(x)$ if $x \in A$ and $h(x) = g(x)$ if $x \in B$.

$\forall s \in [0, \frac{1}{2}]$, $H(s, 0) = F(2s, 0) = f(2s)$ and
 $\forall s \in [\frac{1}{2}, 1]$, $H(s, 0) = G(2s - 1, 0) = g(2s - 1)$.
 $\implies H(s, 0) = (f * g)(s)$, by the definition of $f * g$

$\forall s \in [0, \frac{1}{2}]$, $H(s, 1) = F(2s, 1) = f'(2s)$ and
 $\forall s \in [\frac{1}{2}, 1]$, $H(s, 1) = G(2s - 1, 1) = g'(2s - 1)$.
 $\implies H(s, 1) = (f' * g')(s)$, by the definition of $f' * g'$

$H(0, t) = F(0, t) = f(0) = x_0 = (f * g)(0)$, and
 $H(1, t) = G(1, t) = g'(1) = x_2 = (f' * g')(1)$

Then $H : I \times I \rightarrow X$ is a path-homotopy between $f * g$ and $f' * g'$. \square

Definitions 7.6.6 (Groupoid). Let G be a set and $*$ be a binary operation on G . Then $(G, *)$ is a groupoid if it satisfies the following axioms

- g1 Associativity - $\forall x, y, z \in G$, $(x * y) * z = x * (y * z)$
- g2 Existence of left and right identities - There exist unique elements e_L and e_R such that $\forall x \in G$, $x * e_R = x$ and $e_L * x = x$.
- g3 Existence of inverse
 $\forall x \in G$, $\exists x^{-1} \in G$ such that $x * x^{-1} = e_L$ and $x^{-1} * x = e_R$

Definitions 7.6.7 (Positive Linear Map). A positive linear map $p : [a, b] \rightarrow [c, d]$ is the unique map of the form $p(x) = mx + k$ such that $p(a) = c$ and $p(b) = d$. Clearly, scaling factor, $m = \frac{d-c}{b-a}$ as we want to transform an interval of length $b - a$ into an interval of length $d - c$. And offset k is given by,

$$p(a) = \frac{d-c}{b-a}a + k = c \implies k = c - \frac{a(d-c)}{b-a} = \frac{bc-ad}{b-a}$$

But, we won't fix m and k in $p(x) = mx + k$, instead we will focus on the unique map with graph of positive slope and passing through required end points. The graph of a positive linear map from $[a, b]$ to $[c, d]$ is always a straight-line with positive slope.

Remark. The inverse of a positive linear map is also a positive linear map. Given $p : [a, b] \rightarrow [c, d]$, $p(x) = mx + k$, where $m = \frac{d-c}{b-a}$, $k = \frac{bc-ad}{b-a}$. Then its inverse, $\bar{p} : [c, d] \rightarrow [a, b]$ is given by $\bar{p}(y) = m'y + k'$, where $m' = \frac{b-a}{d-c} = \frac{1}{m}$, $k' = \frac{ad-bc}{d-c} = \frac{-k}{m}$. Clearly $m > 0 \implies m' = \frac{1}{m} > 0$.

Remark. The composite of two positive linear maps is also a (piece-wise) positive linear map. Let f, g be two positive linear maps. Then their composite map $f * g$ is given by

$$(f * g)(x) = \begin{cases} f(2x) & x \in [a, \frac{a+b}{2}] \\ g(2(x - \frac{b-a}{2})) & x \in [\frac{a+b}{2}, b] \end{cases}$$

Remember $f * g$ exists only if $f(b) = g(a)$. Therefore, $f * g$ is a well-defined, continuous (by pasting lemma) and (piecewise) positive linear map.

Lemma 7.6.1. *Let f, f' be two paths in X and $k : X \rightarrow Y$ be a continuous function. Let F be the path homotopy in X between the paths f and f' . Then $k \circ F$ is a path homotopy in Y between the paths $k \circ f$ and $k \circ f'$. That is, path homotopy is preserved under a continuous function.*

Lemma 7.6.2. *Let f, g be two paths in X with $f(1) = g(0)$ and $k : X \rightarrow Y$ be a continuous function. Then $k \circ (f * g) = (k * f) \circ (k * g)$*

Theorem 7.6.3. *Let f, g, h be paths in a topological space X , and $[f], [g], [h]$ be respective path-homotopy classes. Suppose the operation product, $*$ is defined by*

$$[f] * [g] = [f * g] \text{ where } (f * g)(s) = \begin{cases} f(2s) & s \in [0, \frac{1}{2}] \\ g(2s - 1) & s \in [\frac{1}{2}, 1] \end{cases}$$

Then the product, $$ has the following properties :*

1. *Associativity*

$$\forall [f], [g], [h] \in [I, X], ([f] * [g]) * [h] = [f] * ([g] * [h])$$

2. *Existence of left and right identities*

*Let $e_x : I \rightarrow X$ defined by $\forall s \in [0, 1], e_x(s) = x$. Let f be a path from x_0 to x_1 , then there exist unique paths e_{x_0} and e_{x_1} such that $[f] * [e_{x_1}] = [f]$ and $[e_{x_0}] * [f] = [f]$. That is, e_{x_0}, e_{x_1} are respectively the left and right path-homotopy-identities.*

3. *Existence of inverse*

*Let f be a path in X . The path \bar{f} , defined by $\bar{f}(s) = f(1 - s)$ is the reverse path of f . Then $[f] * [\bar{f}] = [e_{x_0}]$ and $[\bar{f}] * [f] = [e_{x_1}]$. That is, the inverse of class of f is the class of reverse path of f .*

In other words, Set of all path-homotopy classes together with binary operation product, $$ is a groupoid. ie, $([I, X], *)$ is a groupoid.*

Proof. Step 1 : Properties 2&3

Let $e_0 : I \rightarrow I$ such that $e_0(t) = 0, \forall t \in I$. And $i : I \rightarrow I$ such that $i(t) = t, \forall t \in I$. Then $e_0 * i$ is also a path. Since I is convex, there is a path homotopy (\star^{33}) G between i and $e_0 * i$. Let $f : I \rightarrow X$ be continuous path in X from x_0 to x_1 . Then $f \circ G$ is a path homotopy (by Lemma 2) in X between $f \circ i$ and $f \circ e_0 * i$ where $f \circ i$ and $f \circ e_0$ are paths from x_0 to x_1 in X .

$$f \circ (e_0 * i) = (f \circ e_0) * (f \circ i), \text{ by Lemma 1}$$

$$= e_{x_0} * f, \text{ since } \forall s \in I, f(e_0(s)) = x_0 = e_{x_0}(s) \text{ and } f * i \simeq_p f$$

Therefore $[e_{x_0}] * [f] \simeq_p [f]$, since $e_0 * i \simeq_p i$, and $f \circ (e_0 * i) \simeq_p f \circ i = f$.

Similarly, $e_1 : I \rightarrow I$ such that $e_1(t) = 1$. Let H be a path homotopy (\star^{34}) between $i * e_1$ and i . Thus, $f \circ H$ is a path homotopy in X from $f \circ (i * e_1)$ and $f \circ i$.

$$f \circ (i * e_1) = (f \circ i) * (f \circ e_1), \text{ by Lemma 1}$$

$$= f * e_{x_1}, \text{ since } f * i \simeq_p f, i * e_1 \simeq_p e_1, \forall s \in I, (f(e_1(s))) = x_1 = e_{x_1}(s)$$

³³ $G : I \times I \rightarrow I, G(s, 0) = i(s), G(s, 1) = (e_0 * i)(s), G(0, t) = 0, G(1, t) = 1.$

³⁴ $H : I \times I \rightarrow I, H(s, 0) = (i * e_1)(s), H(s, 1) = i(s), H(0, t) = 0, H(1, t) = 1$

Since $i * e_1 \simeq_p i$, we have $f \circ (i * e_1) \simeq_p f \circ i = f$. Therefore $[f] * [e_{x_1}] \simeq_p [f]$. Thus, $[f] * [e_{x_1}] \simeq_p [f] \simeq_p [e_{x_0}] * [f]$. Therefore, $[f]$ has left and right inverses ie, property 2 holds.

Consider inverse path $\bar{i} : I \rightarrow I$, $\bar{i}(s) = 1 - s$. Then $i * \bar{i}$ is a path in I with both end points at 0. We have, $e_0 : I \rightarrow I$, $e_0(s) = 0$ is also a path with both end points at 0. Since I is convex, there is a path homotopy H in I between e_0 and $i * \bar{i}$. Then $f \circ H$ is a path homotopy between $f \circ e_0 = e_{x_0}$ and $f \circ (i * \bar{i}) = (f \circ i) * (f \circ \bar{i}) = f * \bar{f}$. Therefore, $[e_{x_0}] \simeq_p [f] * [\bar{f}]$.

Similarly $\bar{i} * i$ and e_1 are paths with both end points at 1. Since I is convex, there is a path homotopy G in I between $\bar{i} * i$ and e_1 . Then $f \circ G$ is a path homotopy between $f \circ (\bar{i} * i) = (f \circ \bar{i}) * (f \circ i) = \bar{f} * f$ and $f \circ e_1 = e_{x_1}$. Therefore, $[\bar{f}] * [f] \simeq_p [e_{x_1}]$. Thus the path $\bar{f} : I \rightarrow X$, $\bar{f}(s) = f(1 - s)$, $\forall s \in I$ is reverse of f . Also $[f] * [\bar{f}] = [e_{x_0}]$ and $[\bar{f}] * [f] = [e_{x_1}]$. ie, property 3 holds.

Step 2 : Property 1

Let f, g, h be three paths in X and $f(1) = g(0) = x_1$ and $g(1) = h(0) = x_2$. Then $f * (g * h)$ is defined by

$$(g * h)(s) = \begin{cases} g(2s) & s \in [0, \frac{1}{2}] \\ h(2s - 1) & s \in [\frac{1}{2}, 1] \end{cases}$$

$$(f * (g * h))(s) = \begin{cases} f(2s) & s \in [0, \frac{1}{2}] \\ (g * h)(2s - 1) & s \in [\frac{1}{2}, 1] \end{cases}$$

$$= \begin{cases} f(2s) & s \in [0, \frac{1}{2}] \\ g(2(2s - 1)) & s \in [\frac{1}{2}, \frac{3}{4}] \\ h(2(2s - 1) - 1) & s \in [\frac{3}{4}, 1] \end{cases}$$

Similarly, $(f * g) * h$ is defined by,

$$(f * g)(s) = \begin{cases} f(2s) & s \in [0, \frac{1}{2}] \\ g(2s - 1) & s \in [\frac{1}{2}, 1] \end{cases}$$

$$((f * g) * h)(s) = \begin{cases} (f * g)(2s) & s \in [0, \frac{1}{2}] \\ h(2s - 1) & s \in [\frac{1}{2}, 1] \end{cases}$$

$$= \begin{cases} f(2(2s)) & s \in [0, \frac{1}{4}] \\ g(2(2s - 1)) & s \in [\frac{1}{4}, \frac{1}{2}] \\ h(2s - 1) & s \in [\frac{1}{2}, \frac{3}{4}] \end{cases}$$

Clearly, $f * (g * h)$ and $(f * g) * h$ are distinct path with common endpoints. ie, $(f * (g * h))(0) = f(0) = ((f * g) * h)(0)$. And $(f * (g * h))(1) = h(1) = ((f * g) * h)(1)$.

We need to define a path homotopy G between $f * (g * h)$ and $(f * g) * h$. Let $[a, b], [c, d] \subset I$. Consider the path $p : I \rightarrow I$ defined by the following

three unique⁽³⁵⁾ positive linear maps $p_1 : [0, a] \rightarrow [0, c]$, $p_2 : [a, b] \rightarrow [c, d]$ and $p_3 : [b, 1] \rightarrow [d, 1]$.

$$p(t) = \begin{cases} p_1(t) & t \in [0, a] \\ p_2(t) & t \in [a, b] \\ p_3(t) & t \in [b, 1] \end{cases}$$

We can easily construct, a path homotopy P between identity map $i : I \rightarrow I$, $i(s) = s$ and p as follows :

$$P(s, t) = \begin{cases} t + (p_1(t) - t) \frac{s}{a} & s \in [0, a] \\ t + (p_2(t) - t) \frac{(s-a)}{(b-a)} & s \in [a, b] \\ t + (p_3(t) - t) \frac{(s-b)}{(1-b)} & s \in [b, 1] \end{cases}$$

Therefore, we have $f * (g * h) \simeq_p i$ since there exists a path homotopy P corresponding to $[a, b] = [\frac{1}{2}, \frac{3}{4}]$ and $[c, d] = [x_1, x_2]$. Similarly $(f * g) * h \simeq_p i$ since there exists a path homotopy P where $[a, b] = [\frac{1}{4}, \frac{1}{2}]$ and $[c, d] = [x_1, x_2]$. ie, $[f * (g * h)] \simeq_p [(f * g) * h]$. \square

Theorem 7.6.4. *Let f be a path in X , and a_0, a_1, \dots, a_n be numbers such that $0 = a_0 < a_1 < \dots < a_n = 1$. Let $f_i : I \rightarrow X$ be the path that equals the positive linear map of I onto $[a_{i-1}, a_i]$ followed by f . Then $[f] = [f_1] * [f_2] * \dots * [f_n]$. In other words, every path is path-homotopic to a piecewise-linear path.*

Proof. Let f be a piece-wise positive linear map such that

$$f(t) = \begin{cases} f_1(t) & t \in [0 = a_0, a_1] \\ f_2(t) & t \in [a_1, a_2] \\ \vdots & \vdots \\ f_n(t) & t \in [a_{n-1}, a_n] \end{cases}$$

where $f_i : I \rightarrow [a_{i-1}, a_i]$ such that $f_i(t)$ are a positive linear maps.

Consider the path $p : I \rightarrow I$ defined by the unique positive linear maps on the subintervals of any partition $\{0 = x_0, x_1, \dots, x_n\}$ of I . ie, $0 = x_0 < x_1 < \dots < x_n = 1$

$$\begin{aligned} p_1 : [x_0, x_1] &\rightarrow [a_0, a_1] \\ p_2 : [x_1, x_2] &\rightarrow [a_1, a_2] \\ &\vdots \\ p_n : [x_{n-1}, x_n] &\rightarrow [a_{n-1}, a_n] \end{aligned}$$

$$\text{Define, } p(t) = \begin{cases} p_1(t) & t \in [x_0, x_1] \\ p_2(t) & t \in [x_1, x_2] \\ \vdots & \vdots \\ p_n(t) & t \in [x_{n-1}, x_n] \end{cases}$$

³⁵ $p_1(t) = \frac{ct}{a}$, $p_2(t) = \frac{(d-c)t}{b-a} + \frac{bc-ad}{b-a}$, $p_3(t) = \frac{(1-d)t}{1-b} + \frac{d-b}{1-b}$

Then there exists a path homotopy P between identity map $i : I \rightarrow I$, $i(t) = t$ and p given by

$$P(s, t) = \begin{cases} t + (p_1(t) - t) \frac{a_1}{x_1} & s \in [0, x_1] \\ t + (p_2(t) - t) \frac{s - x_1}{x_2 - x_1} & s \in [x_1, x_2] \\ \vdots & \\ t + (p_n(t) - t) \frac{s - x_{n-1}}{x_n - x_{n-1}} & s \in [x_{n-1}, x_n] \end{cases}$$

Since any product of f_1, f_2, \dots, f_n is a path p for some partition decided by the order of associativity. This partition can be constructed as follows : Let the last product operation (by associativity) corresponds to $\frac{1}{2}$. The expression on its left corresponds to $[0, \frac{1}{2}]$ and expression on the right corresponds to $[\frac{1}{2}, 1]$. If there are any operations on any of these parts, the last operation (by associativity) in them corresponds to the midpoint the respective subinterval and so on.

For examples : Consider, $(f_1 * (f_2 * f_3)) * (f_4 * f_5)$. Suppose we number the operations, $(f_1 *_1 (f_2 *_2 f_3)) *_3 (f_4 *_4 f_5)$. Then we have, $*_3 \rightarrow \frac{1}{2} \implies *_1 \rightarrow \frac{1}{4} \implies *_2 \rightarrow \frac{3}{8}$. Again $*_3 \rightarrow \frac{1}{2} \implies *_4 \rightarrow \frac{3}{4}$. Thus, we have $\{0, \frac{1}{4}, \frac{3}{8}, \frac{1}{2}, \frac{3}{4}, 1\}$.

Thus given two paths f and f' with distinct order of associativity of n paths : f_1, f_2, \dots, f_n . We have path homotopy P, P' given by the $P(s, t)$ for the respective partition constructed according to the order of associativity. Then, we have $f \simeq_p i$ and $f' \simeq_p i$. Thus, irrespective of the order of associativity all these paths are path homotopic. ie, $[f] = [f_1] * [f_2] * \dots * [f_n]$ \square

Subject 8

ME010203 Numerical Analysis with Python

8.1 Interpolation and Curve Fitting

Definitions 8.1.1. Given $(n + 1)$ data points (x_k, y_k) , $k = 0, 1, \dots, n$, the problem of estimating $y(x)$ using a function $y : \mathbb{R} \rightarrow \mathbb{R}$ that satisfy the data points is the interpolation problem. ie, $y(x_k) = y_k$, $k = 0, 1, \dots, n$.

Definitions 8.1.2. Given $(n + 1)$ data points (x_k, y_k) , $k = 0, 1, \dots, n$, the problem of estimating $y(x)$ using a function $y : \mathbb{R} \rightarrow \mathbb{R}$ that is sufficiently close to the data points is the curve-fitting problem.
ie, Given $\epsilon > 0$, $|y(x_k) - y_k| < \epsilon$, $k = 0, 1, \dots, n$.

Remark. The data could be from scientific experiments or computations on mathematical models. The interpolation problem assumes that the data is accurate. But, curve-fitting problem assumes that there are some errors involved which are sufficiently small.

Definitions 8.1.3. Given $(n + 1)$ data points (x_k, y_k) , $k = 0, 1, \dots, n$, the problem of estimating $y(x)$ using a polynomial function of degree n that satisfy the data points is the polynomial interpolation problem.

Remark. Polynomial is the ‘simplest’ interpolant. [Kiusalaas, 2013, 3.2]

8.2 Polynomial Interpolation

There exists a unique polynomial of degree n that satisfy $(n + 1)$ distinct data points. There are a few methods to find this polynomial : 1. Lagrange’s method
2. Newton’s method.

8.2.1 Lagrange's Method

Interpolation polynomial ¹ is given by,

$$P(x) = \sum_{i=0}^n y_i l_i(x), \text{ where } l_i(x) = \prod_{j=0, j \neq i}^n \frac{x - x_j}{x_i - x_j} \quad (8.1)$$

Remark. Lagrange's cardinal functions l_i , are polynomials of degree n and

$$l_i(x_j) = \delta_{ij} = \begin{cases} 0, & i \neq j \\ 1, & i = j \end{cases}$$

Proposition 8.2.1. Error in polynomial interpolation is given by

$$f(x) - P(x) = \frac{(x - x_0)(x - x_1) \cdots (x - x_n)}{(n + 1)!} f^{(n+1)}(\xi) \quad (8.2)$$

where $\xi \in (x_0, x_n)$

Remark. The error increases as x moves away from the unknown value ξ .

8.2.2 Newton's Method

The interpolation polynomial is given by,

$$P(x) = a_0 + a_1(x - x_0) + \cdots + a_n(x - x_0)(x - x_1) \cdots (x - x_{n-1}) \quad (8.3)$$

where $a_i = \nabla^i y_i$, $i = 0, 1, \dots, n$.

Remark. For Newton's Method, usually it is assumed that $x_0 < x_1 < \cdots < x_n$.

Remark. Lagrange's method is conceptually simple. But, Newton's method is computationally more efficient than Lagrange's method.

Computing coefficients a_i of the polynomial

The coefficients are given by,

$$a_0 = y_0, \quad a_1 = \nabla y_1, \quad a_2 = \nabla^2 y_2, \quad a_3 = \nabla^3 y_3, \dots, a_n = \nabla^n y_n \quad (8.4)$$

Remark. The divided difference $\nabla^i y_i$ are computed as follows:

$$\begin{aligned} \nabla y_1 &= \frac{y_1 - y_0}{x_1 - x_0} \\ \nabla y_2 &= \frac{y_2 - y_1}{x_2 - x_1} & \nabla^2 y_2 &= \frac{\nabla y_2 - \nabla y_1}{x_2 - x_1} \\ \nabla y_3 &= \frac{y_3 - y_2}{x_3 - x_2} & \nabla^2 y_3 &= \frac{\nabla y_3 - \nabla y_2}{x_3 - x_2} & \nabla^3 y_3 &= \frac{\nabla^2 y_3 - \nabla^2 y_2}{x_3 - x_2} \end{aligned}$$

Remark. Practise Problems

Find interpolation polynomial for the following data points :

¹Using P_n to represent some polynomial of degree n . It is quite a confusing a notation when it comes to Newton's method as author construct a psuedo-recursive definition.

1. $\{(0, 7), (2, 11), (3, 28)\}$
 Ans : $5x^2 - 8x + 7$
 [Kiusalaas, 2013, Example 3.1]
2. $\{(-2, -1), (1, 2), (4, 59), (-1, 4), (3, 24), (-4, -53)\}$
 Ans : $x^3 - 2x + 3$
 [Kiusalaas, 2013, Example 3.2]
3. $\{(-1.2, -5.76), (0.3, -5.61), (1.1, -3.69)\}$
 Ans : $x^2 + x - 6$
 [Kiusalaas, 2013, Problem Set 3.1.1]
4. $\{(-3, 0), (2, 5), (-1, -4), (3, 12), (1, 0)\}$
 Ans : $x^2 + 2x - 3$
 [Kiusalaas, 2013, Problem Set 3.1.7]
5. $\{(0, 1.225), (3, 0.905), (6, 0.652)\}$
 Ans : $0.0037x^2 - 0.1178x + 1.225$
 [Kiusalaas, 2013, Problem Set 3.1.9]

Remark. In Lagrange's Method, we can interpolate at the given point even without computing the polynomial. In Newton's method, we have to compute polynomial and then interpolate for the given point.

That is, evaluate the value of cardinal polynomials at the point and substitute in Equation 8.1 as shown in Section 3.2. [Kiusalaas, 2013, Example 3.1]

8.2.3 Implementation of Newton's Method

Program 8.2.1. Computing Coefficients

```
def coefficients(xData, yData):
    m = len(xData)
    a = yData.copy()
    for k in range(1, m):
        a[k:m] = (a[k:m] - a[k-1]) / (xData[k:m] - xData[k-1])
    return a
```

Line 1 `def coefficients(xData, yData):`

Defines a function which takes two arguments/parameters, named *xData* and *yData*. In [Kiusalaas, 2013, 3.2], you will find coeffs which I have changed to coefficients. *xData, yData* are numpy array objects. *xData*

x_0	y_0				
x_1	y_1	∇y_1			
x_2	y_2	∇y_2	$\nabla^2 y_2$		
\dots	\dots	\dots	\dots	\ddots	
x_n	y_n	∇y_n	$\nabla^2 y_n$	\dots	$\nabla^n y_n$

Table 8.1: The $\nabla^i y_i$ Computation Table

is a array with values x_0, x_1, \dots, x_n . And $yData$ is array with values y_0, y_1, \dots, y_n . For example, the value of x_3 can be accessed as $xData[3]$.

Line 2 `m = len(xData)`

The function `len()` is extended by numpy to give the length of array objects. In this context, `len(xData)` will return the value $n+1$, since there are $n+1$ values in $xData$ array.

Line 3 `a = yData.copy()`

We need a copy of $yData$ to work with. Unlike other programming languages like java, in python `a = yData` will assign a new label a to the same memory location and manipulating a will corrupt the original data in $yData$ as well. In order to avoid this, we are **making a copy of the array object using the array method provided by the numpy library.**

Line 4 `for k in range(1,m):`

This is a python loop statement. This ask python interpreter to repeat the following sub-block $m - 1$ times.² In this context, Line 5 will be executed n times, since the `range(1,m)` object is a list-type object with values $1, 2, \dots, m - 1$. And interpreter executes Line 5 for each values in the `range()` object, ie, $k = 1, 2, \dots, m - 1$ before interpreting Line 6.

Line 5 `a[k:m] = (a[k:m]-a[k-1])/(xData[k:m]-xData[k-1])`

This is very nice feature available in python. **This statement, evaluates $m - k$ values in a single step. ie, $a[k], a[k + 1], \dots, a[m]$. This calculation corresponds to subsequent columns of the divided difference table, that we are familiar with.** For example, executing Line 5 with $k = 3$ is same as evaluating the $\nabla^3 y_j$ column. Note that the value $a[0]$ is never updated and similarly $a[2]$ changes when Line 5 is executed with $k = 1, 2$. From column 3 onward, $a[2]$ is not updated. Therefore, **after completing n th executing of the Line 5, we have** $a[0] = y_0$, $a[1] = \nabla y_1$, $a[2] = \nabla^2 y_2, \dots$, $a[n] = \nabla^n y_n$.

Line 6 `return a`

This returns the array a which is the array of coefficients.

The logic of this program is in Line 4 and Line 5. So they need more explanation/understanding than anything else.

Program 8.2.2. Interpolating using Newton's Method

```
def interpolate(a, xData, x):
    n = len(xData)-1
    p = a[n]
    for k in range(1, n+1):
        p = a[n-k]*(x-xData[n-k])*p
    return p
```

The logic this program is in Line 3, Line 4 and Line 5.

²Python block is a group of statement with same level of indentation. A sub-block is a block with an additional indentation.

Line 3 : We initialize the polynomial with the coefficient $a[n] = \nabla^n y_n = a_n$.

Line 4 : We are going to define the polynomial recursively. This takes exactly n steps further. So we use a loop which repeats n times.

Line 5 : The value of p and k changes each time Line 5 is executed. Let P_j be the value in p after executing Line 5 with $k = j$. Then,

$$P_0 = p = a[n]$$

$$P_1 = a[n-1] + (x - x_{n-1})P_0$$

$$P_2 = a[n-2] + (x - x_{n-2})P_1$$

\vdots

$P_n = a[0] + (x - x_0)P_{n-1}$. Clearly, P_n is the unique n degree polynomial given by the Newton's method.

Program 8.2.3. How to interpolate ?

```
from numpy import array
xDData = array([-2,1,4,-1,3,-4])
yData = array([-1,2,59,4,24,-53])
a = coefficients(xDData,yData)
print(interpolate(a,xDData,2))
```

You will have to define both the functions (coefficients, interpolate) before doing this.

Line 1 `from numpy import array`

For defining array objects, we need to import them from numpy library.

Line 2 `xDData = array([-2,1,4,-1,3,-4])`

You can change this line according to the first component of the given data points.

Line 3 `yData = array([-1,2,59,4,24,-53])`

You can change this line according to the second component of the given data points.

Line 4 `a = coefficients(xDData,yData)`

Call function `coefficients` and store the array returned into `a`

Line 5 `print(interpolate(a,xDData,2))`

Call function `interpolate` to interpolate at $x = 2$ and print the value $P(2)$

Program 8.2.4 (Just for Fun). We can do more using sympy !

```
from numpy import array
from sympy import Symbol
xDData = array([-2,1,4,-1,3,-4])
yData = array([-1,2,59,4,24,-53])
a = coefficients(xDData,yData)
x = Symbol('x')
p = interpolate(a,xDData,x)
p.subs({x:2})
```

Remark. Programming Problems

1. $\{(0.15, 4.79867), (2.30, 4.49013), (3.15, 4.2243), (4, 85, 3.47313), (6.25, 2.66674), (7.95, 1.51909)\}$ [Kiusalaas, 2013, Example 3.4]
2. $\{(0, -0.7854), (0.5, 0.6529), (1, 1.7390), (1.5, 2.2071), (2, 1.9425)\}$ [Kiusalaas, 2013, Problem Set 3.1.5]

8.2.4 Limitations of Polynomial Interpolation

1. Inaccuracy - The error in interpolation increases as the point moves away from most of the data points.
2. Oscillation - As the number of data points considered for polynomial interpolation increases, the degree of the polynomial increases. And the graph of the interpolant tend to oscillate excessively. In such cases, the error in interpolation is quite high.
3. The best practice is to consider four to six data points nearest to the point of interest and ignore the rest of them.

Remark. The interpolant obtained by joining cubic polynomials corresponding to four nearest data points each, is a cubic spline³.

8.3 Roots of a Function

Definitions 8.3.1. Let $f : \mathbb{R} \rightarrow \mathbb{R}$, then $x \in \mathbb{R}$ is a root of f if $f(x) = 0$.

Remark. Suppose $a < b$ and $f(a), f(b)$ are nonzero and are of different signs. If f is continuous in $[a, b]$, then there is a point $c \in [a, b]$ such that $f(c) = 0$.

Thus given $a < b$ and $f(a), f(b)$ are nonzero values of different sign, then there may be a bracketed root in $[a, b]$.

Note : There is no guarantee that there exists a root in $[a, b]$ as we are not sure about the continuity of f .

Remark. Given a bracketed root, we can find it using

1. Bisection Method or
2. Newton-Raphson Method

8.3.1 Bisection Method

Suppose $a < b$ and $f(a), f(b)$ are nonzero values of different signs. We evaluate $f(c)$ where $c = \frac{a+b}{2}$. If $f(c)$ is a nonzero value, then at least one of the pairs $f(a), f(c)$ or $f(c), f(b)$ are of different signs. WLOG suppose that $f(a), f(c)$ are of different signs, then set $b = c$ and $c = \frac{a+b}{2}$. And continue this process until we get sufficiently accurate value of a root.

³Cubic spline is a function, the graph of which is piece-wise cubic

Remark. Suppose $f(x) = x^5 - 2$. Then $f(0) = -2$, $f(1) = -1$, $f(2) = 30$. Since f is known to be continuous, there is a bracketed root in $[1, 2]$. Now

$$\begin{aligned} f(1.5) &> 0 \implies [1, 1.5] \\ f(1.25) &> 0 \implies [1, 1.25] \\ f(1.125) &< 0 \implies [1.125, 1.25] \\ f(1.1875) &> 0 \implies [1.125, 1.1875] \\ f(1.15375) &> 0 \implies [1.125, 1.15375] \\ f(1.139375) &< 0 \implies [1.139375, 1.15375] \\ f(1.1465625) &< 0 \implies [1.1465625, 1.15375] \\ f(1.150156250) &> 0 \implies [1.1465625, 1.15015625] \\ f(1.148359375) &< 0 \implies [1.1483594, 1.15015625] \\ f(1.149257825) &> 0 \implies [1.1483594, 1.14925783] \end{aligned}$$

Thus, we have 1.14 is a root of f with accuracy upto two decimal points.

8.3.2 Newton-Raphson Method

Suppose f is differentiable at $x \in \mathbb{R}$ and $f(x) \neq 0$. Then compute $x = x - \frac{f(x)}{df(x)}$ and evaluate $f(x)$. Repeat this process to get more accurate value of a root near x .

Remark. Suppose $f(x) = x^5 - 2$. Then $df(x) = 5x^4$. Let $x = 2$. Then

$$\begin{aligned} x &= 2 - \frac{30}{80} \implies f(1.625) = 9.330 \\ x &= 1.625 - \frac{9.330}{34.86} \implies f(1.35735) = 2.6074 \\ x &= 1.35735 - \frac{2.6074}{16.9721} \implies f(1.20373) = 0.52733 \\ x &= 1.20373 - \frac{0.52733}{10.4975} \implies f(1.15351) = 0.04224 \\ x &= 1.15351 - \frac{0.042245}{8.85225} \implies f(1.148738) = 0.00034312 \end{aligned}$$

Thus we have 1.1487 is quite close to a root of f .

8.4 Matrix Operations

Consider a system of n linear, simultaneous equations in n unknowns,

$$\begin{aligned} A_{11}x_1 + A_{12}x_2 + \cdots + A_{1n}x_n &= b_1 \\ A_{21}x_1 + A_{22}x_2 + \cdots + A_{2n}x_n &= b_2 \\ &\vdots \\ A_{n1}x_1 + A_{n2}x_2 + \cdots + A_{nn}x_n &= b_n \end{aligned}$$

We may represent them using matrices as $Ax = b$. That is,

$$\begin{bmatrix} A_{11} & A_{12} & \dots & A_{1n} \\ A_{21} & A_{22} & \dots & A_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ A_{n1} & A_{n2} & \dots & A_{nn} \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{bmatrix} = \begin{bmatrix} b_1 \\ b_2 \\ \vdots \\ b_n \end{bmatrix}$$

Gauss Elimination and Doolittle Decomposition are two methods to solve system of equations, $Ax = b$.

8.5 Gauss elimination method

Gauss elimination method has of two phases 1. elimination and 2. back substitution. In elimination phase, system $Ax = b$ is transformed into an equivalent system $Ux = c$ where U is an upper-triangular⁴ matrix. And in back substitution phase, $Ux = c$ is solved. Since $Ax = b$ and $Ux = c$ are equivalent, they have the same solution x .

8.5.1 Elimination Phase

We can eliminate unknowns from an equation by adding a scalar multiple of an equation to another equation of the system. In matrices, this is equivalent to adding a scalar multiple of one row to another row, say $R_i \leftarrow R_i + \lambda R_k$.

$$\begin{array}{r} A_{k1}x_1 + A_{k2}x_2 + \dots + A_{kn}x_n = b_k + \\ \lambda(A_{i1}x_1 + A_{i2}x_2 + \dots + A_{in}x_n = b_i) \\ \hline (A_{k1} + \lambda A_{i1})x_1 + (A_{k2} + \lambda A_{i2})x_2 + \dots + (A_{kn} + \lambda A_{in})x_n = b_k + \lambda b_i \end{array}$$

$$\begin{bmatrix} \vdots & \vdots & \vdots & \vdots \\ A_{k1} & A_{k2} & \dots & A_{kn} \\ \vdots & \vdots & \ddots & \vdots \\ A_{i1} & A_{i2} & \dots & A_{in} \\ \vdots & \vdots & \vdots & \vdots \end{bmatrix} \xrightarrow{R_i \leftarrow R_i + \lambda R_k} \begin{bmatrix} \vdots & \vdots & \vdots & \vdots \\ A_{k1} & A_{k2} & \dots & A_{kn} \\ \vdots & \vdots & \ddots & \vdots \\ A_{i1} + \lambda A_{k1} & A_{i2} + \lambda A_{k2} & \dots & A_{in} + \lambda A_{kn} \\ \vdots & \vdots & \vdots & \vdots \end{bmatrix}$$

$$\begin{bmatrix} \vdots \\ b_k \\ \vdots \\ b_i \\ \vdots \end{bmatrix} \xrightarrow{R_i \leftarrow R_i + \lambda R_k} \begin{bmatrix} \vdots \\ b_k \\ \vdots \\ b_i + \lambda b_k \\ \vdots \end{bmatrix}$$

⁴upper triangular - all the entries below the main diagonal are zero. ie $U_{ij} = 0$, if $i < j$

8.5.2 Back substitution

Let $Ux = c$ be a system of n linear equations in n unknowns and U is an upper triangular matrix. Then we can solve the system of equations from the back.

$$\begin{bmatrix} U_{1,1} & U_{1,2} & \cdots & U_{1,n-1} & U_{1,n} \\ 0 & U_{2,2} & \cdots & U_{2,n-1} & U_{2,n} \\ \vdots & \vdots & \ddots & \vdots & \vdots \\ 0 & 0 & \cdots & U_{n-1,n-1} & U_{n-1,n} \\ 0 & 0 & \cdots & 0 & U_{n,n} \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ \vdots \\ x_{n-1} \\ x_n \end{bmatrix} = \begin{bmatrix} c_1 \\ c_2 \\ \vdots \\ c_{n-1} \\ c_n \end{bmatrix}$$

$$U_{n,n} x_n = c_n \implies x_n = \frac{c_n}{U_{n,n}}$$

$$\sum_{i=n-1}^n U_{n-1,i} x_i = c_{n-1} \implies x_{n-1} = \frac{c_{n-1} - U_{n-1,n} x_n}{U_{n-1,n-1}}$$

...

$$\sum_{i=1}^n U_{1,i} x_i = c_1 \implies x_1 = \frac{c_1 - \sum_{i=2}^n U_{1,i} x_i}{U_{1,1}}$$

8.5.3 Illustrative example

Consider the following system of linear equations,

$$\begin{aligned} 4x_1 - 2x_2 + x_3 &= 11 \\ -2x_1 + 4x_2 - 2x_3 &= -16 \\ x_1 - 2x_2 + 4x_3 &= 17 \end{aligned}$$

We may represent the above system of linear equations using matrices,

$$\begin{bmatrix} 4 & -2 & 1 \\ -2 & 4 & -2 \\ 1 & -2 & 4 \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix} = \begin{bmatrix} 11 \\ -16 \\ 17 \end{bmatrix}$$

Phase 1 : Elimination Process

Using eq.1, the unknown x_1 is eliminated from all subsequent equations. An equivalent operation can be performed on both the matrices A and b by adding a suitable scalar multiples of row R_1 to row R_2 and R_3 .

$$\begin{bmatrix} 4 & -2 & 1 \\ -2 & 4 & -2 \\ 1 & -2 & 4 \end{bmatrix} \xrightarrow{\substack{R_2 \leftarrow R_2 + 0.5R_1 \\ R_3 \leftarrow R_3 - 0.25R_1}} \begin{bmatrix} 4 & -2 & 1 \\ 0 & 3 & -1.5 \\ 0 & -1.5 & 3.75 \end{bmatrix}$$

$$\begin{bmatrix} 11 \\ -16 \\ 17 \end{bmatrix} \xrightarrow{\substack{R_2 \leftarrow R_2 + 0.5R_1 \\ R_3 \leftarrow R_3 - 0.25R_1}} \begin{bmatrix} 11 \\ -10.5 \\ 14.25 \end{bmatrix}$$

And using eq.2, x_2 is eliminated from all subsequent equations(only those rows below it). Again, we perform this by adding suitable scalar multiples of row 2 to row R_3 .

$$\begin{bmatrix} 4 & -2 & 1 \\ 0 & 3 & -1.5 \\ 0 & -1.5 & 3.75 \end{bmatrix} \xrightarrow{R_3 \leftarrow R_3 + 0.5R_2} \begin{bmatrix} 4 & -2 & 1 \\ 0 & 3 & -1.5 \\ 0 & 0 & 3 \end{bmatrix}$$

$$\begin{bmatrix} 11 \\ -16 \\ 17 \end{bmatrix} \xrightarrow{R_3 \leftarrow R_3 + 0.5R_2} \begin{bmatrix} 11 \\ -10.5 \\ 9 \end{bmatrix}$$

The elimination process is complete when all entries below the diagonal elements are reduced to zero. ie, upper triangular matrix.

Phase 2 : Back substitution Process

The unknowns are easily found from the equations by solving them in the reverse order. The unknowns are solved from the bottom and solved variables are used to solve the remain unknowns.

$$\begin{bmatrix} 4 & -2 & 1 & 11 \\ 0 & 3 & -1.5 & -10.5 \\ 0 & 0 & 3 & 9 \end{bmatrix} \rightarrow \begin{cases} 4x_1 - 2x_2 + x_3 = 11 \\ 3x_2 - 1.5x_3 = -10.5 \\ 3x_3 = 9 \end{cases}$$

$$x_3 = \frac{9}{3} = 3$$

$$x_2 = \frac{-10.5 + 1.5x_3}{3} = -2$$

$$x_1 = \frac{11 - x_3 + 2x_2}{4} = 1$$

Remark. Why don't they use row-reduced echelon matrix of A to simplify the back substitution phase ?

This doesn't have much advantage from algorithmic point of view. That is, the time complexity (number of steps for computation) is unaffected. And algorithms always prefer methods even with slight advantage in time or memory. And they won't consider complications in the manual execution of the method. Therefore, programmers won't consider alternate algorithm for the sake of computational simplicity.

8.5.4 Python : Gauss elimination method

Program 8.5.1 (Gauss elimination).

```
from numpy import dot
def gaussElimination(a, b):
    n = len(b)
    for k in range(0, n-1):
        for i in range(k+1, n):
            if a[i, k] != 0.0:
                lam = a[i, k]/a[k, k]
                a[i, k+1:n] = a[i, k+1:n] - lam*a[k, k+1:n]
                b[i] = b[i] - lam*b[k]
        for k in range(n-1, -1, -1):
            x[k] = (b[k] - dot(a[k, k+1:n], x[k+1:n]))/a[k, k]
```

```
return b
```

- Line 1 `from numpy import dot`
Imports the “dot()” function for numpy arrays which takes two ‘numpy arrays’ as input arguments and returns the dot product of them.
- Line 2 `def gaussElimination(a,b):`
it defines “gaussElimination()” as a function which takes two arguments (inputs). First argument is the coefficient matrix A and second argument is the constant matrix b of the linear system of the form $Ax = b$.
- Line 3 `n = len(b)`
it assigns the length of the list b into variable n which is obviously the number of equations.
- Line 4 `for k in range(0,n-1):`
it is a loop construct. Five instructions following it are part of this loop body, which are executed for each values of k ie, $k = 0, 1, \dots, n-1$. For each value of k , the unknown x_{k+1} is selected for elimination process.
- Line 5 `for i in range(k+1,n):`
it is a loop inside another loop. Four instructions following it are part of this loop body, which are executed for each values of i , ie, $i = k+1, k+2, \dots, n$. This eliminates x_{k+1} from all the equations after the $k+1$ th equation of the system. Value of $i+1$ is the equation⁵ from which x_{k+1} is eliminated.
- Line 6 `if a[i,k] != 0.0:`
If $a[i,k] = A_{i+1,k+1} \neq 0$, then those three instruction following it are executed. Otherwise, it skips the execution of those three statements. If $x_{k+1} = x[k]$ is not there in the i th equation, it doesn’t need to be eliminated.
- Line 7 `lam = a[i,k]/a[k,k]`
In this step, λ is computed so that $\text{equ.}(i+1) - \lambda \text{ equ.}(k+1)$ doesn’t have x_{k+1} term in it.
- Line 8 `a[i,k+1:n] = a[i,k+1:n] - lam * a[k,k+1:n]`
Coefficients of $(i+1)$ th equation are updated.
Equivalent to $a[i,0:n] = a[i,0:n] - lam * a[k,0:n]$, since zeroes need not be subtracted. This is same as $\text{equ.}(i+1) \leftarrow \text{equ.}(i+1) - \lambda \text{ equ.}(k+1)$
- Line 9 `b[i] = b[i] - lam * b[k]`
The same row operations are performed on the matrix b instead of using an augmented matrix.
- Line 10 `for k in range(n-1,-1,-1):`
This is another loop construct. The following statement is executed n times for values of $k = n-1, n-2, \dots, 0$. Value of $k+1$ gives the unknown x_{k+1} which is solved by the back substitution process.

⁵Python starts counting from zero. For example : $A_{11} = a[0,0]$, $x_1 = x[0]$ and $b_1 = b[0]$

Line 11 $x[k] = (b[k] - \text{dot}(a[k, k+1 : n], x[k+1 : n]))/a[k, k]$

This is the back substitution process. After elimination phase we have k equation in the form $A_{k,k}x_k + A_{k,k+1}x_{k+1} + \dots + A_{k,n}x_n = b_k$. And we already have values of $x_{k+1}, x_{k+2}, \dots, x_n$. Then

$$x_k = \frac{b_k - (A_{k,k+1}x_{k+1} + A_{k,k+2}x_{k+2} + \dots)}{A_{k,k}}$$

This is equivalent to

$$b_k \leftarrow \frac{b_k - [A_{k,k+1} \quad A_{k,k+2} \quad \dots \quad A_{k,n}] \begin{bmatrix} x_{k+1} \\ x_{k+2} \\ \vdots \\ x_n \end{bmatrix}}{A_{k,k}}$$

Remember : The values of x_k are updated into b_k as they are computed. Thus x_k, x_{k+1}, \dots, x_n are stored in b for next back substitution ie, for evaluating x_{k-1} . We start with x_{n-1} , as $x_n = b_n$ is already solved.

Line 12 **return b**

It returns the new b matrix as output of the “gaussElimination()” function where $x_k = b_k, \forall k$.

8.6 LU Decomposition Method : Doolittle

Let $Ax = b$ be a linear system of n equations in n unknowns and let $A = LU$ for some lower triangular matrix L and upper triangular matrix U . Then we have $LUx = Ly = b$ where $y = Ux$. There are two phases for this method : 1. LU decomposition and 2. substitution.

First, we compute L and U such that $A = LU$ using Gauss elimination. Then We can solve $Ly = b$ using forward substitution process and then solve $Ux = y$ using back substitution process.

For Doolittle decomposition, we prefer to write A as a product LU as shown below:

$$A = \begin{bmatrix} 1 & 0 & \dots & 0 \\ L_{2,1} & 1 & \dots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ L_{n,1} & L_{n,2} & \dots & 1 \end{bmatrix} \begin{bmatrix} U_{1,1} & U_{1,2} & \dots & U_{1,n} \\ 0 & U_{2,2} & \dots & U_{2,n} \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \dots & U_{n,n} \end{bmatrix}$$

$$A = \begin{bmatrix} U_{1,1} & U_{1,2} & \dots & U_{1,n} \\ L_{2,1}U_{1,1} & L_{2,1}U_{1,2} + U_{2,2} & \dots & L_{2,1}U_{1,n} + U_{2,n} \\ \vdots & \vdots & \ddots & \vdots \\ L_{n,1}U_{1,1} & L_{n,1}U_{1,2} + L_{n,2}U_{2,2} & \dots & \sum_{k=1}^{n-1} L_{n,k}U_{k,n} + U_{n,n} \end{bmatrix}$$

Note that in Doolittle's decomposition method, the diagonal entries of the lower triangular matrix L are all 1. ie, $L_{ii} = 1, \forall i$. Thus, we can use an $n \times n$ matrix to represent both L and U by overwriting trivial entries(zeroes and ones) of both the matrices. And this matrix is represented by $[L \setminus U]$.⁶

$$[L \setminus U] = \begin{bmatrix} U_{1,1} & U_{1,2} & \cdots & U_{1,n} \\ L_{2,1} & U_{2,2} & \cdots & U_{2,n} \\ \vdots & \vdots & \ddots & \vdots \\ L_{n,1} & L_{n,2} & \cdots & U_{n,n} \end{bmatrix}$$

is the combined matrix made from both the triangular matrices L and U .

The triangular matrices L and U such that $LU = A$ can be computed the variables in the Gauss elimination method.

$$A = \begin{bmatrix} 1 & 0 & \cdots & 0 \\ L_{2,1} & 1 & \cdots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ L_{n,1} & L_{n,2} & \cdots & 1 \end{bmatrix} \begin{bmatrix} U_{1,1} & U_{1,2} & \cdots & U_{1,n} \\ 0 & U_{2,2} & \cdots & U_{2,n} \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & U_{n,n} \end{bmatrix}$$

We can break down this matrix multiplication into the following row operations on the rows of the upper triangular matrix⁷

$$\begin{aligned} U_{R1} &\leftarrow U_{R1} \\ U_{R2} &\leftarrow L_{2,1} \cdot U_{R1} + U_{R2} \\ &\vdots \\ U_{Rn} &\leftarrow L_{n,1} \cdot U_{R1} + L_{n,2} \cdot U_{R2} + \cdots + L_{n,n-1} \cdot U_{R(n-1)} + U_{Rn} \end{aligned}$$

Clearly, λ we use to eliminate x_k from row i are $L_{i,k}$. And the matrix obtained after Gauss elimination is the upper triangular matrix U .

8.6.1 Illustrative example

$$\text{Solve } \begin{bmatrix} -3 & 6 & -4 \\ 9 & -8 & 24 \\ -12 & 24 & -26 \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix} = \begin{bmatrix} -3 \\ 65 \\ -42 \end{bmatrix}$$

Phase 1 : LU Decomposition

Suppose, we have a system of three linear equations, then

$$A = LU = \begin{bmatrix} 1 & 0 & 0 \\ L_{21} & 1 & 0 \\ L_{31} & L_{32} & 1 \end{bmatrix} \begin{bmatrix} U_{1,1} & U_{1,2} & U_{1,3} \\ 0 & U_{2,2} & U_{2,3} \\ 0 & 0 & U_{3,3} \end{bmatrix}$$

$$A = \begin{bmatrix} U_{1,1} & U_{1,2} & U_{1,3} \\ L_{2,1}U_{1,1} & L_{2,1}U_{1,2} + U_{2,2} & L_{2,1}U_{1,3} + U_{2,3} \\ L_{3,1}U_{1,1} & L_{3,1}U_{1,2} + L_{3,2}U_{2,2} & L_{3,1}U_{1,3} + L_{3,2}U_{2,3} + U_{3,3} \end{bmatrix}$$

⁶algorithmic implementation all decomposition algorithms prefer to use a combined matrix

⁷ U_{Rk} : k th row of the matrix U

We can compute L and U using the Gauss elimination process⁸ The matrix obtained after Gauss elimination on A is U and the values of the variable lam used in Gauss elimination are the entries in L . That is, in order to eliminate x_k from row i , we use $lam = L_{i,k}$.

$$\text{Given, } A = \begin{bmatrix} -3 & 6 & -4 \\ 9 & -8 & 24 \\ -12 & 24 & -26 \end{bmatrix}$$

$$\begin{bmatrix} -3 & 6 & -4 \\ 9 & -8 & 24 \\ -12 & 24 & -26 \end{bmatrix} \xrightarrow[R_3 \leftarrow R_3 - 4R_1]{R_2 \leftarrow R_2 + 3R_1} \begin{bmatrix} -3 & 6 & -4 \\ 0 & 10 & 12 \\ 0 & 0 & -10 \end{bmatrix} \implies L_{2,1} = 3, L_{3,1} = -4$$

We store these non-trivial entries of L into A itself.
That is, $A_{2,1} = L_{2,1}$, $A_{3,1} = L_{3,1}$.

“ In this case, $A_{3,2}$ became zero (this is not a trivial zero yet), and we won't eliminate x_2 from row 3 to save computation time. Thus, we are not computing $L_{3,2} = 0$ or storing it. However, the variable representing $L_{3,2}$ is $A_{3,2}$, which is already zero after Gauss elimination and we are quite happy with that. ”

Clearly, $L_{3,2} = 0$. Therefore, we have

$$U = \begin{bmatrix} -3 & 6 & -4 \\ 0 & 10 & 12 \\ 0 & 0 & -10 \end{bmatrix}, L = \begin{bmatrix} 1 & 0 & 0 \\ 3 & 1 & 0 \\ -4 & 0 & 1 \end{bmatrix}$$

Since we are already stored those two non-trivial entries of L into A . We get,

$$[L \setminus U] = \begin{bmatrix} -3 & 6 & -4 \\ 3 & 10 & 12 \\ -4 & 0 & -10 \end{bmatrix}$$

Phase 2 : Substitution

Suppose $Ly = b$,

$$\begin{bmatrix} 1 & 0 & 0 \\ L_{2,1} & 1 & 0 \\ L_{3,1} & L_{3,2} & 1 \end{bmatrix} \begin{bmatrix} y_1 \\ y_2 \\ y_3 \end{bmatrix} = \begin{bmatrix} b_1 \\ b_2 \\ b_3 \end{bmatrix} \rightarrow \begin{cases} y_1 & = b_1 \\ L_{2,1}y_1 + y_2 & = b_2 \\ L_{3,1}y_1 + L_{3,2}y_2 + y_3 & = b_3 \end{cases}$$

Now, we can find the values of y_k and store them into the matrix b itself.

$$b_1 \leftarrow y_1 = b_1$$

$$b_2 \leftarrow y_2 = b_2 - [L_{2,1}] [b_1], \text{ since } b_1 = y_1$$

$$b_3 \leftarrow y_3 = b_3 - [L_{3,1} \quad L_{3,2}] \begin{bmatrix} b_1 \\ b_2 \end{bmatrix}, \text{ since } b_1 = y_1, b_2 = y_2$$

⁸We usually need a proof for such a strong statement. In this paper, they are more focussed on the application side and therefore we will don't present any vigorous proof.

In general,

$$b_k \leftarrow y_k = b_k - [L_{k,1} \quad L_{k,2} \quad \dots \quad L_{k,k-1}] \begin{bmatrix} b_1 \\ b_2 \\ \vdots \\ b_{k-1} \end{bmatrix}, \text{ since } b_j = y_j, j = 1, 2, \dots, (k-1)$$

We have $A = LU \implies LUx = b$. Suppose $Ux = y$, then we get $Ly = b$. First of all, we will solve $Ly = b$ using forward substitution.

$$\begin{bmatrix} 1 & 0 & 0 \\ 3 & 1 & 0 \\ -4 & 0 & 1 \end{bmatrix} \begin{bmatrix} y_1 \\ y_2 \\ y_3 \end{bmatrix} = \begin{bmatrix} -3 \\ 65 \\ -42 \end{bmatrix} \rightarrow \begin{cases} y_1 & = -3 \\ 3y_1 + y_2 & = 65 \\ -4y_1 + y_3 & = -42 \end{cases}$$

$$\begin{aligned} y_1 &= -3 \\ y_2 &= 65 - 3y_1 = 74 \\ y_3 &= -42 + 4y_1 = 54 \end{aligned}$$

Suppose $Ux = y$,

$$\begin{bmatrix} U_{1,1} & U_{1,2} & U_{1,3} \\ 0 & U_{2,2} & U_{2,3} \\ 0 & 0 & U_{3,3} \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ x_3 \end{bmatrix} = \begin{bmatrix} y_1 \\ y_2 \\ y_3 \end{bmatrix}$$

Now, we can find the values of x_k and store them into the matrix y itself.

$$\begin{aligned} y_3 \leftarrow x_3 &= \frac{y_3}{U_{3,3}} \\ y_2 \leftarrow x_2 &= \frac{y_2 - [y_3] [U_{2,3}]}{U_{2,2}}, \text{ since } y_3 = x_3 \\ y_1 \leftarrow x_1 &= \frac{y_1 - \begin{bmatrix} y_2 \\ y_3 \end{bmatrix} [U_{1,2} \quad U_{1,3}]}{U_{1,1}}, \text{ since } y_2 = x_2, y_3 = x_3 \end{aligned}$$

In general,

$$x_k = \frac{y_k - \begin{bmatrix} y_{k+1} \\ y_{k+2} \\ \vdots \\ y_n \end{bmatrix} [U_{k,k+1} \quad U_{k,k+2} \quad \dots \quad U_{k,n}]}{U_{k,k}}, \text{ since } y_j = x_j, j = k+1, k+2, \dots, n$$

8.6.2 Python : Doolittle's LU Decomposition method

Program 8.6.1.

```
from numpy import dot
def LUdecomposition(a):
    n = len(a)
    for k in range(0, n-1):
        for i in range(k+1, n):
```

```

        if a[i, k] != 0.0:
            lam = a[i, k]/a[k, k]
            a[i, k+1:n] = a[i, k+1:n] - lam*a[k, k+1:n]
            a[i, k] = lam
    return a
def LUsolve(a, b):
    n = len(a)
    for k in range(1, n):
        b[k] = b[k] - dot(a[k, 0:k], b[0:k])
        b[n-1] = b[n-1]/a[n-1, n-1]
    for k in range(n-2, -1, -1):
        b[k] = (b[k] - dot(a[k, k+1:n], b[k+1:n]))/a[k, k]
    return b

```

This program mainly uses the Gauss elimination algorithm. Thus, the explanation for Lines 3-8 are not repeated here.

But remember the loop at Line 4 has inner loop at Line 5 and Line 7-9 are at same level of indentation which means they all are either executed or skipped depending on the truthness of the condition in Line 6. And Line 6-9 are executed for each instance of inner loop. Again, Line 5-9 are executed for each instance of the outer loop.

This time the gaussElimination() function which you have seen earlier is split into two functions 1. LUdecomposition() and 2. LUsolve(). And forward substitution is also added to LUsolve().

Line 2 **def LUdecomposition(a):**

LUdecomposition(A) computes L and U such that $A = LU$ and combine both triangular matrices into a single matrix $[L/U]$, by over-writing their trivial entries. And returns this combined matrix.

Line 9 $a[i, k] = lam$

Clearly, lam used for eliminating x_k from row i , $\lambda_{i,k} = a[i, k]/a[k, k] = L[i, k]$, $\forall k, \forall i$, ($i > k$). Also $a[i, k]$ which is reduced zero by Gauss elimination process is not used anymore⁹ in Gauss elimination process. Thus $L[i, k]$ can stored at $a[i, k]$ straight away. And $U[i, j]$, $j \leq i$ are already the entries of the matrix obtained from Gauss elimination. Thus for each iterations of k , the matrix a is updated ($k+1$ th row and $k+1$ th column) with respective entries of the combined matrix $[L/U]$.

Line 10 **return a**

Matrix a is already $[L/U]$, and thus LUdecomposition(A) returns $[L/U]$ such that $A = LU$.

Line 11 **def LUsolve(a, b):**

LUsolve() function does both forward substitution and back substitution. Suppose $Ax = b$ is the system to be solved. Then the inputs of LUsolve() are $a = [L/U]$ where $A = LU$.

⁹ $A[i, k]$ is not used after elimination of x_k from row i - It turns out that the trivial zeroes which are ignored on the row operations in Gauss elimination not only save time, but also provide a variable to store our intermediate result $L_{i,k}$ in Doolittle method.

Line 12 $n = \text{len}(a)$

We have to compute this again as this function starts fresh and thus value of the variable n from `LUdecomposition()` is lost.

Line 13 **for** k **in** `range(1,n):`

This is the loop for forward substitution.

Line 14 $b[k] = b[k] - \text{dot}(a[k, 0 : k], b[0 : k])$

updating b_{k^*} with y_{k^*} such that $Ly = b$ where $k^* = k - 1$.

$$b_k \leftarrow b_k - \begin{bmatrix} L_{k,1} & L_{k,2} & \dots & L_{k,k-1} \end{bmatrix} \begin{bmatrix} b_1 \\ b_2 \\ \vdots \\ b_{k-1} \end{bmatrix}$$

Line 15 $b[n-1] = b[n-1]/a[n-1, n-1]$

Computing ¹⁰ y_n and storing it into b_n .

$$b_n \leftarrow \frac{b_n}{U_{n,n}}$$

Line 16 **for** k **in** `range(n-2, -1, -1):`

This is the loop for back substitution.

Line 17 $b[k] = (b[k] - \text{dot}(a[k, k+1 : n], b[k+1 : n]))/a[k, k]$

updating b_{k^*} with x_{k^*} such that $Ux = y$ where $k^* = k - 1$.

$$b_k \leftarrow \frac{b_k - \begin{bmatrix} x_{k+1} \\ x_{k+2} \\ \vdots \\ x_n \end{bmatrix} \begin{bmatrix} U_{k,k+1} & U_{k,k+2} & \dots & U_{k,n} \end{bmatrix}}{U_{k,k}}$$

Line 18 **return** b

`LUsolve([L\U], b)` returns b where $b[i+1] = x_i$.

Programmer's Tip : There are few things to remember when splitting a function into two functions.

1. These functions are completely independent of one another.
2. Variables defined inside a function are not available outside.
3. The best way to give/take data to/from a function is through arguments/return-value

Beginner's Tip : In any programming language, we reuse variable. Thus, same variable may represent different values at different points of time. In Doolittle LU Decomposition, the variable 'a' initially represent matrix A , this variable is passed into `LUdecomposition()` function. In that function, A is changed to $[L\backslash U]$ in a step-by-step fashion. The value of 'a' is updated in step 7, 8 and 9. This is bit hard to imagine this transition of 'a' from A to $[L\backslash U]$ for a beginner at programming. Similarly, in `LUsolve()` function, the variable 'b' changes from matrix b to matrix y , and then to matrix x .

¹⁰Mathematically, you can define dot product of empty matrices as zero, but numpy dot function can't handle such a situation. Therefore, we have to do this step separately.

8.7 Numerical Integraion

Numerical integration/Quadrature is the numerical approximation of $\int_a^b f(x)dx$ by $\sum_{i=0}^n A_i f(x_i)$ where x_i are nodal abscissas, and A_i are weights. There are two methods to determine these nodal abscissas and suitable weights so that the sum is sufficiently accurate to the value of the integral.

1. Newton-Cotes forumulas
2. Gauss quadrature

Newton-Cotes formulas are useful when $f(x)$ can be evaluated without much computation. And using those values $f(x)$ can be interpolated to a piecewise-polynomial function. Then using equally spaces nodal abscissas and suitable weights $\int_a^b f(x)dx$ can be numerically approximated.

Gauss quadrature rules require lesser evaluations of f . And therefore are quite useful when evaluation of $f(x)$ has much computational complexity. Also, this method can manage integrable singularities where as Newton-Cote formulas can't numerically integrate function with singularities.

8.7.1 Newton-Cotes formulas

We divide the interval of integral (a, b) into n subintervals of equal length, ie, $h = (b - a)/n$. Let $x_0 = a, x_1, x_2, \dots, x_{n-1}, x_n = b$ be the end points of these subintervals. Then we can find an n degree polymonial interpolant satisfying f at those points, using Lagrange's method.

$$\text{Polynomial, } P(x) = \sum_{i=0}^n f(x_i) l_i(x) \text{ where } l_i(x) = \prod_{\substack{j=0 \\ j \neq i}}^n \frac{x - x_j}{x_i - x_j}$$

Thus the integral $I = \int_a^b f(x)dx$ can be numerically evaluated as follows:

$$\begin{aligned} I &= \int_a^b P_n(x)dx = \sum_{i=0}^n \left(f(x_i) \int_a^b l_i(x)dx \right) \\ &= \sum_{i=0}^n A_i f(x_i), \text{ where } A_i = \int_a^b l_i(x)dx \end{aligned}$$

The simplest cases of Newton-Cotes formulas are when $n = 1, 2, \text{ and } 3$

Trapezoidal rule $n = 1 \implies A_0 = \frac{h}{2}, A_1 = \frac{h}{2}$ and

$$\int_a^b f(x)dx = A_0 f(x_0) + A_1 f(x_1) = \frac{h}{2}(f(a) + f(b))$$

Simpson's 1/3 rule $n = 2 \implies A_0 = \frac{h}{3}, A_1 = \frac{4h}{3}, A_2 = \frac{h}{3}$ and

$$\int_a^b f(x)dx = \sum_{i=0}^2 A_i f(x_i) = \frac{h}{3} \left(f(a) + 4f\left(\frac{a+b}{2}\right) + f(b) \right)$$

Simpson's 3/8 rule $n = 3 \implies A_0 = \frac{3h}{8}, A_1 = \frac{9h}{8}, A_2 = \frac{9h}{8}, A_3 = \frac{3h}{8}$ and

$$\int_a^b f(x)dx = \sum_{i=0}^3 A_i f(x_i) = \frac{3h}{8}(f(a) + 3f(a+h) + 3f(a+2h) + f(b))$$

Trapezoidal Rule : $n = 1$

Consider interval (a, b) . Since $n = 1$, we have $x_0 = a$ and $x_1 = b$.

$$l_0(x) = \frac{x - x_1}{x_0 - x_1}$$

$$l_1(x) = \frac{x - x_0}{x_1 - x_0}$$

$$\begin{aligned} A_0 &= \int_a^b l_0(x)dx = \int_a^b \frac{x - x_1}{x_0 - x_1}dx = \frac{-1}{h} \int_a^b (x - b)dx \\ &= \frac{-1}{h} \left(\frac{(x - b)^2}{2} \right)_a^b = \frac{-1}{h} \left(\frac{0 - (a - b)^2}{2} \right) = \frac{h}{2} \\ A_1 &= \int_a^b l_1(x)dx = \int_a^b \frac{x - x_0}{x_1 - x_0}dx = \frac{1}{h} \int_a^b (x - a)dx \\ &= \frac{1}{h} \left(\frac{(x - a)^2}{2} \right)_a^b = \frac{1}{h} \left(\frac{(b - a)^2 - 0}{2} \right) = \frac{h}{2} \end{aligned}$$

Therefore,

$$\int_a^b f(x)dx = A_0 f(x_0) + A_1 f(x_1) = \frac{h}{2}(f(a) + f(b))$$

Simpon's 1/3 Rule : $n = 2$

Consider interval (a, b) divided into two subintervals of equal length $h = \frac{a+b}{2}$.

We have $x_0 = a, x_1 = \frac{a+b}{2}, x_2 = b$.

$$l_0(x) = \frac{x - x_1}{x_0 - x_1} \frac{x - x_2}{x_0 - x_2}$$

$$l_1(x) = \frac{x - x_0}{x_1 - x_0} \frac{x - x_2}{x_1 - x_2}$$

$$l_2(x) = \frac{x - x_0}{x_2 - x_0} \frac{x - x_1}{x_2 - x_1}$$

$$\begin{aligned} A_0 &= \int_a^b l_0(x)dx \\ &= \int_a^b \frac{x - x_1}{x_0 - x_1} \frac{x - x_2}{x_0 - x_2}dx \\ &= \int_a^b \left(\frac{x - \frac{a+b}{2}}{a - \frac{a+b}{2}} \right) \left(\frac{x - b}{a - b} \right) dx \end{aligned}$$

Changing variable of integration $y = x - \frac{a+b}{2}$

$$y = x - \frac{a+b}{2} \implies dy = dx$$

$$x = a \implies y = -h$$

$$x = b \implies y = h$$

Continuing with the value of A_0 ,

$$\begin{aligned} A_0 &= \int_{-h}^h \left(\frac{y}{-h} \right) \left(\frac{y-h}{-2h} \right) dy \\ &= \frac{1}{2h^2} \int_{-h}^h y^2 - \frac{1}{2h} \int_{-h}^h y dy \\ &= \frac{1}{2h^2} \left(\frac{h^3}{3} - \frac{(-h)^3}{3} \right) - \frac{1}{2h} \left(\frac{h^2}{2} - \frac{(-h)^2}{2} \right) \\ &= \frac{h}{3} \end{aligned}$$

$$\begin{aligned} A_1 &= \int_a^b l_1(x) dx \\ &= \int_a^b \frac{x-x_0}{x_1-x_0} \frac{x-x_2}{x_1-x_2} dx \\ &= \int_a^b \left(\frac{x-a}{h} \right) \left(\frac{x-b}{-h} \right) dx \end{aligned}$$

Applying change of variable, $y = x - \frac{a+b}{2}$

$$\begin{aligned} &= \int_{-h}^h \left(\frac{y+h}{h} \right) \left(\frac{y-h}{-h} \right) dy \\ &= \frac{-1}{h^2} \int_{-h}^h y^2 dy + \int_{-h}^h 1 dy \\ &= \frac{-1}{h^2} \left(\frac{h^3}{3} - \frac{(-h)^3}{3} \right) + (h - (-h)) \\ &= \frac{-2h^3}{3h^2} + 2h \\ &= \frac{4h}{3} \end{aligned}$$

$$\begin{aligned} A_2 &= \int_a^b l_2(x) dx \\ &= \int_a^b \left(\frac{x-x_0}{x_2-x_0} \right) \left(\frac{x-x_1}{x_2-x_1} \right) dx \\ &= \int_a^b \left(\frac{x-a}{2h} \right) \left(\frac{x-\frac{a+b}{2}}{h} \right) dx \end{aligned}$$

Applying change of variable, $y = x - \frac{a+b}{2}$

$$\begin{aligned}
 &= \int_{-h}^h \left(\frac{y+h}{2h} \right) \left(\frac{y}{h} \right) dy \\
 &= \frac{1}{2h^2} \int_{-h}^h y^2 dy + \frac{1}{2h} \int_{-h}^h y dy \\
 &= \frac{1}{2h^2} \left(\frac{h^3}{3} - \frac{(-h)^3}{3} \right) + \frac{1}{2h} \left(\frac{h^2}{2} - \frac{(-h)^2}{2} \right) \\
 &= \frac{2h^3}{6h^2} \\
 &= \frac{h}{3}
 \end{aligned}$$

Therefore,

$$\int_a^b f(x) dx = \sum_{i=0}^2 A_i f(x_i) = \frac{h}{3} \left(f(a) + 4f\left(\frac{a+b}{2}\right) + f(b) \right)$$

8.7.2 Composite Trapezoidal Rule

Suppose an interval (a, b) is divided into n subintervals. In Composite Trapezoidal Rule, Trapezoidal Rule is applied to each subinterval. Thus we have,

$$\begin{aligned}
 I &= I_0 + I_1 + \cdots + I_{n-1} \text{ where } I_k \text{ is the integral over } (x_k, x_{k+1}) \\
 &= \frac{h(f(x_0) + f(x_1))}{2} + \frac{h(f(x_1) + f(x_2))}{2} + \cdots + \frac{h(f(x_{n-1}) + f(x_n))}{2} \\
 &= \frac{h}{2} (f(x_0) + 2f(x_1) + \cdots + 2f(x_{n-1}) + f(x_n))
 \end{aligned}$$

8.7.3 Recursive Trapezoidal Rule

Suppose an interval (a, b) is divided into 2^{k-1} subintervals. In Recursive Trapezoidal Rule, we apply Trapezoidal Rule on each subinterval. And there is a recursive formula since the number of intervals doubles as value of k increases by one. Let $H = b - a$

$$k = 1 \implies 2^0 = 1 \text{ and}$$

$$I_1 = \frac{H}{2} (f(a) + f(b))$$

$$k = 2 \implies 2^1 = 2 \text{ and}$$

$$\begin{aligned}
 I_2 &= \frac{H}{4} (f(a) + 2f\left(a + \frac{H}{2}\right) + f(b)) \\
 &= \frac{H}{4} (f(a) + f(b)) + \frac{H}{2} f\left(a + \frac{H}{2}\right) \\
 &= \frac{I_1}{2} + \frac{H}{2} f\left(a + \frac{H}{2}\right)
 \end{aligned}$$

$k = 3 \implies 2^2 = 4$ and

$$\begin{aligned} I_3 &= \frac{H}{8} \left(f(a) + 2f\left(a + \frac{H}{4}\right) + 2f\left(a + \frac{2H}{4}\right) + 2f\left(a + \frac{3H}{4}\right) + f(b) \right) \\ &= \frac{H}{8} \left(f(a) + 2f\left(a + \frac{2H}{4}\right) + 2f\left(a + \frac{4H}{4}\right) + 2f\left(a + \frac{6H}{4}\right) + f(b) \right) \\ &\quad + \frac{H}{4} \left(f\left(a + \frac{H}{4}\right) + f\left(a + \frac{3H}{4}\right) \right) \\ &= \frac{I_2}{2} + \frac{H}{4} \left(f\left(a + \frac{H}{4}\right) + f\left(a + \frac{3H}{4}\right) \right) \end{aligned}$$

Consider interval $(0, 64)$. We have $b - a = H = 64$.

$$I_1 = 32(f(0) + f(64))$$

$$I_2 = 16(f(0) + 2f(32) + f(64))$$

$$I_3 = 8(f(0) + 2f(16) + 2f(32) + 2f(48) + f(64))$$

$$I_4 = 4(f(0) + 2f(8) + 2f(16) + \cdots + 2f(56) + f(64))$$

$$I_5 = 2(f(0) + 2f(4) + 2f(8) + \cdots + 2f(60) + f(64))$$

$$I_6 = f(0) + 2f(2) + 2f(4) + \cdots + 2f(62) + f(64)$$

\vdots

The values corresponding to the intervals in I_{k-1} appear in I_k as alternate terms. Other terms, corresponds to the odd multiples of $\frac{H}{2^k}$. We separate them into two sums and represent the first sum as $\frac{I_{k-1}}{2}$.

$$\begin{aligned} \text{Clearly, } I_k &= \frac{H}{2^k} \sum_{i=0}^{2^{k-2}} f\left(a + \frac{2iH}{2^k}\right) + \frac{2H}{2^k} \sum_{i=1}^{2^{k-2}} f\left(a + \frac{(2i-1)H}{2^{k-1}}\right) \\ &= \frac{I_{k-1}}{2} + \frac{H}{2^{k-1}} \sum_{i=1}^{2^{k-2}} f\left(a + \frac{(2i-1)H}{2^{k-1}}\right) \end{aligned}$$

8.7.4 Python : Recursive Trapezoidal Rule

Program 8.7.1.

```
def recursiveTrapezoidalRule(f,a,b,Iold,k):
    if k == 1 :
        Inew = (f(a)+f(b))*(b-a)/2.0
    else :
        n = 2**(k-2)
        h = (b-a) * 1.0 / n
        x = a + h/2.0
        sum = 0.0
        for i in range(n):
            sum = sum + f(x)
            x = x + h
        Inew = (Iold + h * sum) / 2.0
    return Inew
```

Line 1 `def recursiveTrapezoidalRule(f,a,b,Iold,k):`

This function has five input arguments. (a) f is a real function (b) a, b are start and end of interval in which f is going to be integrated (c) $Iold$ is

the value of the integral for 2^{k-1} subintervals using recursive trapezoidal method (d) k is a variable such that 2^k is the number of subintervals considered for Integration.

- Line 2 **if** $k == 1$:
 If $k = 1$, we proceed to Line 3, otherwise we go to Line 4.
- Line 3 $I_{new} = (f(a) + f(b)) * (b - a) / 2.0$
 For $k = 1$, we use trapezoidal rule $I_1 = \frac{b-a}{2}(f(a) + f(b))$. When writing this in python, we use 2.0 so that the python won't ignore the decimal part of this fraction. In python, $5/2 = 2$. And $5/2.0 = 2.5$
- Line 4 **else** :
 If Line 2 is false, (ie $k \neq 1$) python executes Line 5-8. These line implements the recursive formula for I_k .
- Line 5 $n = 2 * (k - 2)$
 Equivalent to $n \leftarrow 2^{k-2}$.
- Line 6 $h = (b - a) * 1.0 / n$
 This the length of a subinterval when we divide (a, b) into 2^k subintervals/panels. Equivalent to $h \leftarrow \frac{b-a}{2^{k-2}}$
- Line 7 $x = a + h/2.0$
 This the parameter of f in the first term in the sum $\sum_{i=1}^{2^{k-2}} f\left(a + \frac{(2i-1)H}{2^{k-1}}\right)$ in the recursive formula for I_k . Equivalent to $x \leftarrow a + \frac{h}{2}$.
- Line 8 $sum = 0.0$
 We are going to use this variable to find that sum. To start with, we will make it 0 and will add each term to it one-by-one. Equivalent to $sum \leftarrow 0$.
- Line 9 **for** i **in** $range(n)$:
 This the variable i in the recursive formula for I_k . For each value of $i = 1, 2, \dots, 2^{k-2}$, Lines 10 and 11 are executed. That is, for each value of i , the corresponding term in the sum is computed and added to the variable sum .
- Line 10 $sum = sum + f(x)$
 Value of f at x is computed and added to the partial sum. Equivalent to $sum \leftarrow sum + f(x)$. However, the value of x is changed for each i in Line 11. Thus, for next value of i , x and sum have the new values to use.
- Line 11 $x = x + h$
 Equivalent to $x \leftarrow x + h$. For $i = 1$, $x = a + \frac{h}{2}$ before Line 11. At Line 11, $x \leftarrow a + \frac{3h}{2}$. And this is the value of x for $i = 2$ before Line 11 next time. Thus, x iterates through $a + \frac{h}{2}, a + \frac{3h}{2}, \dots, a + \frac{(2n-1)h}{2}$. This x is updated and used for next execution of Line 10 and 11.
- Line 12 $I_{new} = (I_{old} + h * sum) / 2.0$
 We reach here only after executing Line 10-11 for all values of i . That is, sum in the recursive formula is already computed. This line, implements the recursive formula and stores that value into the variable I_{new} . Equivalent to $I_{new} \leftarrow \frac{I_{old} + h * sum}{2}$.

Line 13 **return Inew**

It returns the value of I_k , the integral of f over (a, b) using recursive trapezoidal rule for 2^k subintervals.

Subject 9

ME010204 Complex Analysis

Module 1

9.1 Complex Numbers

The Spherical Representation

Definitions 9.1.1 (extended complex plane). Let \mathbb{C} be the set of all complex numbers of the form $z = x + iy$ where x, y are real numbers. And,

$$z_1 + z_2 = (x_1 + iy_1) + (x_2 + iy_2) = (x_1 + x_2) + i(y_1 + y_2)$$

$$z_1 z_2 = (x_1 + iy_1)(x_2 + iy_2) = x_1 x_2 + ix_1 y_2 + iy_1 x_2 + i^2 y_1 y_2 = (x_1 x_2 - y_1 y_2) + i(x_1 y_2 + x_2 y_1)$$

We also have,

$$z_1/z_2 = \frac{x_1 + iy_1}{x_2 + iy_2} = \frac{(x_1 + iy_1)(x_2 - iy_2)}{(x_2 + iy_2)(x_2 - iy_2)} = \frac{x_1 x_2 + y_1 y_2}{x_2^2 + y_2^2} + i \frac{x_2 y_1 - x_1 y_2}{x_2^2 + y_2^2}, (z_2 \neq 0)$$

Define ∞ such that

$$z/0 = \infty, \forall z \in \mathbb{C}, (z \neq 0)$$

$$z \pm \infty = \infty, \forall z \in \mathbb{C}$$

$$z\infty = \infty, \forall z \in \mathbb{C} (z \neq 0)$$

$$z/\infty = 0, \forall z \in \mathbb{C}$$

$$\infty/z = \infty, \forall z \in \mathbb{C} (z \neq 0)$$

Then set \mathbb{C} together with ∞ is the extended complex plane, \mathbb{C}^* .

Definitions 9.1.2 (Riemann Sphere). The function which maps $z \rightarrow (x_1, x_2, x_3)$ such that

$$z = \frac{x_1 + ix_2}{1 - x_3} \quad (9.1)$$

And maps $\infty \rightarrow (0, 0, 1)$. Now we have a bijection from extended complex plane into the unit sphere in \mathbb{R}^3 . This spherical representation of extended complex plane is the Riemann sphere.

Motivation The significance of this projection is that on the unit sphere, point $(0, 0, 1)$ represents ∞ , the point at infinity.

Riemann Sphere : Justification Let S be the unit sphere in three dimensional space.

$$S : x_1^2 + x_2^2 + x_3^2 = 1 \quad (9.2)$$

Consider the map given by (eqn.9.1). Then,

$$|z|^2 = \frac{x_1^2 + x_2^2}{(1 - x_3)^2} = \frac{1 - x_3^2}{(1 - x_3)^2} = \frac{1 + x_3}{1 - x_3} \quad (\because \text{eqn.9.2})$$

$$\frac{|z|^2 - 1}{|z|^2 + 1} = \frac{(1 + x_3) - (1 - x_3)}{(1 + x_3) + (1 - x_3)} = x_3$$

$$1 - x_3 = \frac{(|z|^2 + 1) - (|z|^2 - 1)}{|z|^2 + 1} = \frac{2}{|z|^2 + 1}$$

Therefore,

$$x_1 = \frac{(z + \bar{z})(1 - x_3)}{2} = \frac{z + \bar{z}}{|z|^2 + 1} \quad (\because z + \bar{z} = \frac{2x_1}{1 - x_3}) \quad (9.3)$$

$$x_2 = \frac{(z - \bar{z})(1 - x_3)}{2i} = \frac{z - \bar{z}}{i(|z|^2 + 1)} \quad (\because z - \bar{z} = \frac{2ix_2}{1 - x_3}) \quad (9.4)$$

$$x_3 = \frac{|z|^2 - 1}{|z|^2 + 1} \quad (9.5)$$

Clearly, the map $z \rightarrow \frac{x_1 + ix_2}{1 - x_3}$ is a one-one correspondence. And when $x_3 = 0$, we have $|z| = 1$. Consider the hemisphere, $x_3 < 0$. Then $|z| < 1$ since $|z|^2 < 1$. Similarly, when $x_3 > 0$ we have $|z| > 1$.

Riemann Sphere : Sterographic Projection Let $z = x + iy$. Consider the map $z \rightarrow (x, y)$. Then,

$$x : y : -1 = \frac{x_1}{1 - x_3} : \frac{x_2}{1 - x_3} : -1 = x_1 : x_2 : x_3 - 1$$

Thus we have,

$$\begin{vmatrix} 0 & 0 & 1 \\ x & y & 0 \\ x_1 & x_2 & x_3 \end{vmatrix} = \begin{vmatrix} 0 & 0 & 1 \\ x & y & -1 \\ x_1 & x_2 & x_3 - 1 \end{vmatrix} = 0$$

Clearly, the points $(x, y, 0)$, $(0, 0, 1)$ and (x_1, x_2, x_3) lies on a straight line.

Definitions 9.1.3 (collinear). Let $(x_1, x_2, x_3), (x'_1, x'_2, x'_3), (x''_1, x''_2, x''_3) \in \mathbb{R}^3$. These points are collinear if

$$\begin{vmatrix} x_1 & x_2 & x_3 \\ x'_1 & x'_2 & x'_3 \\ x''_1 & x''_2 & x''_3 \end{vmatrix} = 0$$

If you consider the complex plane as the plane $x_3 = 0$ in the three dimensional space, then the straight lines through $(0, 0, 1)$ will geometrically give you the image and preimage under the bijection given by (eqn.9.1). Therefore, this is a stereographic projection.

From $(0, 0, 1)$, you could draw a line which gives you the correspondence between the sphere and the complex plane. This projection is stereographic in the sense that you could see points on the unit sphere projected on the plane as well as the points on the plane projected on the unit sphere using the same straight line.

Circles on the Riemann Sphere

Remark. The stereographic projection $z = \frac{x_1 + ix_2}{1 - x_3}$ transforms straight lines in the complex plane into a circle on the unit sphere S through $(0, 0, 1)$. And any circle on the unit sphere transforms into a circle or straight line in the complex plane.

Proof. Any circle on the unit sphere is an intersection of the unit sphere $S : x_1^2 + x_2^2 + x_3^2 = 1$ and a plane $P : \alpha_1 x_1 + \alpha_2 x_2 + \alpha_3 x_3 = \alpha_0$.

Without loss of generality, $\alpha_1^2 + \alpha_2^2 + \alpha_3^2 = 1$ and $0 \leq \alpha_0 < 1$. Suppose $P : \alpha'_1 x_1 + \alpha'_2 x_2 + \alpha'_3 x_3 = \alpha'_0$. Suppose $\alpha'_0 > 0$. If $\alpha'_0 < 0$, then consider $P : -\alpha'_1 x_1 - \alpha'_2 x_2 - \alpha'_3 x_3 = -\alpha'_0$. And define

$$\alpha_1 = \frac{\alpha'_1}{|\alpha'|} \quad \alpha_2 = \frac{\alpha'_2}{|\alpha'|} \quad \alpha_3 = \frac{\alpha'_3}{|\alpha'|} \quad \alpha_0 = \frac{\alpha'_0}{|\alpha'|}$$

where $|\alpha'| = \sqrt{\alpha'^2_1 + \alpha'^2_2 + \alpha'^2_3}$. Thus, $\alpha_1^2 + \alpha_2^2 + \alpha_3^2 = 1$ and $0 \leq \alpha_0 < 1$.

For any circle on the unit sphere, we have a plane intersecting the sphere at a positive distance α_0 from the origin which is parallel to the tangent plane at the point $(\alpha_1, \alpha_2, \alpha_3)$ on the unit sphere. Clearly, the plane won't touch the sphere if $\alpha_0 > 1$ and when $\alpha_0 = 1$ the plane is tangential to the sphere. Thus, $0 \leq \alpha_0 < 1$.

Substituting the values of x_1, x_2, x_3 , we get

$$\alpha_1 \frac{z + \bar{z}}{|z|^2 + 1} + \alpha_2 \frac{z - \bar{z}}{i(|z|^2 + 1)} + \alpha_3 \frac{|z|^2 - 1}{|z|^2 + 1} = \alpha_0$$

$$\alpha_1(z + \bar{z}) - i\alpha_2(z - \bar{z}) + \alpha_3(|z|^2 - 1) = \alpha_0(|z|^2 + 1)$$

We have, $z = x + iy$. Then $|z|^2 = x^2 + y^2$, $z + \bar{z} = 2x$ and $z - \bar{z} = 2iy$.

$$2x\alpha_1 + 2y\alpha_2 + x^2\alpha_3 + y^2\alpha_3 - \alpha_3 = x^2\alpha_0 + y^2\alpha_0 + \alpha_0$$

$$x^2(\alpha_3 - \alpha_0) + y^2(\alpha_3 - \alpha_0) + 2x\alpha_1 + 2y\alpha_2 - (\alpha_3 + \alpha_0) = 0$$

Suppose $\alpha_3 \neq \alpha_0$. Then, we have

$$x^2(\alpha_3 - \alpha_0) + y^2(\alpha_3 - \alpha_0) + 2x\alpha_1 + 2y\alpha_2 - (\alpha_3 + \alpha_0) = 0 \quad (9.6)$$

Clearly, (eqn.9.6) represents a circle in the complex plane.

Definitions 9.1.4 (circle). Circle with center (h, k) and radius r is given by

$$(x - h)^2 + (y - k)^2 = r^2$$

$$x^2 + y^2 - 2xh - 2ky + (h^2 + k^2 - r^2) = 0$$

In the above case, we have

$$x^2 + y^2 + 2x \frac{\alpha_1}{\alpha_3 - \alpha_0} + 2y \frac{\alpha_2}{\alpha_3 - \alpha_0} - \frac{\alpha_3 + \alpha_0}{\alpha_3 - \alpha_0} = 0$$

is a circle with center $\left(\frac{\alpha_1}{\alpha_0 - \alpha_3}, \frac{\alpha_2}{\alpha_0 - \alpha_3}\right)$ and radius $\frac{\sqrt{1 - \alpha_0^2}}{\alpha_0 - \alpha_3}$.

Suppose $\alpha_3 = \alpha_0$. Then (eqn.9.6) becomes $x\alpha_1 + y\alpha_2 - \alpha_0 = 0$. Clearly, it represents a straight line in the complex plane. \square

Any straight line transforms into a circle through $(0, 0, 1)$ on the Riemann sphere. Thus, straight lines are circles through the point at infinity.

Distance between two points on the Riemann Sphere Let $z, z' \in \mathbb{C}$. Let $z \rightarrow (x_1, x_2, x_3)$ and $z' \rightarrow (x'_1, x'_2, x'_3)$. We have,

$$|z - z'|^2 = (z - z')(\bar{z} - \bar{z}') = |z|^2 + |z'|^2 - (z\bar{z}' + \bar{z}z')$$

And,

$$(|z|^2 + 1)(|z'|^2 + 1) - (|z|^2 - 1)(|z'|^2 - 1) = 2(|z|^2 + |z'|^2)$$

Thus,

$$z\bar{z}' + \bar{z}z' = |z|^2 + |z'|^2 - |z - z'|^2$$

$$2(z\bar{z}' + \bar{z}z') = \{(|z|^2 + 1)(|z'|^2 + 1) - (|z|^2 - 1)(|z'|^2 - 1)\} - 2|z - z'|^2$$

Rearranging the terms, we get

$$2z\bar{z}' + 2\bar{z}z' + (|z|^2 - 1)(|z'|^2 - 1) = |z|^2 + 1 + |z'|^2 + 1 - 2|z - z'|^2 \quad (9.7)$$

Let $d(z, z')$ denote the distance between their images on the sphere. Then,

$$(d(z, z'))^2 = (x_1 - x'_1)^2 + (x_2 - x'_2)^2 + (x_3 - x'_3)^2$$

$$= (x_1^2 + x_2^2 + x_3^2) + (x'^2_1 + x'^2_2 + x'^2_3) - 2(x_1x'_1 + x_2x'_2 + x_3x'_3)$$

$$= 2 - 2(x_1x'_1 + x_2x'_2 + x_3x'_3)$$

Substituting the values, we get

$$x_1x'_1 + x_2x'_2 + x_3x'_3 = \frac{(z + \bar{z})(z' + \bar{z}') + i^2(z - \bar{z})(z' - \bar{z}') + (|z|^2 - 1)(|z'|^2 - 1)}{(|z|^2 + 1)(|z'|^2 + 1)}$$

$$= \frac{2z\bar{z}' + 2\bar{z}z' + (|z|^2 - 1)(|z'|^2 - 1)}{(|z|^2 + 1)(|z'|^2 + 1)}$$

$$= \frac{(|z|^2 + 1)(|z'|^2 + 1) - 2|z - z'|^2}{(|z|^2 + 1)(|z'|^2 + 1)} \quad (\because \text{eqn.9.7})$$

$$= 1 - \frac{2|z - z'|^2}{(|z|^2 + 1)(|z'|^2 + 1)}$$

$$(d(z, z'))^2 = \frac{4|z - z'|^2}{(|z|^2 + 1)(|z'|^2 + 1)}$$

Therefore, we have

$$d(z, z') = \frac{2|z - z'|}{\sqrt{(1 + |z|^2)(1 + |z'|^2)}} \quad (9.8)$$

Suppose $z' = \infty$. Then $\frac{|z - z'|}{\sqrt{1 + |z'|^2}} \rightarrow 1$ as $z' \rightarrow \infty$. Therefore,

$$d(z, \infty) = \frac{2}{\sqrt{1 + |z|^2}} \quad (9.9)$$

Exercises

1. z, z' are diametrically opposite points on the Riemann Sphere if and only if $zz' = -1$.

Proof. Suppose $zz' = -1$. Then $z' = \frac{-z}{|z|^2}$.

$$|z - z'| = \left| z + \frac{z}{|z|^2} \right| = \frac{|z|^2 + 1}{|z|}$$

$$1 + |z'|^2 = 1 + \frac{1}{|z|^2} = \frac{|z|^2 + 1}{|z|^2}$$

Then,

$$d(z, z') = \frac{2|z - z'|}{\sqrt{(1 + |z|^2)(1 + |z'|^2)}} = \frac{2 \frac{|z|^2 + 1}{|z|}}{\sqrt{(1 + |z|^2) \frac{|z|^2 + 1}{|z|^2}}} = 2$$

Clearly, the points are diametrically opposite on the unit sphere.

Suppose the points are diametrically opposite on the Riemann sphere. Then

$$|z - z'|^2 = (1 + |z|^2)(1 + |z'|^2)$$

$$\begin{aligned} |z - z'|^2 &= (z - z')(\bar{z} - \bar{z}') \\ &= |z|^2 + |z'|^2 - z\bar{z}' - \bar{z}z' \\ &= 1 + |z|^2 + |z'|^2 + |z|^2|z'|^2 \\ 0 &= 1 + z\bar{z}' + \bar{z}z' + |z|^2|z'|^2 \\ &= (1 + z\bar{z}')(1 + \bar{z}z') \end{aligned}$$

Then either $z\bar{z}' + 1 = 0$ or $\bar{z}z' + 1 = 0$. In either case, we have $zz' = -1$ since $z\bar{z}' = \bar{z}z'$ when they are purely real. \square

2. Find vertices of the cube inscribed in the Riemann sphere with edges parallel to the coordinate axes.

Answer. Let a be the length of the sides of the cube inscribed in the Riemann sphere. Then the main diagonal is of length $a\sqrt{3} = 2$, the diameter of the unit sphere. Thus, $a = 2/\sqrt{3}$. The vertices of the cube on the Riemann sphere are $(\pm a/2, \pm a/2, \pm a/2)$. The corresponding images are given by

$$z = \frac{a}{(2 \pm a)}(\pm 1 \pm i) = \frac{\pm 1 \pm i}{\sqrt{3} \pm 1}$$

Therefore, the vertices are $\frac{1+i}{\sqrt{3}+1}$, $\frac{1-i}{\sqrt{3}+1}$, $\frac{-1+i}{\sqrt{3}+1}$, $\frac{-(1+i)}{\sqrt{3}+1}$, $\frac{1+i}{\sqrt{3}-1}$, $\frac{1-i}{\sqrt{3}-1}$, $\frac{-1+i}{\sqrt{3}-1}$, and $\frac{-(1+i)}{\sqrt{3}-1}$.

□

3. Find the vertices of the regular tetrahedron inscribed in Riemann Sphere.

Answer. Let a be the length of the sides of the regular tetrahedron inscribed in the Riemann sphere. Radius of circumsphere of regular tetrahedron is $\frac{\sqrt{6}}{4}a = 1$. Thus, $a = 4/\sqrt{6}$.

One vertex at $(0, 0, 1)$ is the point at infinity. The other three points are at a distance of $4/\sqrt{6}$ from $(0, 0, 1)$. We have,

$$d(z, \infty) = \frac{2}{\sqrt{1+|z|^2}} = \frac{4}{\sqrt{6}} = \frac{2}{\sqrt{\frac{6}{4}}}$$

Thus, $1 + |z|^2 = 1 + \frac{1}{4}$. Therefore, $|z| = \frac{1}{2}$.

In its general position the projection of one of the base vertex on the regular tetrahedron is purely imaginary $(\frac{i}{2})$ and the other vertices makes 120° with the imaginary axis. Therefore, the vertices are ∞ , $\frac{i}{2}$, $\frac{1}{4}(-\sqrt{3}-i)$, and $\frac{1}{4}(\sqrt{3}-i)$ □

4. Find the radius of the spherical image of the circle $|z - a| = r$.

Answer. The spherical image is the intersection of the unit sphere $x_1^2 + x_2^2 + x_3^2 = 1$ and $\alpha_1 x_1 + \alpha_2 x_2 + \alpha_3 x_3 = \alpha_0$, where $\alpha_1^2 + \alpha_2^2 + \alpha_3^2 = 1$ and $0 \leq \alpha_0 < 1$. The intersection is a circle with center $(\frac{\alpha_1}{\alpha_0 - \alpha_3}, \frac{\alpha_2}{\alpha_0 - \alpha_3})$ and radius $\frac{\sqrt{1-\alpha_0^2}}{\alpha_0 - \alpha_3}$. Clearly, the radius of the spherical image lies on this plane of intersection which is at α_0 distance from origin.

Suppose $a = x + iy$. Let $k = \alpha_0 - \alpha_3$. Then,

$$\alpha_1 = kx, \quad \alpha_2 = ky, \quad \alpha_3 = \sqrt{1 - k^2|a|^2}$$

And,

$$\begin{aligned} \alpha_0^2 &= (k + \alpha_3)^2 = k^2 + \alpha_3^2 + 2k\alpha_3 = k^2 + (1 - k^2|a|^2) + 2k(\alpha_0 - k) \\ \alpha_0^2 - 2k\alpha_0 + k^2 + k^2|a|^2 - 1 &= 0 \end{aligned} \quad (9.10)$$

We have $r^2 = \frac{1-\alpha_0^2}{k^2}$. Thus,

$$\alpha_0^2 + k^2 r^2 - 1 = 0 \quad (9.11)$$

From equations (9.10,9.11), we have

$$\begin{aligned} 2k\alpha_0 - k^2|a|^2 + k^2r^2 - k^2 &= 0 \\ \alpha_0 &= \frac{k(1 + |a|^2 - r^2)}{2} \end{aligned}$$

Thus,

$$\alpha_3^2 = 1 - k^2|a|^2 = (\alpha_0 - k)^2 = \frac{k^2(1 + r^2 - |a|^2)^2}{4}$$

Therefore,

$$\begin{aligned} k &= \frac{2}{\sqrt{(1 + r^2 - |a|^2)^2 + 4|a|^2}} \\ \alpha_0 &= \frac{1 + |a|^2 - r^2}{\sqrt{(1 + r^2 - |a|^2)^2 + 4|a|^2}} = \frac{\lambda + 2}{\sqrt{\lambda^2 + 4|a|^2}} \end{aligned}$$

where $\lambda = |a|^2 - r^2 - 1$. Now, we have right triangle with sides α_0 , R and 1. Thus, the radius R of the circle of intersection is given by

$$R = \sqrt{1 - \alpha_0^2} = \frac{2\sqrt{|a|^2 - \lambda - 1}}{\sqrt{\lambda^2 + 4|a|^2}} = \frac{2r}{\sqrt{(|a|^2 - r^2 - 1)^2 + 4|a|^2}}$$

□

Exercise 4 is a bit complicated. For $a = 3 + 4i$ and $r = 2$, I have substituted the value of $k = 0.04\sqrt{5}$ to find the equation for the plane. I find the value of $|\alpha_0| = 0.44\sqrt{5} = 0.98 < 1$. And $R = 0.18$.

There is an alternate solution for the radius of the spherical image of a circle at <https://math.stackexchange.com/q/190270>. Let me know if you have an opinion different from mine.

Let $z = a + r$ and $z' = a - r$. Then $|z - z'| = 2r$. Let R be the radius of the spherical image. We claim that the points z, z' are diametrically opposite on the spherical image. Then,

$$R = \frac{d(z, z')}{2} = \frac{|z - z'|}{\sqrt{(1 + |z|^2)(1 + |z'|^2)}} = \frac{2r}{\sqrt{(1 + |a + r|^2)(1 + |a - r|^2)}}$$

However, I think that the claim doesn't stand when $a \neq 0$. That is, the projections of diametrically opposite points of a circle need not be diametrically opposite on the spherical image of it.

Suppose the claim is true. Consider, $z = a + ir$ and $z' = a - ir$ are diametrically opposite points of the circle. Then, by above relation the radius of the spherical image assumes a different value, which is not possible.

9.2 Complex Functions

9.2.1 Analytic Function

Definitions 9.2.1 (region). A region Ω is an open, connected subset of \mathbb{C} .

Limit and Continuity

Definitions 9.2.2 (limit). Let Ω be a region. Let function $f : \Omega \rightarrow \mathbb{C}$. Function f has limit A as x tends to a if for every $\varepsilon > 0$, there exists $\delta > 0$ such that $|f(x) - A| < \varepsilon$ for any x such that $|x - a| < \delta$ and $x \neq a$.

Notation: $f(x) \rightarrow A$ as $x \rightarrow a$

Properties of Limit Suppose $f(x) \rightarrow A$ and $g(x) \rightarrow B$ as $x \rightarrow a$.

1. $(f + g)(x) \rightarrow A + B$ as $x \rightarrow a$
2. $(fg)(x) \rightarrow AB$ as $x \rightarrow a$
3. $(f/g)(x) \rightarrow A/B$ as $x \rightarrow a$ ^{†1}
4. $\overline{f(x)} \rightarrow \bar{A}$ as $x \rightarrow a$

Definitions 9.2.3 (continuity). f is continuous at a if $f(x) \rightarrow f(a)$ as $x \rightarrow a$. And $f : \Omega \rightarrow \mathbb{C}$ is continuous if it is continuous at each point of Ω .

Properties of Continuous Functions Let f, g be two continuous function.

1. $f \pm g, fg$ and f/g are continuous ($g \neq 0$).
2. $Re(f), Im(f), |f|$ are continuous.

Definitions 9.2.4 (derivative). The derivative of f at a is

$$f'(a) = \lim_{x \rightarrow a} \frac{f(x) - f(a)}{x - a} \quad (9.12)$$

Properties of Derivatives

1. $(f + g)'(a) = f'(a) + g'(a)$
2. $(fg)'(a) = f'(a)g(a) + f(a)g'(a)$
3. $(f/g)'(a) = \frac{g(a)f'(a) - f(a)g'(a)}{g^2(a)}$
4. $(f \circ g)'(a) = f'(g(a))g'(a)$

A real function of complex variable either has the derivative zero or else the derivative does not exist. And complex function of real variable can be reduced to the real case.

$$f(x) = g(x) + ih(x) \quad f'(x) = g'(x) + ih'(x)$$

^{†1}If g is bounded away from zero.

Analytic Functions

Definitions 9.2.5 (analytic). An analytic function is a complex function of complex variable which has derivative wherever it is defined.

Definitions 9.2.6 (Cauchy-Riemann). Cauchy-Riemann differential equations for a complex function $f : \Omega \rightarrow \mathbb{C}$, $f(z) = u(z) + iv(z)$ are

$$\frac{\partial f}{\partial x} = -i \frac{\partial f}{\partial y} \quad \text{OR} \quad \frac{\partial u}{\partial x} = \frac{\partial v}{\partial y} \quad \text{and} \quad \frac{\partial u}{\partial y} = -\frac{\partial v}{\partial x} \quad (9.13)$$

Then the pair (u, v) is said to satisfy Cauchy-Riemann differential equations. And $f = u + iv$ is said to satisfy Cauchy-Riemann conditions for analyticity.

Definitions 9.2.7 (harmonic). A real function $u : \Omega \rightarrow \mathbb{R}$ of complex variable is a harmonic function if $u(z)$ satisfies the Laplace equation for every $z \in \Omega$.

$$\Delta u = \frac{\partial^2 u}{\partial x^2} + \frac{\partial^2 u}{\partial y^2} = 0 \quad (9.14)$$

Definitions 9.2.8 (conjugate harmonic). Two harmonic functions u, v are conjugate harmonic if the pair (u, v) satisfy Cauchy-Riemann differential equations.

Properties of Analytic Functions Let f, g be analytic functions.

$$f'(z) = \lim_{h \rightarrow 0} \frac{f(z+h) - f(z)}{h}$$

1. $f \pm g, fg, f/g$ are analytic functions ($g \neq 0$).
2. Analytic functions are continuous.

Proof. Let $f : \Omega \rightarrow \mathbb{C}$ be an analytic function. Let $z \in \mathbb{C}$. Then for sufficiently small h , we have $z+h \in \Omega$. And,

$$f(z+h) - f(z) = h \frac{f(z+h) - f(z)}{h}$$

Thus, $f(z+h) - f(z) \rightarrow 0$ as $h \rightarrow 0$. In other words, $f(z+h) \rightarrow f(z)$ as $h \rightarrow 0$. Clearly, f is continuous. \square

3. Real and imaginary parts of analytic functions satisfies Cauchy-Riemann differential equations.

Proof. Let $f : \Omega \rightarrow \mathbb{C}$ be an analytic function. Let $z = x + iy \in \Omega$. Then $f'(z)$ exists. And the value of $f'(z)$ is independent of the path along which $z+h$ approaches z .

Suppose h is real, then $f(z+h) = f(x+h+iy)$. Thus,

$$f'(z) = \lim_{h \rightarrow 0} \frac{f(x+h+iy) - f(x+iy)}{h} = \frac{\partial f}{\partial x} \quad (9.15)$$

$$\begin{aligned} &= \lim_{h \rightarrow 0} \frac{u(x+h, y) - u(x, y)}{h} + \lim_{h \rightarrow 0} i \frac{v(x+h, y) - v(x, y)}{h} \\ &= \frac{\partial u}{\partial x} + i \frac{\partial v}{\partial x} \end{aligned} \quad (9.16)$$

Suppose h is imaginary, then $f(z+h) = f(x+i(y+k))$. Thus,

$$f'(z) = \lim_{k \rightarrow 0} \frac{f(x+i(y+k)) - f(x+iy)}{ik} = -i \frac{\partial f}{\partial y} \quad (9.17)$$

$$\begin{aligned} &= \lim_{k \rightarrow 0} \frac{u(x, y+k) - u(x, y)}{ik} + \lim_{k \rightarrow 0} i \frac{v(x, y+k) - v(x, y)}{ik} \\ &= -i \frac{\partial u}{\partial y} + \frac{\partial v}{\partial y} \end{aligned} \quad (9.18)$$

Comparing the equations, we get

$$\frac{\partial f}{\partial x} = -i \frac{\partial f}{\partial y} \quad \text{OR} \quad \frac{\partial u}{\partial x} = \frac{\partial v}{\partial y} \quad \text{and} \quad \frac{\partial u}{\partial y} = -\frac{\partial v}{\partial x}$$

Therefore, real and imaginary parts of f satisfies Cauchy-Riemann differential equations. In other words, f satisfying the Cauchy-Riemann conditions for analyticity. \square

4. $|f'(z)|^2$ is the Jacobian of u and v with respect to x and y .

Proof. Let $f = u + iv$ be an analytic function. Then,

$$\begin{aligned} f'(z) &= \frac{\partial u}{\partial x} + i \frac{\partial v}{\partial x} \\ |f'(z)|^2 &= \left(\frac{\partial u}{\partial x} \right)^2 + \left(\frac{\partial v}{\partial x} \right)^2 \\ &= \frac{\partial u}{\partial x} \frac{\partial v}{\partial y} - \frac{\partial u}{\partial y} \frac{\partial v}{\partial x} = \begin{vmatrix} \frac{\partial u}{\partial x} & \frac{\partial u}{\partial y} \\ \frac{\partial v}{\partial x} & \frac{\partial v}{\partial y} \end{vmatrix} \end{aligned}$$

\square

5. Derivative of an analytic function is analytic. ([proof : 9.4.2](#))
6. Analyticity implies existence of the partial derivatives $\frac{\partial u}{\partial x}, \frac{\partial u}{\partial y}, \frac{\partial v}{\partial x}, \frac{\partial v}{\partial y}$ of real and imaginary parts. Since derivative of an analytic function is also an analytic function, real and imaginary parts of an analytic function has continuous partial derivatives of all orders. And mixed partial derivatives are equal since the second order partial derivatives are continuous.
7. If $f = u + iv$ is analytic, then u, v are harmonic functions. And v is the unique conjugate harmonic function of u upto an additive constant.

Theorem 9.2.1 (conjugate harmonic). *Function $f = u + iv$ is an analytic function if and only if u, v are conjugate harmonic functions.*

Proof. Suppose v is the conjugate harmonic function of u . Then the four first order partial derivatives of u, v exists, they are continuous and they satisfy Cauchy-Riemann differential equations.

$$\begin{aligned} u(x+h, y+k) - u(x, y) &= u(x+h, y+k) - u(x, y+k) + u(x, y+k) - u(x, y) \\ &= h \frac{u(x+h, y+k) - u(x, y+k)}{h} + k \frac{u(x, y+k) - u(x, y)}{k} \end{aligned}$$

By existence and continuity of the partial derivatives,

$$u(x+h, y+k) - u(x, y) \rightarrow h \frac{\partial u}{\partial x} + k \frac{\partial u}{\partial y} \text{ as } h+ik \rightarrow 0$$

For sufficiently small values of h, k we have,

$$u(x+h, y+k) - u(x, y) = h \frac{\partial u}{\partial x} + k \frac{\partial u}{\partial y} + \varepsilon_1 \text{ where } \frac{\varepsilon_1}{h+ik} \rightarrow 0 \text{ as } h+ik \rightarrow 0$$

Similarly,

$$v(x+h, y+k) - v(x, y) = h \frac{\partial v}{\partial x} + k \frac{\partial v}{\partial y} + \varepsilon_2 \text{ where } \frac{\varepsilon_2}{h+ik} \rightarrow 0 \text{ as } h+ik \rightarrow 0$$

Substituting the values,

$$\begin{aligned} f(z+h+ik) - f(z) &= u(x+h, y+k) - u(x, y) + i(v(x+h, y+k) - v(x, y)) \\ &= h \frac{\partial u}{\partial x} + k \frac{\partial u}{\partial y} + \varepsilon_1 + i(h \frac{\partial v}{\partial x} + k \frac{\partial v}{\partial y} + \varepsilon_2) \\ &= \left(\frac{\partial u}{\partial x} + i \frac{\partial v}{\partial x} \right) (h+ik) + (\varepsilon_1 + i\varepsilon_2) \\ \implies f'(z) &\rightarrow \frac{\partial u}{\partial x} + i \frac{\partial v}{\partial x} \text{ as } h+ik \rightarrow 0 \end{aligned}$$

Thus, f is analytic.

Suppose $f = u + iv$ is an analytic function. We know that,

$$f'(z) = \frac{\partial u}{\partial x} + i \frac{\partial v}{\partial x} = -i \frac{\partial u}{\partial y} + \frac{\partial v}{\partial y}$$

Suppose u, v are harmonic functions. Comparing real and imaginary parts, we have v is the conjugate harmonic function of u upto an additive constant.

It remains to prove that u, v are harmonic functions. We have,

$$\begin{aligned} \frac{\partial u}{\partial x} = \frac{\partial v}{\partial y} &\implies \frac{\partial^2 u}{\partial x^2} = \frac{\partial^2 v}{\partial x \partial y} \\ \frac{\partial u}{\partial y} = -\frac{\partial v}{\partial x} &\implies \frac{\partial^2 u}{\partial y^2} = -\frac{\partial^2 v}{\partial y \partial x} \end{aligned}$$

Since the second order partial derivatives of v are continuous, these mixed partial derivatives of v are equal.

$$\Delta u = \frac{\partial^2 u}{\partial x^2} + \frac{\partial^2 u}{\partial y^2} = \frac{\partial^2 v}{\partial x \partial y} - \frac{\partial^2 v}{\partial y \partial x} = 0$$

Therefore, u is harmonic and similarly v is also harmonic. \square

Remark. An analytic function can't be reduced to a function of two real variables.

$$\begin{aligned}
f(z) &= f(x + iy) \\
&= f\left(\frac{z + \bar{z}}{2} - i\frac{z - \bar{z}}{2}\right) \\
\Rightarrow \frac{\partial f}{\partial \bar{z}} &= \frac{1}{2} \frac{\partial f}{\partial x} + \frac{i}{2} \frac{\partial f}{\partial y} = 0 \quad \left(\because f \text{ is analytic, } \frac{\partial f}{\partial x} = -i \frac{\partial f}{\partial y}\right)
\end{aligned}$$

Thus, f is independent of \bar{z} . Therefore, an analytic function $f(x + iy)$ can't be truly represented by $f(x, y)$.

Computing f from rational function $u(x, y)$

$$f(z) = 2u(z/2, z/2i) - u(0, 0) + c \quad (9.19)$$

Exercise

1. Composition of analytic functions is also analytic.

Proof.

$$(g \circ f)'(z) = \lim_{h \rightarrow 0} \frac{(g \circ f)(z + h) - (g \circ f)(z)}{h} \quad ?$$

□

2. (a) $z^2 = x^2 - y^2 + 2xyi$

$$\frac{\partial f}{\partial x} = 2x + 2iy = -i(2ix - 2y) \quad \frac{\partial f}{\partial y} = -2y + 2ix$$

- (b) $z^3 = x^3 - 3xy^2 + i(3x^2y - y^3)$

$$\frac{\partial f}{\partial x} = 3x^2 - 3y^2 + 6ixy = -i(3ix^2 - 3iy^2 - 6xy) \quad \frac{\partial f}{\partial y} = -6xy + 3ix^2 - 3iy^2$$

3. $u = ax^3 + bx^2y + cxy^2 + dy^3$

$$\frac{\partial u}{\partial x} = 3ax^2 + 2bxy + cy^2$$

$$\frac{\partial^2 u}{\partial x^2} = 6ax + 2by$$

$$\frac{\partial u}{\partial y} = bx^2 + 2cxy + 3dy^2$$

$$\frac{\partial^2 u}{\partial y^2} = 2cx + 6dy$$

$$\Delta u = 0 \implies c = -3a, d = -b/3$$

$$u = ax^3 + bx^2y - 3axy^2 - y^2/3$$

$$\frac{\partial v}{\partial y} = 3ax^2 + 2bxy - 3ay^2$$

$$\frac{\partial v}{\partial x} = -bx^2 + 6axy + by^2$$

4.

5.

6.

7.

Polynomials

Polynomials are analytic functions Polynomials are the *simplest* analytic functions.

Proof. Constant functions are analytic with derivative 0 and z is the simplest non-constant analytic function and it has derivative 1. Since sum and product of analytic functions are analytic, the polynomial

$$P(z) = a_0 + a_1z + \cdots + a_nz^n, \quad (a_n \neq 0) \quad (9.20)$$

with degree n is analytic with derivative

$$P'(z) = a_1 + 2a_2z + \cdots + na_nz^{n-1} \quad (9.21)$$

□

By fundamental theorem of algebra (theorem 9.4.4), every polynomial $P(z)$ with degree $n > 0$ has at least one root, say α_1 . Thus, $P(z) = (z - \alpha_1)P_1(z)$ where $P_1(z)$ is a polynomial of degree $n - 1$. Repeating the process, we get $P(z) = a_n(z - \alpha_1)(z - \alpha_2) \cdots (z - \alpha_n)$ where α_j 's are not necessarily distinct.

Order of a zero These α_j 's are the zeroes of $P(z)$. If there are h number of α_j 's such that $\alpha = \alpha_j$. Then α is a **zero of order h** .

Also order of a zero α is the order of the first nonvanishing derivative of $P(z)$ at α . That is, the least positive integer m such that the derivative

$$P^{(h)}(\alpha) = \left. \frac{d^h}{dz^h} P(z) \right|_{z=\alpha} \neq 0$$

Thus, **simple zero** of $P(z)$ is an α such that $P(\alpha) = 0$ and $P'(\alpha) \neq 0$.

Theorem 9.2.2 (Lucas). *If all zeroes of a polynomial $P(z)$ lies in a half plane, then all zeroes of the derivative $P'(z)$ lie in the same half plane.*

Proof. Let H be the half plane defined by

$$H = \left\{ z : \operatorname{Im} \left(\frac{z - a}{b} \right) < 0 \right\}$$

Suppose every zero α_k of $P(z)$ belong to H . Then,

$$\operatorname{Im} \left(\frac{\alpha_k - a}{b} \right) < 0, \quad \forall k$$

Suppose z belongs to the other half plane. Then, $P(z) \neq 0$ and

$$\operatorname{Im} \left(\frac{z - a}{b} \right) > 0$$

Thus,

$$\operatorname{Im} \left(\frac{z - \alpha_k}{b} \right) = \operatorname{Im} \left(\frac{z - a}{b} - \frac{\alpha_k - a}{b} \right) = \operatorname{Im} \left(\frac{z - a}{b} \right) - \operatorname{Im} \left(\frac{\alpha_k - a}{b} \right) > 0$$

Since imaginary part of reciprocals have opposite sign, we have

$$\operatorname{Im}\left(\frac{b}{z - \alpha_k}\right) < 0, \forall k$$

We have,

$$\begin{aligned} \log P(z) &= \log a_n + \log(z - \alpha_1) + \log(z - \alpha_2) + \cdots + \log(z - \alpha_n) \\ \Rightarrow \frac{P'(z)}{P(z)} &= \frac{1}{z - \alpha_1} + \frac{1}{z - \alpha_2} + \cdots + \frac{1}{z - \alpha_n} = \sum_{k=1}^n \frac{1}{z - \alpha_k} \\ \Rightarrow \operatorname{Im}\left(\frac{bP'(z)}{P(z)}\right) &= \sum_{k=1}^n \operatorname{Im}\left(\frac{b}{z - \alpha_k}\right) < 0 \end{aligned}$$

Therefore, $P'(z) \neq 0$ for any $z \notin H$. □

Theorem 9.2.3 (Lucas). *The smallest convex polygon containing the zeroes of $P(z)$ contains all the zeroes of $P'(z)$. In other words, it contains all critical points of $P(z)$.*

Proof. Let C be the smallest convex polygon containing all zeroes of $P(z)$. Without loss of generality, all the zeroes of $P(z)$ lies on the same half plane defined by each edge of this polygon. By above proof, we know that the critical points of $P(z)$ which are the zeroes of $P'(z)$ belongs to the same half plane. □

Rational Functions

Definitions 9.2.9. A function of the form $R(z) = \frac{P(z)}{Q(z)}$ where $P(z), Q(z)$ are polynomials such that they don't have any common zeroes.

The zeroes of $P(z)$ are the **zeroes** of $R(z)$ and the zeroes of $Q(z)$ are the **poles** of $R(z)$. The **order of a pole** of $R(z)$ is the order of the respective zero of $Q(z)$.

Derivative The derivative of $R(z)$ exists only when $Q(z) \neq 0$.

$$R'(z) = \frac{P'(z)Q(z) - P(z)Q'(z)}{Q(z)^2}$$

Order of a rational function Rational functions have common number of zeroes and poles in the extended plane. This number is the **order** of the rational function.

Let n, m be the order of the polynomials $P(z), Q(z)$. The order of the rational function $R(z) = \frac{P(z)}{Q(z)}$ is $\max\{m, n\}$. If $n > m$, then $R(z)$ has a pole of order $n - m$ at ∞ . If $m < n$, then $R(z)$ has a zero of order $m - n$ at ∞ .

If $R(z)$ is a rational polynomial of order n . Then order of $R(z) - a$ is n . Thus, $R(z) = a$ has n zeroes.

Definitions 9.2.10. A linear transformation is a rational function of order 1.

$$S(z) = \frac{\alpha z + \beta}{\gamma z + \delta}, \quad \alpha\delta - \beta\gamma \neq 0$$

parallel translations are linear transformations of the form $z + a$.

inversion is the linear transformation $1/z$.

Definitions 9.2.11. Fixed point of a linear transformation S is a complex number z in the extended plane for which $S(z) = z$.

Fixed point of parallel translation is ∞ and that of inversion are $1, -1$.

Partial Fraction Representation

$$R(z) = \frac{P(z)}{Q(z)} = G(z) + H(z)$$

where $H(z)$ is finite at ∞ and $G(z)$ is a polynomial without constant term. In other words, $H(z) = \frac{P_1(z)}{Q(z)}$ where $P_1(z)$ and $Q(z)$ are of the same degree. Then, there exists a unique $G(z)$ such that $P(z) = G(z)Q(z) + P_1(z)$. Clearly, $G(z)$ is **singular part** of $R(z)$ at ∞ .

Suppose $R(z)$ has q distinct, finite poles, say $\beta_1, \beta_2, \dots, \beta_q$. Then,

$$R(z) = G(z) + \sum_{j=1}^q G_j \left(\frac{1}{z - \beta_j} \right) \quad (9.22)$$

where G_j is the singular part of $R(z)$ at β_j .

9.2.2 Elementary Theory of Power Series

Sequences

Definitions 9.2.12. A sequence of complex numbers $\{a_n\}$ converges to A if

$$\forall \varepsilon > 0, \exists N \in \mathbb{N} \text{ such that } \forall n > N, |a_n - A| < \varepsilon \quad (9.23)$$

Remark. Sequence diverges to infinity if $\lim_{n \rightarrow \infty} a_n = \infty$.

$$\forall \varepsilon > 0, \exists N \in \mathbb{N} \text{ such that } \forall n > N, |a_n| > \frac{1}{\varepsilon} \quad (9.24)$$

Definitions 9.2.13 (Cauchy). A sequence $\{a_n\}$ is Cauchy/Fundamental if

$$\forall \varepsilon > 0, \exists N \in \mathbb{N} \text{ such that } \forall m, n > N, |a_m - a_n| < \varepsilon \quad (9.25)$$

Theorem 9.2.4. A sequence $\{a_n\}$ is convergent if and only if Cauchy.

Proof. Suppose $\{a_n\}$ converges to A . Then

$$\forall \varepsilon > 0, \exists N \in \mathbb{N}, \forall m, n > N, |a_n - A| < \varepsilon, |a_m - A| < \varepsilon$$

Consider,

$$|a_m - a_n| = |a_m - A + A - a_n| \leq |a_m - A| + |a_n - A| < 2\varepsilon$$

Therefore, the sequence $\{a_n\}$ is Cauchy.

Suppose $\{x_n\}$ is Cauchy. Consider the sequence of real parts,

$$\forall \varepsilon > 0, \exists N \in \mathbb{N}, \forall m, n > N, |\alpha_m - \alpha_n| < \varepsilon$$

where $x_n = \alpha_n + i\beta_n$. Define $a_n = \max\{\alpha_1, \alpha_2, \dots, \alpha_n\}$. Then the sequence $\{a_n\}$ is a bounded, monotone sequence. Thus, the $\{a_n\}$ has a limit, say A . Therefore, $A = \overline{\lim} \alpha_n$. Similarly, let $a_n = \min\{\alpha_1, \alpha_2, \dots, \alpha_n\}$. Then, the $\{a_n\}$ has a limit, $a = \underline{\lim} \alpha_n$.

Suppose $a \neq A$. Let $\varepsilon = \frac{A-a}{3}$. Then there exists $\alpha_m < A - \varepsilon$ and exists $\alpha_n > a + \varepsilon$. Therefore,

$$A - a = A - \alpha_m + \alpha_m - \alpha_n + \alpha_n - a < 3\varepsilon$$

□

Limit Superior and Limit Inferior

Definitions 9.2.14. Let $\{\alpha_n\}_{n=1}^{\infty}$ be a real sequence. Define $a_n = \max\{\alpha_1, \alpha_2, \dots, \alpha_n\}$.
?

Properties of Limit Superior/Inferior

1. The hierarchy

$$\underline{\lim} \alpha_n + \underline{\lim} \beta_n \leq \underline{\lim}(\alpha_n + \beta_n) \leq \underline{\lim} \alpha_n + \overline{\lim} \beta_n \leq \overline{\lim}(\alpha_n + \beta_n) \leq \overline{\lim} \alpha_n + \overline{\lim} \beta_n$$

Series

Definitions 9.2.15. A sequence $\{b_n\}$ is a **contraction** of sequence $\{a_n\}$ if $|b_m - b_n| \leq |a_m - a_n|$, $\forall m, n \in \mathbb{N}$.

Definitions 9.2.16. An infinite series is a formal infinite sum

$$\sum_{n=1}^{\infty} a_n = a_1 + a_2 + \dots + a_n + \dots$$

Definitions 9.2.17. Series $\sum a_n$ converges if and only if the sequence of partial sums $\{s_n\}$ where $s_n = \sum_{j=1}^n a_j$ converges. The limit of the sequence of partial sum is the **sum** of the series.

Cauchy Condition The series converges if and only if for every $\varepsilon > 0$ there exists a natural number n_0 such that $|a_n + a_{n+1} + \dots + a_{n+p}| < \varepsilon$.

For $p = 0$, we have $|a_n| < \varepsilon$. In other words, $a_n \rightarrow 0$.

If a finite number of terms are omitted from a series, then the new series converges/diverges simultaneously with the first series. And the sum of the original(first) series $S = s_n + R_n$ where s_n is the sum of the omitted terms and R_n is the sum of the second series.

Absolute Convergence Let $\sum a_n$ be a series. The series converges absolutely if $\sum |a_n|$ converges. Absolute convergence is stronger than convergence in the case of series. In other words, absolute convergence implies convergence.

Uniform Convergence

Definitions 9.2.18. A sequence of functions $\{f_n(x)\}$ defined on E is said to converge uniformly to a function $f(x)$ if

$$\forall \varepsilon > 0, \exists n_0 \in \mathbb{N} \text{ such that } \forall x \in E, \forall n > n_0, |f_n(x) - f(x)| < \varepsilon$$

Convergence is pointwise if n_0 depends on $x \in E$.

Properties of Uniform Convergence

1. The limit function of a uniformly convergent sequence of continuous functions is continuous.

Cauchy Condition A sequence of functions $\{f_n(x)\}$ is uniformly convergence if and only if

$$\forall \varepsilon > 0, \exists n_0 \in \mathbb{N} \text{ such that } \forall x \in E, \forall m, n > n_0, |f_m(x) - f_n(x)| < \varepsilon$$

Weierstrass M Test If $\{f_n(x)\}$ is a contraction of a convergent sequence $\{a_n\}$, then $\{f_n(x)\}$ is uniformly convergent.

$$|f_n(x) - f_m(x)| \leq |a_n - a_m|, \forall n, m \in \mathbb{N} \implies \sum f_n(x) = f(x)$$

Suppose $\sum a_n$ is convergent and $\sum f_n(x)$ is a minorant of $\sum Ma_n$ for some constant M .

$$\exists n_0 \in \mathbb{N} \text{ such that } \forall n > n_0, |f_n(x)| \leq M|a_n|$$

Then $\sum f_n(x)$ is uniformly convergent.

Power Series

Definitions 9.2.19. A power series is a series of the form

$$\sum a_n(z - z_0)^n = a_0 + a_1(z - z_0) + \cdots + a_n(z - z_0)^n + \cdots$$

where $a_n, z_0 \in \mathbb{C}$.

Theorem 9.2.5. For every power series about origin there exists a number R called **radius of convergence** such that $0 \leq R \leq \infty$ and

1. The series converges absolutely for every z with $|z| < R$. If $0 \leq \rho < R$, the convergence is uniform for $|z| \leq \rho$
2. If $|z| > R$, the terms of the series are unbounded, and the series is divergent.
3. In $|z| < R$, the sum of the series is an analytic function. The derivative can be obtained by termwise differentiation. And the derived series has the same radius of convergence.

Proof.

□

The circle $|z| = R$ is the **circle of convergence**.

Hadamard's Formula

$$\frac{1}{R} = \overline{\lim} \sqrt[n]{|a_n|} \quad (9.26)$$

Taylor-Maclaurin Development

$$f(z) = f(0) + f'(0)z + \frac{f''(0)}{2!}z^2 + \cdots + \frac{f^{(n)}(0)}{n!}z^n + \cdots \quad (9.27)$$

Abel's Limit Theorem

Theorem 9.2.6 (Abel). *If $\sum a_n$ converges, then $f(z) = \sum a_n z^n$ tends to $f(1)$ as z approaches 1 in such a way that $|1 - z|/(1 - |z|)$ remains bounded.*

Proof. □

In other words, the series converges at $z = 1$ if the approach is in Stolz angle.

9.2.3 The exponential and trigonometric functions**The Exponential**

Definitions 9.2.20. The exponential function is the function $f(z)$ of power series form $\sum a_n z^n$ such that $f'(z) = f(z)$.

$$e^z = \sum_{n=0}^{\infty} \frac{z^n}{n!} = 1 + z + \frac{z^2}{2!} + \cdots + \frac{z^n}{n!} + \cdots$$

9.2.4 The Trigonometric Functions

All the trigonometric functions are rational functions of e^{iz} .

$$\cos z = \frac{e^{iz} + e^{-iz}}{2} = 1 - \frac{z^2}{2!} + \frac{z^4}{4!} + \cdots$$

$$\sin z = \frac{e^{iz} - e^{-iz}}{2i} = z - \frac{z^3}{3!} + \frac{z^5}{5!} + \cdots$$

$$\tan z = -i \frac{e^{iz} - e^{-iz}}{e^{iz} + e^{-iz}}$$

The Periodicity

The function $f(z)$ has period c if $f(z + c) = f(z)$, $\forall z$. Then, period of e^z is $c = 2\pi i$. And period of e^{iz} is $c = 2\pi$. The smallest positive period of e^{iz} is 2π .

The Logarithm

The inverse function of the exponential function. There exists only one elementary transcendental function, say e^z .

9.3 Analytic Functions as Mappings

9.3.1 Elementary point set topology

Sets and Elements

Point, Set, Intersection, Union, Complement, De Morgan laws

Metric Spaces

Metric, Open ball, Open set, Neighbourhood characterisation of open set, Closed set, Interior, Exterior, Boundary, Accumulation Point

Connectedness

Connected, Path connectedness, Connected components

Definitions 9.3.1. A region is a nonempty, connected, open set.

In a locally connected, separable space every open set is a countable union of disjoint regions.

Compactness

complete, compact, bounded, Sequential compactness

Continuous Functions

compactness, connectedness, extrema, uniform continuity

Topological Spaces

Hausdorff

9.3.2 Conformality

Arcs and Closed Curves

Definitions 9.3.2 (Arc). A continuous function γ defined on a closed, finite interval $[\alpha, \beta]$.

$$z = \gamma(t), \text{ where } \gamma : [\alpha, \beta] \rightarrow \mathbb{C}, \alpha \leq t \leq \beta, \text{ and } \gamma \text{ continuous} \quad (9.28)$$

Remark. Arc is a point set, with a succession of points.

Definitions 9.3.3 (change of parameter). Let $\gamma : [\alpha, \beta] \rightarrow \mathbb{C}$ be an arc. Let $\varphi : [\alpha', \beta'] \rightarrow [\alpha, \beta]$ be a non-decreasing onto function. Then $\gamma \circ \varphi$ defines the same succession of points as the arc γ .

differentiable $z'(t)$ exist and is continuous at all points.

regular differentiable and $z'(t) \neq 0$ at points on γ .

piecewise differentiable differentiable except for finitely many points

piecewise regular regular except for finitely many points.

opposite arc set of points $z = z(-t)$, $-b \leq t \leq -a$.

Analytic Functions in Regions

Conformal Mapping

Length and Area

9.3.3 Linear Transformations

The Linear Group

The Cross Ratio

Symmetry

Oriented Circles

Family of Circles

Module 2

9.4 Complex Integration

9.4.1 Fundamental Theorems

Line Integrals

Rectifiable Arcs

Line Integrals as Functions of Arcs

Cauchy's Theorem for a Rectangle

Cauchy's Theorem in a Disk

9.4.2 Cauchy's Integral Formula

The Index of a Point wrt a Closed Curve

The Integral Formula

Module 3

Higher Derivatives

Lemma 9.4.1. *Suppose $\phi(\zeta)$ is continuous on the closed curve arc γ . Then the function*

$$F_n(z) = \int_{\gamma} \frac{\phi(\zeta) d\zeta}{(\zeta - z)^n} \quad (9.29)$$

is analytic in each of the regions determined by the closed curve γ and its derivative is $F'_n(z) = nF_{n+1}(z)$.

Theorem 9.4.2 (Morera). *If $f(z)$ is defined and continuous in a region Ω , and if $\int_{\gamma} f(z) dz = 0$ for all closed curves γ in Ω , then $f(z)$ is analytic in Ω .*

Theorem 9.4.3 (Liouville). *A function which is analytic and bounded in the whole plane must reduce to a constant.*

Theorem 9.4.4 (fundamental theorem). *Every non-constant polynomial with complex coefficients has a zero.*

9.4.3 Local Properties of Analytical Functions

Removable Singularities, Taylor's Theorem

Zeros and Poles

The Local Mapping

The Maximum Principle

Module 4

9.4.4 The General Form of Cauchy's Theorem

Chains and Cycles

Definitions 9.4.1 (chain). A chain is a formal sum of arcs.

For example, $\gamma = \gamma_1 + \gamma_2 + \cdots + \gamma_n$ is a chain if $\gamma_1, \gamma_2, \dots, \gamma_n$ are arcs.

Properties of Chains

1. Different formal sums can represent the same chain.
2. Two chains are identical if they yield the same line integral for all functions f .

$$\int_{\gamma_1 + \gamma_2 + \cdots + \gamma_n} f dz = \int_{\gamma_1} f dz + \int_{\gamma_2} f dz + \cdots + \int_{\gamma_n} f dz$$

3. The following operations does not affect any chain
 - (a) Permutation of two arcs. $\gamma_1 + \gamma_2 + \gamma_3 = \gamma_1 + \gamma_3 + \gamma_2$
 - (b) Subdivision of an arc. $\gamma_1 + \gamma_2 = \gamma_1 + \gamma_{21} + \gamma_{22}$
 - (c) Fusion of subarcs into a single arc. $\gamma_1 + \gamma_{21} + \gamma_{22} = \gamma_1 + \gamma_2$

- (d) Reparametrisation of an arc. $\gamma(\tau(t)) = \gamma(t)$
 - (e) Cancellation of opposite arcs. $\gamma_1 + \gamma_2 - \gamma_2 = \gamma_1$
4. Sum of two chains defined to preserve the additive property of the line integrals.

$$\gamma = \alpha_1\gamma_1 + \alpha_2\gamma_2 + \cdots + \alpha_n\gamma_n$$

where α_j are integers and γ_j are arcs. That is, instead of $\gamma_1 + \gamma_1 = 2\gamma_1$. And $-\gamma_1 - \gamma_1 = -2\gamma_1 = 2(-\gamma_1)$. Also, $\gamma_1 - \gamma_1 = 0\gamma_1$. Terms with zero coefficients can be added/removed without altering the chain.

Definitions 9.4.2 (cycle). A cycle is a formal sum of closed curves.

Definitions 9.4.3 (index of a point). Index of a point with respect to a cycle $\gamma = \alpha_1\gamma_1 + \alpha_2\gamma_2 + \cdots + \alpha_n\gamma_n$ is given by

$$n(\alpha_1\gamma_1 + \alpha_2\gamma_2 + \cdots + \alpha_n\gamma_n, a) = \alpha_1n(\gamma_1, a) + \alpha_2n(\gamma_2, a) + \cdots + \alpha_nn(\gamma_n, a) \quad (9.30)$$

Simple Connectivity

Definitions 9.4.4 (simply connected). A region is simply connected if its complement with respect to the extended plane is connected.

Theorem 9.4.5. A region Ω is simply connected if and only if $n(\gamma, a) = 0$ for all cycle γ in Ω and all points a which does not belong to Ω .

Proof. □

Homology

Definitions 9.4.5. A cycle γ in an open set Ω is **homologous to zero with respect to Ω** if $n(\gamma, a) = 0$ for all points a in the complement of Ω .

Notation : $\gamma \sim 0 \pmod{\Omega}$

Remark.

$$\begin{aligned} \gamma_1 \sim \gamma_2 \pmod{\Omega} &\iff \gamma_1 - \gamma_2 \sim 0 \pmod{\Omega} \\ \Omega \subset \Omega', \gamma \sim 0 \pmod{\Omega} &\implies \gamma \sim 0 \pmod{\Omega'} \end{aligned}$$

The General Statement of Cauchy's Theorem

Theorem 9.4.6. If $f(z)$ is analytic in Ω , then

$$\int_{\gamma} f(z) dz = 0$$

for every cycle γ which is homologous to zero in Ω .

Corollary 9.4.6.1. If $f(z)$ is analytic in a simply connected region Ω , then $\int_{\gamma} f(z) dz = 0$ for all cycles γ in Ω .

Proof. Every cycle γ in a simply connected region Ω is homologous to zero. Thus by general Cauchy's theorem, $\int_{\gamma} f(z) dz = 0$. □

Corollary 9.4.6.2. If $f(z)$ is analytic and $f \neq 0$ in a simply connected region Ω then it is possible to define a single valued analytic branches of $\log f(z)$ and $\sqrt[n]{f(z)}$ in Ω .

Proof. □

Proof of Cauchy's Theorem**Locally Exact Differentials**

Definitions 9.4.6. A differential $pdx + qdy$ is exact if there exists an analytic function U such that $\frac{\partial U}{\partial x} = p$ and $\frac{\partial U}{\partial y} = q$.

Remark. If $Mdx + Ndy$ is exact differential then

$$\frac{\partial M}{\partial y} = \frac{\partial^2 U}{\partial y \partial x} = \frac{\partial^2 U}{\partial x \partial y} = \frac{\partial N}{\partial x}$$

Definitions 9.4.7. A differential $pdx + qdy$ is locally exact in Ω if it is exact in some neighbourhood of every point in Ω .

Theorem 9.4.7. If $pdx + qdy$ is locally exact in Ω , then

$$\int_{\gamma} pdx + qdy = 0$$

for every cycle $\gamma \sim 0(\text{mod } \Omega)$.

Multiply Connected Regions

Definitions 9.4.8. A multiply connected region is a region which is not simply connected.

9.4.5 The Calculus of Residues**The Residue Theorem**

Definitions 9.4.9. The residue of $f(z)$ at an isolated singularity of the unique complex number R which makes $f(z) - R/(z - a)$ the derivative of a single valued analytic function in the annulus $0 < |z - a| < \delta$.

Theorem 9.4.8. Let $f(z)$ be analytic except for a isolated singularities a_j in a region Ω . Then,

$$\frac{1}{2\pi i} \int_{\gamma} f(z) dz = \sum_j n(\gamma, a_j) \text{Res}_{z=a_j} f(z) \quad (9.31)$$

for any cycle $\gamma \sim 0(\text{mod } \Omega)$ and γ doesn't pass through any of the points a_j .

Proof.

□

Definitions 9.4.10. A cycle γ bounds the region Ω if and only if $n(\gamma, a)$ is defined and equal to 1 for all points $a \in \Omega$ and either undefined or equal to zero for all points $a \notin \Omega$.

In other words, γ is the boundary of the region Ω .

The Argument Principle

Theorem 9.4.9. *If $f(z)$ is meromorphic in Ω with the zeroes a_j and the poles b_j , then*

$$\frac{1}{2\pi i} \int_{\gamma} \frac{f'(z)}{f(z)} dz = \sum_j n(\gamma, a_j) - \sum_k n(\gamma, b_k) \quad (9.32)$$

Proof. □

Corollary 9.4.9.1 (Rouché). *Let $\gamma \sim 0 \pmod{\Omega}$ such that $n(\gamma, z)$ is either 0 or 1 for any point $z \notin \gamma$. Suppose $f(z), g(z)$ are analytic in Ω and satisfies the inequality $|f(z) - g(z)| < |f(z)|$ on γ . Then $f(z)$ and $g(z)$ have the same number of zeroes enclosed by γ .*

Proof. □

Exercise

- 1.
- 2.
- 3.

Evaluation of Definite Integrals

Type 1

$$\int_0^{2\pi} R(\cos \theta, \sin \theta) d\theta = -i \int_{|z|=1} R\left(\frac{1}{2}\left(z + \frac{1}{z}\right), \frac{1}{2i}\left(z - \frac{1}{z}\right)\right) \frac{dz}{z} \quad (9.33)$$

Type 2

$$\int_{-\infty}^{\infty} R(x) dx = 2\pi i \sum_{y>0} \text{Res } R(z) \quad (9.34)$$

Type 3

$$\int_{-\infty}^{\infty} R(x) e^{ix} dx = 2\pi i \sum_{y>0} \text{Res } R(z) e^{iz} \quad (9.35)$$

$$\int_{-\infty}^{\infty} R(x) \sin x dx = \text{Re} \left(2\pi i \sum_{y>0} \text{Res } R(z) e^{iz} \right) \quad (9.36)$$

$$\int_{-\infty}^{\infty} R(x) \cos x dx = \text{Im} \left(2\pi i \sum_{y>0} \text{Res } R(z) e^{iz} \right) \quad (9.37)$$

Type 4

$$\int_0^{\infty} x^{\alpha} R(x) dx = (1 - e^{2\pi i \alpha}) \int_0^{\infty} z^{2\alpha+1} R(z^2) dz \quad (9.38)$$

Type 5

$$\int_0^{\pi} \log \sin \theta d\theta = -\pi \log 2 \quad (9.39)$$

Exercise

1. Find the poles and residues of the following functions

(a)

$$\begin{aligned} \operatorname{Res}_{z=-2} \frac{1}{z^2 + 5z + 6} &= \lim_{z \rightarrow -2} \frac{z+2}{(z+2)(z+3)} = 1 \\ \operatorname{Res}_{z=-3} \frac{1}{z^2 + 5z + 6} &= \lim_{z \rightarrow -3} \frac{z+3}{(z+2)(z+3)} = -1 \end{aligned}$$

(b)

$$\begin{aligned} \operatorname{Res}_{z=-1} \frac{1}{(z^2 - 1)^2} &= \lim_{z \rightarrow -1} \frac{d}{dz} \frac{(z+1)^2}{(z+1)^2(z-1)^2} = \frac{1}{4} \\ \operatorname{Res}_{z=1} \frac{1}{(z^2 - 1)^2} &= \lim_{z \rightarrow 1} \frac{d}{dz} \frac{(z-1)^2}{(z+1)^2(z-1)^2} = -\frac{1}{4} \end{aligned}$$

(c)

$$\begin{aligned} \operatorname{Res}_{z=2n\pi} \frac{1}{\sin z} &= \lim_{z \rightarrow 2n\pi} \frac{z - 2n\pi}{\sin z} = 1 \\ \operatorname{Res}_{z=(2n+1)\pi} \frac{1}{\sin z} &= \lim_{z \rightarrow (2n+1)\pi} \frac{z - 2n\pi - \pi}{\sin z} = -1 \end{aligned}$$

(d)

$$\operatorname{Res}_{z=n\pi} \frac{\cos z}{\sin z} = \lim_{z \rightarrow n\pi} \frac{(z - n\pi) \cos z}{\sin z} = 1$$

(e)

$$\operatorname{Res}_{z=n\pi} \frac{1}{\sin^2 z} = \lim_{z \rightarrow n\pi} \frac{d}{dz} \frac{(z - n\pi)^2}{\sin^2 z} = 0$$

(f)

$$\begin{aligned} \operatorname{Res}_{z=0} \frac{1}{z^m(1-z)^n} &= \lim_{z \rightarrow 0} \frac{1}{m!} \frac{d^{m-1}}{dz^{m-1}} \frac{1}{(1-z)^n} = \\ \operatorname{Res}_{z=1} \frac{1}{z^m(1-z)^n} &= \lim_{z \rightarrow 1} \frac{1}{n!} \frac{d^{n-1}}{dz^{n-1}} \frac{1}{z^m} = \end{aligned}$$

2.

3. (a) hint :
- $\int_0^{2\pi} \frac{1}{a + \sin^2 x} dx = 4 \int_0^{\pi/2} \frac{1}{a + \sin^2 x} dx$
- .

$$R(\cos \theta, \sin \theta) = \frac{1}{a + \sin^2 x}$$

$$\int_0^{\pi/2} \frac{1}{a + \sin^2 x} dx = -i \int_{|z|=1} \frac{z}{z^4 + (a-2)z^2 + 1} dz$$

- (b) hint :
- $\int_{-\infty}^{\infty} R(x) dx = 2 \int_0^{\infty} R(x) dx$
- if
- $R(-x) = R(x)$

$$R(x) = \frac{x^2}{x^4 + 5x^2 + 6}$$

$$\int_0^{\infty} \frac{x^2}{x^4 + 5x^2 + 6} dx = \pi i \sum_{y>0} \operatorname{Res} R(z) = \frac{(\sqrt{3} - \sqrt{2})\pi}{2}$$

$$\operatorname{Res}_{z=\sqrt{2}i} R(z) = \lim_{z \rightarrow \sqrt{2}i} \frac{z^2(z - \sqrt{2}i)}{(z^2 + 2)(z^2 + 3)} = \frac{i}{\sqrt{2}}$$

$$\operatorname{Res}_{z=\sqrt{3}i} R(z) = \lim_{z \rightarrow \sqrt{3}i} \frac{z^2(z - \sqrt{3}i)}{(z^2 + 2)(z^2 + 3)} = -\frac{\sqrt{3}i}{2}$$

(c)

$$R(x) = \frac{x^2 - x + 2}{x^4 + 10x^2 + 9}$$

$$\int_{-\infty}^{\infty} \frac{x^2 - x + 2}{x^4 + 10x^2 + 9} = 2\pi i \sum_{y>0} \operatorname{Res} R(z) = \frac{5\pi}{12}$$

$$\operatorname{Res}_{z=i} R(z) = \lim_{z \rightarrow i} \frac{(z^2 - z + 2)(z - i)}{(z^2 + 1)(z^2 + 9)} = -\frac{1+i}{16}$$

$$\operatorname{Res}_{z=3i} R(z) = \lim_{z \rightarrow 3i} \frac{(z^2 - z + 2)(z - 3i)}{(z^2 + 1)(z^2 + 9)} = \frac{3-7i}{48}$$

(d)

$$R(x) = \frac{x^2}{(x^2 + a^2)^3}$$

(e)

$$R(x) \cos x = \frac{1}{x^2 + a^2} \cos x$$

(f)

$$R(x) \sin x = \frac{x}{x^2 + a^2} \sin x$$

(g)

$$x^\alpha R(x) = x^{\frac{1}{3}} \frac{1}{1+x^2}$$

(h)

(i)

Subject 10

ME010205 Measure & Integration

10.2 Lebesgue Measure

10.2.1 Introduction

set function A function which maps sets into (extended) real numbers.

σ -algebra A family \mathcal{A} of subsets of a nonempty set X such that

1. \mathcal{A} contains the empty set,
2. \mathcal{A} contains complement of each of its members and
3. \mathcal{A} is closed under countable unions.

From these 3 axioms, we can deduce the following,

4. \mathcal{A} is closed under countable intersections (by de Morgan's laws).

$$\left(\bigcup_{k=1}^{\infty} E_k^c \right)^c = \bigcap_{k=1}^{\infty} E_k \in \mathcal{A}$$

5. $E, F \in \mathcal{A} \implies E - F \in \mathcal{A}$ since $E - F = E \cap F^c$

Definitions 10.2.1 (Length of an interval). Length is a real valued set function. Let I be a bounded interval say $[a, b)$. Then its length $l(I) = b - a$ is the difference between endpoints. If an interval I is unbounded say (a, ∞) , then its length, $l(I) = \infty$.

Exercise

Techniques in Measure Theory Let \mathcal{A} be a σ -algebra. Let Lebesgue Measure $m : \mathcal{A} \rightarrow [0, \infty]$ be countably additive over disjoint collection of sets in \mathcal{A} .

- Lebesgue Measure m has monotonicity.
 $A \subseteq B \implies B = A \cup (B - A)$ is a disjoint union
 $\implies m(B) = m(A) + m(B - A) \geq m(A)$

- If exists $E \in \mathcal{A}$ such that $m(E) < \infty$, then $m(\phi) = 0$
Suppose $m(\phi) = c$ and $m(E) = k$ where $k < \infty$. If $c \neq 0$, then $m(E \cup \phi) = m(E) + m(\phi) = c + k > k = m(E)$ is a contradiction.
- $m\left(\bigcup_{k=1}^{\infty} E_k\right) \leq \sum_{k=1}^{\infty} m(E_k)$
Define $\{F_k : k \in \mathbb{N}\}$ by $F_k = E_k - \bigcup_{j=1}^{k-1} E_j$
Then $F_1 = E_1$, $F_2 = E_2 - E_1$, $F_3 = E_3 - (E_1 \cup E_2), \dots$
Also $F_k \in \mathcal{A}$ and $F_k \subseteq E_k$, $\forall k \in \mathbb{N}$. Thus $m(F_k) \leq m(E_k)$, $\forall k$ However,
 $\bigcup_{k=1}^{\infty} E_k = \bigcup_{k=1}^{\infty} F_k$
 $\implies m\left(\bigcup_{k=1}^{\infty} E_k\right) = m\left(\bigcup_{k=1}^{\infty} F_k\right) = \sum_{k=1}^{\infty} m(F_k) \leq \sum_{k=1}^{\infty} m(E_k)$

Counting Measure The counting measure $c : \mathcal{A} \rightarrow [0, \infty]$ is a set function which maps sets to their cardinality. For example, if $E = \{2, 3, 4\}$, then $c(E) = 3$.

- The counting measure is **translation invariant** since translation never increases the cardinality of the set.
For example, $5 + E = \{7, 8, 9\}$. And $m(5 + E) = 3 = m(E)$.
- The counting measure is **countably additive** over disjoint collections since the cardinality of disjoint union of two sets is the sum of their cardinalities.
- However, counting measure of (non-degenerate) intervals are ∞ which is **not the same as their length** for bounded intervals.

10.2.2 Lebesgue Outer Measure

G_δ A set which is countable intersection of open subsets.

F_σ A set which is countable union of closed subsets.

Caratheodory Construction of Lebesgue Measure

1. Construct Lebesgue Outer Measure m^* (with Axiom 3 relaxed)
ie, Obtain the underlying relation of the set function
2. Restrict m^* to the σ -algebra of our interest
ie, Choose a domain so that set function is well defined.

Definitions 10.2.2 (Lebesgue Outer Measure). Let $A \subset \mathbb{R}$. Let $\mathcal{C} = \{I_k : k \in \mathbb{N}\}$ be an open cover of A such that I_k are non-empty, bounded, open intervals. Consider the sum of length of intervals for such covers of A . (Lebesgue) Outer Measure $m^*(A)$ is the infimum of all such sums.

$$m^*(A) = \inf \left\{ \sum_{k=1}^{\infty} l(I_k) : A \subset \bigcup_{k=1}^{\infty} I_k \right\} \quad (10.1)$$

Properties of Lebesgue Outer Measure

1. Outer Measure of the empty set is zero

Let $\varepsilon > 0$. Then $\mathcal{C}_\varepsilon = \{(0, \frac{\varepsilon}{2^n}) : n \in \mathbb{N}\}$ is an open cover of ϕ containing nonempty, bounded, open intervals. Clearly, sum of length of intervals in $\mathcal{C}_\varepsilon = \varepsilon$. Suppose $m^*(\phi) = \delta$ and $\delta > 0$. There exists ε such that $0 < \varepsilon < \delta$. The sum of intervals of \mathcal{C}_ε is less than δ , which is a contradiction by the definition of Outer Measure.

2. Outer Measure is monotone

Suppose $A \subset B$. Then every cover of B is also an cover of A . Let \mathcal{U} be the set of all open covers of A with nonempty, bounded intervals and \mathcal{V} be the set of all such open covers of A . Clearly, $\mathcal{V} \subset \mathcal{U}$. We know that, if $A \subset B$, then $\inf B \leq \inf A$. Therefore,

$$A \subset B \implies m^*(A) \leq m^*(B) \quad (10.2)$$

3. Outer Measure of Countable Sets is zero

Let C be a countable set. That is, $C = \{c_k\}_{k=1}^\infty$.

Then $\{(c_k - \frac{\varepsilon}{2^k}, c_k + \frac{\varepsilon}{2^k})\}_{k=1}^\infty$ is cover of C with sum of length of intervals ε . Thus, for any $\varepsilon > 0$, we have $m^*(C) \leq \varepsilon$. Thus, $m^*(C) = 0$.

4. Outer Measure of an Interval is its length

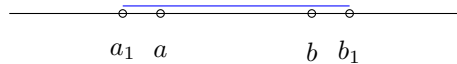
Proof. Case 1 : Closed, Bounded Interval Let $[a, b]$ be a closed, bounded interval. Then for any $\varepsilon > 0$, $(a - \varepsilon, b + \varepsilon)$ is a cover of $[a, b]$. Thus, by the definition of Lebesgue outer measure $m^*([a, b]) \leq b - a + 2\varepsilon$ since $[a, b] \subset (a - \varepsilon, b + \varepsilon)$ and m^* is monotonic. Therefore,

$$m^*([a, b]) \leq b - a \quad (10.3)$$

Since $[a, b]$ is closed and bounded, $[a, b]$ is compact. And by Heine-Borel theorem, every open cover of $[a, b]$ has a finite subcover. Thus, it is sufficient to prove the theorem for finite covers of $[a, b]$.

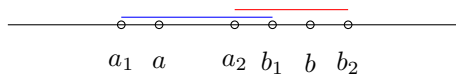
Let \mathcal{C} be a finite cover of $[a, b]$ with n open intervals. Let (a_1, b_1) be an open interval containing a in \mathcal{C} . Then $a_1 < a < b_1$. If $b_1 > b$ then

$$l(a_1, b_1) > l(a, b). \text{ And } \sum_{i=1}^k l(I_k) \geq l(a_1, b_1) \geq b - a.$$

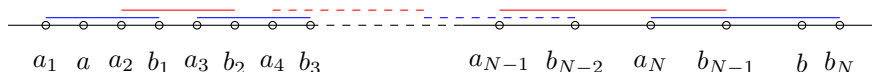


Suppose $b_1 < b$. Clearly, $a < b_1$. And the cover \mathcal{C} must have an open interval containing b_1 . Otherwise \mathcal{C} is not a cover of $[a, b]$. That is, there exists (a_2, b_2) containing $b_1 \in (a, b)$ such that $a_2 < b_1 < b_2$. If $b_2 > b$, then

$$\sum_{i=1}^k l(I_k) \geq l(a_1, b_1) + l(a_2, b_2) \geq l(a_1, b_2) \geq b - a.$$



Suppose $b_2 < b$. Continuing like this we get, N open intervals in \mathcal{C} , $\{(a_k, b_k) : k = 1, 2, \dots, N\}$ such that $a_1 < a < b_1$ and $a_N < b < b_N$ and $a_k < b_{k-1} < b_k$ for all k . The process should terminate in finite steps as \mathcal{C} is a finite cover of $[a, b]$. Then $\sum_{k=1}^N l(I_k) \geq \sum_{k=1}^N l(a_k, b_k) \geq l(a_1, b_N) \geq b - a$.



Clearly, every open cover of $[a, b]$ contains a finite subcover \mathcal{C} , which contains a finite subcover of the form $\{(a_k, b_k) : k = 1, 2, \dots, N\}$ such that $\sum_{k=1}^N l(I_k) \geq b - a$. Thus, for any open cover $\sum_{k=1}^{\infty} l(I_k) \geq b - a$. And thus,

$$m^*([a, b]) \geq b - a \quad (10.4)$$

Case 2 : Bounded Interval Let I be a bounded interval. Then there exists bounded closed intervals J_1 and J_2 such that $J_1 \subsetneq I \subsetneq J_2$ such that $l(I) - \varepsilon < l(J_1)$ and $l(J_2) < l(I) + \varepsilon$. Suppose $I = (a, b]$, then $J_1 = [a + \frac{\varepsilon}{2}, b - \frac{\varepsilon}{2}]$ and $J_2 = [a - \frac{\varepsilon}{2}, b + \frac{\varepsilon}{2}]$.

By monotonicity of Lebesgue outer measure, we have $m^*(J_1) \leq m^*(I) \leq m^*(J_2)$. However $m^*(J_1) = l(I) - \varepsilon$ and $m^*(J_2) = l(I) + \varepsilon$. Thus, $l(I) - \varepsilon \leq m^*(I) \leq l(I) + \varepsilon$. Therefore, $m^*(I) = l(I)$.

Case 3 : Unbounded Interval Let I be an unbounded interval. Then for any natural number n , there exists a closed bounded interval J such that $J \subset I$ and $l(J) = n$. And $n = m^*(J) \leq m^*(I)$, $\forall n \in \mathbb{N}$. Therefore, $m^*(I) = \infty = l(I)$. \square

5. Outer Measure is translation invariant

Proof. Let A be any set and $y \in \mathbb{R}$. Let $\{I_k : k = 1, 2, \dots\}$ be a cover of A . Then $\{I_k + y : k = 1, 2, \dots\}$ is a cover of $A + y$. And $l(I_k) = l(I_k + y)$ for every natural number k and real number y . Thus, $\sum_{k=1}^{\infty} l(I_k) = \sum_{k=1}^{\infty} l(I_k + y)$.

Clearly, for each cover $\{I_k\}_{k=1}^{\infty}$ of A , there exists a cover $\{I_k + y\}_{k=1}^{\infty}$ of $A + y$ containing intervals of same length. Therefore, $m^*(A) = m^*(A + y)$. \square

6. Outer Measure is countably subadditive

$$m^*\left(\bigcup_{k=1}^{\infty} E_k\right) \leq \sum_{k=1}^{\infty} m^*(E_k) \quad (10.5)$$

Proof. Let $\{E_k\}_{k=1}^\infty$ be a countable collection of sets. Without loss of generality $m^*(E_k) < \infty$ for any k . Suppose $m^*(E_k) = \infty$ for some k . Then the result is trivial. Since the outer measure is finite, for each natural number k , we have a cover of E_k , say $\{I_{k,i}\}_{i=1}^\infty$ such that

$$\sum_{i=1}^{\infty} l(I_{k,i}) < m^*(E_k) + \frac{\varepsilon}{2^k}$$

Suppose that, for some $\varepsilon > 0$, E_k doesn't have such a cover, then $m^*(E_k) + \frac{\varepsilon}{2^k}$ is an upper bound contradicting the assumption that $m^*(E_k)$ is the least upper bound.

The union of all such open intervals $I_{k,i}$ is a countable cover of $\bigcup E_k$. Thus, ‘

$$\begin{aligned} m^*\left(\bigcup_{k=0}^{\infty} E_k\right) &\leq \sum_{k,i=1}^{\infty} l(I_{k,i}) \\ &\leq \sum_{k=1}^{\infty} \sum_{i=1}^{\infty} l(I_{k,i}) \\ &= \sum_{k=1}^{\infty} \left(m^*(E_k) + \frac{\varepsilon}{2^k}\right) \\ &= \sum_{k=1}^{\infty} m^*(E_k) + \varepsilon \end{aligned}$$

□

Note : Finite subadditivity is a weaker notion than countable subadditivity. Since every finite collection is a countable collection.

Exercise

5. Closed Interval $[0, 1]$ is uncountable.

Suppose $[0, 1]$ is countable, then Lebesgue outer measure of any countable set is zero, $m^*([0, 1]) = 0$. But, $[0, 1]$ is an interval and Lebesgue outer measure of an interval is its length, $m^*([0, 1]) = l([0, 1]) = 1$ which is a contradiction.

6. $m^*([0, 1] - \mathbb{Q}) = 1$

$$[0, 1] = ([0, 1] \cap \mathbb{Q}) \cup ([0, 1] \cap \mathbb{Q}^c)$$

Clearly, $m^*([0, 1]) = 1$. And $[0, 1] \cap \mathbb{Q}$ is a countable set since \mathbb{Q} is countable. And thus has Lebesgue outer measure zero. Thus by countable subadditivity, we have

$$1 = m^*([0, 1]) \leq m^*([0, 1] \cap \mathbb{Q}^c) + 0$$

Thus, $m^*([0, 1] \cap \mathbb{Q}^c) \geq 1$. And $[0, 1] \cap \mathbb{Q}^c \subset [0, 1]$. By monotonicity, $m^*([0, 1] \cap \mathbb{Q}^c) \leq m^*([0, 1]) = 1$. Therefore, $m^*([0, 1] \cap \mathbb{Q}^c) = 1$.

7. Construction of a G_δ set containing E

Proof. Let E be a bounded subset of \mathbb{R} . Then there exists a bounded, open interval A containing E since E is bounded. And,

$$m^*(E) \leq m^*(A) = l(A) < \infty$$

Since, E is of finite outer measure for any $n \in \mathbb{N}$ there exists an open cover $I_{n,k}$ such that

$$m^*(E) \leq \sum_{k=1}^{\infty} l(I_{n,k}) + \frac{1}{n}$$

Define

$$G = \bigcap_{n \in \mathbb{N}} \bigcup_{k=1}^{\infty} I_{n,k}$$

Clearly, $E \subset G$ and G is a G_δ set. And $m^*(G) \leq m^*(E) + \frac{1}{n}$ for any $n \in \mathbb{N}$. Therefore, $m^*(G) = m^*(E)$. \square

8. Let $B = [0, 1] \cap \mathbb{Q}$. Let $\{I_k\}_{k=1}^n$ be any finite cover of B . Let (a_1, b_1) be the open interval containing 0. If $b_1 < 1$, let (a_2, b_2) be the open interval containing b_1 . Since rational numbers are dense, even if b_1 is irrational there must exist an open interval containing b_1 as the number of open intervals covering B is finite. Continuing like this we get open intervals $(a_1, b_1), (a_2, b_2), \dots, (a_n, b_n)$ such that $1 \in (a_n, b_n)$ and

$$\sum_{k=1}^n l(a_k, b_k) \leq l(a_1, b_n) = b_n - a_1 > 1$$

Thus, for any finite cover of B , the length of cover is greater than 1.

9. hint : $A \cup B = A \cup (B - A) = A \cup (B \cap A^c)$
 10. hint : A and B are separated by distance α , thus are disjoint.

10.2.3 σ -algebra of Lebesgue Measurable Sets

Lebesgue Outer Measure is defined for any subset of real numbers and Lebesgue outer measure of an interval is its length. However, it isn't countable additive.

There exists disjoint sets A, B such that $m^*(A \cup B) < m^*(A) + m^*(B)$.

Since countable additivity is a favourable property over countable subadditivity. We restrict the family of subsets of real numbers to those subsets that allow countable additivity.

Lebesgue Measurable Set

Definitions 10.2.3 (Measurable Set). Let E be a subset of \mathbb{R} . Then E is Lebesgue measurable if

$$m^*(A) = m^*(A \cap E) + m^*(A \cap E^c) \quad (10.6)$$

for any subset A of \mathbb{R} .

In other words, E is Lebesgue measurable if E doesn't affect countable additivity of Lebesgue Outer Measure.

We will consider only those subset of real numbers, which won't affect countable additivity. These subsets are **Lebesgue Measurable**. And we could show that the collection of all Lebesgue Measurable sets forms a σ -algebra. Clearly, intervals allow countable additivity, thus the Borel Algebra is contained in this σ -algebra of Lebesgue measurable sets.

Simplified Condition for Lebesgue Measurability

We know that Lebesgue Outer Measure has countable subadditivity.

$$m^*(A) \leq m^*(A \cap E) + m^*(A \cap E^c)$$

Thus, for condition (10.6), it is sufficient to check the following condition,

$$m^*(A) \geq m^*(A \cap E) + m^*(A \cap E^c) \quad (10.7)$$

Properties of Lebesgue Measure

1. Any set of Lebesgue outer measure zero is Lebesgue measurable.

Proof. Let E be a subset of real numbers with Lebesgue outer measure zero. Let A be any subset of real numbers. Then $A = (A \cap E) \cup (A \cap E^c)$. By countable additivity, $m^*(A) \leq m^*(A \cap E) + m^*(A \cap E^c)$. Since $A \cap E \subset E$, we have by monotonicity $m^*(A \cap E) \leq m^*(E) = 0$.

Again, $A \cap E^c \subset A$ and by monotonicity, $m^*(A) \geq m^*(A \cap E^c) = 0 + m^*(A \cap E^c) = m^*(A \cap E) + m^*(A \cap E^c)$. Thus, E is Lebesgue measurable by the simplified condition for Lebesgue measurability. \square

2. Countable sets are Lebesgue measurable.

Proof. Countable sets are of Lebesgue outer measure zero. And sets of Lebesgue outer measure zero are Lebesgue measurable. Thus, they are Lebesgue measurable. \square

3. Finite union of Lebesgue measurable sets is Lebesgue measurable.

Proof. It is enough to prove that if E_1 and E_2 are Lebesgue measurable, then their union is also Lebesgue measurable. Then, by finite mathematical induction, we can prove that the result is true for any finite collection of Lebesgue measurable sets.

Suppose E_1, E_2 are Lebesgue measurable sets. Since E_1 is Lebesgue measurable,

$$m^*(A) = m^*(A \cap E_1) + m^*(A \cap E_1^c) \quad (10.8)$$

And consider $A \cap E_1^c$ instead of A . Since E_2 is Lebesgue measurable, we get

$$m^*(A \cap E_1^c) = m^*(A \cap E_1^c \cap E_2) + m^*(A \cap E_1^c \cap E_2^c) \quad (10.9)$$

We have $(A \cap E_1^c) \cap E_2^c = A \cap (E_1^c \cap E_2^c) = A \cap (E_1 \cup E_2)^c$. And $(A \cap E_1) \cup (A \cap E_1^c \cap E_2) = (A \cap E_1) \cup [A \cap (E_2 \cap E_1^c)] = A \cap (E_1 \cup E_2)$.



$$\begin{aligned} m^*(A) &= m^*(A \cap E_1) + m^*(A \cap E_1^c) \\ &= m^*(A \cap E_1) + m^*(A \cap E_1^c \cap E_2) + m^*(A \cap E_1^c \cap E_2^c) \\ &\geq m^*[A \cap (E_1 \cup E_2)] + m^*[A \cap (E_1 \cup E_2)^c] \end{aligned}$$

Therefore $E_1 \cup E_2$ is Lebesgue measurable. And by finite induction, finite union of Lebesgue measurable sets is also Lebesgue measurable. \square

4. Lebesgue Measure is finitely additive.

In other words, Suppose $\{E_k\}_{k=1}^n$ be a finite collection of disjoint, Lebesgue measurable sets. Then Lebesgue measure of their union is the sum of Lebesgue measures.

Proof. Let A be any subset of \mathbb{R} and $\{E_k\}_{k=1}^n$ be a finite collection of disjoint, Lebesgue measurable subsets of \mathbb{R} .

$$\text{Claim : } m^*\left(A \cap \left[\bigcup_{k=1}^{\infty} E_k\right]\right) = \sum_{k=1}^{\infty} m^*(A \cap E_k) \quad (10.10)$$

Trivially, the claim is true for $n = 1$. Suppose the claim is true for $n - 1$. That is,

$$m^*\left(A \cap \left[\bigcup_{k=1}^{n-1} E_k\right]\right) = \sum_{k=1}^{n-1} m^*(A \cap E_k) \quad (10.11)$$

From set theory we have,

$$A \cap \left[\bigcup_{k=1}^n E_k\right] \cap E_n = A \cap E_n \quad (10.12)$$

$$A \cap \left[\bigcup_{k=1}^n E_k\right] \cap E_n^c = A \cap \left[\bigcup_{k=1}^{n-1} E_k\right] \quad (10.13)$$

By Lebesgue measurability of E_n , we have

$$\begin{aligned} m^* \left(A \cap \left[\bigcup_{k=1}^n E_k \right] \right) &= m^* \left(A \cap \left[\bigcup_{k=1}^n E_k \right] \cap E_n \right) + m^* \left(A \cap \left[\bigcup_{k=1}^n E_k \right] \cap E_n^c \right) \\ &= m^* (A \cap E_n) + m^* \left(A \cap \left[\bigcup_{k=1}^{n-1} E_k \right] \right) \\ &= \sum_{k=1}^n m^* (A \cap E_k), \text{ by mathematical induction} \end{aligned}$$

Taking $A = \mathbb{R}$, we get Lebesgue measure is finitely additive. That is,

$$m^* \left(\bigcup_{k=1}^n E_k \right) = \sum_{k=1}^n m^* (E_k) \quad (10.14)$$

□

5. Countable union of Lebesgue measurable sets is Lebesgue measurable

Proof. Let A be any subset of \mathbb{R} . And $\{E_k\}_{k=1}^\infty$ be a countable collection of disjoint, Lebesgue measurable subsets of \mathbb{R} . Define $F_n = \bigcup_{k=1}^n E_k$ and $E = \bigcup_{k=1}^\infty E_k$. Clearly, $F \subset E$ and $F^c \supset E^c$. Thus, $m^*(A \cap F_n^c) \geq m^*(A \cap E^c)$.

$$\begin{aligned} m^*(A) &= m^*(A \cap F_n) + m^*(A \cap F_n^c) \\ &\geq m^* \left(A \cap \left[\bigcup_{k=1}^n E_k \right] \right) + m^*(A \cap E^c) \\ &\geq \sum_{k=1}^n m^*(A \cap E_k) + m^*(A \cap E^c) \\ \lim_{n \rightarrow \infty} m^*(A) &\geq \lim_{n \rightarrow \infty} \sum_{k=1}^n m^*(A \cap E_k) + m^*(A \cap E^c) \\ m^*(A) &\geq \sum_{k=1}^\infty m^*(A \cap E_k) + m^*(A \cap E^c) \\ &\geq m^* \left(A \cap \left[\bigcup_{k=1}^\infty E_k \right] \right) + m^*(A \cap E^c) \\ \implies m^*(A) &\geq m^*(A \cap E) + m^*(A \cap E^c) \end{aligned}$$

By above inequality, $E = \bigcup_{k=1}^\infty E_k$ is Lebesgue measurable.

And, any countable collection of Lebesgue measurable sets can be expressed as a countable collection of disjoint, Lebesgue measurable sets. Let $\{E_k\}_{k=1}^\infty$ be a collection of Lebesgue measurable sets. Then, the countable collection, $\{E'_k\}_{k=1}^\infty$ defined by $E'_k = E_k - \bigcup_{j=1}^{k-1} E_k$ contains

disjoint, Lebesgue measurable^{†1} subsets of \mathbb{R} . Therefore, countable union of Lebesgue measurable sets is Lebesgue measurable. \square

6. Every interval is Lebesgue measurable.

Proof. It is sufficient to prove that (a, ∞) is Lebesgue measurable. Suppose (a, ∞) is Lebesgue measurable for every $a \in \mathbb{R}$. Then interval (a, b) is Lebesgue measurable, since $(a, b) = [(a, \infty) \cap (\mathbb{R} - (b, \infty))] - \{b\}$.

Let A be any subset of \mathbb{R} . Define $A_1 = A \cap (-\infty, a)$ and $A_2 = A \cap (a, \infty)$ such that $A - \{a\} = A_1 \cup A_2$ and $A_1 \cap A_2 = \emptyset$. And the interval (a, ∞) is Lebesgue measurable only if

$$m^*(A) \geq m^*(A \cap (a, \infty)) + m^*(A \cap (a, \infty)^c) = m^*(A_1) + m^*(A_2) \quad (10.15)$$

Let collection $\{I_k\}_{k=1}^\infty$ be a countable cover of A . Then collection $\{I'_k\}_{k=1}^\infty$ defined by $I'_k = I_k \cap (a, \infty)$ is a cover of A_1 . And collection $\{I''_k\}_{k=1}^\infty$ defined by $I''_k = I_k \cap (-\infty, a)$ is a cover of A_2 .

Clearly, $m^*(A_1) \leq \sum_{k=1}^\infty l(I'_k)$ and $m^*(A_2) \leq \sum_{k=1}^\infty l(I''_k)$.

$$\begin{aligned} m^*(A_1) + m^*(A_2) &\leq \sum_{k=1}^\infty l(I'_k) + \sum_{k=1}^\infty l(I''_k) \\ &\leq \sum_{k=1}^\infty l(I'_k) + l(I''_k) \\ &\leq \sum_{k=1}^\infty l(I_k) \leq m^*(A) \end{aligned}$$

Thus, (a, ∞) is Lebesgue measurable. Therefore, every interval is Lebesgue measurable. \square

7. σ -algebra of Lebesgue measurable sets \mathcal{M} contains Borel Sets \mathcal{B} .

$$\mathcal{B} \subset \mathcal{M} \quad (10.16)$$

Proof. The Borel algebra \mathcal{B} is the σ -algebra containing all intervals. We have proved that, every intervals are Lebesgue measurable. Also, we have proved that the set of all Lebesgue measurable subsets of \mathbb{R} is a σ -algebra as complements of Lebesgue measurable sets are Lebesgue measurable by duality of the condition and countable union of Lebesgue measurable sets are also Lebesgue measurable. Therefore, every Borel set is Lebesgue measurable. Clearly, G_δ and F_σ are Borel sets and are Lebesgue measurable. \square

^{†1}Suppose E_1, E_2 are measurable, then $E_2^c = \mathbb{R} - E_2$ is measurable by the duality of measurability condition. And $E_1 \cap E_2^c = E_1 - E_2$ is Lebesgue measurable since countable intersection of Lebesgue measurable sets is Lebesgue measurable (by Property 4 and de Morgan's Law).

²We have, $m^*(A \cap (-\infty, a]) = m^*(A \cap (-\infty, a))$ since removing finite number of points from a subset of \mathbb{R} won't affect its Lebesgue measure.

8. Lebesgue Measurability is translation invariant.

Proof. Let E be a Lebesgue measurable set. Let A be any subset of \mathbb{R} and $y \in \mathbb{R}$. Then,

$$\begin{aligned} m^*(A) &= m^*(A - y) \\ &= m^*((A - y) \cap E) + m^*((A - y) \cap E^c) \\ &= m^*(A \cap (E + y)) + m^*(A \cap (E + y)^c) \end{aligned}$$

Thus, $E + y$ is Lebesgue measurable. And Lebesgue measurability is translation invariant. \square

Exercise

11. Let \mathcal{A} be the σ -algebra containing all intervals of the form (a, ∞) . Every interval has one of the four forms,

$$(a, b] = (a, \infty) \cap (b, \infty)^c \quad (10.17)$$

$$(a, b) = (a, \infty) \cap \left[\bigcap_{k=1}^{\infty} \left(b - \frac{1}{k}, \infty \right) \right]^c \quad (10.18)$$

$$[a, b] = \left[\bigcap_{k=1}^{\infty} \left(a - \frac{1}{k}, \infty \right) \right] \cap (b, \infty)^c \quad (10.19)$$

$$[a, b) = \left[\bigcap_{k=1}^{\infty} \left(a - \frac{1}{k}, \infty \right) \right] \cap \left[\bigcap_{k=1}^{\infty} \left(b - \frac{1}{k}, \infty \right) \right]^c \quad (10.20)$$

12. **Borel sets** is the σ -algebra containing all open intervals. The equations from previous equations are sufficient. However, we have a simpler form for closed intervals, $[a, b]$.

$$[a, b] = [(-\infty, a) \cup (b, \infty)]^c \quad (10.21)$$

Clearly, every interval is a Borel set.

13.
 - $C \in F_\sigma \implies C = \bigcup_{k=1}^{\infty} C_k \implies \bigcup_{k=1}^{\infty} (C_k + y) = C + y \in F_\sigma$
 - $O \in G_\delta \implies O = \bigcap_{k=1}^{\infty} O_k \implies \bigcup_{k=1}^{\infty} (O_k + y) = O + y \in G_\delta$
 - $m^*(E) = 0 \implies 0 = \inf\{\sum_{k=1}^{\infty} l(I_k)\} = \inf\{\sum_{k=1}^{\infty} l(I_k + y)\} \implies m^*(E + y) = 0$
14. A subset E has positive Lebesgue measure if and only if it has a bounded subset of positive Lebesgue measure.

$$m^*(E) > 0 \iff \exists \text{ bounded subset } F \subset E, \text{ such that } m^*(F) > 0$$

Sufficient Part : By monotonicity, $F \subset E \implies m^*(F) \leq m^*(E)$. And $m^*(F) > 0 \implies 0 < m^*(F) \leq m^*(E)$.

Necessary Part :

15.

10.2.4 Outer and Inner Approximation

Definitions 10.2.4 (Excision Property). Let A be a Lebesgue measurable set of finite measure and $A \subset B$. Then,

$$m^*(B \sim A) = m^*(B) - m^*(A) \quad (10.22)$$

Proof.

$$\begin{aligned} m^*(B) &= m^*(B \cap A) + m^*(B \cap A^c) \\ &= m^*(A) + m^*(B \sim A) \\ \implies m^*(B \sim A) &= m^*(B) - m^*(A) \end{aligned}$$

□

Theorem 10.2.1 (approximation). *The following conditions are equivalent to Lebesgue measurability of E*

1. $\forall \varepsilon > 0$, there is an open set \mathcal{O} containing E for which $m^*(\mathcal{O} \sim E) < \varepsilon$.
2. There is a G_δ set G containing E such that $m^*(G \sim E) = 0$.
3. $\forall \varepsilon > 0$, there is a closed set F contained in E such that $m^*(E \sim F) < \varepsilon$.
4. There is an F_σ set F contained in E such that $m^*(E \sim F) = 0$.

Note : Equivalent conditions 1 & 2 are about outer approximation of a Lebesgue measurable set and 3 & 4 are about inner approximation of a Lebesgue measurable set.

Proof. Measurability \implies Open set - Outer approximation

Suppose that E is Lebesgue measurable. And let $\varepsilon > 0$.

Case 1 : Suppose $m^*(E) < \infty$. By definition of Lebesgue outer measure, there exists an open cover $\{I_k\}_{k=1}^\infty$ such that $\sum_{k=1}^\infty l(I_k) < m^*(E) + \varepsilon$.

Define $\mathcal{O} = \bigcup_{k=1}^\infty I_k$. Then \mathcal{O} is an open set containing E and thus Lebesgue measurable. Also, $m^*(\mathcal{O}) \leq \sum_{k=1}^\infty l(I_k) < m^*(E) + \varepsilon$. Therefore, by excision property we have $m^*(\mathcal{O} \sim E) = m^*(\mathcal{O}) - m^*(E) < \varepsilon$.

Case 2 : Suppose $m^*(E) = \infty$. Without loss of generality, E may be written as countable union Lebesgue measurable sets $\{E_k\}_{k=1}^\infty$ of finite measure.

By case 1, for every k , there exists \mathcal{O}_k for each E_k of finite measure such that $m^*(\mathcal{O}_k \sim E_k) < \frac{\varepsilon}{2^k}$. Define $\mathcal{O} = \bigcup_{k=1}^{\infty} \mathcal{O}_k$. Then \mathcal{O} is open, contains E and

$$\begin{aligned} m^*(\mathcal{O} \sim E) &= m^*\left(\bigcup_{k=1}^{\infty} \mathcal{O}_k \sim E\right) \\ &\leq m^*\left(\bigcup_{k=1}^{\infty} \mathcal{O}_k \sim E_k\right) \\ &\leq \sum_{k=1}^{\infty} m^*(\mathcal{O}_k \sim E_k) = \varepsilon \sum_{k=1}^{\infty} \frac{1}{2^k} = \varepsilon \end{aligned}$$

Open, Outer approximation $\implies G_\delta$, Outer approximation

Let E be a subset of real numbers such that Lebesgue measure of E has an open set inner approximation. That is, for every $\varepsilon > 0$, there exists an open set \mathcal{O} such that $m^*(\mathcal{O} \sim E) < \varepsilon$.

Let \mathcal{O}_k be open sets such that $m^*(\mathcal{O}_k \sim E) < \frac{1}{k}$. Define $G = \bigcap_{k=1}^{\infty} \mathcal{O}_k$. Then, G is a G_δ set containing E . And $G \sim E \subset \mathcal{O}_k \sim E$. Thus, by monotonicity, $m^*(G \sim E) \leq m^*(\mathcal{O}_k \sim E) < \frac{1}{k}$. Thus, we have a G_δ set G containing E such that $m^*(G \sim E) = 0$.

G_δ -Outer approximation \implies Measurability

We have, $m^*(G \sim E) = 0$. Since every set of Lebesgue measure zero is Lebesgue measurable, $G \sim E$ is Lebesgue measurable. And its complement $(G \sim E)^c$ is also Lebesgue measurable. Also we have, G is a G_δ set, thus a Borel set and hence Lebesgue measurable.

Clearly, we have $E = G \cap (G \sim E)^c$. And therefore, E is Lebesgue measurable.

Open, Outer approximation \iff Closed, Inner approximation

By duality of Lebesgue measurability, E is Lebesgue measurable if and only if E^c is Lebesgue measurable. And by de Morgan's Law, we have E^c has an open set - outer approximation \mathcal{O} if and only if E has a closed set - inner approximation \mathcal{O}^c .

$$m^*(\mathcal{O} \sim E^c) < \varepsilon \iff m^*(E \sim \mathcal{O}^c) < \varepsilon$$

G_δ -Outer approximation $\iff F_\sigma$ -Inner approximation

Again, by duality of Lebesgue measurability and de Morgan's Law, we have E^c has a G_δ -outer approximation G if and only if E has an F_σ -inner approximation F .

$$m^*(G \sim E^c) = 0 \iff m^*(E \sim F) = 0$$

□

Theorem 10.2.2. *Let E be a Lebesgue measurable set of finite measure. Then for any $\varepsilon > 0$, there exists a finite collection of disjoint open sets $\{I_k\}_{k=1}^n$ such that $\mathcal{O} = \bigcup_{k=1}^n I_k$ and $m^*(\mathcal{O} \sim E) + m^*(E \sim \mathcal{O}) < \varepsilon$.*

Proof. Since E is Lebesgue measurable, by outer approximation theorem we have an open set U such that $E \subset U$ and

$$m^*(U \sim E) < \frac{\varepsilon}{2}$$

$$U = (U \cap E) \cup (U \cap E^c)$$

Since E is Lebesgue measurable, $m^*(E) < \infty$

$$\begin{aligned} m^*(U) &= m^*(U \cap E) + m^*(U \cap E^c) \\ &= m^*(E) + m^*(U \sim E) \end{aligned}$$

Since E has finite measure

$$\begin{aligned} m^*(U) &= m^*(E) + m^*(U \sim E) \\ &< \infty + \frac{\varepsilon}{2} < \infty \end{aligned}$$

That is, U is of finite measure.

Since U is open, U is countable^{†3} union of a disjoint collection of open intervals, say $\{I_k\}_{k=1}^\infty$. Clearly,

$$\sum_{k=1}^n l(I_k) \leq \sum_{k=1}^\infty l(I_k) \leq m^*(U) < \infty$$

By characterisation of series convergence, there exists an integer n such that,

$$\sum_{k=n+1}^\infty l(I_k) < \frac{\varepsilon}{2}$$

Define $\mathcal{O} = \bigcup_{k=1}^n I_k$. Since $\mathcal{O} \sim E \subset U \sim E$, by monotonicity we have

$$m^*(\mathcal{O} \sim E) \leq m^*(U \sim E) < \frac{\varepsilon}{2} \quad (10.23)$$

Since $E \subset U$, we have $E \sim \mathcal{O} \subset U \sim \mathcal{O} = \bigcup_{k=1}^n I_k$. And clearly,

$$U \sim \mathcal{O} = \bigcup_{k=1}^\infty I_k \sim \bigcup_{k=1}^n I_k \subseteq \bigcup_{k=n+1}^\infty I_k$$

³By definition, **Open sets** are countable union of disjoint, open intervals

$$\text{Thus, } m^*(E \sim \mathcal{O}) \leq m^*(U \sim \mathcal{O}) \leq \sum_{k=n+1}^{\infty} l(I_k) < \frac{\varepsilon}{2} \quad (10.24)$$

Therefore,

$$m^*(E \sim \mathcal{O}) + m^*(\mathcal{O} \sim E) < \varepsilon$$

□

Exercise

17. Let $\varepsilon > 0$ and E is Lebesgue measurable. Then there exists open set \mathcal{O} and closed set F such that $F \subset E \subset \mathcal{O}$, $m^*(E \sim F) < \frac{\varepsilon}{2}$ and $m^*(\mathcal{O} \sim E) < \frac{\varepsilon}{2}$. Clearly, $\mathcal{O} \sim E$ and $E \sim F$ are disjoint and $\mathcal{O} \sim F = (\mathcal{O} \sim E) \cup (E \sim F)$. Thus by monotonicity of Lebesgue outer measure, we have $m^*(\mathcal{O} \sim F) \leq m^*(\mathcal{O} \sim E) + m^*(E \sim F) < \varepsilon$.

18. Suppose E has finite outer measure. ^{†4}

G_δ set : Let $\varepsilon > 0$. Then by the definition of Lebesgue outer measure, there exists a cover $\{I_k\}_{k=1}^{\infty}$ of E such that $\sum_{k=1}^{\infty} l(I_k) < m^*(E) - \frac{\varepsilon}{2}$. Define

$$G = \bigcup_{k=1}^{\infty} I_k. \text{ Then } G \text{ is a } G_\delta \text{ set and } m^*(G) \leq \sum_{k=1}^{\infty} l(I_k) < m^*(E) - \frac{\varepsilon}{2}.$$

F_σ set :

19. Let E be a set of finite outer measure. Suppose E is not Lebesgue measurable. And \mathcal{O} be an open set containing E . Then $\mathcal{O} = (\mathcal{O} \sim E) \cup E$. By monotonicity, $m^*(\mathcal{O}) \leq m^*(\mathcal{O} \sim E) + m^*(E)$. Since $m^*(E)$ is finite, we have $m^*(\mathcal{O}) - m^*(E) \leq m^*(\mathcal{O} \sim E)$.
20. Let E be a set of finite outer measure. Suppose E is Lebesgue measurable. Let (a, b) be any open, bounded interval. Then by the definition of Lebesgue measurability, $b - a = m^*(a, b) = m^*((a, b) \cap E) + m^*((a, b) \cap E^c)$.
21. A subset E is Lebesgue measurable if there exists a G_δ set G containing E such that $m^*(G \sim E) = 0$. Suppose E_1 and E_2 are Lebesgue measurable sets. Then, we have G_δ sets G_1 and G_2 . And two countable family of open intervals $\{\mathcal{O}_{1,k}\}_{k=1}^{\infty}$ and $\{\mathcal{O}_{2,k}\}_{k=1}^{\infty}$ such that $\cap \mathcal{O}_{1,k} = G_1$ and $\cap \mathcal{O}_{2,k} = G_2$. Let $\mathcal{O} = \{\mathcal{O}_k\}_{k=1}^{\infty}$ be collection of open intervals in \mathcal{O}_1 and \mathcal{O}_2 . Define $G = \cap_{k=1}^{\infty} \mathcal{O}_k$. Then, $E = E_1 \cup E_2 \subset G_1 \cup G_2 = G$. Since $G \sim E = (G_1 \sim E_1) \cup (G_2 \sim E_2)$, by monotonicity of Lebesgue outer measure we have $m^*(G \sim E) \leq m^*(G_1 \sim E_1) + m^*(G_2 \sim E_2) = 0$.
22. Let m^{**} be a non-negative set function defined by $m^{**}(A) = \inf\{m^*(\mathcal{O}) : A \subset \mathcal{O}, \mathcal{O} \text{ is open}\}$. Suppose E is Lebesgue measurable. Then, by open set - outer approximation theorem we have $m^*(E) \leq m^{**}(E) < m^*(E) + \varepsilon$ for any $\varepsilon > 0$. Thus, $m^{**}(E) = m^*(E)$.

In other words, for Lebesgue measurable sets m^* and m^{**} are the same.

⁴ E has finite outer measure does not imply E is bounded or Lebesgue measurable.

23. Let m^{***} be a non-negative set function defined by $m^{***}(E) = \sup\{m^*(F) : F \subset E, F \text{ is closed}\}$. Let E be Lebesgue measurable set. Then, by closed set - inner approximation theorem we have $m^*(E) \geq m^{***}(E) > m^*(E) - \varepsilon$.

In other words, for Lebesgue measurable sets m^* and m^{***} are the same.

10.2.5 Further Properties

The Lebesgue measure has the following properties.

1. Every Borel set is Lebesgue measurable.
2. Lebesgue Measure of an interval is its length.
3. Lebesgue Measure is translation invariant.
4. Lebesgue Measure is countably additive.
5. There exists non-measurable(Lebesgue) sets. eg. $C_E \subset E$
6. There exists uncountable set of zero measure. eg. Cantor set

Countable Subadditivity

Theorem 10.2.3. *The set function Lebesgue measure defined on σ algebra of Lebesgue measurable sets 1. assigns length to any interval, 2. is translation invariant, and 3. is countably additive.*

Proof. Length of interval

Let E be an interval, then E belongs to the σ algebra of Lebesgue measurable sets as Borel sets are Lebesgue measurable. Also, we have $m^*(E) = m(E)$ for any Lebesgue measurable set E . And Lebesgue outer measure m^* of an interval is its length. Therefore, Lebesgue measure of any interval is its length.

Translation invariant

Let E be a Lebesgue measurable set. We have, $E + y$ is also Lebesgue measurable. Since $E + y$ is Lebesgue measurable and Lebesgue outer measure is translation invariant, we have $m^*(E) = m^*(E + y) = m(E + y)$. Clearly, $m(E) = m(E + y)$.

Countably additive ⁵

Let $\{E_k\}_{k=1}^{\infty}$ be a countable family of disjoint, Lebesgue measurable sets. Lebesgue outer measure is countably subadditive and countable union of Lebesgue measurable sets is also Lebesgue measurable. Thus we have,

$$m\left(\bigcup_{k=1}^{\infty} E_k\right) \leq \sum_{k=1}^{\infty} m(E_k)$$

⁵A real-valued set function m is countably additive if $m(\bigcup_{k=1}^{\infty} E_k) = \sum_{k=1}^{\infty} m(E_k)$ for any disjoint family of sets $\{E_k\}$.

Since Lebesgue measure is finitely additive we have,

$$\begin{aligned} m\left(\bigcup_{k=1}^{\infty} E_k\right) &\geq m\left(\bigcup_{k=1}^n E_k\right) = \sum_{k=1}^n m(E_k) \\ \lim_{n \rightarrow \infty} m\left(\bigcup_{k=1}^{\infty} E_k\right) &\geq \lim_{n \rightarrow \infty} m\left(\bigcup_{k=1}^n E_k\right) = \lim_{n \rightarrow \infty} \sum_{k=1}^n m(E_k) \\ &\implies m\left(\bigcup_{k=1}^{\infty} E_k\right) \geq \sum_{k=1}^{\infty} m(E_k) \end{aligned}$$

Therefore, Lebesgue measure is countably additive.

$$m\left(\bigcup_{k=1}^{\infty} E_k\right) = \sum_{k=1}^{\infty} m(E_k)$$

□

Continuity of Lebesgue measure

Theorem 10.2.4 (continuity). *Let m be Lebesgue measure.*

1. Suppose $\{A_k\}_{k=1}^{\infty}$ be an ascending⁶ collection of Lebesgue measurable sets. Then,

$$m\left(\bigcup_{k=1}^{\infty} A_k\right) = \lim_{k \rightarrow \infty} m(A_k) \quad (10.25)$$

2. Suppose $\{B_k\}_{k=1}^{\infty}$ be an descending⁷ collection of Lebesgue measurable sets and $m(B_1) < \infty$. Then,

$$m\left(\bigcap_{k=1}^{\infty} B_k\right) = \lim_{k \rightarrow \infty} m(B_k) \quad (10.26)$$

Proof. Ascending Collection

Let $\{A_k\}_{k=1}^{\infty}$ be an ascending collection of Lebesgue measurable sets. Define $A_0 = \phi$.

Case 1 : $\exists k' \in \mathbb{N}$, $m(A_{k'}) = \infty$

Suppose the collection has a Lebesgue measurable set $A_{k'}$ of infinite measure. Then for $\forall k \geq k'$, $m(A_k) = \infty$. Clearly,

$$\lim_{k \rightarrow \infty} m(A_k) = \infty = m(A_k) \leq m\left(\bigcup_{k=1}^{\infty} A_k\right)$$

Case 2 : $\forall k \in \mathbb{N}$, $m(A_k) < \infty$

Suppose that every Lebesgue measurable set in the collection is of finite measure. Consider the ascending collection of disjoint Lebesgue measurable sets, $\{C_k\}_{k=1}^{\infty}$

⁶ $\{A_k\}$ is ascending if $A_1 \subset A_2 \subset \dots$

⁷ $\{A_k\}$ is descending if $B_1 \supset B_2 \supset \dots$

given by $C_k = A_k \sim A_{k-1}$. Clearly, $\cup_{k=1}^{\infty} A_k = \cup_{k=1}^{\infty} C_k$. By countable additivity of Lebesgue measure, we have

$$m\left(\bigcup_{k=1}^{\infty} A_k\right) = m\left(\bigcup_{k=1}^{\infty} C_k\right) = \sum_{k=1}^{\infty} m(A_k \sim A_{k-1})$$

Also we have,

$$\begin{aligned} \sum_{k=1}^{\infty} m(A_k \sim A_{k-1}) &= \sum_{k=1}^{\infty} m(A_k) - m(A_{k-1}) \\ &= \lim_{n \rightarrow \infty} \sum_{k=1}^n m(A_k) - m(A_{k-1}) \\ &= \lim_{n \rightarrow \infty} m(A_n) - m(A_0) \end{aligned}$$

Therefore, $m\left(\bigcup_{k=1}^{\infty} A_k\right) = \lim_{n \rightarrow \infty} m(A_n)$.

Descending Collection Let $\{B_k\}$ be a descending collection of Lebesgue measurable sets and $m(B_1) < \infty$. Consider the ascending collection of Lebesgue measurable sets, $\{D_k\}_{k=1}^{\infty}$ given by $D_k = B_1 \sim B_k$. By the continuity of Lebesgue measure for ascending collection of sets, we have

$$m\left(\bigcup_{k=1}^{\infty} D_k\right) = \lim_{n \rightarrow \infty} m(D_n) = m(B_1) - \lim_{n \rightarrow \infty} m(B_n)$$

By de Morgan's law, we have $B_1 - \cap_k B_k = \cup_k (B_1 - B_k)$. Since B_1 is of finite measure, by excision property we have,

$$m\left(\bigcap_{k=1}^{\infty} B_k\right) = m\left(B_1 - \bigcup_{k=1}^{\infty} D_k\right) = \lim_{n \rightarrow \infty} m(B_n)$$

□

Borel-Cantelli Lemma

Definitions 10.2.5 (ae). A property of real numbers is true except for a set of zero measure, then it is true **almost everywhere**.

Lemma 10.2.5 (Borel-Cantelli). Let $\{E_k\}_{k=1}^{\infty}$ be a countable collection of Lebesgue measurable sets for which $\sum_{k=1}^{\infty} m(E_k) < \infty$. Then, almost all $x \in \mathbb{R}$ belongs to at most finitely many of the E_k 's.

Proof. We have $\sum_{k=1}^{\infty} m(E_k) < \infty$. Then, by convergence of the series

$$\lim_{n \rightarrow \infty} \sum_{k=n}^{\infty} m(E_k) = 0$$

Also, $\left\{\bigcup_{k=n}^{\infty} E_k\right\}_{n=1}^{\infty}$ is a descending collection of Lebesgue measurable sets. By continuity of Lebesgue measure for descending collection of Lebesgue measurable

sets, we have

$$m\left(\bigcap_{n=1}^{\infty} \bigcup_{k=n}^{\infty} E_k\right) = \lim_{n \rightarrow \infty} m\left(\bigcup_{k=n}^{\infty} E_k\right) = 0$$

Clearly, $\lim_{n \rightarrow \infty} \bigcup_{k=n}^{\infty} E_k$ is a set of zero measure. Suppose $x \in \mathbb{R}$ belongs to countably many E_k 's. Then, for any $m \in \mathbb{N}$, there exists $k > m$ such that $x \in E_k$. Clearly, $x \in \lim_{n \rightarrow \infty} \bigcup_{k=n}^{\infty} E_k$. That is, x belongs to a set of zero measure. Therefore by contraposition, if x does not belong to a set of measure zero, then x belongs to at most finitely many E_k 's. In other words, almost every x in \mathbb{R} belongs to at most finitely many E_k 's. \square

Exercise

24. $m(E_1 \cup E_2) + m(E_1 \cap E_2) = m(E_1) + m(E_2)$

25. $m(B_1) < \infty$ is necessary for continuity property of measure for descending collection of measurable sets.

26.

$$m^*\left(A \cap \bigcup_{k=1}^{\infty} E_k\right) = \sum_{k=1}^{\infty} m^*(A \cap E_k)$$

27. Let m' be set function on a σ -algebra and m' is countably additive.

- (a) m' is finitely additive, monotone, countably monotone, and has excision property
- (b) m' has continuity properties

28. continuity + finite additivity \implies countable additivity

10.2.6 Non-measurable sets

Every measurable set of positive measure contains a non-measurable set.

Definitions 10.2.6 (Rational Equivalence). Let E be any subset of \mathbb{R} . The relation $xRy \iff x - y \in \mathbb{Q}$ is an equivalence^{†8} relation on \mathbb{R} .

Definitions 10.2.7 (Choice set, C_E). Let E be any subset of \mathbb{R} and R be an equivalence relation on E . By axiom of choice, there exists a choice set $C_E \subset E$ containing an exactly an element from each equivalence class.

Definitions 10.2.8 (translate). Let E be a subset of \mathbb{R} . Let $\lambda \in \mathbb{R}$. Then $\lambda + E = \{\lambda + x : x \in E\}$ is a translate of E .

With the help of following lemma, we prove that for any measurable set E of positive measure, the subset C_E is non-measurable.

⁸ $x - x = 0 \in \mathbb{Q}$, $(y - x) = -(x - y) \in \mathbb{Q}$ and $x - z = (x - y) + (y - z) \in \mathbb{Q}$

Lemma 10.2.6. *Let E be a bounded, measurable set. Suppose there exists a bounded, countably infinite set Λ for which the collection of translates of E under Λ , $\{\lambda + E\}_{\lambda \in \Lambda}$ are disjoint. Then $m(E) = 0$.*

In other words, if a bounded measurable set has countably many disjoint translates, then it is of measure zero. That is, there doesn't exist a bounded set of positive measure which has countably many disjoint translates.

Proof. The Lebesgue measure is translation invariant. Thus, the translates, $\lambda + E$ are measurable and $m(\lambda + E) = m(E)$.

We have, E and Λ are bounded, $\bigcup_{\lambda \in \Lambda} \lambda + E$ is bounded and is of finite measure. The Lebesgue measure is countably additive. And since translates are disjoint,

$$m \left[\bigcup_{\lambda \in \Lambda} \lambda + E \right] = \sum_{\lambda \in \Lambda} m(\lambda + E) < \infty$$

Clearly, $m(E) = 0$. Suppose $m(E) = \varepsilon$. Then $\sum_{\lambda \in \Lambda} m(\lambda + E) = \sum_{\lambda \in \Lambda} \varepsilon = \infty$ since Λ has countably infinite elements. \square

Theorem 10.2.7 (Vitali). *Any set E of real numbers with positive outer measure contains a subset that fails to be measurable.*

More importantly, every measurable set of positive measure contains a non-measurable set.

Proof. Case 1 : E is bounded and non-measurable

Suppose E is not measurable. Then $E \subset \mathbb{R}$ is a non-measurable set and the result is trivial.

Case 2 : E is bounded and measurable

Suppose E is a bounded, measurable subset of positive measure. Let C_E be a choice set of E under rational equivalence. Let Λ be any bounded, countably infinite set of rational numbers. Clearly, the translates of C_E under Λ are disjoint.

Suppose $x \in (\lambda_1 + C_E) \cap (\lambda_2 + C_E)$. Then, $x = \lambda_1 + y = \lambda_2 + z$. Clearly, $y = z$ since $y - z \in \mathbb{Q}$ and we chose precisely one element from each equivalence class. Again, $y = z \implies \lambda_1 = \lambda_2$. In other words, the intersecting the translates are identical. That is, two distinct translates will be disjoint.

Suppose C_E is measurable. Since C_E and Λ are bounded and C_E has countably many disjoint translates under Λ , by lemma $m(C_E) = 0$. However, $E \subset \bigcup_{\lambda \in [-2b, 2b] \cap \mathbb{Q}} \lambda + C_E$ for sufficiently large⁹ $b \in \mathbb{R}$.

$$m(E) \leq m \left(\bigcup_{\lambda \in [-2b, 2b] \cap \mathbb{Q}} \lambda + C_E \right) = \sum_{\lambda \in [-2b, 2b] \cap \mathbb{Q}} m(\lambda + C_E) = 0$$

⁹Since E is bounded there exists $b \in \mathbb{R}$ such that $E \subset [-b, b]$

which is a contradiction since E has positive measure.

Case 3 : E is unbounded and measurable

Suppose E is an unbounded subset of positive outer measure, then by the definition of Lebesgue outer measure, E has a bounded subset of positive outer measure. And by case 1 & 2, this set has a non-measurable subset. \square

Theorem 10.2.8. *There are disjoint subsets A, B of real numbers for which*

$$m^*(A \cup B) < m^*(A) + m^*(B) \quad (10.27)$$

Proof. Suppose that for every disjoint pair of subsets $A, B \subset \mathbb{R}$, $m^*(A \cup B) = m^*(A) + m^*(B)$. Then, by the definition of Lebesgue measurability, every subset of real numbers is measurable. By, Vital's theorem there does exist non-measurable subsets of real numbers which is a contradiction. Therefore, there does exist disjoint subsets A, B such that $m^*(A \cup B) \neq m^*(A) + m^*(B)$. By subadditivity of Lebesgue outer measure, we have $m^*(A \cup B) \leq m^*(A) + m^*(B)$. Therefore, $m^*(A \cup B) < m^*(A) + m^*(B)$. \square

Exercise

29. (a)
- (b) Rational Equivalence on \mathbb{Q} gives singleton choice set as difference two rational numbers is always rational. $\frac{a}{b} - \frac{c}{d} = \frac{ad-bc}{bd}$. Thus, $\{0\}$ is a choice set.
- (c) Difference two numbers being irrational is not an equivalence relation as it violates transitivity. $x-y, y-z \notin \mathbb{Q} \not\Rightarrow x-z \notin \mathbb{Q}$. For example, $\sqrt{2} - \sqrt{3}, \sqrt{3} - \sqrt{2} \notin \mathbb{Q}$. However, $\sqrt{2} - \sqrt{2} = 0 \in \mathbb{Q}$.
- 30.
- 31.
- 32.
- 33.

10.2.7 Cantor Set and Cantor-Lebesgue Function

Cantor set, C

We know that every countable set is of measure zero. However, subsets of zero measure are not necessarily countable. Cantor set is an uncountable set of zero measure.

Definitions 10.2.9 (Cantor set). Consider unit interval, $I = [0, 1]$. Let $C_0 = I$. Let $\{I_k\}$ be the collection of subintervals in C_k . Construct C_{k+1} recursively by removing subintervals of length $\frac{l(I_k)}{3}$ from the middle of each I_k in C_k . Cantor set C is given by,

$$C = \bigcap_{k=1}^{\infty} C_k \quad (10.28)$$

Note that C_k are descending collection of 2^k disjoint, closed intervals, each of length $\frac{1}{3^k}$. Thus, effective length of C_k is $(\frac{2}{3})^k$.

Theorem 10.2.9. *Cantor set C is a closed, uncountable set of measure zero.*

Proof. Step 1 : C is measurable

By construction, Cantor set is an intersection of (countably many) closed subsets of real numbers. Since intersections of closed sets are closed, Cantor set is also closed. Also every closed subset is a Borel set and therefore measurable.

Step 2 : $m(C) = 0$

From the construction of C , we have $m(C_k) = (\frac{2}{3})^k$. Clearly, $\{C_k\}_{k=1}^\infty$ is a descending collection of measurable sets and $m(C_1) < \infty$. Thus by continuity of Lebesgue measure,

$$m(C) = m\left(\bigcap_{k=1}^\infty C_k\right) = \lim_{k \rightarrow \infty} m(C_k) = \lim_{k \rightarrow \infty} \left(\frac{2}{3}\right)^k = 0$$

Step 3 : C is uncountable

Suppose C is countable. Then elements of C can be enumerated. That is, $C = \{x_k\}_{k=1}^\infty$. We have, $x_1 \in C \implies x_1 \in C_1$. There are two disjoint intervals in C_1 . Clearly, C_1 has an interval F_1 which doesn't contain x_1 . Similarly, there exists a closed interval, F_2 in C_2 such that $x_2 \notin F_2$ and $F_2 \subset F_1$. Continuing like this, we get a descending collection of closed intervals $\{F_k\}_{k=1}^\infty$ such that $x_k \notin F_k$.

By nested set theorem, intersection of descending collection of closed and bounded intervals is non-empty. Thus, $\bigcap_{k=1}^\infty F_k \neq \phi$. Let $x \in \bigcap_{k=1}^\infty F_k$. Clearly $x \neq c_k$ for any k as $c_k \notin F_k \implies c_k \notin \bigcap_{k=1}^\infty F_k$. However, $x \in C$ which is a contradiction to the assumption that $C = \{c_k : k = 1, 2, \dots\}$. Therefore, elements of C can't be enumerated. In other words, C is uncountable. \square

Cantor-Lebesgue Function, φ

Definitions 10.2.10 (increasing). A real-valued function f is increasing if $f(x) \geq f(y)$ for every $x > y$.

Definitions 10.2.11 (strictly increasing). A real-valued function f is strictly increasing if $f(x) > f(y)$ for every $x > y$.

Definitions 10.2.12 (Cantor-Lebesgue function φ). Let C be the Cantor set. Define open set, $\mathcal{O} = I \sim C$.

$$\mathcal{O} = I \sim C = I \sim \left(\bigcap_{k=1}^\infty C_k\right) = \bigcup_{k=1}^\infty (I \sim C_k) = \bigcup_{k=1}^\infty \mathcal{O}_k$$

We know that \mathcal{O}_k has 2^{k-1} open intervals $I_{k,1}, I_{k,2}, \dots, I_{k,2^{k-1}}$. Define Cantor-Lebesgue function φ on \mathcal{O} by $\varphi(x) = \frac{m}{2^k}$ for every $x \in I_{k,m}$. We may extend φ to unit interval such that $\varphi(0) = 0$ and $\varphi(x) = \sup \{\varphi(t) : t \in [0, x] \cap \mathcal{O}\}$.

Clearly, by the construction ϕ is an increasing real-valued function.

Theorem 10.2.10 (Properties of φ). *Cantor-Lebesgue function $\varphi : I \rightarrow I$ is a surjection. And φ is differentiable in \mathcal{O} .*

Proof. We know from the definition of φ on \mathcal{O} , it is increasing on \mathcal{O} . And for extending φ from \mathcal{O} to $[0, 1]$ we are considering the supremum(least upper bound) of all the previous values of φ on \mathcal{O} . Thus, function φ is increasing on the unit interval, $[0, 1]$.

For continuity of φ , it is enough to prove that φ doesn't have any jump discontinuities^{†10} as it is an increasing function. Suppose $x \in C$ and $x \neq 0$. Then for sufficiently large k , x lies between two consecutive intervals of \mathcal{O}_k . Let a_k and b_k be the upper and lower bounds of intervals to the left and right of x in \mathcal{O}_k . Then $x \in (a_k, b_k)$.

We know that, $\varphi(b_k) - \varphi(a_k) = \frac{1}{2^k}$. And $\varphi(a_k) < \varphi(x) < \varphi(b_k)$, by the construction of φ . As $k \rightarrow \infty$, the jump $\lim_{k \rightarrow \infty} \varphi(b_k) - \varphi(a_k) = \lim_{k \rightarrow \infty} \frac{1}{2^k} = 0$. Thus, φ doesn't have any jump discontinuities. Therefore, φ is continuous on $[0, 1]$.

Clearly, φ is constant on each interval of \mathcal{O} . Therefore, its derivative exists and equals to zero everywhere on \mathcal{O} .

We have, $\mathcal{O} = [0, 1] \sim C$. Thus, $m(\mathcal{O}) = 1$, since $m(C) = 0$ and $m([0, 1]) = 1$. Also, φ is an increasing, continuous function from $[0, 1]$ into $[0, 1]$ where $\varphi(0) = 0$ and $\varphi(1) = 1$. Therefore, Cantor-Lebesgue function, φ is onto unit interval $[0, 1]$. \square

Measurable, non-Borel set

We already saw that there exists non-measurable subsets. Using the properties of the function ψ , we assert that there exist a measurable set $\psi^{-1}(W)$ which is not a Borel set.

Theorem 10.2.11. *The function $\psi : [0, 1] \rightarrow [0, 2]$ defined by $\psi(x) = \varphi(x) + x$*

1. *is a strictly increasing^{†11} function which maps $[0, 1]$ onto $[0, 2]$ and*
2. *ψ maps Cantor set, C onto a set of positive measure and*
3. *ψ maps a measurable^{†12} subset of C onto a non-measurable set.*

Proof. By definition, $\psi(x) = \varphi(x) + id(x)$. We know that the sum of continuous functions is continuous. Thus, ψ is continuous. Again, the sum of an increasing function and strictly increasing function is always strictly increasing. Thus, ψ is a strictly increasing function.

^{†10}If the function φ has unbounded variation(jump) at some points of $[0, 1]$. That is $\exists \varepsilon > 0$ such that $\forall \delta > 0, \exists x \in [0, 1]$ where $f(x + \varepsilon/2) - f(x - \varepsilon/2) > \delta$. —need revisit—

^{†11}Strictly increasing functions are one-to-one. Thus, ψ is a bijection.

^{†12}Cantor set is a subset of zero measure. Thus every subset of Cantor set is measurable.

We know that ψ is only translating open intervals in \mathcal{O} . That is, $\psi(I_{k,m}) = I_{k,m} + \frac{m}{2^k}$. In other words, ψ translates every interval I_k in \mathcal{O} into an interval of same length since φ is constant in each subinterval of \mathcal{O} . Thus,

$$m(\psi(\mathcal{O})) = \sum_{k=1}^{\infty} l(\psi(I_k)) = \sum_{k=1}^{\infty} l(I_k) = m(\mathcal{O}) = 1$$

We know that $[0, 2] = \psi(\mathcal{O}) \cup \psi(C)$. Since $m(\psi(\mathcal{O})) = 1$, $m(\psi(C)) = 2 - 1 = 1$.

Clearly, $\psi(C)$ is a subset of positive measure. Thus, by Vitali's theorem $\psi(C)$ has a subset W which is not measurable. However, $\psi^{-1}(W)$ is a subset of C which is of measure zero. Therefore, C has a measurable subset which ψ maps onto a non-measurable set. \square

Theorem 10.2.12. *Cantor set has a measurable subset which is not a Borel set.*

Proof. By Vitali's theorem, $\psi(C)$ has a non-measurable subset, say W . Clearly W is not a Borel set, since every Borel set is measurable.

Now $\psi : \psi^{-1}(W) \rightarrow W$ is a bijection. And $W \subset \psi(C) \implies \psi^{-1}(W) \subset C$. Also we know that, every subset of zero measure sets are measurable. Thus, $\psi^{-1}(W)$ is measurable subset of Cantor set, C .

Suppose $\psi^{-1}(W)$ is a Borel set. Then, W must be a Borel set, since continuous functions maps Borel sets onto Borel sets. This leads to a contradiction since W is not a Borel set. Therefore, Cantor set C has a measurable, non-Borel subset $\psi^{-1}(W)$. \square

10.3 Lebesgue Measurable Functions

10.3.1 Sum, Products, and Compositions

Theorem 10.3.1. *Let function f have a measurable domain. Then the following statements are equivalent :*

1. $\forall c \in \mathbb{R}, \{x \in E : f(x) > c\}$ is measurable
2. $\forall c \in \mathbb{R}, \{x \in E : f(x) \geq c\}$ is measurable
3. $\forall c \in \mathbb{R}, \{x \in E : f(x) < c\}$ is measurable
4. $\forall c \in \mathbb{R}, \{x \in E : f(x) \leq c\}$ is measurable

And each statement above imply that $\{x \in E : f(x) = c\}$ is measurable.

Proof. Step 1 : 1 \iff 4 and 2 \iff 3

The sets considered in statements 1 and 4 are complementary. Thus, if one set is measurable, then the other set is also measurable. Similarly, the sets in 2 and 3 are complementary. Thus, one statement implies the other.

Step 2 : 1 \implies 2

Suppose $\{x \in E : f(x) > c\}$ is measurable for any real number c . Clearly, the

sets $\{x \in E : f(x) > c - \frac{1}{k}\}$ is measurable for every natural number k . And countable intersection of measurable sets is measurable. Thus,

$$\bigcap_{k=1}^{\infty} \left\{ x \in E : f(x) > c - \frac{1}{k} \right\} = \{x \in E : f(x) \geq c\} \text{ is measurable}$$

Step 3 : $2 \implies 1$ Suppose $\{x \in E : f(x) \geq c\}$ is measurable for any real number c . Clearly, the sets $\{x \in E : f(x) \geq c + \frac{1}{k}\}$ is measurable for every natural number k . Again, countable union of measurable sets is measurable. Thus,

$$\bigcup_{k=1}^{\infty} \left\{ x \in E : f(x) \geq c + \frac{1}{k} \right\} = \{x \in E : f(x) > c\} \text{ is measurable}$$

Step 4 : $\forall c \in \mathbb{R}, \{x \in E : f(x) = c\}$ is measurable

Suppose one of the statements is true. Then other three statements are also true. The sets $\{x \in E : f(x) \geq c\}$ and $\{x \in E : f(x) \leq c\}$ are both measurable. Thus, their intersection $\{x \in E : f(x) = c\}$ is also measurable. \square

Definitions 10.3.1 (measurable function). A function f is measurable if

1. the domain of f is measurable and
2. one of the following statements is true
 - (a) $\forall c \in \mathbb{R}, \{x \in E : f(x) > c\}$ is measurable
 - (b) $\forall c \in \mathbb{R}, \{x \in E : f(x) \geq c\}$ is measurable
 - (c) $\forall c \in \mathbb{R}, \{x \in E : f(x) < c\}$ is measurable
 - (d) $\forall c \in \mathbb{R}, \{x \in E : f(x) \leq c\}$ is measurable

Note : If f is a measurable function, then its domain is measurable and $f^{-1}(c)$ is measurable for any real number c . (why ?)

Study of measurable functions

1. Inverse images of open sets measurable

Suppose function f is defined on a measurable set E . Then f is a measurable function $\iff \forall$ open set \mathcal{O} , $f^{-1}(\mathcal{O})$ is measurable.

Proof. Given the domain of f is measurable.

Step 1 : $f^{-1}(\mathcal{O})$ is measurable $\implies f$ is measurable

Suppose inverse image of every open set is measurable. Then, $f^{-1}(c, \infty) = \{x \in E : f(x) > c\}$ is measurable. Then, by the definition, f is a measurable function.

Step 2 : f measurable $\implies f^{-1}(\mathcal{O})$ is measurable

Suppose function f is measurable. Let \mathcal{O} be an open set. Then it can be expressed as a countable union of open, bounded intervals, I_k . That

is, $\mathcal{O} = \bigcup_{k=1}^{\infty} I_k$. We know that, $\forall k \in \mathbb{N}$, $I_k = (a_k, b_k) = (-\infty, b_k) \cap$

(a_k, ∞) . By the definition of measurability $\{x \in E : f(x) < b_k\} = f^{-1}(-\infty, b_k)$ and $\{x \in E : f(x) > a_k\} = f^{-1}(a_k, \infty)$ are measurable. And $f^{-1}(a_k, b_k) = f^{-1}(-\infty, b_k) \cap f^{-1}(a_k, \infty)$ is measurable. Again, inverse image of the open set $f^{-1}(\mathcal{O})$ which is a countable union of measurable sets $\{f^{-1}(I_k)\}_{k=1}^{\infty}$ is also measurable. Therefore, inverse image of every open set is measurable. \square

2. Continuous, real-valued functions are measurable

Suppose the function f has a measurable domain. And f is a real-valued, continuous function. Then f is measurable.

Proof. Let E be a measurable set. And f be a continuous function on E . Let \mathcal{O} be an open set, then by open set characterisation of continuous function we have $f^{-1}(\mathcal{O}) = E \cap U$ where U is an open set. Clearly, $f^{-1}(\mathcal{O})$ is a union of two measurable sets and thus measurable. Therefore, f is measurable since \forall open set \mathcal{O} , $f^{-1}(\mathcal{O})$ is measurable. \square

3. Monotone function on an interval is measurable

Proof. Without loss of generality, suppose that f is an increasing function defined on an interval I . Since, every interval is measurable, f is defined on a measurable set.

Consider $\{g_n\}_{n=1}^{\infty}$ defined by $g_n(x) = f(x) + \frac{x}{n}$. Since g_n are strictly increasing functions, $g_n^{-1}(c, \infty) = I \cap U$ where U is an interval. Thus, $g_n^{-1}(c, \infty) = \{x \in I : g_n(x) > c\}$ is always measurable. Therefore, $\{g_n\}$ is a family of measurable functions.

Now, from the construction of g_n , we have $\lim_{n \rightarrow \infty} g_n = f$. We have, $\{g_n\}$ is a sequence of measurable functions converging pointwise to the limit function f (a.e.) on the interval I . Therefore, f is measurable. ^{†13} \square

4. f measurable and $f = g$ (a.e.) $\implies g$ measurable

Proof. Suppose f, g are functions defined on a measurable set E . Suppose f is measurable. Let $A \subset E$ such that $A = \{x \in \mathbb{R} : f(x) \neq g(x)\}$. We have $f = g$ (a.e.). Thus, $m(A) = 0$. And we have,

$$\begin{aligned} \{x \in E : g(x) > c\} &= \{x \in A : g(x) > c\} \cup \{x \in E \sim A : g(x) > c\} \\ &= \{x \in A : g(x) > c\} \cup \{x \in E \sim A : f(x) > c\} \\ &= \{x \in A : g(x) > c\} \cup [\{x \in E : f(x) > c\} \cap (E \sim A)] \end{aligned}$$

Clearly, $\{x \in A : g(x) > c\}$ is a subset of a set of measure zero and thus measurable. Also, $E \sim A$ measurable since both E and A are measurable. And since f is measurable, $\{x \in E : f(x) > c\}$ is measurable. Thus, $\{x \in E : g(x) > c\}$ is measurable for any $c \in \mathbb{R}$. Therefore, g is a measurable function. \square

¹³Limit of measurable functions under pointwise convergence(a.e.) is measurable. [We will prove this result in the upcoming subsection .](#)

5. **f measurable $\iff f|_D, D \sim E$ measurable ($\forall D$ measurable)**

Suppose function f is an extended real-valued function on a measurable set E . For measurable subset D of E , f is measurable on E if and only if the restriction of f to D and $D \sim E$ are measurable.

Proof. Part 1 : f measurable $\implies f|_D$ measurable

Suppose f is a measurable function defined on E and D is a measurable subset of E . Since, f is measurable, E is measurable. Clearly $E \sim D$ is measurable.

$$\begin{aligned}\{x \in D : f|_D(x) > c\} &= \{x \in D : f(x) > c\} \\ &= \{x \in E : f(x) > c\} \cap D\end{aligned}$$

Since f is measurable, $\{x \in E : f(x) > c\}$ is measurable for any $c \in \mathbb{R}$. And intersection of measurable sets are measurable. Thus, $\{x \in D : f|_D(x) > c\}$ is measurable for any $c \in \mathbb{R}$. Therefore, $f|_D$ is a measurable function. \square

Definitions 10.3.2 (characteristic function). Let A be a subset of \mathbb{R} . The characteristic function χ_A of A is given by

$$\chi_A(x) = \begin{cases} 1 & x \in A \\ 0 & x \notin A \end{cases} \quad (10.29)$$

Definitions 10.3.3. Let $\{f_1, f_2, \dots, f_n\}$ be a finite family of measurable functions on the same domain E . Then $\max\{f_1, f_2, \dots, f_n\}$ on E is given by

$$\max\{f_1, f_2, \dots, f_n\}(x) = \max\{f_1(x), f_2(x), \dots, f_n(x)\}, \quad \forall x \in E \quad (10.30)$$

And function $\min\{f_1, f_2, \dots, f_n\}$ on E is given by

$$\min\{f_1, f_2, \dots, f_n\}(x) = \min\{f_1(x), f_2(x), \dots, f_n(x)\}, \quad \forall x \in E \quad (10.31)$$

Properties of Measurable Functions

Suppose f, g are measurable functions on E and are finite a.e. on E .

1. **Linearity :** $\alpha f + \beta g$ is measurable $\forall \alpha, \beta$

Proof. Step 1 : f measurable $\implies \alpha f$ measurable

Suppose $\alpha = 0$. Then $\alpha f = 0$. And 0 function is trivially measurable.

Suppose $\alpha \neq 0$. Then $\{x \in E : f(x) > c/\alpha\} = \{x \in E : \alpha f(x) > c\}$ is measurable for any $c \in \mathbb{R}$. Therefore αf is measurable for any $\alpha \in \mathbb{R}$.

In other words, if f is measurable, then αf is measurable and if g is measurable, then βg is measurable for any $\alpha, \beta \in \mathbb{R}$. Therefore, it is enough to prove that f, g measurable $\implies f + g$ measurable.

Step 2 : f, g are measurable $\implies f + g$ is measurable

$$f(x) + g(x) < c \iff f(x) < c - g(x)$$

Since \mathbb{Q} is dense, there exists a rational number $q \in \mathbb{Q}$ between any two distinct real numbers

$$\begin{aligned} f(x) + g(x) < c &\iff \exists q \in \mathbb{Q}, f(x) < q < c - g(x) \\ &\iff f(x) < q \text{ AND } g(x) < c - q \end{aligned}$$

$$\begin{aligned} \{x \in E : f + g(x) < c\} &= \{x \in E : f(x) + g(x) < c\} \\ &= \bigcup_{q \in \mathbb{Q}} [\{x \in E : f(x) < q\} \cap \{x \in E : g(x) < c - q\}] \end{aligned}$$

Since, f, g are measurable, each set in the union is measurable. Thus, $\{x \in E : f + g(x) < c\}$ is measurable for every $c \in \mathbb{R}$, since rational numbers are countable, and countable union of measurable sets is measurable. Therefore, $f + g$ is measurable. \square

2. Product : fg is measurable

Proof. Suppose f, g are measurable functions which are finite (a.e.) on a measurable set E . And we have,

$$fg = \frac{1}{2} [(f + g)^2 - f^2 - g^2] \quad (10.32)$$

Step 1 : f is measurable $\implies f^2$ is measurable

Suppose $c \geq 0$. If $c < 0$, then $\{x \in E : f^2(x) > c\} = E$ is measurable.

$$\{x \in E : f^2(x) > c\} = \{x \in E : f(x) > \sqrt{c}\} \cup \{x \in E : f(x) < -\sqrt{c}\}$$

is a union of two measurable sets and is measurable.

Step 2 : fg is measurable

We have f, g are measurable. Thus, $f + g$ is measurable, since linear combination of measurable sets is measurable. And, $f^2, g^2, (f + g)^2$ are measurable by Step 1. We know that fg is a linear combination of these measurable sets and therefore fg is measurable. \square

3. Composition of measurable functions is not necessarily measurable

We know that, Cantor set C has subset $A = \psi^{-1}(W)$ such that ψ maps A onto a non-measurable set W .^{†14} Then A is a measurable subset of open interval $(0, 1)$ and $\psi(A)$ is non-measurable.

Let χ_A be the characteristic function of A . Then $\chi_A \circ \psi^{-1}$ is not measurable since $\{x \in (0, 1) : \chi_A \circ \psi^{-1}(x) \geq 1\} = \psi(A)$ is not measurable. But, both the functions χ_A and ψ^{-1} are measurable functions since $\chi_A^{-1}(y)$ is either ϕ, A or A^c and ψ^{-1} is a continuous real-valued function defined on $(0, 1)$.

^{†14}From the proof of Vitali's theorem and associated lemma, $\psi(C)$ has such a subset W which is a choice set of $\psi(C)$ under rational equivalence.

4. **If f is continuous and g is measurable, then $f \circ g$ is measurable**

Proof. Suppose f be a continuous function and g be a measurable function both defined on a common set E which is measurable. Since f is continuous, we know that f is measurable.

Let \mathcal{O} be an open set. Then $(f \circ g)^{-1}(\mathcal{O}) = g^{-1}(f^{-1}(\mathcal{O}))$. By characterisation of continuity, $U = f^{-1}(\mathcal{O})$ is an open set since \mathcal{O} is open. And by characterisation of measurability of g , we have $g^{-1}(U)$ is an open set since g is measurable and U is open. Now we have $(f \circ g)^{-1}(\mathcal{O})$ is an open set for any open set \mathcal{O} . Therefore by the characterisation of measurability, $f \circ g$ is measurable. \square

5.

For a finite family of measurable functions $\{f_1, f_2, \dots, f_n\}$ with common domain E , both the functions $\max\{f_1, f_2, \dots, f_n\}$ and $\min\{f_1, f_2, \dots, f_n\}$ are measurable.

Proof. Let $\{f_1, f_2, \dots, f_n\}$ be a finite family of measurable functions defined on a common measurable set E . We have,

$$\{x \in E : \max\{f_1, f_2, \dots, f_n\}(x) > c\} = \bigcup_{k=1}^n \{x \in E : f_k(x) > c\}$$

is a finite union of measurable sets since each f_k is measurable. Therefore, $\max\{f_1, f_2, \dots, f_n\}$ is measurable. Similarly, we have,

$$\{x \in E : \min\{f_1, f_2, \dots, f_n\}(x) < c\} = \bigcup_{k=1}^n \{x \in E : f_k(x) < c\}$$

Therefore, $\min\{f_1, f_2, \dots, f_n\}$ is also measurable. \square

Note : Let f be a measurable function on a measurable set E . Then $-f$ is measurable since

$$\{x \in E : -f(x) > c\} = \{x \in E : f(x) < -c\}$$

And $|f|, f^+, f^-$ are also measurable since

$$|f| = \max\{f, -f\}, \quad f^+ = \max\{f, 0\}, \quad f^- = \max\{-f, 0\}$$

These functions f^+, f^- are quite important as we may write $f = f^+ - f^-$ where both the functions f^+ and f^- are non-negative. And Lebesgue integral for non-negative functions are sufficient to integrate any function f .

$$\int_E f = \int_E f^+ - \int_E f^-$$

10.3.2 Sequential Pointwise Limits and Simple Approximation

Definitions 10.3.4 (pointwise convergence). A sequence of functions $\{f_n\}$ on a common domain E convergence pointwise to f if

$$\forall x \in E, \lim_{n \rightarrow \infty} f_n(x) = f(x)$$

Definitions 10.3.5 (pointwise convergence(a.e.)). Let E_0 be a set of zero measure. A sequence of functions $\{f_n\}$ on a common domain E convergence pointwise a.e. on $E_0 \subset E$ to f if

$$\forall x \in E \sim E_0, \lim_{n \rightarrow \infty} f_n(x) = f(x)$$

Definitions 10.3.6 (uniform convergence). A sequence of functions $\{f_n\}$ on a common domain E convergence uniformly to f if

$$\forall \varepsilon > 0, \exists N \in \mathbb{N}, \text{ such that } \forall n \geq N, |f_n - f| < \varepsilon$$

Theorem 10.3.2 (Pointwise Convergence(a.e.) preserves measurability). *Let $\{f_n\}$ be a sequence measurable functions on E . If $\{f_n\}$ converges to f pointwise a.e. on E , then f is measurable.*

Proof. Let $\{f_n\}$ be a sequence of measurable functions converging to f pointwise a.e. on E . Then $\{f_n\}_{n=1}^{\infty}$ converges pointwise on $E \sim E_0$ where E_0 is a set of measure zero. Without loss of generality, suppose that $\{f_n\}$ converges pointwise everywhere on E .¹⁵

Let $c \in \mathbb{R}$. We claim that,

$$f(x) < c \iff \exists n, k \text{ such that } f_j(x) < c - \frac{1}{n}, \forall j \geq k \quad (10.33)$$

Suppose that for every natural numbers n, k , there exists $j \geq k$ for which $f_j(x) \geq c - \frac{1}{n}$. Since $f(x)$ is the limit function, $f(x) \not< c$ which is a contradiction. Thus, we may write,

$$\{x \in E : f(x) < c\} = \bigcup_{n,k=1}^{\infty} \left[\bigcap_{j=k}^{\infty} \left\{ x \in E : f_j(x) < c - \frac{1}{n} \right\} \right] \quad (10.34)$$

Countable union of countable intersections of measurable sets is also measurable. Thus $\{x \in E : f(x) < c\}$ is a measurable set for any $c \in \mathbb{R}$. Therefore, f is measurable. \square

Definitions 10.3.7 (simple function). A real-valued function φ on a measurable set E is simple if it is measurable and assumes at most finitely many values.

Note : $xRy \iff \varphi(x) = \varphi(y)$ is an equivalence relation which partitions E into k disjoint subsets E .

Note : Suppose φ is a simple function. Then each equivalent class of E under above equivalence is also measurable by the definition of measurable function.(Why ?)

¹⁵Consider $E' = E \sim E_0$

Definitions 10.3.8 (canonical representation). The canonical representation of a simple function φ on E is given by,

$$\varphi = \sum_{k=1}^n c_k \cdot \chi_{E_k} \text{ where } E_k = \{x \in E : \varphi(x) = c_k\} \quad (10.35)$$

Definitions 10.3.9 (bounded). Let f be a function on E . The function f is bounded if there exists $m \in \mathbb{R}$, $m \geq 0$ such that $|f(x)| < m$ for every $x \in E$.

Lemma 10.3.3 (simple approximation). Let f be a measurable, real-valued function which is bounded on E . Then for any $\varepsilon > 0$, there exist simple functions φ_ε and ψ_ε approximating f (from below and above) such that $0 \leq \psi_\varepsilon - \varphi_\varepsilon < \varepsilon$ on E .

Proof. Let f be a bounded, measurable function on E . Then there exists open, bounded interval (c, d) such that $f(E) \subset (c, d)$. Let $P = \{y_0, y_1, \dots, y_n\}$ be a partitions of open interval (c, d) such that $y_0 = c < y_1 < \dots < y_{n-1} < y_n = d$ and $y_k - y_{k-1} < \varepsilon$ for every k . Define $I_k = [y_{k-1}, y_k)$ and $E_k = f^{-1}(I_k)$. Then E_k are measurable since I_k are intervals and f is measurable. (Why ?)

Define $\varphi_\varepsilon = \sum_{k=1}^n y_{k-1} \cdot \chi_{E_k}$ and $\psi_\varepsilon = \sum_{k=1}^n y_k \cdot \chi_{E_k}$. Then for any $x \in E$, we have $\varphi_\varepsilon(x) = y_{k-1} \leq f(x) \leq y_k = \psi_\varepsilon(x)$. Clearly, $\varphi_\varepsilon \leq \psi_\varepsilon$ by the construction. And for each k , we have $\psi_\varepsilon(y) - \varphi_\varepsilon(y) = y_k - y_{k-1} < \varepsilon$ for any $y \in [y_{k-1}, y_k)$. Therefore, $0 \leq \psi_\varepsilon - \varphi_\varepsilon < \varepsilon$. \square

Theorem 10.3.4 (simple approximation). An extended real-valued function f on a measurable set E is measurable if and only if there exists a sequence of simple functions $\{\varphi_n\}$ converging to f pointwise on E and $|\varphi_n| \leq |f|$ on E for every $n \in \mathbb{N}$.

And if f is non-negative, there exists an increasing sequence of functions $\{\varphi_n\}$ with the same property.

Proof. Part 1

Suppose there exists a sequence of simple functions $\{\varphi_n\}$ converging pointwise to f on E . Since simple functions are measurable and pointwise limit of measurable functions are measurable, f is measurable.

Part 2

Suppose f is measurable. Without loss of generality, suppose that $f \geq 0$ on E . Otherwise, $f = f^+ - f^-$ where $f^+, f^- \geq 0$. And f is measurable since both f^+ and f^- are measurable.

Step 1 : Construction of simple functions on E_n

Let $n \in \mathbb{N}$. Define $E_n = \{x \in E : f(x) \leq n\}$. Since f is measurable, E_n is a measurable subset of E . And we know that $f|_{E_n}$ is also a non-negative, bounded, measurable function.

Then by simple approximation lemma, for $\varepsilon = \frac{1}{n}$ there exists φ_n and ψ_n such that $0 \leq \varphi_n \leq f \leq \psi_n$ and $0 \leq \psi_n - \varphi_n < \frac{1}{n}$ on E_n .

Step 2 : Extending simple function to E

Extend φ_n to E by setting $\varphi_n(x) = n$ if $f(x) > n$. Again, φ_n is a simple function on E . And $0 \leq \varphi_n \leq f$ on E .

Step 3 : Sequence of simple functions converging to f

We claim that the sequence of simple functions $\{\varphi_n\}$ converges to f pointwise on E . Suppose $x \in E$. Suppose $f(x)$ is finite. Let $m \in \mathbb{N}$. Then $0 \leq f(x) - \varphi_n(x) \leq \frac{1}{n}$ for all $n \geq m$. Therefore, $\lim_{n \rightarrow \infty} \varphi_n(x) = f(x)$. Suppose $f(x) = \infty$. Then $\varphi_n(x) = n, \forall n$. Therefore, $\lim_{n \rightarrow \infty} \varphi_n(x) = f(x)$.

Replace each φ_n with $\max\{\varphi_1, \varphi_2, \dots, \varphi_n\}$ ¹⁶, we have an increasing sequence of simple functions φ_n converging to f pointwise on E . \square

Exercise

- 12.
- 13.
- 14.
- 15.
- 16.
- 17.
- 18.
- 19.
- 20.
- 21.
- 22.
- 23.
- 24.

10.3.3 Littlewood, Egoroff, Lusin

Out of syllabus - But important for our understanding of the subject.

¹⁶We replace each simple function φ_k with the maximum of the finite subsequence from φ_1 upto φ_k . That is, $\varphi'_1 = \max\{\varphi_1\} = \varphi_1$, and $\varphi'_2 = \max\{\varphi_1, \varphi_2\}$, and $\varphi'_3 = \max\{\varphi_1, \varphi_2, \varphi_3\}, \dots$

Littlewoods' three principles

1. Every measurable set is **nearly** a finite union of intervals.
2. Every measurable function is **nearly** continuous.
3. Every convergent sequence of measurable functions is **nearly** uniform.

Theorem 10.3.5 (Egoroff). *Suppose a sequence $\{f_n\}$ of measurable functions converges to f pointwise on a set E of finite measure. Then for any $\varepsilon > 0$, there exists a closed set $F \subset E$ such that $m(E \setminus F) < \varepsilon$ and $\{f_n\}$ converges to f uniformly on F .*

Theorem 10.3.6. *Let f be a simple function on E . Then for any $\varepsilon > 0$, there exists a continuous function g on \mathbb{R} and a closed set $F \subset E$ such that $f = g$ on F and $m(E \setminus F) < \varepsilon$.*

Theorem 10.3.7 (Lusin). *Let f be a real-valued measurable function on E . Then for any $\varepsilon > 0$, there exists a continuous function g on \mathbb{R} and closed set $F \subset E$ such that $f = g$ on F and $m(E \setminus F) < \varepsilon$.*

10.4 Lebesgue Integration

10.4.1 Riemann integral

Definitions 10.4.1 (partition). A set $P = \{a, x_1, x_2, \dots, x_{n-1}, b\}$ is a partition of an interval $[a, b]$ into n subintervals if $a < x_1 < x_2 < \dots < x_{n-1} < b$.

Definitions 10.4.2 (lower/upper Riemann integral). Let f be a bounded, real-valued function on a closed interval $[a, b]$ and P be a partition of $[a, b]$. Let M_i, m_i be the supremum and infimum of f in each subinterval of the partition. Then upper and lower Riemann integrals are the infimum of upper Darboux sums and supremum of lower Darboux sums.

Riemann upper integral,

$$(R) \overline{\int_a^b} f = \inf_P \{U(f, P)\} = \inf \left\{ \sum_{i=1}^n M_i (x_i - x_{i-1}) \right\} \quad (10.36)$$

Riemann lower integral,

$$(R) \underline{\int_a^b} f = \sup_P \{L(f, P)\} = \sup \left\{ \sum_{i=1}^n m_i (x_i - x_{i-1}) \right\} \quad (10.37)$$

Definitions 10.4.3 (Riemann integrable). A function f is Riemann integrable over $[a, b]$ if lower and upper Riemann integrals are equal. And Riemann integral of f over $[a, b]$ is given by,

$$(R) \int_a^b f = (R) \overline{\int_a^b} f = (R) \underline{\int_a^b} f \quad (10.38)$$

Definitions 10.4.4 (step function). A function ψ is a step function if it assumes constant values in each subinterval of some partition of its domain.

$$\psi(x) = \sum_{i=1}^n c_i \cdot \chi_{(x_i, x_{i+1})}$$

Note : $(R) \int_a^b \psi = \sum_{i=1}^n c_i (x_i - x_{i-1})$

Definitions 10.4.5 (Dirchlet's function). Dirchlet's function is given by,

$$f(x) = \begin{cases} 1 & x \in \mathbb{Q} \\ 0 & x \notin \mathbb{Q} \end{cases}$$

Note : Dirchlet's function is not Riemann integrable(why ?). But it is Lebesgue integrable since \mathbb{Q} is Lebesgue measurable.

10.4.2 Lebesgue integral of a bounded, measurable function over a set of finite measure

Definitions 10.4.6. Let ψ be a simple function on a finite measure set E . Then,

$$\int_E \psi = \sum_{i=1}^n a_i \cdot m(E_i) \text{ where } \psi = \sum_{i=1}^n a_i \cdot \chi_{E_i} \quad (10.39)$$

Note : Step functions are simple functions. And the Riemann integral and Lebesgue integral of step functions are the same.

Note : Dirchlet's function is a simple function. And thus Lebesgue integrable.

Lemma 10.4.1. Let $\{E_k\}_{k=1}^n$ be a disjoint collection of measurable subsets of a set E of finite measure. If $\varphi = \sum_{k=1}^n a_k \cdot \chi_{E_k}$, then $\int_E \varphi = \sum_{k=1}^n a_k \cdot m(E_k)$

Proof. Suppose $\varphi = \sum_{k=1}^n a_k \cdot \chi_{E_k}$ where E_k are of finite measure. We consider distinct values of a_i , $\{\lambda_1, \lambda_2, \dots, \lambda_m\}$. Then, the function has the canonical representation, $\varphi = \sum_{r=1}^m \lambda_r \cdot \chi_{A_r}$ where $A_r = \bigcup_{a_k=\lambda_r} E_k$.

Then, we have $\varphi^{-1}(\lambda_k) = A_k$ is a union of finite measurable set and thus measurable. Thus, φ is a measurable function which assumes at most m different

values. And

$$\begin{aligned}
 \int_E \varphi &= \sum_{r=1}^m \lambda_r \cdot m(A_r) \\
 &= \sum_{r=1}^m \lambda_r \cdot m\left(\bigcup_{\substack{k \\ a_k = \lambda_r}} E_k\right) \\
 &= \sum_{r=1}^m \sum_{\substack{k \\ a_k = \lambda_r}} a_k \cdot m(E_k) \\
 &= \sum_{k=1}^n a_k \cdot m(E_k)
 \end{aligned}$$

□

Theorem 10.4.2 (linearity + monotonicity of integral of simple functions). *Let φ and ψ be simple functions defined a set E of finite measure. Then for any $\alpha, \beta \in \mathbb{R}$,*

$$\int_E (\alpha\varphi + \beta\psi) = \alpha \int_E \varphi + \beta \int_E \psi \quad (10.40)$$

Moreover, if $\varphi \leq \psi$ on E , then

$$\int_E \varphi \leq \int_E \psi \quad (10.41)$$

Proof. Suppose φ, ψ are simple functions and $\alpha, \beta \in \mathbb{R}$. Since φ, ψ are simple, there are n disjoint subset of E in which both φ and ψ are constant¹⁷. Then,

$$\begin{aligned}
 \int_E (\alpha\varphi + \beta\psi) &= \int_E \alpha \left(\sum_{i=1}^n a_i \cdot \chi_{E_i} \right) + \beta \left(\sum_{i=1}^n b_i \cdot \chi_{E_i} \right) \\
 &= \int_E (\alpha a_i + \beta b_i) \cdot \chi_{E_i} \\
 &= \sum_{i=1}^n (\alpha a_i + \beta b_i) \cdot m(E_i) \\
 &= \alpha \sum_{i=1}^n a_i \cdot m(E_i) + \beta \sum_{i=1}^n b_i \cdot m(E_i) \\
 &= \alpha \int_E \varphi + \beta \int_E \psi
 \end{aligned}$$

Suppose $\varphi \leq \psi$. Then $\eta = \psi - \varphi$ is also a simple function. And $0 \leq \eta$. Then by linearity of Lebesgue integral of simple functions,

$$\begin{aligned}
 \int_E \eta &= \int_E \psi - \varphi \\
 &= \int_E \psi - \int_E \varphi
 \end{aligned}$$

We have, η is a non-negative, measurable function. And thus, $\int_E \eta \geq 0$. Therefore, $\int_E \varphi \leq \int_E \psi$. □

¹⁷We don't assume that the values assume by ψ are distinct for distinct subsets.

Lebesgue integral for bounded functions

Now we define Lebesgue integral of bounded functions, the same way we have constructed Riemann integral. We use simple approximations and the definition of Lebesgue integral for simple functions for this construction.

Definitions 10.4.7 (upper/lower Lebesgue integral). Let f be a bounded, real-valued function on a set E of finite measure. Any simple approximation lemma, we have two families of simple functions $\{\varphi_\varepsilon\}$ and $\{\psi_\varepsilon\}$ such that $\forall \varepsilon > 0$, $\varphi \leq f \leq \psi$ and $0 \leq \psi - \varphi < \varepsilon$. Then

Upper Lebesgue integral,

$$\overline{\int_E} f = \inf_{\psi_\varepsilon} \left\{ \int_E \psi_\varepsilon \right\} \quad (10.42)$$

Lower Lebesgue integral,

$$\underline{\int_E} f = \sup_{\varphi_\varepsilon} \left\{ \int_E \varphi_\varepsilon \right\} \quad (10.43)$$

Definitions 10.4.8 (Lebesgue integrability of bounded functions). Let f be a bounded, real-valued function defined on a set E of finite measure. Function f is Lebesgue integrable on E if both lower and upper Lebesgue integrals of f over E are equal.

And the Lebesgue integral of a bounded, real-valued function f over E is given by,

$$\int_E f = \sup_{\varphi} \int_E \varphi = \inf_{\psi} \int_E \psi \quad (10.44)$$

Now we can prove that Lebesgue integral is a generalisation of Riemann integral for bounded functions.

Theorem 10.4.3. *Let f be a bounded, real-valued function defined on a closed, bounded interval $[a, b]$. If f is Riemann integrable over $[a, b]$, then it is Lebesgue integrable over $[a, b]$. And both Riemann integral and Lebesgue integral of f over $[a, b]$ are equal.*

Proof. Suppose f is Riemann integrable over $[a, b]$. Then,

$$\sup \left\{ (R) \int_a^b \varphi : \varphi \text{ step}, \varphi \leq f \right\} = \inf \int \left\{ \int_a^b \psi : \psi \text{ step}, f \leq \psi \right\} \quad (10.45)$$

However, every step function is simple and we have

$$\sup \left\{ \int_a^b \varphi : \varphi \text{ simple}, \varphi \leq f \right\} = \inf \int \left\{ \int_a^b \psi : \psi \text{ simple}, f \leq \psi \right\} \quad (10.46)$$

Therefore, every bounded, real-valued function on closed, bounded interval is Lebesgue integrable function if it is Riemann integrable. \square

Theorem 10.4.4. *Every bounded, measurable function f defined on a set E of finite measure is Lebesgue integrable over E .*

Proof. Let E be a set of finite measure. And f be a bounded, measurable function on E . Let $n \in \mathbb{N}$. By simple approximation lemma, for $\varepsilon = \frac{1}{n}$ we have simple functions φ_n, ψ_n such that $\varphi_n \leq f \leq \psi_n$ and $\psi_n - \varphi_n < \frac{1}{n}$ on E .

$$0 \leq \psi_n - \varphi_n \leq \frac{1}{n} \implies 0 \leq \int_E \psi_n - \varphi_n = \int_E \psi_n - \int_E \varphi_n \leq \frac{1}{n} \int_E 1$$

We know that, $\inf\{r_1, r_2, \dots\} \leq r_k$ and $-\sup\{s_1, s_2, \dots\} \leq -s_k$.

$$\begin{aligned} 0 &\leq \inf \left\{ \int_E \psi : \psi \text{ simple}, f \leq \psi \right\} - \sup \left\{ \int_E \varphi : \varphi \text{ simple}, \varphi \leq f \right\} \\ &\leq \int_E \psi_n - \int_E \varphi_n \leq \frac{1}{n} m(E) \end{aligned}$$

This inequality is true for any $n \in \mathbb{N}$. Since $m(E)$ is finite, $\frac{1}{n}m(E) \rightarrow 0$ as $n \rightarrow \infty$. Thus, upper and lower Lebesgue integrals of f are equal. Therefore, f is Lebesgue integrable. \square

Theorem 10.4.5 (linearity + monotonicity of integral of bounded functions).
Let f and g be bounded, measurable functions defined a set E of finite measure. Then for any $\alpha, \beta \in \mathbb{R}$,

$$\int_E (\alpha f + \beta g) = \alpha \int_E f + \beta \int_E g \quad (10.47)$$

Moreover, if $f \leq g$ on E , then

$$\int_E f \leq \int_E g \quad (10.48)$$

Proof. Linear combination of bounded, measurable functions is measurable and bounded. Thus, if f, g are bounded, measurable functions, then $\alpha f + \beta g$ is also a bounded, measurable function. And $\alpha f + \beta g$ is integrable as every bounded, measurable function on E is integrable over E .

Step 1 : $\int_E \alpha f = \alpha \int_E f$

$$\begin{aligned} \int_E \alpha f &= \inf_{\psi \geq \alpha f} \left\{ \int_E \psi : \psi \text{ simple} \right\} \\ &= \alpha \inf_{\psi/\alpha \geq f} \left\{ \int_E \psi/\alpha : \psi/\alpha \text{ simple} \right\} \\ &= \alpha \inf_{\psi' \geq f} \left\{ \int_E \psi' : \psi' \text{ simple} \right\} \\ &= \alpha \int_E f \end{aligned}$$

Step 2 : $\int_E f + g = \int_E f + \int_E g$

Since $f \leq \psi_f$ and $g \leq \psi_g$, we have $f + g \leq \psi_f + \psi_g$ and $\psi_f + \psi_g$ is a simple function. And

$$\int_E f + g = \inf \left\{ \int_E \psi : \psi \text{ simple, } f + g \leq \psi \right\} \leq \int_E \psi_f + \psi_g$$

Since integral of simple functions have linearity

$$\int_E f + g \leq \int_E \psi_f + \psi_g = \int_E \psi_f + \int_E \psi_g$$

This inequality is true for simple functions dominating f, g . Thus, it is true for the infimum of a family of such simple functions.

$$\int_E f + g \leq \inf \left\{ \int_E \psi_f : \psi_f \text{ simple, } f \leq \psi_f \right\} + \inf \left\{ \int_E \psi_g : \psi_g \text{ simple, } g \leq \psi_g \right\}$$

Therefore,

$$\int_E f + g \leq \int_E f + \int_E g \quad (10.49)$$

Similarly,

$$\begin{aligned} \int_E f + g &= \sup \left\{ \int_E \varphi : \varphi \text{ simple, } \varphi \leq f + g \right\} \\ &\geq \int_E \varphi_f + \varphi_g = \int_E \varphi_f + \int_E \varphi_g \\ &\geq \sup \left\{ \int_E \varphi_f : \varphi_f \text{ simple, } \varphi_f \leq f \right\} + \sup \left\{ \int_E \varphi_g : \varphi_g \text{ simple, } \varphi_g \leq g \right\} \end{aligned}$$

Therefore,

$$\int_E f + g \geq \int_E f + \int_E g \quad (10.50)$$

Thus, for any bounded, measurable function f and g , $\int_E f + g = \int_E f + \int_E g$.

Therefore, $\int_E \alpha f + \beta g = \int_E \alpha f + \int_E \beta g = \alpha \int_E f + \beta \int_E g$. \square

Theorem 10.4.6 (additivity over the domain of integration). *Let f be a bounded, measurable function on a set E of finite measure. Suppose A, B are disjoint, measurable subsets of E . Then*

$$\int_{A \cup B} f = \int_A f + \int_B f \quad (10.51)$$

Proof. We have $\chi_{A \cup B} = \chi_A + \chi_B$. And,

$$f \cdot \chi_{A \cup B} = f \cdot (\chi_A + \chi_B) = f \cdot \chi_A + f \cdot \chi_B \quad (10.52)$$

And if $E_1 \subset E$, then by the definition of Lebesgue integral

$$\int_E f \cdot \chi_{E_1} = \inf \left\{ \int_E \psi : \psi \text{ simple, } f \cdot \chi_{E_1} \leq \psi \right\}$$

The collection of simple function ψ such that $f \cdot \chi_{E_1} \leq \psi$ on E is same as the collection of simple functions ψ such that $f \leq \psi$ on E_1 .

$$= \inf \left\{ \int_{E_1} \psi : \psi \text{ simple, } f \leq \psi \right\} = \int_{E_1} f$$

Since integral has linearity, we have

$$\int_{A \cup B} f = \int_E f \cdot \chi_{A \cup B} = \int_E (f \cdot \chi_A + f \cdot \chi_B) = \int_E f \cdot \chi_A + \int_E f \cdot \chi_B = \int_A f + \int_B f$$

□

Corollary 10.4.6.1. *Let f be a bounded, measurable function on a set of finite measure E . Then,*

$$\left| \int_E f \right| \leq \int_E |f| \quad (10.53)$$

Proof. We have, $-|f| \leq f \leq |f|$ on E . Therefore by linearity and monotonicity,

$$-\int_E |f| = \int_E -|f| \leq \int_E f \leq \int_E |f|$$

Therefore,

$$\left| \int_E f \right| \leq \int_E |f|$$

□

Theorem 10.4.7 (passage of limit under the integral sign). *Let $\{f_n\}$ be a sequence of bounded, measurable functions on a set E of finite measure. If $\{f_n\}$ converges to f uniformly on E , then the sequence of the integrals $\{\int_E f_n\}$ converges to the integral of the limit function $\int_E f$.*

$$\lim_{n \rightarrow \infty} \int_E f_n = \int_E f \quad (10.54)$$

Proof. Suppose a sequence $\{f_n\}$ of bounded, measurable functions converges uniformly to f on a set of finite measure E . Uniform limit of bounded functions is bounded. And limit function of measurable functions is measurable. Therefore, f is also a bounded, measurable function.

Let $\varepsilon > 0$. Choose an N such that $|f - f_n| \leq \varepsilon/m(E)$, $\forall n \geq N$. (such an N exists since the convergence is uniform)

By linearity and monotonicity,

$$\left| \int_E f - \int_E f_n \right| = \left| \int_E f - f_n \right| \leq \int_E |f - f_n| \leq \frac{\varepsilon}{m(E)} \cdot m(E) = \varepsilon$$

Clearly, $\left\{ \int_E f_n \right\}$ is sequence of functions converging pointwise to $\int_E f$. Therefore we have, $\lim_{n \rightarrow \infty} \int_E f_n = \int_E f$. □

Definitions 10.4.9 (passage of limit under integral sign). Suppose $\{f_n\}$ is a sequence of bounded, measurable functions converging to f pointwise (a.e.) on E . If

$$\lim_{n \rightarrow \infty} \int_E f_n = \int_E \lim_{n \rightarrow \infty} f_n = \int_E f \quad (10.55)$$

then this equality is the passage^{†18} of the limit under the integral sign.

Note : Pointwise convergence is not sufficient for passage of limit under integral sign. In other words, there exists a sequence $\{f_n\}$ of bounded, measurable functions such that

$$\lim_{n \rightarrow \infty} \int_0^1 f_n \neq \int_0^1 f$$

where $\{f_n\}$ converges to limit function f pointwise on $[0, 1]$.

Proof. Consider a family of functions defined by $f_n(\frac{1}{n}) = n$ and f_n increases from $f_n(0) = 0$ to $f_n(\frac{1}{n}) = n$ linearly and decreases from $f_n(\frac{1}{n}) = n$ to $f_n(\frac{2}{n}) = 0$ linearly. And vanishes for any $x \geq \frac{2}{n}$ on $[0, 1]$.

$$f_n(x) = \begin{cases} 0 & x \in [\frac{2}{n}, 1] \\ nx(2-n) & x \in [\frac{1}{n}, \frac{2}{n}] \\ n^2x & x \in [0, \frac{1}{n}] \end{cases}$$

We know that these functions are bounded by their construction as $0 \leq f_n \leq \frac{1}{n}$ for every $n \in \mathbb{N}$. Each f_n is continuous, and thus measurable. Clearly, $\{f_n\}$ is a sequence of bounded, measurable functions defined on $[0, 1]$.

$$\text{We have, } \lim_{n \rightarrow \infty} f_n = 0. \text{ And, } \lim_{n \rightarrow \infty} \int_0^1 f_n = 1 \neq 0 = \int_0^1 0. \quad \square$$

Note : However, the bounded convergence theorem gives us an additional constraint (uniformly pointwise bounded) for the passage of limit under integral sign.

Theorem 10.4.8 (bounded convergence). Let $\{f_n\}$ be a sequence of measurable, functions on a set of finite measure E . Suppose $\{f_n\}$ is uniformly pointwise bounded on E . If $\{f_n\}$ converges to f pointwise on E , then $\lim_{n \rightarrow \infty} \int_E f_n = \int_E f$.

Proof. Suppose sequence $\{f_n\}$ of bounded, measurable functions converges to f pointwise on E . Then the limit function f is also measurable. And since $\{f_n\}$ is uniformly pointwise bounded on E , f is bounded on E .

Let A be a measurable subset of E and $n \in \mathbb{N}$. By linearity and monotonicity,

$$\int_E f_n - \int_E f = \int_E f_n - f = \int_A f_n - f + \int_{E \sim A} f_n + \int_{E \sim A} -f$$

By monotonicity of integral,

$$\left| \int_E f_n - \int_E f \right| \leq \int_A |f_n - f| + 2M \cdot m(E \sim A) \quad (10.56)$$

¹⁸Passage of limit under integral' is used by I.P. Natanson, in his "Theory of functions of real variable"

Let $\varepsilon > 0$. Since E a set of finite measure and f is a real-valued function. By Egoroff's theorem, there exists a measurable subset $A \subset E$ on which $\{f_n\}$ converges to f uniformly and $m(E \setminus A) < \varepsilon/4M$. By uniform convergence on A , there exists $N \in \mathbb{N}$ such that

$$|f_n - f| < \frac{\varepsilon}{2 \cdot m(E)} \text{ on } A, \forall n \geq N$$

Thus, by monotonicity of integration we have,

$$\left| \int_E f_n - \int_E f \right| \leq \frac{\varepsilon}{2 \cdot m(E)} m(A) + 2M \cdot m(E \setminus A) < \varepsilon$$

Therefore, $\lim_{n \rightarrow \infty} \int_E f_n = \int_E f$. □

Exercise

- 1.
- 2.
- 3.
- 4.
- 5.
- 6.
- 7.
- 8.
- 9.
10. Let f be a bounded, measurable function on E . Let A be measurable subset of E . Then,

$$\begin{aligned} \int_E f \cdot \chi_A &= \sup \left\{ \int_E \psi : \psi \text{ simple, } 0 \leq \psi \leq f \cdot \chi_A \right\} \\ &= \sup \left\{ \int_A \psi : \psi \text{ simple, } 0 \leq \psi \leq f \right\} = \int_A f \end{aligned}$$

- 11.
- 12.
- 13.
- 14.
- 15.
- 16.

10.4.3 Lebesgue integral of a measurable non-negative function

Definitions 10.4.10 (vanishing). A function f on E vanishes outside $A \subset E$ if $f(x) = 0$ for every $x \in E \setminus A$.

Definitions 10.4.11 (support). Let f be a function on E . Then support of f is the set $\{x \in E : f(x) \neq 0\}$ of all non-vanishing points of f in E .

Warning : The definition of support is slightly different for measure spaces and topological spaces.^{†19}

Definitions 10.4.12 (finite support). A function f on E has finite support if support of f is of finite measure.

Definitions 10.4.13. Let f be a non-negative measurable function on E . Then the integral of f over E is the supremum of the integrals of bounded, measurable functions of finite support h such that $0 \leq h \leq f$ in E .

$$\int_E f = \sup \left\{ \int_E h : h \text{ bounded, measurable, finite support, } 0 \leq h \leq f \right\} \quad (10.57)$$

Theorem 10.4.9 (Chebychev's inequality). Let f be a non-negative, measurable function on E . Then for any $\lambda > 0$, we have

$$m\{x \in E : f(x) \geq \lambda\} \leq \frac{1}{\lambda} \cdot \int_E f \quad (10.58)$$

In other words $\lambda \cdot m(E_\lambda) \leq \int_E f$ where $E_\lambda = \{x \in E : f(x) \geq \lambda\}$.

Proof. Suppose f is a non-negative, measurable function on E . Let $\lambda > 0$ be a real number. Define $E_\lambda = \{x \in E : f(x) \geq \lambda\}$

Case 1 : $m(E_\lambda) = \infty$

Let $n \in \mathbb{N}$. Define $E_{\lambda,n} = E_\lambda \cap [-n, n]$ and $\psi_n = \lambda \cdot \chi_{E_{\lambda,n}}$. Clearly, ψ_n is a bounded, measurable function of finite support.

$$\lambda \cdot m(E_{\lambda,n}) = \int_E \psi_n \text{ and } 0 \leq \psi_n \leq f \text{ on } E \text{ for every } n$$

By continuity of measure we have,

$$\infty = \lambda \cdot m(E_\lambda) = \lambda \lim_{n \rightarrow \infty} m(E_{\lambda,n}) = \lim_{n \rightarrow \infty} \int_E \psi_n \leq \int_E f$$

Therefore,

$$\lambda \cdot m(E_\lambda) = \int_E f$$

¹⁹“Support of a function f on a topological space is the **closure** of the set of all non-vanishing points of f .” We use topological version for differential forms in multivariate real analysis and above definition for Lebesgue integration in integral transforms. Luckily, it is not apparent in our syllabus.

Case 2 : $m(E_\lambda) < \infty$

Define $h = \lambda \cdot \chi_{E_\lambda}$. Clearly, h is a bounded, measurable function of finite support and $0 \leq h \leq f$ on E . By definition of integral,

$$\lambda \cdot m(E_\lambda) = \int_E h \leq \int_E f$$

Therefore,

$$m(E_\lambda) = \frac{1}{\lambda} \int_E h \leq \frac{1}{\lambda} \int_E f$$

□

Theorem 10.4.10. *Let f be a non-negative measurable function on E . Then,*

$$\int_E f = 0 \iff f = 0 \text{ (a.e.) on } E \quad (10.59)$$

Proof. Part 1

Suppose $\int_E f = 0$. Then by Chebychev's inequality,

$$\forall n \in \mathbb{N}, m\left(\left\{x \in E : f(x) \geq \frac{1}{n}\right\}\right) \leq n \int_E f = 0$$

We have,

$$\{x \in E : f(x) > 0\} = \bigcup_{n \in \mathbb{N}} \left\{x \in E : f(x) \geq \frac{1}{n}\right\} = \emptyset$$

Part 2

Suppose $f = 0$ (a.e.) on E . Let φ be a simple function and h be a bounded, measurable function of finite support such that $0 \leq \varphi \leq h \leq f$.

We have, $f = 0$ (a.e.) on E , thus $\varphi = 0$ (a.e.) on E . Then $\varphi = 0$ (a.e.) on E , and we have $\int_E \varphi = 0$. Since $\int_E \varphi = 0$ for every simple function $\varphi \leq h$, we have $\int_E h = 0$. Again, $\int_E h = 0$ for every bounded, measurable function $h \leq f$ with finite support. Therefore, $\int_E f = 0$. □

Theorem 10.4.11 (linearity + monotonicity of integral of non-negative functions). *Let f, g be non-negative, measurable functions on E . Then for any real numbers $\alpha, \beta > 0$ ²⁰, we have*

$$\int_E (\alpha f + \beta g) = \alpha \int_E f + \beta \int_E g \quad (10.60)$$

And we have,

$$\text{If } f \leq g \text{ on } E, \text{ then } \int_E f \leq \int_E g \quad (10.61)$$

Proof. Step 1 : $\int_E \alpha f = \alpha \int_E f$

Suppose $\alpha > 0$. For any bounded, measurable function h with finite support,

²⁰We know that $\alpha f + \beta g$ is non-negative only when α and β are non-negative.

dominated by f , we have αh is also a bounded, measurable function with finite support such that $0 \leq \alpha h \leq \alpha f$ on E .

$$0 \leq h \leq f \iff 0 \leq \alpha h \leq \alpha f$$

Thus by the definition of Lebesgue integral for non-negative, measurable functions,

$$0 \leq \int_E \alpha h = \alpha \int_E h \leq \alpha \int_E f$$

Therefore, $\int_E \alpha f = \alpha \int_E f$, $\forall \alpha > 0$.

$$\textbf{Step 2 : } \int_E f + g = \int_E f + \int_E g$$

Suppose $0 \leq h \leq f$ and $0 \leq k \leq g$ on E . Then, $h + k$ is also a bounded, measurable function with finite support such that $0 \leq h + k \leq f + g$ on E .

$$0 \leq h + k \leq f + g \implies \int_E h + \int_E k = \int_E h + k \leq \int_E f + g \quad (10.62)$$

Let l be the supremum of all the bounded, measurable functions with finite support such that $0 \leq l \leq f + g$ on E . Then, it is enough to prove that,

$$\int_E f + g = \int_E l \leq \int_E f + \int_E g$$

Define $h = \min\{f, l\}$ and $k = l - h$ on E . Let $x \in E$.

If $l(x) \leq f(x)$, then $h(x) = l(x) \leq f(x)$ and $k(x) = h(x) - l(x) = 0 \leq g(x)$.

If $l(x) > f(x)$, then $l(x) < g(x)$ since $l \leq f + g$ on E .

Thus, $h(x) = f(x) < l(x) < g(x)$ and $k(x) = h(x) - l(x) < 0 < f(x)$. Clearly, h, k are bounded, measurable functions with finite support.

We have, $0 \leq h \leq f$, $0 \leq k \leq g$, and $l = h + k$ on E . Then,

$$\int_E l = \int_E h + k = \int_E h + \int_E k \leq \int_E f + \int_E g \quad (10.63)$$

Therefore, by Step 1

$$\int_E \alpha f + \beta g = \int_E \alpha f + \int_E \beta g = \alpha \int_E f + \beta \int_E g$$

Suppose $f \leq g$. Let h be any bounded, measurable function with finite support such that $0 \leq h \leq f$. Then $h \leq g$. Thus, for any bounded, measurable function h with finite support and $h \leq f$, we have $\int_E h \leq \int_E g$. Therefore,

$$\int_E f = \sup \left\{ \int_E h : \text{bounded, measurable, finite support, } h \leq f \right\} \leq \int_E g$$

□

Theorem 10.4.12 (additivity over domain of integration of non-negative functions). *Let f be a non-negative, measurable function on E . If A, B are disjoint subsets of E , then*

$$\int_{A \cup B} f = \int_A f + \int_B f \quad (10.64)$$

In particular, if $E_0 \subset E$ is of measure zero, then

$$\int_E f = \int_{E \sim E_0} f \quad (10.65)$$

Proof. Let f be a non-negative, measurable function defined on a measurable set E and A, B are two disjoint measurable subsets of E .

We have $\int_E f \cdot \chi_A = \int_A f$ and $f \cdot \chi_{A \cup B} = f \cdot (\chi_A + \chi_B) = f \cdot \chi_A + f \cdot \chi_B$. And if f is a bounded, measurable function with finite support, then $f \cdot \chi_A$ is also a bounded, measurable function with finite support for any measurable subset $A \subset E$. Therefore, by linearity of the Lebesgue integral of non-negative functions,

$$\int_{A \cup B} f = \int_E f \cdot \chi_{A \cup B} = \int_E f \cdot \chi_A + f \cdot \chi_B = \int_E f \cdot \chi_A + \int_E f \cdot \chi_B = \int_A f + \int_B f$$

Suppose E_0 is a subset of E of measure zero. Then, we have the **excision formula** by the additivity of integral over domain

$$\int_{(E \sim E_0) \cup E_0} f = \int_{E \sim E_0} f + \int_{E_0} f = \int_{E \sim E_0} f \quad (10.66)$$

since $m(E_0) = 0$ and $\int_{E_0} f = 0$. □

Fatou's Lemma and Passage of limit under Integral sign

Fatou's Lemma is a important tool for establishing passage of limit under integral sign.

Lemma 10.4.13 (Fatou). *Let $\{f_n\}$ be a sequence of non-negative, measurable functions on E . If $\{f_n\}$ converges to f pointwise (a.e.) on E , then*

$$\int_E f \leq \liminf \int_E f_n$$

Proof. Let $\{f_n\}$ be a sequence of functions converging to f pointwise a.e. on E . Let h be a bounded, measurable function with finite support and $0 \leq h \leq f$. Then $0 \leq h \leq M$, h has support, $E_0 = \{x \in E : h(x) \neq 0\}$ and $m(E_0) < \infty$.

Step 1 : Construction of h_n

Define $h_n = \min\{h, f_n\}$ on E . And $0 \leq h_n \leq M$. Then, h_n are measurable since each h_n is a bounded function defined on a set of finite measure. And h_n has finite support $E_n \subset E_0$.

Step 2 : Bounded Convergence Theorem

For each $x \in E$, $\{h_n(x)\}$ converges to $h(x)$ since $\{f_n(x)\}$ converges to $f(x)$. Since h_n are uniformly bounded, by Bounded Convergence theorem we have

$$\lim_{n \rightarrow \infty} \int_E h_n = \lim_{n \rightarrow \infty} \int_{E_0} h_n = \int_{E_0} h$$

However $h_n \leq f_n$. Therefore, $\int_E h_n \leq \int_E f_n \leq \liminf \int_E f_n$. \square

Theorem 10.4.14 (monotone convergence). *Let $\{f_n\}$ be an increasing sequence of non-negative, measurable functions on E . If $\{f_n\}$ converges to f pointwise (a.e.) on E , then $\lim_{n \rightarrow \infty} \int_E f_n = \int_E f$.*

Proof. Let $\{f_n\}$ be an increasing sequence of non-negative, measurable functions on E . Then, by Fatou's lemma

$$\int_E f \leq \liminf \int_E f_n \quad (10.67)$$

However, $\forall n \in \mathbb{N}$, $f_n \leq f$. Otherwise, $\{f_n\}$ won't converge to f . Thus, by monotonicity of Lebesgue integral for non-negative functions,

$$\forall n \in \mathbb{N}, \int_E f_n \leq \int_E f \implies \limsup \int_E f_n \leq \int_E f \quad (10.68)$$

Thus, limit inferior and limit superior are equal by sandwich lemma.

Therefore, the sequence $\left\{ \int_E f_n \right\}$ converges and $\int_E f = \lim_{n \rightarrow \infty} \int_E f_n$. \square

Corollary 10.4.14.1. *Let $\{u_n\}$ be a sequence of non-negative, measurable functions on E . If $f = \sum_{n=1}^{\infty} u_n$ pointwise (a.e.) on E , then $\int_E f = \sum_{n=1}^{\infty} \int_E u_n$.*

Proof. Let $\{u_n\}$ be a sequence of non-negative, measurable functions on E and $f = \sum_{n=1}^{\infty} u_n$. Define $f_n = \sum_{k=1}^n u_k$. Then $\{f_n\}$ is an increasing sequence of non-negative, measurable functions. And by linearity and monotone convergence theorem,

$$\lim_{n \rightarrow \infty} \sum_{k=1}^n \int_E u_k = \lim_{n \rightarrow \infty} \int_E \sum_{k=1}^n u_k = \lim_{n \rightarrow \infty} \int_E f_n = \int_E f$$

\square

Definitions 10.4.14 (integrable function). A non-negative, measurable function f on a measurable set E is integrable over E if $\int_E f < \infty$.

Theorem 10.4.15. *Let f be non-negative, measurable function which is integrable over E . Then f is finite a.e. on E .*

Proof. Suppose f is a non-negative measurable function defined on a measurable set E . And f is integrable over E . Then,

$$\int_E f = M < \infty$$

However by monotonicity and Chebychev's inequality,

$$m\{x \in E : f(x) = \infty\} \leq m\{x \in E : f(x) \geq n\} \leq \frac{1}{n} \int_E f = \frac{M}{n}, \forall n \in \mathbb{N}$$

As $n \rightarrow \infty$, we have $m\{x \in E : f(x) = \infty\} \rightarrow 0$.

Therefore, f is finite a.e. on E . \square

Theorem 10.4.16 (Beppo Levi's Lemma). *Let $\{f_n\}$ be an increasing sequence of non-negative, measurable functions on E . If the sequence of integrals $\left\{\int_E f_n\right\}$ are bounded, then $\{f_n\}$ converges pointwise on E to a measurable function f which is finite a.e. on E and*

$$\lim_{n \rightarrow \infty} \int_E f_n = \int_E f < \infty$$

Proof. Let $\{f_n\}$ be an increasing, sequence of non-negative, measurable functions on E . Since every monotone sequence of extended real numbers converges to an extended real number, $\{f_n\}$ converges to extended real valued function f on E . And by monotone convergence theorem, $\{\int_E f_n\}$ converges to $\int_E f$.

Suppose $\left\{\int_E f_n\right\}$ is a bounded sequence, then $\int_E f$ is also bounded. Thus, $\int_E f < \infty$. Therefore, f is finite a.e. on E , since f is integrable over E . \square

Exercise

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- 18.
- 19.
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- 25.
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- 27.

10.4.4 General Lebesgue integral

Theorem 10.4.17. *Let f be a measurable function on E . Then f^+, f^- are integrable over E if and only if $|f|$ is integrable over E .*

Proof. Let f be a measurable function on E . Suppose f^+, f^- are non-negative, integrable functions over E . Then $|f| = f^+ + f^-$ is integrable over E by linearity.

Suppose $|f|$ is a non-negative, integrable function over E . Then, we have $0 \leq f^+ \leq |f|$ and $0 \leq f^- \leq |f|$. Thus, by monotonicity $\int_E f^+ \leq \int_E |f| < \infty$. Similarly, $\int_E f^- < \infty$. Therefore, f^+, f^- are integrable over E . \square

Note : A measurable function f is integrable over E if $|f|$ is integrable over E .

Definitions 10.4.15 (general Lebesgue integral). Let f be a measurable function such that $|f|$ is integrable over E . Then integral of f over E is

$$\int_E f = \int_E f^+ - \int_E f^- \quad (10.69)$$

Theorem 10.4.18. *Let f be integrable over E . Then f is finite a.e. on E and*

$$\int_E f = \int_{E \sim E_0} f \quad (10.70)$$

where E_0 is a subset of E and is of measure zero.

Proof. Suppose f is an integrable function. Then $f = f^+ - f^-$ where f^+, f^- are non-negative, integrable functions. And $|f| = f^+ + f^-$ is a non-negative integrable function on E , thus $|f|$ is finite a.e. on E . Therefore f is finite a.e. on E .

Suppose $E_0 \subset E$ such that $m(E_0) = 0$. Then by linearity and additivity over domain of the integral of non-negative measurable functions,

$$\begin{aligned} \int_E f &= \int_E f^+ - f^- = \int_E f^+ + \int_E -f^- \\ &= \int_{E_0} f^+ + \int_{E \sim E_0} f^+ + \int_{E_0} -f^- + \int_{E \sim E_0} -f^- \\ &= \int_{E \sim E_0} f^+ + \int_{E \sim E_0} -f^- = \int_{E \sim E_0} f^+ - f^- = \int_{E \sim E_0} f \end{aligned}$$

\square

Theorem 10.4.19 (integral comparison test). *Let f be a measurable function on E . Suppose there exists a non-negative function g such that g integrable over E and g dominates f on E . Then f is integrable over E and*

$$\left| \int_E f \right| \leq \int_E |f| \quad (10.71)$$

Proof. Let f be a measurable function. And g be a non-negative, measurable function such that g is integrable over E and $|f| \leq g$ on E . Clearly, by monotonicity of integral $|f|$ is integrable over E . Therefore, f is also integrable over E .

Also we have,

$$\left| \int_E f \right| = \left| \int_E f^+ - \int_E f^- \right| = \left| \int_E f^+ - \int_E f^- \right|$$

By triangular inequality,

$$\begin{aligned} &\leq \left| \int_E f^+ \right| + \left| \int_E f^- \right| = \int_E f^+ + \int_E f^- \text{ since } 0 \leq \int_E f^+ \\ &\leq \int_E f^+ + \int_E f^- = \int_E |f| \end{aligned}$$

□

Linearity, Monotonicity, and Additivity over the domain of integration

Theorem 10.4.20 (linearity+monotonicity of general integral). *Let functions f, g be integrable over E . Then for any $\alpha, \beta \in \mathbb{R}$, function $\alpha f + \beta g$ is integrable over E and*

$$\int_E (\alpha f + \beta g) = \alpha \int_E f + \beta \int_E g \quad (10.72)$$

Moreover,

$$\text{If } f \leq g \text{ a.e. on } E, \text{ then } \int_E f \leq \int_E g \quad (10.73)$$

Proof. Step 1 : $\int_E \alpha f = \alpha \int_E f$

If $\alpha > 0$, then $[\alpha f]^+ = \alpha f^+$ and $[\alpha f]^- = \alpha f^-$.

If $\alpha < 0$, then $[\alpha f]^+ = -\alpha f^-$ and $[\alpha f]^- = -\alpha f^+$.

Thus,

$$\begin{aligned} \int_E \alpha f &= \int_E [\alpha f]^+ - \int_E [\alpha f]^- \\ &= \int_E \alpha f^+ - \int_E \alpha f^- = \alpha \int_E f^+ - \alpha \int_E f^- \\ &= \alpha \left(\int_E f^+ - \int_E f^- \right) = \alpha \int_E f \end{aligned}$$

Step 2 : $\int_E f + g = \int_E f + \int_E g$

We have,

$$[f + g]^+ - [f + g]^- = f + g = (f^+ - f^-) + (g^+ - g^-)$$

Therefore,

$$[f + g]^+ + f^- + g^- = [f + g]^- + f^+ + g^+ \quad (10.74)$$

And,

$$\int_E [f + g]^+ + f^- + g^- = \int_E [f + g]^- + f^+ + g^+$$

By linearity of non-negative, measurable functions,

$$\int_E [f + g]^+ + \int_E f^- + \int_E g^- = \int_E [f + g]^- + \int_E f^+ + \int_E g^+$$

Since all the terms are finite, rearranging the terms we get,

$$\int_E [f + g]^+ - \int_E [f + g]^- = \int_E f^+ - \int_E f^- + \int_E g^+ - \int_E g^-$$

Therefore,

$$\int_E f + g = \int_E f + \int_E g$$

Let f, g be measurable functions such that $f \leq g$ on E . Then $h = g - f$ is a non-negative, measurable function. And

$$\int_E h = \int_E g - f = \int_E g - \int_E f \geq 0$$

Therefore,

$$\int_E f \leq \int_E g$$

□

Theorem 10.4.21 (additivity over domain of integration). *Let function f be integrable over E . Let A, B are disjoint, measurable subsets of E . Then,*

$$\int_{A \cup B} f = \int_A f + \int_B f \quad (10.75)$$

Proof. We have, $f \cdot \chi_{A \cup B} = f \cdot \chi_A + f \cdot \chi_B$.

Also by monotonicity, $|f \cdot \chi_A| \leq |f| \implies \int_E |f \cdot \chi_A| \leq \int_E |f| < \infty$.

Thus, $|f \cdot \chi_A|$ and $|f \cdot \chi_B|$ are integrable.

$$\int_{A \cup B} f = \int_E f \cdot (\chi_{A \cup B}) = \int_E f \cdot \chi_A + \int_E f \cdot \chi_B = \int_A f + \int_B f$$

□

Sufficient Condition for Passage of limit under integral sign

Lebesgue dominated convergence theorems shows that **the existence of an integrable function or a sequence of functions (with integrable limit function) dominating every function in the sequence** is sufficient for the passage of limit under the integral sign.

Theorem 10.4.22 (Lebesgue dominated convergence). *Let $\{f_n\}$ be a sequence of measurable functions on E . Suppose there exists a function g which is integrable over E and dominates every function in the sequence $\{f_n\}$ on E . If $\{f_n\}$ converges to f pointwise a.e. on E , then f is integrable on E and*

$$\lim_{n \rightarrow \infty} \int_E f_n = \int_E f \quad (10.76)$$

Proof. Suppose $\{f_n\}$ is a sequence of measurable functions on E . And g is a function which is integrable over E and $|f_n| \leq g$ on E . Suppose $\{f_n\}$ converges to f pointwise on E . Clearly, $|f| \leq g$ a.e. on E . And by monotonicity, f, f_n are integrable over E . We know that, since f, f_n are integrable over E , they are finite a.e. on E .

We have, $g - f, g - f_n$ are non-negative, measurable functions on E . And $\{g - f_n\}$ converges to $g - f$ pointwise a.e. on E . By Fatou's lemma,

$$\begin{aligned} \int_E (g - f) &\leq \liminf \int_E (g - f_n) \\ \int_E g - \int_E f &\leq \int_E g - \limsup \int_E f_n \end{aligned}$$

Then,

$$\limsup \int_E f_n \leq \int_E f$$

Now consider, $\{g + f_n\}$ converging to $g + f$ pointwise a.e. on E . Again by Fatou's lemma,

$$\begin{aligned} \int_E (g + f) &\leq \liminf \int_E (g + f_n) \\ \int_E g + \int_E f &\leq \int_E g + \liminf \int_E f_n \end{aligned}$$

Then,

$$\int_E f \leq \liminf \int_E f_n$$

Therefore,

$$\int_E f = \lim_{n \rightarrow \infty} \int_E f_n$$

□

Theorem 10.4.23 (General Lebesgue dominated convergence). *Let $\{f_n\}$ be a sequence of measurable functions on E . And $\{f_n\}$ converges pointwise to f a.e. on E . Suppose there exists a sequence $\{g_n\}$ of non-negative, measurable functions on E that converges pointwise to g a.e. on E and g dominates every function in the sequence $\{f_n\}$.*

$$\text{If } \lim_{n \rightarrow \infty} \int_E g_n = \int_E g < \infty, \text{ then } \lim_{n \rightarrow \infty} \int_E f_n = \int_E f \quad (10.77)$$

Proof. Suppose $\{f_n\}$ is a sequence of measurable functions on E converging to f pointwise a.e. on E . Also suppose $\{g_n\}$ is another sequence of measurable functions on E converging to g pointwise a.e. on E such that $|f_n| \leq g_n$ and g is integrable on E . Clearly, $|f_n| \leq g_n \implies |f| \leq g$. And by monotonicity, f_n, f are integrable over E .

Consider, $\{g_n - f_n\}$ converging to $g - f$ pointwise a.e. on E . By Fatou's lemma,

$$\begin{aligned} \int_E (g - f) &\leq \liminf \int_E (g_n - f_n) \\ \int_E g - \int_E f &\leq \liminf \int_E g_n - \limsup \int_E f_n \dagger^{21} \end{aligned}$$

Then,

$$\limsup \int_E f_n \leq \int_E f$$

Now consider, $\{g_n + f_n\}$ converging to $g + f$ pointwise a.e. on E . Again, by Fatou's lemma,

$$\begin{aligned} \int_E (g + f) &\leq \liminf \int_E (g_n + f_n) \\ \int_E g + \int_E f &\leq \liminf \int_E g_n + \liminf \int_E f_n \end{aligned}$$

Then,

$$\int_E f \leq \liminf \int_E f_n$$

Therefore,

$$\int_E f = \lim_{n \rightarrow \infty} \int_E f_n$$

□

Exercise

- 28.
- 29.
- 30.
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- 36.

10.17 General Measure Spaces

Let X be any non-empty set. Let μ be any non-negative set function defined on a family of subsets of X . Then there exists a σ -algebra \mathcal{M} of measurable functions with respect to the outer measure μ^* induced by μ .

10.17.1 Measures and Measurable Sets

Definitions 10.17.1 (measure space). Let X be non-empty set and \mathcal{M} be a σ -algebra of subset of X . Then (X, \mathcal{M}) is a measure space. And $E \subset X$ is a measurable subset of X if $E \in \mathcal{M}$.

Definitions 10.17.2 (measure). A set function $\mu : \mathcal{M} \rightarrow [0, \infty]$ is a measure if $\mu(\emptyset) = 0$ and μ is countably additive^{†22}.

A few examples of measure space

1. $(\mathbb{R}, \mathcal{L}, m)$
The set of real numbers \mathbb{R} together with the σ -algebra \mathcal{L} of Lebesgue measurable functions and the Lebesgue measure m is a measure space.
2. $(\mathbb{R}, \mathcal{B}, m)$
The set of real numbers together with the σ -algebra \mathcal{B} of Borel sets and Lebesgue measure m is a measure space.
3. $(X, 2^X, \eta)$
A non-empty set X together with its powerset 2^X and counting measure, $\eta : \mathcal{M} \rightarrow [0, \infty)$ given by $\eta(A) = |A|$ is a measure space.
4. $(X, \mathcal{M}, \delta_{x_0})$
A non-empty set X together with a σ -algebra \mathcal{M} of subset of X and Dirac measure concentrated about $x_0 \in X$, $\delta_{x_0} : \mathcal{M} \rightarrow \{0, 1\}$ given by

$$\delta_{x_0}(A) = \begin{cases} 1 & x_0 \in A \\ 0 & x_0 \notin A \end{cases}$$

is a measure space.

5. (X, \mathcal{C}, μ)
A non-empty set X together with the family of all countable subsets and their complements, \mathcal{C} and measure $\mu : \mathcal{C} \rightarrow \{0, 1\}$ given by

$$\mu(A) = \begin{cases} 0 & A \text{ is countable} \\ 1 & A^c \text{ is countable} \end{cases}$$

is a measure space.

Note : Lebesgue measure offers an addition property, that is measure of an interval is its length.

²²We have dropped an axiom - "length of the interval is its measure."

Measure Subspace

Let (X, \mathcal{M}, μ) be a measure space. Let $X_0 \in \mathcal{M}$. Then $(X_0, \mathcal{M}_0, \mu_0)$ is also a measure space where \mathcal{M}_0 is \mathcal{M} restricted to X_0 and μ_0 is μ restricted to X_0 .

For example, let $I = [0, 1]$. Clearly, $I \in \mathcal{L}$ since I is a Borel set, and measurable. Then $(I, \mathcal{L}|_I, m|_I)$ is a measure space on the closed interval I obtained from the Lebesgue measure Space $(\mathbb{R}, \mathcal{L}, m)$.

Properties : Additivity, Monotonicity, Excision

1. Finite Additivity

For any finite, disjoint collection of measurable sets $\{E_k\}_{k=1}^n$, we have

$$\mu \left(\bigcup_{k=1}^n E_k \right) = \sum_{k=1}^n \mu(E_k)$$

Proof. Countably additive implies finitely additive. \square

2. Monotonicity

Suppose A, B are measurable subsets and $A \subset B$, then $\mu(A) \leq \mu(B)$.

Proof. $\mu(B) = \mu(A) + \mu(B \sim A) \implies \mu(B) \geq \mu(A)$, since $\mu(B \sim A) \geq 0$. \square

3. Excision

Suppose A, B are measurable subsets, $A \subset B$ and $\mu(A) < \infty$, then $\mu(B \sim A) = \mu(B) - \mu(A)$. And if $\mu(A) = 0$, then $\mu(B \sim A) = \mu(B)$.

Proof. $\mu(B) = \mu(A) + \mu(B \sim A) \implies \mu(B \sim A) = \mu(B) - \mu(A)$. \square

4. Countable Monotonicity

Suppose a countable collection of measurable sets $\{E_k\}_{k=1}^\infty$ covers another measurable sets E . Then, $\mu(E) \leq \sum_{k=1}^\infty \mu(E_k)$.

Proof. Suppose $\{E_k\}_{k=1}^\infty$ is a countable collection of measurable sets covering E . Then, we can construct $\{F_k\}_{k=1}^\infty$ is a countable collection of disjoint, measurable sets where $F_k = E_k \sim \bigcup_{n=1}^{k-1} E_n$.

Clearly, $\bigcup_{k=1}^\infty E_k = \bigcup_{k=1}^\infty F_k$ and $\forall k, \mu(F_k) \leq \mu(E_k)$ since $F_k \subset E_k$.

Therefore $\mu(E) \leq \mu \left(\bigcup_{k=1}^\infty E_k \right) = \mu \left(\bigcup_{k=1}^\infty F_k \right) = \sum_{k=1}^\infty \mu(F_k) \leq \sum_{k=1}^\infty \mu(E_k)$. \square

Properties : Continuity of Measure

1. If $\{A_k\}_{k=1}^{\infty}$ is an ascending sequence of measurable sets, then

$$\mu\left(\bigcup_{k=1}^{\infty} A_k\right) = \lim_{k \rightarrow \infty} \mu(A_k)$$

2. If $\{B_k\}_{k=1}^{\infty}$ is a descending sequence of measurable sets and $\mu(B_1) < \infty$, then

$$\mu\left(\bigcap_{k=1}^{\infty} B_k\right) = \lim_{k \rightarrow \infty} \mu(B_k)$$

Proof. Refer proof on page 255 □

Properties : Borel-Cantelli Lemma

Let (X, \mathcal{M}, μ) be a measure space. And $\{E_k\}_{k=1}^{\infty}$ a countable collection of measurable sets such that $\sum_{k=1}^{\infty} \mu(E_k) < \infty$. Then almost all $x \in X$ belongs to at most a finite number of E_k 's.

Proof. Refer proof on page 256 □

 σ -finite Measure

Definitions 10.17.3 (σ -finite). Let (X, \mathcal{M}, μ) be a measure space. Measure μ is finite if $\mu(X) < \infty$. Measure μ is σ -finite if X is a countable union of measurable sets of finite measure. A measurable subset E is of finite measure if $\mu(E) < \infty$. And E is σ -finite if E is a countable union of measurable sets of finite measure.

For example, Lebesgue measure is σ -finite since $\mathbb{R} = \bigcup_{n \in \mathbb{Z}} [n, n+1]$. But, counting measure on $[0, 1]$ is not σ -finite since $[0, 1]$ is uncountably infinite and countable union of finite subsets is countable.

Complete Measure Space

Definitions 10.17.4 (complete measure space). Let (X, \mathcal{M}, μ) be a measure space. Then the measure space is complete if every subset of sets of measure zero are measurable.

Note : $(\mathbb{R}, \mathcal{L}, m)$ is complete, however, $(\mathbb{R}, \mathcal{B}, m)$ is not complete since cantor set C of measure zero has a non-borel subset $\psi^{-1}(W)$ where W is a choice set of $\psi(C)$ under rational equivalence.

Theorem 10.17.1 (completion). Let (X, \mathcal{M}, μ) be a measure space. Let $\mathcal{M}_0 = \{A \cup B : A \subset C \in \mathcal{M}, \mu(C) = 0, B \in \mathcal{M}\}$. And $\mu_0(E) = \mu(B)$ where $E = A \cup B$. Then, $(X, \mathcal{M}_0, \mu_0)$ is a completion of (X, \mathcal{M}, μ) .

Proof. Let (X, \mathcal{M}, μ) be a measure space.

Step 1 : \mathcal{M}_0 is a σ -algebra containing \mathcal{M}

Define a family of subsets of X , \mathcal{M}_0 in such a way that a subset $E \in \mathcal{M}_0$ if it can be expressed as union of a subset C such that $\mu(C) = 0$ and a set $B \in \mathcal{M}$. Clearly, $E \in \mathcal{M} \implies E = E \cup \phi \in \mathcal{M}_0$. Thus, $\mathcal{M} \subset \mathcal{M}_0$.

Suppose $E = A \cup B$, then

$$X \sim E = X \sim (A \cup B) = (X \sim D) \cup (X \sim (C \cup B))$$

where $D = C \sim (A \cup B)$. Clearly $D \subset C$ and $C \cup B$ is a union of measurable sets and therefore measurable. Therefore, for any set $E \in \mathcal{M}_0$, we have $X \sim E \in \mathcal{M}_0$.

Suppose $\{E_k\}$ be a collection of sets in \mathcal{M}_0 . Let $E_k = A_k \cup B_k$ where A_k 's are subsets of some set C_k such that $\mu(C_k) = 0$. Clearly, $\mu(C) = \mu(\bigcup_{k=1}^{\infty} C_k) = 0$. And $A = \bigcup_{k=1}^{\infty} A_k$ is a subset of C . Also $B = \bigcup_{k=1}^{\infty} B_k \in \mathcal{M}$ since \mathcal{M} is a σ -algebra. Therefore,

$$E = \bigcup_{k=1}^{\infty} E_k = \bigcup_{k=1}^{\infty} (A_k \cup B_k) = \bigcup_{k=1}^{\infty} A_k \cup \bigcup_{k=1}^{\infty} B_k = A \cup B \in \mathcal{M}_0$$

Therefore, by de Morgan's law, \mathcal{M}_0 is a σ -algebra.

Step 2 : μ_0 is well-defined

Let $E \in \mathcal{M}_0$. We have $E = A \cup B$ where A is a subset C such that $\mu(C) = 0$, $B \in \mathcal{M}$ and $\mu_0(E) = \mu(B)$. Suppose there exists A', B' such that $E = A' \cup B'$ satisfying the conditions. Then $\mu(B) = \mu(B')$. Suppose $\mu(B) < \mu(B')$, without loss of generality. Then $\mu(B' \sim B) > 0 \implies 0 < \mu(A \sim A') \leq \mu(C) = 0$ which is a contradiction. Therefore, $\mu_0(E)$ is well-defined irrespective of the representation $E = A \cup B$.

Step 3 : $(X, \mathcal{M}_0, \mu_0)$ is complete

Suppose A is a subset of some set C such that $\mu(C) = 0$. Then there exists $E \in \mathcal{M}_0$ such that $E = A \cup \phi$ where $\phi \in \mathcal{M}$. Therefore, $(X, \mathcal{M}_0, \mu_0)$ is complete. \square

Exercise

Definitions 10.17.5 (semi-finite). Measure μ is semi-finite if every subset of infinite measure contains measurable subsets of arbitrarily large finite measure.

Definitions 10.17.6 (locally measurable). A subset E is locally measurable if $B \cap E$ is measurable for any subset B of finite measure.

Definitions 10.17.7 (saturated). A measure space is saturated if every locally measurable subset is measurable.

- 1.
- 2.

- 3.
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- 6.
7. (a) If μ, ν are measures on (X, \mathcal{M}) , then $\lambda = \mu + \nu$ is also a measure on (X, \mathcal{M}) .
 (b) Suppose ν, μ are measures on (X, \mathcal{M}) . If $\nu \leq \mu$, then there exists measure λ such that $\nu + \lambda = \mu$.
 (c) Suppose $\nu \leq \mu$ and μ is σ -finite, then λ is unique.
 (d) Suppose $\nu \leq \mu$. Then there exists a minimal λ such that $\nu + \lambda = \mu$.
8. (a) Every σ -finite measure is semi-finite.
 (b) Any measure $\mu = \mu_1 + \mu_2$ where measure μ_1 is semifinite and measure μ_2 assumes either 0 or ∞ .
9. The completion $(X, \mathcal{M}_0, \mu_0)$ of (X, \mathcal{M}, μ) is minimal with respect to ?
10. (a) Every σ -finite measure is saturated.
 (b) Locally measurable sets is a σ -algebra.
 (c) $(X, 2^X, \bar{\mu})$ is a saturation of (X, \mathcal{M}, μ) where $\bar{\mu}(E) = \infty$ if $E \notin \mathcal{M}$.
 (d) $(X, \mathcal{C}, \bar{\mu})$ is a saturation of (X, \mathcal{M}, μ) where \mathcal{C} is the σ -algebra of locally measurable sets and $\bar{\mu}(E) = \sup\{\mu(B) : B \subset E, B \in \mathcal{M}\}$
11. If μ, η are measures on (X, \mathcal{M}) , then $\max\{\mu, \eta\}$ is also a measure on \mathcal{M} .

10.17.2 Signed measures

Definitions 10.17.8 (signed measure). A set function $\nu : \mathcal{M} \rightarrow [-\infty, \infty]$ is a signed measure on X, \mathcal{M} if

1. ν assumes at most one of the infinities
2. $\nu(\emptyset) = 0$ and
3. If ν is countably additive for a disjoint collection of measurable sets $\{E_k\}_{k=1}^\infty$ and $\nu(\cup_{k=1}^\infty E_k) < \infty$, then the series $\sum_{k=1}^\infty \nu(E_k)$ is absolutely convergent.

Note : Difference of two measures, one of which is finite is always a signed measure. And we later prove that every signed measure is of this form.

Definitions 10.17.9. Let ν be a signed measure.

positive set A set is positive with respect to ν if it is measurable and every measurable subset $E \subset A$ has positive measure.

negative set A set is negative with respect to ν if it is measurable and every measurable subset of it has negative measure.

null (measure) set A set is null with respect to ν if it is measurable and every measurable subset of it has measure zero.

Note : Signed measure restricted to any positive set is a measure.

Theorem 10.17.2. *Let ν be a signed measure on (X, \mathcal{M}) . Then every measurable subset of a positive set is positive and union of countable collection of positive sets is positive.*

Proof. Let $\{A_k\}_{k=1}^{\infty}$ be countable collection of positive sets. Let $A = \bigcup_{k=1}^{\infty} A_k$ and E be a measurable subset of A . Consider the disjoint collection of countable subsets, $\{E_k\}_{k=1}^{\infty}$ where

$$E_k = (E \cap A_k) \sim \bigcup_{k=1}^{n-1} E_k$$

Clearly $\bigcup_{k=1}^{\infty} E_k = E$ and $\nu(E) = \sum_{k=1}^{\infty} \nu(E_k) \geq 0$. Since $E \subset A$ is arbitrary, every measurable subset of A is of non-negative measure. Therefore, union of positive sets is also a positive set. \square

Lemma 10.17.3 (Hahn). *Let ν be a signed measure on (X, \mathcal{M}) . And E is a measurable set of finite, positive measure. Then there exists subset $A \subset E$ such that A is positive and is of positive measure.*

Proof. Suppose E is not a positive set, otherwise the proof is complete. Let m_1 be the smallest natural number²³ such that there exists a measurable set with measure less than $-\frac{1}{m}$. Let $E_1 \subset E$ be a measurable subset with $\nu(E_1) < -\frac{1}{m_1}$.

Let n be a natural number and m_1, m_2, \dots, m_n are natural numbers where m_k 's are the smallest natural numbers such that there exists a measurable subset E_k ,

$$E_k \subset E \sim \bigcup_{j=1}^{k-1} E_j \text{ and } \nu(E_k) < -\frac{1}{m_k}$$

If this sequence terminates, then $E \sim \bigcup_{j=1}^n E_j$ is a positive set with positive

measure. Suppose the sequence does not terminate. Define $A = E \sim \bigcup_{j=1}^{\infty} E_j$.

Then A is a positive set by construction. Suppose A has a measurable subset B , then $\nu(B) \geq \lim_{k \rightarrow \infty} -\frac{1}{m_k} = 0$. And measure of A is positive, since we have

$$E = A \cup \bigcup_{j=1}^{\infty} E_j \text{ and } 0 < \nu(E) = \nu(A) + \sum_{j=1}^{\infty} \nu(E_j) \text{ where } \nu(E_j) \text{'s are all negative.}$$

Therefore, every measurable set E with positive measure has a positive subset A with positive measure. \square

²³Since E has measurable subsets of negative measure and negative measures are bounded above by 0, there exists a least upper bound $-M < -\frac{1}{m}$ for a natural number m .

Theorem 10.17.4 (Hahn Decomposition). *Let ν be a signed measure on (X, \mathcal{M}) . Then there exists a positive set A and a negative set B for which $X = A \cup B$ and $A \cap B = \phi$. And we write $[A, B] = X$.*

Proof. Let (X, \mathcal{M}) be measure space with signed measure ν . Let \mathcal{P} be the collection of positive subset of X . Let $\lambda = \sup\{\nu(E) : E \in \mathcal{P}\}$. And $\lambda \geq 0$ since ϕ is trivially a positive set in \mathcal{P} with measure zero.

Let $\{A_n\}$ be a countable collection of positive subsets of X such that $\lim_{n \rightarrow \infty} \nu(A_k) =$

λ . Define $A = \bigcup_{n=1}^{\infty} A_n$. Then A is also a positive set and $\nu(A) \leq \sup\{\nu(E) : E \in \mathcal{P}\} = \lambda$ since $A \in \mathcal{P}$. Also for each k , $A \sim A_k \subset A$ and $\nu(A) = \nu(A \sim A_k) + \nu(A_k) \geq \nu(A_k)$. This inequality is true for every k , and for the limit as well. Thus, $\nu(A) \geq \lambda$.

Let $B = X \sim A$. Then B is a negative set by construction as $X \sim A$ doesn't have any measurable subset of positive measure. Suppose E be a subset of B with positive measure. Then $\nu(A \cap E) = \nu(A) + \nu(E) < \nu(A) = \lambda$ which is a contradiction. \square

Definitions 10.17.10 (singular measures). Two signed measures ν_1, ν_2 are singular if there exists disjoint subsets $A, B \subset X$ such that $X = A \cup B$ and $\nu_1(A) = \nu_2(B) = 0$. If ν_1, ν_2 are singular, then we write $\nu_1 \perp \nu_2$.

Theorem 10.17.5 (Jordan-Hahn Decomposition). *Let ν be a signed measure on (X, \mathcal{M}) . Then there two mutually singular measures ν^+, ν^- on (X, \mathcal{M}) such that $\nu = \nu^+ - \nu^-$. Moreover, there is only one such pair of mutually singular measures.*

Proof. Let (X, \mathcal{M}) be a measure space with signed measure ν . Then by Hahn decomposition theorem, X has a partition $X = A \cup B$, $A \cap B = \phi$ such that A, B are positive and negative sets $[\nu]$ respectively.

Define $\nu^+(E) = \nu(E \cap A)$ and $\nu^-(E) = -\nu(E \cap B)$. Then ν^+, ν^- are measure and $\nu = \nu^+ - \nu^-$. Also, $\nu^+(B) = \nu(B \cap A) = \nu(\phi) = 0$ and $\nu^-(A) = \nu(A \cap B) = \nu(\phi) = 0$. \square

Exercise

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10.17.3 Caratheodory measure

Definitions 10.17.11 (outer measure). A set function $\mu^* : 2^X \rightarrow [0, \infty]$ is an outer measure if $\mu^*(\emptyset) = 0$. and μ^* is countably monotone.^{†24}

Definitions 10.17.12 (measurable set). A set E is measurable if for any subset A of X , we have

$$\mu^*(A) = \mu^*(A \cap E) + \mu^*(A \cap E^c)$$

Since μ^* is monotone, we can use simpler condition

$$\mu^*(A) \geq \mu^*(A \cap E) + \mu^*(A \cap E^c)$$

Theorem 10.17.6. *Union of finite collection of measurable sets is measurable.*

Proof. Refer proof on page 245 □

Theorem 10.17.7. *Let $A \subset X$ and $\{E_k\}_{k=1}^n$ be a finite, disjoint collection of measurable sets. Then*

$$\mu^*\left(A \cap \bigcup_{k=1}^n E_k\right) = \sum_{k=1}^n \mu^*(A \cap E_k)$$

In particular, the restriction of μ^ to the finite collection of measurable sets is finitely additive.*

Proof. —yet to complete— □

Theorem 10.17.8. *Union of countable collection of measurable sets is measurable.*

Proof. Refer proof on page 247 □

Theorem 10.17.9. *Let μ^* be an outer measure on 2^X . Then the collection \mathcal{M} of measurable subsets is a σ -algebra. If $\bar{\mu}$ is the restriction of μ^* to \mathcal{M} , then $(X, \mathcal{M}, \bar{\mu})$ is a complete measure space.*

Proof. —yet to complete— □

10.18 Integration over General measure Spaces

10.18.1 Measurable Functions

Theorem 10.18.1. *Let (X, \mathcal{M}) be a measure space. Let function f be defined on X . Then the following statements are equivalent*

1. $\forall c \in \mathbb{R}, \{x \in X : f(x) < c\}$ is measurable.
2. $\forall c \in \mathbb{R}, \{x \in X : f(x) \leq c\}$ is measurable.
3. $\forall c \in \mathbb{R}, \{x \in X : f(x) > c\}$ is measurable.
4. $\forall c \in \mathbb{R}, \{x \in X : f(x) \geq c\}$ is measurable.

²⁴Further simplifying the axiom countable additivity into countable monotonicity.

Each of the above statements imply that $\forall c \in \mathbb{R}$, $\{x \in X : f(x) = c\}$ is measurable.

Proof. Refer proof on page 262 □

Definitions 10.18.1 (measurable function). Let (X, \mathcal{M}) be a measure space. A function f on X is measurable if one of the equivalent statements above and hence all are true.

Theorem 10.18.2. Let (X, \mathcal{M}) be a measure space and function f be defined on X . Then f is measurable if and only if for every open set \mathcal{O} , $f^{-1}(\mathcal{O})$ is open.

Proof. Refer proof on page 263 □

Theorem 10.18.3. Let (X, \mathcal{M}) be a complete measure space. And X_0 is a measurable subset of X for which $\mu(X \setminus X_0) = 0$. Then f is measurable if and only if its restriction to X_0 is measurable. In particular, let two functions $g = h$ a.e. on X , g is measurable if and only if h is measurable.

Proof. Refer proof on page 264 □

Theorem 10.18.4. Let (X, \mathcal{M}) be a measure space. And f, g are measurable functions on X .

1. Linearity : $\alpha f + \beta g$ is measurable $\forall \alpha, \beta \in \mathbb{R}$.
2. Products : $f \cdot g$ is measurable.
3. minimum/maximum : $\max\{f, g\}, \min\{f, g\}$ are measurable.

Proof. □

Theorem 10.18.5. Let (X, \mathcal{M}) be a measure space. Let f be continuous function and g be measurable, then $f \circ g$ is measurable.

Proof. □

Theorem 10.18.6. Let (X, \mathcal{M}, μ) be a measure space. Let $\{f_n\}$ be a sequence of measurable functions on X for which $\{f_n\}$ converges to f pointwise a.e. on X . If either (X, \mathcal{M}, μ) is complete or convergence is pointwise on X , then f is measurable.

Proof. □

Corollary 10.18.6.1. Let (X, \mathcal{M}) be a measure space. Let $\{f_n\}$ be a sequence of measurable functions on X . Then $\sup\{f_n\}, \inf\{f_n\}, \limsup\{f_n\}, \liminf\{f_n\}$ are measurable.

Proof. □

10.18.2 Integration of non-negative measurable functions

Definitions 10.18.2. Let (X, \mathcal{M}, μ) be a measure space. And ψ be a non-negative simple function on X . Then

$$\int_X \psi \, d\mu = \sum_{k=1}^n c_k \cdot \mu(E_k) \text{ where } \psi = \sum_{k=1}^n c_k \cdot \chi_{E_k} \quad (10.78)$$

And for $E \subset X$, we have

$$\int_E \psi \, d\mu = \int_X \psi \cdot \chi_E \, d\mu = \sum_{k=1}^n c_k \cdot \mu(E_k \cap E) \text{ where } \psi = \sum_{k=1}^n c_k \cdot \chi_{E_k} \quad (10.79)$$

Theorem 10.18.7. Let (X, \mathcal{M}, μ) be a measure space. Let ψ, φ be non-negative simple functions.

1. *Linearity*

Let α, β be positive real numbers.

$$\int_X (\alpha\psi + \beta\varphi) \, d\mu = \alpha \int_X \psi \, d\mu + \beta \int_X \varphi \, d\mu \quad (10.80)$$

2. *Additivity over domain of integration*

Let A, B be disjoint measurable subsets of X

$$\int_{A \cup B} \psi \, d\mu = \int_A \psi \, d\mu + \int_B \psi \, d\mu \quad (10.81)$$

3. *Suppose $X_0 \subset X$ such that $\mu(X \setminus X_0) = 0$, then*

$$\int_X \psi \, d\mu = \int_{X_0} \psi \, d\mu \quad (10.82)$$

4. *Monotonicity*

Suppose $\psi \leq \varphi$ a.e. on X , then

$$\int_X \psi \, d\mu \leq \int_X \varphi \, d\mu \quad (10.83)$$

Proof.

□

Definitions 10.18.3. Let (X, \mathcal{M}, μ) be a measure space. And f be a non-negative, extended real-valued, measurable function on X . Then

$$\int_X f \, d\mu = \sup \left\{ \int_X \varphi \, d\mu : \varphi \text{ simple, } 0 \leq \varphi \leq f \right\} \quad (10.84)$$

And if $E \subset X$, then

$$\int_E f \, d\mu = \int_X f \cdot \chi_E \, d\mu = \sup \left\{ \int_X \varphi \cdot \chi_E \, d\mu : \varphi \text{ simple, } 0 \leq \varphi \leq f \text{ on } E \right\} \quad (10.85)$$

Theorem 10.18.8. *Let (X, \mathcal{M}, μ) be a measure space. Let g, h be non-negative, measurable functions. Then*

1. *Linearity (Part 1)*

Let α be a positive real number.

$$\int_X \alpha f \, d\mu = \alpha \int_X f \, d\mu \quad (10.86)$$

2. *Suppose $X_0 \subset X$ such that $\mu(X \setminus X_0) = 0$, then*

$$\int_X f \, d\mu = \int_{X_0} f \, d\mu \quad (10.87)$$

3. *Monotonicity*

Suppose $f \leq g$ a.e. on X , then

$$\int_X f \, d\mu \leq \int_X g \, d\mu \quad (10.88)$$

Proof.

□

Theorem 10.18.9 (Chebychev's Inequality). *Let (X, \mathcal{M}, μ) be a measure space. Let f be a non-negative measurable function on X . Let $\lambda \in \mathbb{R}$.*

$$\mu\{x \in X : f(x) \geq \lambda\} \leq \frac{1}{\lambda} \int_X f \, d\mu$$

Proof. Refer proof on page 280

□

Theorem 10.18.10. *Let (X, \mathcal{M}, μ) be a measure space. Let f be a non-negative measurable function on X for which $\int_X f \, d\mu < \infty$. Then f is finite a.e. on X and $\{x \in X : f(x) \geq \frac{1}{n}\}$ is σ -finite.*

Proof.

□

Lemma 10.18.11 (Fatou). *Let (X, \mathcal{M}, μ) be a measure space. Let $\{f_n\}$ be a sequence of non-negative, measurable functions converging to f pointwise a.e. on X . If f is measurable, then*

$$\int_X f \, d\mu \leq \liminf \int_X f_n \, d\mu$$

Proof.

□

Theorem 10.18.12 (monotone convergence). *Let (X, \mathcal{M}, μ) be a measure space. And $\{f_n\}$ be an increasing sequence of non-negative measurable functions converging to f pointwise on X . Then*

$$\lim_{n \rightarrow \infty} \int_X f_n \, d\mu = \int_X f \, d\mu$$

Proof.

□

Lemma 10.18.13 (Beppo-Levi). *Let (X, \mathcal{M}, μ) be a measure space. Let $\{f_n\}$ be an increasing sequence of non-negative measurable functions on X . If the sequence of integrals $\left\{\int_X f_n d\mu\right\}$ is bounded, then $\{f_n\}$ converges to a function f pointwise on X which is measurable, finite a.e. on X and*

$$\lim_{n \rightarrow \infty} \int_X f_n d\mu = \int_X f d\mu < \infty$$

Proof. □

Theorem 10.18.14 (simple approximation). *Let (X, \mathcal{M}, μ) be a measure space. Let f be a non-negative measurable function on X . Then there exists an increasing sequence $\{\psi_n\}$ of simple functions on X that converges pointwise on X and*

$$\lim_{n \rightarrow \infty} \int_X \psi_n d\mu = \int_X f d\mu$$

Proof. Refer proof on page 269 □

Theorem 10.18.15. *Let (X, \mathcal{M}, μ) be a measure space. Let f, g be non-negative, measurable functions on X . If $\alpha, \beta > 0$, then*

$$\int_X (\alpha f + \beta g) d\mu = \alpha \int_X f d\mu + \beta \int_X g d\mu$$

Proof. □

Theorem 10.18.16. *Let (X, \mathcal{M}, μ) be a measure space. Let f be a non-negative, measurable function on X . Then there exists an increasing sequence of simple functions $\{\psi_n\}$ that converges to f pointwise on X and*

$$\lim_{n \rightarrow \infty} \int_X \psi_n d\mu = \int_X f d\mu$$

Proof. □

Theorem 10.18.17. *Let (X, \mathcal{M}, μ) be a measure space. Let f, g be non-negative measurable functions on X . If $\alpha, \beta > 0$, then*

$$\int_E (\alpha f + \beta g) d\mu = \alpha \int_X f d\mu + \beta \int_X g d\mu$$

Proof. □

Definitions 10.18.4 (integrable). *Let (X, \mathcal{M}, μ) be a measure space. A non-negative, measurable function f is integrable over X with respect to μ if $\int_X f d\mu < \infty$.*

Exercise

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- 26.

10.18.3 Integration of general measurable functions

Definitions 10.18.5. Let (X, \mathcal{M}, μ) be a measure space. Let f be a measurable function on X . Then f be an integrable function over X with respect to μ , if $|f|$ is integrable over X with respect to μ . And

$$\int_X f \, d\mu = \int_X f^+ \, d\mu - \int_X f^- \, d\mu$$

Theorem 10.18.18 (integral comparison test). *Let (X, \mathcal{M}, μ) be a measure space. Let f be a measurable function on X . If g is integrable over X and g dominates f on X , ie, $|f| \leq g$. Then f is integrable over X and*

$$\left| \int_X f \, d\mu \right| \leq \int_X |f| \, d\mu \leq \int_X g \, d\mu$$

Proof.

□

Theorem 10.18.19. *Let (X, \mathcal{M}, μ) be a measure space. Let f, g be integrable over X with respect to μ .*

1. *Linearity*

$$\int_X (\alpha f + \beta g) \, d\mu = \alpha \int_X f \, d\mu + \beta \int_X g \, d\mu$$

2. *Monotonicity*

If $f \leq g$ a.e. on X , then

$$\int_X f \, d\mu \leq \int_X g \, d\mu$$

3. Additivity over domain of integration

Let A, B be disjoint, measurable subsets of X . Then

$$\int_{A \cup B} f \, d\mu = \int_A f \, d\mu + \int_B f \, d\mu$$

Proof. □

Theorem 10.18.20 (countably additivity over the domain of integration). *Let (X, \mathcal{M}, μ) be a measure space. Let f be a function integrable over X with respect to μ . Let $\{X_n\}_{n=1}^\infty$ be a disjoint, collection of measurable sets whose union is X . Then*

$$\int_X f \, d\mu = \sum_{n=1}^\infty \int_{X_n} f \, d\mu$$

Proof. □

Theorem 10.18.21 (continuity of integration). *Let (X, \mathcal{M}, μ) be a measure space. Let f be a function integrable over X with respect to μ .*

1. *If $\{X_n\}_{n=1}^\infty$ is an ascending collection of measurable sets whose union is X . Then*

$$\int_X f \, d\mu = \lim_{n \rightarrow \infty} \int_{X_n} f \, d\mu$$

2. *If $\{X_n\}_{n=1}^\infty$ is a descending collection of measurable sets, then*

$$\int_{\bigcap_{n=1}^\infty X_n} f \, d\mu = \lim_{n \rightarrow \infty} \int_{X_n} f \, d\mu$$

Proof. □

Theorem 10.18.22. *Let (X, \mathcal{M}, μ) be a measure space. Let f be a measurable function on X . If f is bounded and vanishes outside a set of finite measure, then f is integrable over X .*

Proof. □

Corollary 10.18.22.1. *Let X be a compact, topological space. Let \mathcal{M} be a σ -algebra containing the topology. If f is continuous, real-valued function on X and (X, \mathcal{M}, μ) is a finite measure space, then f is integrable over X with respect to μ .*

Proof. □

Theorem 10.18.23 (Lebesgue Dominated Convergence). *Let (X, \mathcal{M}, μ) be a measure space. Let $\{f_n\}$ be a sequence of measurable functions converging to a measurable function f pointwise a.e. on X . Let g be an integrable, non-negative function dominating the sequence $\{f_n\}$. Then f is integrable over X and,*

$$\lim_{n \rightarrow \infty} \int_X f_n \, d\mu = \int_X f \, d\mu$$

Proof. □

Vitali Convergence Theorem

Definitions 10.18.6 (uniformly integrable). Let (X, \mathcal{M}, μ) be a measure space. Let $\{f_n\}$ be a sequence of functions integrable over X . The sequence $\{f_n\}$ is uniformly integrable over X if for any $\varepsilon > 0$, there exist $\delta > 0$ such that for any measurable subset $E \subset X$ of measure $\mu(E) < \delta$, $\int_E |f_n| d\mu < \varepsilon$, $\forall n \in \mathbb{N}$.

Definitions 10.18.7 (tight sequence). Let (X, \mathcal{M}, μ) be a measure space. A sequence of function $\{f_n\}$ is tight over X if for any $\varepsilon > 0$, there exists a subset of finite measure X_0 such that $\int_{X \sim X_0} |f_n| d\mu < \varepsilon$, $\forall n \in \mathbb{N}$.

Theorem 10.18.24. Let (X, \mathcal{M}, μ) be a measure space. Let function f be integrable over X . Then for any $\varepsilon > 0$, there exists $\delta > 0$ such that for any subset E with $\mu(E) < \delta$, $\int_E |f| d\mu < \varepsilon$.

Furthermore, for any $\varepsilon > 0$, there exists a subset of finite measure $X_0 \subset X$ such that $\int_{X \sim X_0} |f| d\mu < \varepsilon$.

Proof. Without loss of generality, suppose $0 \leq f$ on X . Otherwise, consider $f = f^+ - f^-$ and apply the following proof for those non-negative functions.

Let $\varepsilon > 0$. Since f is integrable, $\int_X f d\mu$ is finite. By the definition of integral of a non-negative functions, there exists a simple function ψ on X such that $0 \leq \psi \leq f$ on X and

$$0 \leq \int_X f d\mu - \int_X \psi d\mu < \varepsilon/2$$

Since ψ assume only finite number of values, we can choose $M > 0$ such that $0 \leq \psi < M$ on X .

$$\int_E f d\mu = \int_E \psi d\mu + \int_E (f - \psi) d\mu \leq M \cdot \mu(E) + \varepsilon/2$$

Choose $\delta = \varepsilon/2M$, then $\mu(E) < \varepsilon/2M$ and $\int_E f d\mu < \varepsilon$.

Let $X_0 = \{x \in X : f(x) > 0\}$. Then $f = f - \psi$ on $X \sim X_0$. Thus,

$$\int_{X \sim X_0} |f| d\mu = \int_{X \sim X_0} f d\mu = \int_{X \sim X_0} (f - \psi) d\mu \leq \int_X (f - \psi) d\mu < \varepsilon$$

□

Theorem 10.18.25 (Vitali Convergence theorem). Let (X, \mathcal{M}, μ) be a measure space. Let $\{f_n\}$ be a sequence of functions which are uniformly integrable and tight over X . Suppose $\{f_n\}$ converges to function f pointwise a.e. on X and f is integrable over X . Then, $\lim_{n \rightarrow \infty} \int_E f_n d\mu = \int_E f d\mu$.

Proof. By integral test, we have

$$\begin{aligned} \left| \int_X (f_n - f) d\mu \right| &\leq \int_X |f_n - f| d\mu \\ &\leq \int_{X_1} |f_n - f| d\mu + \int_{X_0 \sim X_1} |f_n - f| d\mu + \int_{X \sim X_0} |f_n - f| d\mu \\ &\leq \int_{X_1} |f_n - f| d\mu + \int_{X_0 \sim X_1} (|f_n| + |f|) d\mu + \int_{X \sim X_0} (|f_n| + |f|) d\mu \end{aligned}$$

Let $\varepsilon > 0$. Then since $\{f_n\}$ is tight and f is integrable, there exists $X_0 \subset X$ such that X_0 is a measurable subset of finite measure and

$$\int_{X \sim X_0} |f_n| d\mu + \int_{X \sim X_0} |f| d\mu < \varepsilon/3 \quad (10.89)$$

Since f is integrable over X , f is finite a.e. on X and $\mu(X_0) < \infty$. Thus by Egoroff's theorem, X_0 has a subset X_1 with $\mu(X_0 \sim X_1) < \delta$ such that $\{f_n\}$ is uniformly convergent on X_1 . By uniform integrability of $\{f_n\}$, there exists $\delta > 0$ such that for any measurable set E with $\mu(E) < \delta$.

$$\int_{X_0 \sim X_1} (|f_n| + |f|) d\mu = \int_{X_0 \sim X_1} |f_n| d\mu + \int_{X_0 \sim X_1} |f| d\mu < \varepsilon/3 \quad (10.90)$$

By uniform convergence of $\{f_n\}$ into f on X_1 , there exists N such that for every $n \geq N$.

$$\int_{X_1} |f_n - f| d\mu \leq \sup_{x \in X_1} |f_n(x) - f(x)| \cdot \mu(X_1) < \varepsilon/3 \quad (10.91)$$

□

Corollary 10.18.25.1. *Let (X, \mathcal{M}, μ) be a measure space. Let $\{h_n\}$ be sequence on non-negative, integrable functions over X . Suppose $\{h_n\}$ converges to 0 for almost all $x \in X$. Then $\lim_{n \rightarrow \infty} \int_E h_n d\mu = 0$ if and only if $\{h_n\}$ is uniformly integrable and tight over X .*

Proof. Since $\{h_n\}$ is uniformly convergent to 0 almost everywhere on X and is tight, from theorem we have

$$\lim_{n \rightarrow \infty} \int_X h_n d\mu = 0$$

Suppose $\lim_{n \rightarrow \infty} \int_X h_n d\mu = 0$. Then by Fatou's lemma,

$$\int_X \lim_{n \rightarrow \infty} h_n d\mu < \liminf \int_X h_n(x) d\mu = 0$$

□

Exercises

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10.18.4 Radon-Nikodym Theorem

Definitions 10.18.8 (absolutely continuous). Let (X, \mathcal{M}) be a measure space. Let μ, ν be measures on X . The measure ν is absolutely continuous with respect to μ if $\nu(E) = 0$ whenever $\mu(E) = 0$ for any measure subset $E \in \mathcal{M}$. And we write $\nu \ll \mu$.

For example : The set function $\nu(E) = \int_E f \, d\mu$ is a measure on (X, \mathcal{M}) and is absolutely continuous with respect to μ .

Theorem 10.18.26. *Let (X, \mathcal{M}, μ) be a measure space. Let ν be a measure on (X, \mathcal{M}) . Then ν is absolutely continuous with respect to μ if and only if for any $\varepsilon > 0$, there exists $\delta > 0$ such that for any measurable subset $E \in \mathcal{M}$ with $\mu(E) < \delta$, $\nu(E) < \varepsilon$.*

Proof. Suppose for $\varepsilon > 0$, there exists $\delta > 0$ such that $\nu(E) < \varepsilon$ if $\mu(E) < \delta$. Since measure has monotonicity, as $\varepsilon \rightarrow 0$, $\delta \rightarrow 0$. Therefore, ν is absolutely continuous with respect to μ .

Let $\varepsilon > 0$. Suppose that for any $\delta > 0$, there exists some measurable set E with $\mu(E) < \delta$, but $\nu(E) \geq \varepsilon$. Then, we can construct a sequence of measurable sets, $\{E_n\}$ such that $\mu(E_n) \leq \frac{1}{2^n}$ and $\nu(E_n) \geq \varepsilon$. Define a descending sequence $\{A_n\}$ where $A_n = \bigcup_{k=n}^{\infty} E_k$. By the continuity of measure,

$\lim_{n \rightarrow \infty} \mu(A_n) = \mu\left(\bigcap_{n=1}^{\infty} A_n\right) = 0$ which is a contradiction to absolute continuity since $\lim_{n \rightarrow \infty} \nu(A_n) \geq \varepsilon$. Therefore, if $\varepsilon > 0$, then there exists $\delta > 0$ such that $\mu(E) < \delta \implies \nu(E) < \varepsilon$. \square

Theorem 10.18.27 (Radon-Nikodym). *Let (X, \mathcal{M}, μ) be a σ -finite measure space. If there exists a σ -finite measure ν on (X, \mathcal{M}) which is absolutely continuous with respect to μ . Then there exists a non-negative function f such that f is measurable (with respect to \mathcal{M}) and*

$$\nu(E) = \int_E f \, d\mu, \quad \forall E \in \mathcal{M} \quad (10.92)$$

And this function f is unique with respect to μ . That is, for any other non-negative, measurable $[\mu]$ function g defining ν , $f = g$ a.e. $[\mu]$ on X .

Proof. Step 1 : Finite Measures μ, ν

Suppose μ, ν are finite measures. If $\nu(E) = 0$ for any $E \in \mathcal{M}$, then the equation is true for $f = 0$ on X .

Step 2 : Non-vanishing Measure ν

Suppose $\nu(E) \neq 0$ for some $E \in \mathcal{M}$. First, we prove that there exists a non-negative function f on X such that

$$\int_X f \, d\mu > 0 \text{ and } \int_E f \, d\mu \leq \nu(E), \quad \forall E \in \mathcal{M}$$

Then we prove that the supremum of all such function is the unique function satisfying our requirements.

Step 3 : λ_0 such that $\mu(P_{\lambda_0}) > 0$

Let $\lambda > 0$. Then $\nu - \lambda\mu$ is difference of two finite measures and thus a signed measure. According to Hahn decomposition theorem, there exists a Hahn decomposition $\{P_\lambda, N_\lambda\}$ of signed measure $\nu - \lambda\mu$. Thus, for each λ , we have $X = P_\lambda \cup N_\lambda$ and $P_\lambda \cap N_\lambda = \emptyset$. Also the sets P_λ, N_λ are positive and negative with respect to $\nu - \lambda\mu$ respectively.

We claim that there exists $\lambda > 0$ such that $\mu(P_\lambda) > 0$. Suppose that for any $\lambda > 0$, $\mu(P_\lambda) = 0$. Then by monotonicity, $\mu(E) = 0$ for any measurable subset $E \subset P_\lambda$. We have P_λ is a positive set with respect to $\nu - \lambda\mu$. Also ν is absolutely continuous with respect to μ . Thus, we have $\nu(E) = 0$. And since N_λ is negative with respect to $\nu - \lambda\mu$, we have $\nu(E) \leq \lambda\mu(E)$. Clearly, every measurable subset E is of the form $E = E_P \cup E_N$ where $E_P \subset P_\lambda$ and $E_N \subset N_\lambda$. And $\nu(E) = \nu(E_P) + \nu(E_N) \leq 0 + \lambda\mu(E_N) \leq \lambda\mu(E)$. Thus, $\nu \leq \lambda\mu$ for every measurable subset $E \in \mathcal{M}$ and for any real-number $\lambda > 0$. Clearly, $\mu(E) = 0 \implies \nu(E) = 0$ and $\mu(E) < \infty \implies \nu(E) = 0$ by archimedian property. This is a contradiction, since we assumed that $\nu(E) \neq 0$ for some $E \in \mathcal{M}$.

Step 4 : Existence of f such that $\int_E f d\mu \leq \nu(E)$

Let $\lambda_0 > 0$ such that $\mu(P_{\lambda_0}) > 0$. Define $f = \lambda_0 \cdot \chi_{P_{\lambda_0}}$. Then, we have

$\int_X f d\mu = \lambda_0 \cdot \mu(P_{\lambda_0}) > 0$. And for any measurable subset $E \in \mathcal{M}$, we have

$$\int_E f d\mu = \lambda_0 \cdot \mu(P_{\lambda_0} \cap E) \leq \nu(P_{\lambda_0} \cap E) \leq \nu(E)$$

And by construction, f is a non-negative, measurable function on X .

Step 5 : Construction of f such that $\int_E f d\mu = \nu(E)$

Let \mathcal{F} be the collection of all such non-negative, measurable functions on X .

Define $M = \sup_{f \in \mathcal{F}} \left\{ \int_X f d\mu \right\}$. Suppose $g, h \in \mathcal{F}$, then $\max\{g, h\} \in \mathcal{F}$. And,

$$\int_E \max\{g, h\} d\mu = \int_{E_1} h d\mu + \int_{E_2} g d\mu$$

where $E_1 = \{x \in X : h(x) > g(x)\}$ and $E_2 = \{x \in X : h(x) \leq g(x)\}$.

By the definition of supremum, there exists a sequence of functions $\{f_n\}$ in \mathcal{F} such that $\lim_{n \rightarrow \infty} \int_X f_n d\mu = M$. Without loss of generality, this is an increasing sequence. If the sequence is not increasing, then replace f_k with $\max\{f_1, f_2, \dots, f_k\}$.

Define $f(x) = \lim_{n \rightarrow \infty} \int_X f_n d\mu$. By monotone convergence theorem, we have $\int_X f d\mu = M$. Define $\eta(E) = \nu(E) - \int_E f d\mu$. Since integral is countably additive and $\nu(X) < \infty$, η is a signed measure. Also η is absolutely continuous with respect to ν by its construction.

We claim that $\eta = 0$ on \mathcal{M} .

Suppose $\eta \neq 0$. By above proof, replacing η with ν , we get a non-negative measurable function \hat{f} such that

$$\int_X \hat{f} d\mu > 0 \text{ and } \int_E \hat{f} d\mu \leq \eta(E) = \nu(E) - \int_E f d\mu$$

Clearly $f + \hat{f} \in \mathcal{F}$ and $\int_X (f + \hat{f}) d\mu > \int_X f d\mu = M$ which is a contradiction.

Step 6 : Uniqueness of f

Suppose f_1, f_2 are two limit function on \mathcal{F} satisfying the conditions. Then, for every measurable subset $E \in \mathcal{M}$, we have $\int_E (f_1 - f_2) d\mu = 0$ and $f_1 = f_2$ a.e. $[\mu]$ on X . Therefore, the choice of f is unique upto measure μ .

Step 4 : Extending the proof to σ -finite measures

Suppose μ, ν are σ -finite measures. Then from the definition of σ -finite measure, X has a countable partition, $X = \cup_{n=1}^{\infty} X_n$ such that $\mu(X_n) < \infty$ and $\nu(X_n) < \infty$ for every n .

We know that on each X_n there exists a non-negative function f_n such that $\int_{E \cap X_n} f_n d\mu_n = \nu_n(E \cap X_n)$. Define $f = \sum f_n \cdot \chi_{X_n}$.

$$\int_X f d\mu = \sum_{n=1}^{\infty} \int_{X_n} f_n d\mu_n > 0$$

And

$$\int_E f d\mu = \sum_{n=1}^{\infty} \int_{E \cap X_n} f_n d\mu_n = \sum_{n=1}^{\infty} \nu_n(E \cap X_n) = \nu(E)$$

□

Theorem 10.18.28. *Let (X, \mathcal{M}, μ) be a σ -finite measure space. If there exists a finite signed measure ν which is absolutely continuous with respect to μ on X . Then there exist a function f which is integrable over X with respect to μ and*

$$\nu(E) = \int_E f d\mu, \forall E \in \mathcal{M} \quad (10.93)$$

Proof.

□

Theorem 10.18.29 (Lebesgue Decomposition). *Let (X, \mathcal{M}, μ) be a σ -finite measure space. Let ν be a σ -finite measure on (X, \mathcal{M}) . Then there exists unique measures ν_0, ν_1 such that $\nu = \nu_0 + \nu_1$, ν_0 is singular with respect to μ and ν_1 is absolutely continuous with respect to μ .*

Proof.

□

Definitions 10.18.9 (Radon-Nikodym derivative). Let (X, \mathcal{M}) be a measure space. Let μ, ν be two σ -finite measures on (X, \mathcal{M}) such that ν is absolutely continuous with respect to μ . By, Radon-Nikodym theorem, there exists a measurable function f which is unique upto μ such that $\nu = \int_E f d\mu$. This function f is the Radon-Nikodym derivative $f = \frac{d\nu}{d\mu}$.

Exercises

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10.20 The Construction of a particular measure

10.20.1 Product measures

Definitions 10.20.1 (measurable rectangle). Let (X, \mathcal{A}, μ) , (Y, \mathcal{B}, ν) be two measure spaces. A rectangle is a subset of $X \times Y$ of the form $A \times B$ where $A \subset X$ and $B \subset Y$. A measurable rectangle is a subset of $X \times Y$ of the form $A \times B$ where $A \in \mathcal{A}$ and $B \in \mathcal{B}$.

Lemma 10.20.1. Let $\{A_k \times B_k\}_{k=1}^{\infty}$ be a countable disjoint, collection of measurable rectangles such that $\bigcup_{k=1}^{\infty} (A_k \times B_k) = A \times B$ is also a measurable rectangle. Then $\mu(A) \times \nu(B) = \sum_{k=1}^{\infty} \mu(A_k) \times \nu(B_k)$

Proof. Consider $A \times B$. Fix $x \in A$. Now for each $y \in B$, there exists exactly one $A_k \times B_k$ such that $(x, y) \in A_k \times B_k$. Thus, $B = \bigcup \{B_k : x \in A_k\}$. By countable additivity of the measure ν , $\nu(B) = \sum_{k: x \in A_k} \nu(B_k)$.

$$\nu(B) \cdot \chi_A(x) = \sum_{k=1}^{\infty} \nu(B_k) \cdot \chi_{A_k}(x), \quad \forall x \in A$$

For $x \in X - A$, we have $\chi_{A_k}(x) = 0$, $\forall k$ since $A_k \subset A$. Thus,

$$\nu(B) \cdot \chi_A(x) = \sum_{k=1}^{\infty} \nu(B_k) \cdot \chi_{A_k}(x), \quad \forall x \in X \quad (10.94)$$

By monotone convergence theorem,²⁵ the passage of limit under integral sign is

²⁵The sequence of partial sums $\{s_n\}$, $s_n = \sum_{k=1}^n \nu(B_k) \cdot \chi_{A_k}$ is an increasing sequence of non-negative measurable functions.

feasible. Therefore,

$$\begin{aligned}
 \mu(A)\nu(B) &= \nu(B) \int_X \chi_A d\mu = \int_X \nu(B) \cdot \chi_A d\mu \quad (\because \text{linearity of integral}) \\
 &= \int_X \sum_{k=1}^{\infty} \nu(B_k) \cdot \chi_{A_k} d\mu \quad (\because \text{eqn.10.94}) \\
 &= \int_X \lim_{n \rightarrow \infty} \sum_{k=1}^n \nu(B_k) \cdot \chi_{A_k} d\mu \\
 &= \lim_{n \rightarrow \infty} \int_X \sum_{k=1}^n \nu(B_k) \cdot \chi_{A_k} d\mu \quad (\because \text{monotone convergence}) \\
 &= \lim_{n \rightarrow \infty} \sum_{k=1}^n \int_X \nu(B_k) \cdot \chi_{A_k} d\mu \quad (\because \text{linearity of integral}) \\
 &= \sum_{k=1}^{\infty} \nu(B_k) \int_{A_k} d\mu = \sum_{k=1}^{\infty} \mu(A_k) \nu(B_k)
 \end{aligned}$$

□

Definitions 10.20.2 (premeasure). A premeasure μ on X satisfies

1. μ is set function which maps a non-negative (extended) real number to every subset of X

$$\mu : \mathcal{P}(X) \rightarrow [0, \infty]$$

2. μ is finitely additive

$$\mu \left(\bigcup_{k=1}^n A_k \right) = \sum_{k=1}^n \mu(A_k)$$

3. μ is countably monotone

$$A \subset \bigcup_{k=1}^{\infty} A_k \implies \mu(A) \leq \sum_{k=1}^{\infty} \mu(A_k)$$

4. $\mu(\emptyset) = 0$

For example, Lebesgue outer measure m^* is a premeasure on \mathbb{R} .

Definitions 10.20.3 (semiring). A collection sets S is a semiring if it satisfies

1. S is closed under intersection
2. For every pair $A, B \in S$, the relative complement, $A - B$ can be expressed as disjoint union of elements of S .

Theorem 10.20.2. Let (X, \mathcal{A}, μ) , (Y, \mathcal{B}, ν) be two measure spaces. Let \mathcal{R} be the collection of all measurable rectangles in $X \times Y$. Define $\lambda : \mathcal{R} \rightarrow [0, \infty]$ as $\lambda(A \times B) = \mu(A) \cdot \nu(B)$. Then \mathcal{R} is a semiring and $\lambda : \mathcal{R} \rightarrow [0, \infty]$ is a premeasure.

Proof. Step 1 : \mathcal{R} is a semiring

The intersection of two measurable rectangles is also a measurable rectangle.

$$(A \times B) \cap (C \times D) = (A \cap C) \times (B \cap D)$$

And, relative complement of two measurable rectangles is the disjoint union of two measurable rectangles.

$$(A \times C) - (B \times D) = [(A - B) \times C] \cup [(A \cap B) \times (C - D)]$$

Thus, \mathcal{R} is a semiring.

Step 2 : λ is a premeasure

We have, $\lambda(\phi \times \phi) = \mu(\phi)\nu(\phi) = 0$. Suppose $A \times B = \bigcup_{k=1}^{\infty} A_k \times B_k$.

By lemma,

$$\lambda(A \times B) = \mu(A)\nu(B) = \sum_{k=1}^{\infty} \mu(A_k)\nu(B_k) = \sum_{k=1}^{\infty} \lambda(A_k \times B_k)$$

That is, λ is countably additive. Suppose $A \times B = \sum_{k=1}^n A_k \times B_k$. Define $A_k = B_k = \phi$, $\forall k > n$. Then, λ is finitely additive.

Let $E \subset \bigcup_{k=1}^{\infty} E_k$ where E_k are disjoint.²⁶ Then, $E = \bigcup_{k=1}^{\infty} (E \cap E_k)$. Thus, λ is countably monotone since,

$$\lambda(E) = \sum_{k=1}^{\infty} \lambda(E \cap E_k) \leq \sum_{k=1}^{\infty} \lambda(E_k) \quad (\because E \cap E_k \subset E_k)$$

□

Definitions 10.20.4 (product measure). Let (X, \mathcal{A}, μ) , (Y, \mathcal{B}, ν) be two measure spaces. Let \mathcal{R} be the semiring of measurable rectangles in $X \times Y$ and λ is the premeasure defined by $\lambda : \mathcal{R} \rightarrow [0, \infty]$, $\lambda(A \times B) = \mu(A) \cdot \nu(B)$. The **product measure** $\lambda = \mu \times \nu$ is the Caratheodory extension of $\lambda : \mathcal{R} \rightarrow [0, \infty]$ defined by the σ -algebra of $(\mu \times \nu)^*$ -measurable subsets of $X \times Y$.

Definitions 10.20.5 (x -section). Let E be a subset of $X \times Y$ and f be a function on E . Let $x \in X$. Then the x -section of E , $E_x = \{y \in Y : (x, y) \in E\}$. And the x -section of f , $f(x, \cdot)$ is a function on E_x such that $f(x, \cdot)(y) = f(x)(y)$, $\forall y \in E_x$.

10.20.2 Iterated Integration

Lemma 10.20.3. Let $E \subset X \times Y$ be an $\mathcal{R}_{\sigma\delta}$ set for which $(\mu \times \nu)(E) < \infty$. Then for every $x \in X$, the x -section of E , E_x is a ν -measurable subset of Y , the function which maps each $x \in X$ into $\nu(E_x)$ is a μ -measurable function and

$$(\mu \times \nu)(E) = \int_X \nu(E_x) d\mu(x) \quad (10.95)$$

²⁶ E, E_k are measurable rectangles.

Proof. Consider $E = A \times B$. Then,

$$E_x = \begin{cases} B & x \in A \\ \phi & x \notin A \end{cases} \quad \text{and} \quad \nu(E_x) = \begin{cases} \nu(B) & x \in A \\ 0 & x \notin A \end{cases}$$

Thus, $\nu(E_x) = \nu(B) \cdot \chi_A(x)$ And,

$$(\mu \times \nu)(E) = \mu(A) \cdot \nu(B) = \nu(B) \int_X \chi_A \, d\mu = \int_X \nu(B) \cdot \chi_A(x) \, d\mu(x)$$

Let E be an \mathcal{R}_σ set. Then, E_σ can be expressed as disjoint union of measurable rectangles $\{A_k \times B_k : k \in \mathbb{N}\}$ of finite measure. Since \mathcal{R} is a semiring, the x -section of E can be expressed as disjoint, union of the x -sections of the respective measurable rectangles. That is,

$$E_x = \bigcup_{k=1}^{\infty} (A_k \times B_k)_x$$

$$\begin{aligned} \int_x \nu(E_x) \, d\mu(x) &= \int_X \sum_{k=1}^{\infty} \nu([A_k \times B_k]_x) \, d\mu = \int_X \lim_{n \rightarrow \infty} \sum_{k=1}^n \nu([A_k \times B_k]_x) \, d\mu \\ &= \lim_{n \rightarrow \infty} \int_X \sum_{k=1}^n \nu([A_k \times B_k]_x) \, d\mu \quad (\because \text{monotone convergence theorem}) \\ &= \sum_{k=1}^{\infty} \int_X \nu([A_k \times B_k]_x) \, d\mu \\ &= \sum_{k=1}^{\infty} \mu(A_k) \cdot \nu(B_k) = (\mu \times \nu)(E) \end{aligned}$$

Let E be an $\mathcal{R}_{\sigma\delta}$ set with $(\mu \times \nu)(E) < \infty$. Since \mathcal{R} is semiring, E can be expressed as intersection of a descending sequence $\{E_k\}_{k=1}^{\infty}$ in \mathcal{R}_σ . And without loss of generality $(\mu \times \nu)(E_1) < \infty$. By continuity of measure $\mu \times \nu$,

$$\lim_{k \rightarrow \infty} (\mu \times \nu)(E_k) = (\mu \times \nu)(E)$$

Since E_1 is an \mathcal{R}_σ set,

$$(\mu \times \nu)(E_1) = \int_X \nu([E_1]_x) \, d\mu(x)$$

Thus, $\nu([E_1]_x) < \infty$. And for each $x \in X$, E_x is ν measurable. By continuity of measure,

$$\lim_{k \rightarrow \infty} \nu([E_k]_x) = \nu(E_x)$$

□

Definitions 10.20.6. Let \mathcal{S} be a collection of measurable sets. Then \mathcal{S}_σ is the collection of all countable unions of members of \mathcal{S} . And $\mathcal{S}_{\sigma\delta}$ is the collection of all countable intersection of members of \mathcal{S}_σ .

Proposition 10.20.1 (§10.17 pp. 362). *Let \mathcal{S} be a collection of subsets of X . Let $\mu : \mathcal{S} \rightarrow [0, \infty]$ be a set function on \mathcal{S} , $\mu^* : 2^X \rightarrow [0, \infty]$ be the Caratheodory outer measure and $\bar{\mu} : \mathcal{M} \rightarrow [0, \infty]$ be the Caratheodory measure induced by μ . Let E be a subset of X such that $\mu^*(E) < \infty$. Then there exists an $\mathcal{S}_{\sigma\delta}$ set A such that $E \subset A$ and $\mu^*(E) = \mu^*(A)$. Furthermore if E and each set in \mathcal{S} is measurable with respect to μ^* , then $\bar{\mu}(A \sim E) = 0$.*

Proof. is not required. \square

Lemma 10.20.4. *Suppose measure ν is complete. Let $E \subset X \times Y$ be measurable with respect to $\mu \times \nu$. If $(\mu \times \nu)(E) = 0$, then for almost all $x \in X$, the x -section of E , E_x is ν -measurable and $\nu(E_x) = 0$. Therefore,*

$$(\mu \times \nu)(E) = \int_X \nu(E_x) d\mu(x) \quad (10.96)$$

Proof. We have $(\mu \times \nu)(E) < \infty$. Then there exists an $\mathcal{R}_{\sigma\delta}$ set A such that $E \subset A$ and $(\mu \times \nu)(A) = 0$. Then every x -section of A , A_x is ν measurable and $(\mu \times \nu)(A) = \int_X \nu(A_x) d\mu$.

Clearly, $\nu(A_x) = 0$ for almost all $x \in X$ and $E_x \subset A_x$, $\forall x$. Then, $\nu(E_x) = 0$ since ν is complete. \square

Theorem 10.20.5. *Suppose measure ν is complete. Let $E \subset X \times Y$ be measurable with respect to $\mu \times \nu$ and $(\mu \times \nu)(E) < \infty$. Then for almost all $x \in X$, the x -section of E , E_x is a ν -measurable subset of Y , the function which maps each $x \in X$ into $\nu(E_x)$ is a μ -measurable function and*

$$(\mu \times \nu)(E) = \int_X \nu(E_x) d\mu(x) \quad (10.97)$$

Proof. We have $(\mu \times \nu)E < \infty$. Then, there exists $\mathcal{R}_{\sigma\delta}$ set A such that $E \subset A$ and $(\mu \times \nu)(A \sim E) = 0$. By excision property, $(\mu \times \nu)E = (\mu \times \nu)A$. And $\nu(A_x) = \nu(E_x) + \nu((A \sim E)_x) = \nu(E_x)$ for almost all $x \in X$. Thus,

$$\begin{aligned} (\mu \times \nu)E &= (\mu \times \nu)A \\ &= \int_X \nu(A_x) d\mu \\ &= \int_X \nu(E_x) d\mu \end{aligned}$$

\square

Theorem 10.20.6. *Suppose measure ν is complete. Let $\varphi : X \times Y \rightarrow \mathbb{R}$ be a simple function that is integrable over $X \times Y$ with respect to $\mu \times \nu$. Then for almost all $x \in X$, the x -section of φ , $\varphi(x, \cdot)$ is integrable over Y with respect to ν and*

$$\int_{X \times Y} \varphi d(\mu \times \nu) = \int_X \left(\int_Y \varphi(x, y) d\nu(y) \right) d\mu(x) \quad (10.98)$$

Proof. Let $\phi = \chi_E$ be a characteristic function on a measurable rectangle E of $X \times Y$. Then,

$$\begin{aligned} \int_{X \times Y} \phi \, d(\mu \times \nu) &= \int_{X \times Y} \chi_E \, d(\mu \times \nu) \\ &= \int_X \int_Y \chi_{E_x} \, d\mu \, d\nu \\ &= \int_X \left(\int_Y \phi(x, \cdot)(y) \, d\nu \right) d\mu \\ &= \int_X \left(\int_Y \phi(x, y) \, d\nu \right) d\mu \end{aligned}$$

Let $\phi = \sum a_n \chi_{E_n}$ be a simple function where E_n are measurable rectangles. Then by linearity of integration, we have

$$\begin{aligned} \int_{X \times Y} \phi \, d(\mu \times \nu) &= \sum a_n \int_{X \times Y} \phi_n \, d(\mu \times \nu) \\ &= \sum a_n \int_X \int_Y \phi_n(x, y) \, d\mu \, d\nu \\ &= \int_X \int_Y \phi(x, y) \, d(\mu \times \nu) \end{aligned}$$

□

Theorem 10.20.7 (Fubini). *Let (X, \mathcal{A}, μ) , (Y, \mathcal{B}, ν) be two measure spaces where ν is complete. Let f be integrable over $X \times Y$ with respect to the product measure $\mu \times \nu$. Then for almost all $x \in X$, the x -section of f is integrable over Y with respect to ν and*

$$\int_{X \times Y} f \, d(\mu \times \nu) = \int_X \left(\int_Y f(x, y) \, d\nu(y) \right) d\mu(x) \quad (10.99)$$

Proof. Let f be a non-negative, measurable function on $X \times Y$. By simple approximation theorem, there exists an increasing sequence $\{\phi_n\}$ of simple functions such that $0 \leq \phi_n \leq f$. Since f is integrable, ϕ_n are integrable.

$$\int_{X \times Y} \phi_n \, d(\mu \times \nu) \leq \int_{X \times Y} f \, d(\mu \times \nu) < \infty$$

By monotone convergence theorem,

$$\lim_{n \rightarrow \infty} \int_{X \times Y} \phi_n \, d(\mu \times \nu) = \int_{X \times Y} f \, d(\mu \times \nu)$$

□

Theorem 10.20.8 (Tonelli). *Let (X, \mathcal{A}, μ) and (Y, \mathcal{B}, ν) be two σ -finite measure spaces where ν is complete. Let f be a non-negative $(\mu \times \nu)$ -measurable function on $X \times Y$. Then for almost all $x \in X$, the x -section of f , $f(x, \cdot)$ is ν -measurable and the function which maps almost every $x \in X$ into integral of $f(x, \cdot)$ over Y with respect to ν is μ -measurable. Moreover,*

$$\int_{X \times Y} f \, d(\mu \times \nu) = \int_X \left(\int_Y f(x, y) \, d\nu(y) \right) d\mu(x) \quad (10.100)$$

Proof. □

Corollary 10.20.8.1. *Let (X, \mathcal{A}, μ) and (Y, \mathcal{B}, ν) be two σ -finite, complete measure spaces. Let f be a non-negative $(\mu \times \nu)$ -measurable function on $X \times Y$. Then*

1. *for almost all $x \in X$, the x -section of f , $f(x, \cdot)$ is ν -measurable function and the function defined almost everywhere on X by $x \rightarrow$ the integral of $f(x, \cdot)$ over Y with respect to ν is μ -measurable and*
2. *for almost all $y \in Y$, the y -section of f , $f(\cdot, y)$ is μ -measurable and the function defined almost everywhere on Y by $y \rightarrow$ the integral of $f(\cdot, y)$ over X with respect to μ is ν -measurable if*

$$\int_X \left(\int_Y f(x, y) \, d\nu(y) \right) d\mu(x) < \infty \quad (10.101)$$

then f is integrable over $X \times Y$ with respect to $\mu \times \nu$ and

$$\begin{aligned} \int_Y \left(\int_X f(x, y) \, d\mu(x) \right) d\nu(y) &= \int_{X \times Y} f d(\mu \times \nu) \\ &= \int_X \left(\int_Y f(x, y) \, d\nu(y) \right) d\mu(x) \end{aligned} \quad (10.102)$$

Proof. □

Exercises

Definitions 10.20.7 (probability measure space). A probability measure space is a measure space (X, \mathcal{M}, μ) where $\mu(X) = 1$.

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Semester III

Semester 3

Subject 11

ME010301 Advanced Complex Analysis

Module 1

11.1 Harmonic Functions(4.6)

11.1.1 Definition and Basic properties(4.6.1) pp. 162

Definitions 11.1.1 (harmonic). A real-valued function $u(z)$ defined and single-valued in a region Ω is harmonic if it is continuous together with its partial derivatives of first two orders and satisfies Laplace's equation

$$\Delta u = \frac{\partial^2 u}{\partial x^2} + \frac{\partial^2 u}{\partial y^2} = 0$$

In polar form,

$$r \frac{\partial}{\partial r} \left(r \frac{\partial u}{\partial r} \right) + \frac{\partial^2 u}{\partial \theta^2} = 0$$

Remark. The sum of two harmonic functions is harmonic. A constant multiple of a harmonic function is harmonic.

From polar form, $\log r$ is harmonic and any harmonic function that depends only on r must be of the form $a \log r + b$. The argument θ is harmonic whenever it can be uniquely defined.

If u is harmonic in Ω , then $f(z) = \frac{\partial u}{\partial x} - i \frac{\partial u}{\partial y}$ is analytic on Ω .

$$f dz = f dx + i f dy = \left(\frac{\partial u}{\partial x} dx + \frac{\partial u}{\partial y} dy \right) + i \left(-\frac{\partial u}{\partial y} dx + \frac{\partial u}{\partial x} dy \right)$$

The real part is the differential of u and imaginary part is the conjugate differential of du .

$$f dz = du + i^* du$$

—to be continued—

Theorem 11.1.1. *If u_1 and u_2 are harmonic in a region Ω , then*

$$\int_{\gamma} u_1^* du_2 - u_2^* du_1 = 0$$

Proof.

□

11.1.2 The Mean-value Property(4.6.2) pp. 165

Theorem 11.1.2. *The arithmetic mean of a harmonic function over arithmetic circles $|z| = r$ is a linear function of $\log r$,*

$$u(z_0) = \frac{1}{2\pi} \int_0^{2\pi} u \, d\theta = \alpha \log r + \beta \quad (11.1)$$

Proof.

□

Theorem 11.1.3. *A nonconstant harmonic function has neither a maximum nor a minimum in its region of definition. Consequently, the maximum and the minimum on a closed bounded set E are taken on the boundary of E*

Proof.

□

11.1.3 Poisson's Formula(4.6.3) pp. 166

Remark. If $u(z)$ is continuous on a closed bounded set E and harmonic in the interior of E , then it is uniquely determined by its values on the boundary of E .

Theorem 11.1.4. *Suppose that $u(z)$ is harmonic for $|z| < R$, continuous for $|z| \leq R$. Then*

$$u(a) = \frac{1}{2\pi} \int_{|z|=R} \frac{R^2 - |a|^2}{|z - a|^2} u(z) \, d\theta, \quad (|a| < R) \quad (11.2)$$

Proof.

□

11.1.4 Schwarz's Theorem(4.6.4) pp. 168

Definitions 11.1.2 (Poisson integral). Let U be a piece-wise continuous function on $[0, 2\pi]$. The Poisson integral of U is given by,

$$P_U(z) = \frac{1}{2\pi} \int_0^{2\pi} \Re \left(\frac{e^{i\theta} + z}{e^{i\theta} - z} \right) U(\theta) \, d\theta \quad (11.3)$$

Remark. Clearly, Poisson integral is a positive linear functional since the Riemann integral is linear.

$$P_{U+V} = P_U + P_V$$

$$P_{cU} = cP_U$$

$$U \geq 0 \implies P_U \geq 0$$

Proof. □

Remark. Also we have, $P_c = c$. And since P_U is a positive, linear functional, we have

$$m \leq U \leq M \implies m \leq P_U \leq M$$

Proof. □

Theorem 11.1.5 (Schwarz). *The function $P_U(z)$ is harmonic for $|z| < 1$, and*

$$\lim_{z \rightarrow e^{i\theta_0}} P_U(z) = U(\theta_0) \quad (11.4)$$

provided U is continuous at θ_0 and $P_U(z)$ is the Poisson integral of the piecewise continuous function U on $[0, 2\pi]$.

Proof. □

11.1.5 The Reflection Principle(4.6.5) pp. 172

Remark. Let Ω be a region and Ω^* be its reflection about real axis. In other words, if $z \in \Omega$, then $\bar{z} \in \Omega^*$. Let $u(z)$ be a harmonic function defined on Ω , then $u(\bar{z})$ is a harmonic function on Ω^* . And if $f(z)$ is analytic on Ω , then $\overline{f(\bar{z})}$ is analytic on Ω^* .

Proof. □

Definitions 11.1.3 (Symmetric Region). Region Ω is symmetric if it is symmetric about real axis. In other words, for every $z \in \Omega$, \bar{z} is also in Ω .

Theorem 11.1.6 (Reflection Principle). *Let Ω^+ be the part in the upper plane of a symmetric region Ω and σ be its part on the real axis. Let $v(z)$ be continuous on $\Omega^+ \cup \sigma$, analytic on Ω^+ and vanishes on σ . Then v has a harmonic extension to Ω satisfying $v(\bar{z}) = -v(z)$. And if v is the imaginary part of an analytic function $f(z)$ in Ω^+ , then $f(z)$ has an analytic extension which satisfies $f(z) = \overline{f(\bar{z})}$.*

Proof. □

11.2 A Closer look at Harmonic Functions(6.3)

11.2.1 Functions with the Mean-value Property (6.3.1) pp. 242

Definitions 11.2.1. Let $u(z)$ be a real-valued continuous function in a region Ω . Function u satisfies mean-value property if

$$u(z_0) = \frac{1}{2\pi} \int_0^{2\pi} u(z_0 + re^{i\theta}) d\theta$$

when the disk $|z - z_0| \leq r$ is contained in Ω .

Remark. Mean-value property implies maximum principle.

Theorem 11.2.1. *A continuous function $u(z)$ satisfying the mean-value property is necessarily harmonic.*

Proof. □

11.2.2 Harnack's Principle (6.3.2) pp. 243

Theorem 11.2.2 (Harnack's Inequality).

$$\frac{\rho - r}{\rho + r} u(0) \leq u(z) \leq \frac{\rho + r}{\rho - r} u(0)$$

where $u(z)$ is harmonic in $|z| = r < \rho$ and continuous in $|z| \leq \rho$.

Proof. □

Theorem 11.2.3 (Harnack's Principle). *Increasing sequence of harmonic functions on regions Ω_n converges to either $+\infty$ or a harmonic function u uniformly on every compact subset of Ω where every point $z \in \Omega$ belongs to all except finitely many Ω_n .*

Proof. □

11.3 The Dirichlet Problem(6.4)

Definitions 11.3.1 (Dirichlet Problem). Find a harmonic function $u(z)$ on a region Ω with given boundary values.

11.3.1 Subharmonic Functions (6.4.1) pp. 245

Definitions 11.3.2 (subharmonic). A continuous, real-valued function $v(z)$ is subharmonic in Ω if for any harmonic function $u(z)$ in $\Omega' \subset \Omega$, $v - u$ satisfies the maximum principle on Ω' .

Remark. Harmonic functions are subharmonic.

Sufficient condition for subharmonicity : If v has a positive laplacian, then v is subharmonic. The condition is not necessary as there exists subharmonic functions without partial derivatives.

Theorem 11.3.1. *A continuous function $v(z)$ is subharmonic in Ω if and only if*

$$v(z_0) \leq \frac{1}{2\pi} \int_0^{2\pi} v(z_0 + re^{i\theta}) d\theta$$

for every disk $|z - z_0| \leq r$ contained in Ω .

Proof. □

Remark. Elementary Properties of Subharmonic Functions,

1. If v is subharmonic, then kv is subharmonic for any constant $k \geq 0$.

Proof. □

2. If v_1, v_2 are subharmonic, then $v_1 + v_2$ is subharmonic.

Proof. □

3. If v_1, v_2 are subharmonic, then $v = \max(v_1, v_2)$ is subharmonic.

Proof.

□

4. Let v be subharmonic on Ω . Let Δ be a disk whose closure is contained in Ω and P_v be the Poisson integral formed with v on the boundary of Δ . Then v' defined as P_v on Δ and v outside Δ is subharmonic.

Proof.

□

Definitions 11.3.3 (semicontinuous). A function $v(z)$ is **upper semicontinuous** at z_0 if $\limsup_{z \rightarrow z_0} v(z) \leq v(z_0)$. A function $v(z)$ is **lower semicontinuous** at z_0 if $\liminf_{z \rightarrow z_0} v(z) \geq v(z_0)$.

Remark. Upper continuous function attains maximum on any compact set.

11.3.2 Solution of Dirichlet's Problem (6.4.2) pp. 248

Definitions 11.3.4 ($\mathfrak{B}(f)$). Let f be continuous real-valued function defined on the boundary Γ of a bounded region Ω . The family of functions, $\mathfrak{B}(f)$ contains all functions v such that

1. v is subharmonic in Ω and
2. $\overline{\lim}_{z \rightarrow \zeta} v(z) \leq f(\zeta), \forall \zeta \in \Gamma$.

Lemma 11.3.2. The function u defined as $u(z) = \inf\{v(z) : v \in \mathfrak{B}(f)\}$ is harmonic in Ω .

Proof.

□

Remark (Perron's method). Let Ω be a bounded¹ region. Let Γ be the boundary of Ω . Let f be a continuous, real-valued function on Γ . Then, corresponding to each function f , there exists a harmonic function u on Ω which can be obtained by Perron's method.

Let Δ be a disk whose closure is contained in Ω . Let $z_0 \in \Delta$ and sequence $\{v_n\}$ in $\mathfrak{B}(f)$ such that $\{v_n(z_0)\}$ converges. Define $u(z_0) = \lim_{n \rightarrow \infty} v_n(z_0)$. Without loss of generality, $\{v_n\}$ is nondecreasing.² By Harnack's principle, there exists a harmonic function U such that $U \leq u$ and $U(z_0) = u(z_0)$.

Let $z_1 \in \Delta$ and sequence $\{w_n\}$ in $\mathfrak{B}(f)$ converges. Define $u(z_1) = \lim_{n \rightarrow \infty} w_n(z_1)$. Then there exists a harmonic function U_1 such that $U \leq U_1 \leq u$ and $U_1(z_1) = u(z_1)$. However, $U - U_1$ has maximum zero at z_0 . Thus, $U = U_1$ and $\forall z_1 \in \Delta, U(z_1) = U_1(z_1) = u(z_1)$. Clearly, we have constructed the harmonic function $u = U$.

¹Bounded Set : A set Ω of complex numbers is bounded, if there exists $\rho \in \mathbb{R}$ such that Ω is contained in an open disk of radius ρ about origin.

²If not, consider $V_n = \max\{v_1, v_2, \dots, v_n\}$

Lemma 11.3.3. *Suppose there exists a harmonic function $w(z)$ in Ω whose continuous boundary values $w(\zeta)$ are strictly positive except at one point $\zeta_0 = 0$. Then, if $f(\zeta)$ is continuous at ζ_0 , the corresponding function u determined by Perron's method satisfies $\lim_{z \rightarrow \zeta_0} u(z) = f(\zeta_0)$.*

Proof.

□

Theorem 11.3.4. *The Dirichlet problem can be solved for any region Ω such that each boundary point is the end point of a line segment whose other points are exterior to Ω .*

Proof.

□

Module 2

11.4 Power Series Expansions(5.1)

11.4.1 Weierstrass's Theorem(5.1.1) pp.175

Weierstrass's theorem gives a sufficient condition for convergence to preserve analyticity, that is uniform convergence on every compact subset of the region. And Hurwitz's theorem characterises such analytic, limit functions.

Theorem 11.4.1 (Weierstrass). *Suppose $f_n(z)$ is analytic in the region Ω_n , and the sequence $\{f_n(z)\}$ converges to $f(z)$ in a region Ω uniformly on every compact subset of Ω . Then the limit function $f(z)$ is analytic in Ω . Moreover, $f'_n(z)$ converges to $f'(z)$ on every compact subset of Ω .*

Proof. □

Remark. A series with analytic terms

$$f(z) = f_1(z) + f_2(z) + \cdots + f_n(z) + \cdots$$

converges uniformly on every compact subset of a region Ω , then the sum $f(z)$ is analytic in Ω . And the series can be differentiated term by term.

Proof. □

Remark. Uniform convergence on the boundary of a compact set A implies uniform convergence on A .

Proof. Hint : maximum principle □

Remark. If $f_n(z)$ are analytic in $|z| < 1$ and convergence is uniform on each circle $C : |z| = r_m$ where $r_m \rightarrow 1$ as $m \rightarrow \infty$, then limit function $f(z)$ is analytic.

Proof. □

Theorem 11.4.2 (Hurwitz). *If the functions $f_n(z)$ are analytic in a region Ω , $f_n(z) \neq 0, \forall z \in \Omega$ and $f_n(z)$ converges to $f(z)$ uniformly on every compact subset of Ω , then $f(z)$ is either identically zero in Ω or never assumes zero in Ω .*

Proof. □

11.4.2 The Taylor Series (5.1.2) pp 179

Every analytic function has a convergent Taylor series representation.

Theorem 11.4.3. *If $f(z)$ is analytic in a region Ω containing z_0 , then the representation*

$$f(z) = f(z_0) + \frac{f'(z_0)}{1!}(z - z_0) + \cdots + \frac{f^{(n)}(z_0)}{n!}(z - z_0)^n + \cdots \quad (11.5)$$

is valid in the largest open disk of center z_0 contained in Ω .

Proof.

□

Taylor series representation for a few analytic functions,

1. $e^z = 1 + z + \frac{z^2}{2!} + \cdots + \frac{z^n}{n!} + \cdots$
2. $\cos z = 1 - \frac{z^2}{2!} + \frac{z^4}{4!} + \cdots +$
3. $\sin z = z - \frac{z^3}{3!} + \frac{z^5}{5!} + \cdots +$
4. $(1+z)^\mu = 1 + \mu z + \binom{\mu}{2} z^2 + \cdots$ where $\binom{\mu}{n} = \frac{\mu(\mu-1)(\mu-2)\cdots(\mu-n+1)}{1\cdot 2\cdot 3\cdots n}$
5. $\log(1+z) = z - \frac{z^2}{2} + \frac{z^3}{3} - \frac{z^4}{4} + \cdots$
6. $\arctan z = z - \frac{z^3}{3} + \frac{z^5}{5} - \frac{z^7}{7} + \cdots$
7. $\frac{1}{\sqrt{1-z^2}} = 1 + \frac{1}{2}z^2 + \frac{1}{2}\frac{3}{4}z^4 + \frac{1}{2}\frac{3}{4}\frac{5}{6}z^6 + \cdots$
8. $\arcsin z = z + \frac{1}{2}\frac{z^3}{3} + \frac{1}{2}\frac{3}{4}\frac{z^5}{5} + \frac{1}{2}\frac{3}{4}\frac{5}{6}\frac{z^7}{7} + \cdots$
9. $\tan z = z + \frac{z^3}{3} + \frac{2}{15}z^5 + \cdots$
10. $\log \frac{\sin z}{z} =$

The procedure to find Taylor representation of an analytic function,

1. Choose a well-defined branch so that we have a single-valued function.
2. Choose a center z_0 which is not a branch point about which the series is to be developed.

Computational techniques to obtain Taylor series representation,

1. Comparing Real and Imaginary Parts of a Taylor Series
2. Differentiating/Integrating a Taylor Series
3. $f(z)g(z) = P_n(z)Q_n(z) + [z^{n+1}]$
where $f(z) = P_n(z) + [z^{n+1}]$ and $g(z) = Q_n(z) + [z^{n+1}]$.
4. $f(z)/g(z) = R_n(z) + [z^{n+1}]$
such that $P_n(z) + [z^{n+1}] = Q_n(z)R_n(z) + [z^{n+1}]$.
5. $f(g(z)) = P_n(Q_n(z)) + [z^{n+1}]$
6. $g^{-1}(w) = P_n(w) + [w^{n+1}]$
such that $P_n(Q_n(z)) + [z^{n+1}] = z$.

Definitions 11.4.1 (Legendre Polynomials). Legendre polynomials in α , $P_n(\alpha)$ are the coefficients of the Taylor series of $(1 - 2\alpha z + z^2)^{-\frac{1}{2}}$ about origin.

$$(1 - 2\alpha z + z^2)^{-\frac{1}{2}} = 1 + P_1(\alpha)z + P_2(\alpha)z^2 + \cdots \quad (11.6)$$

11.4.3 The Laurent Series (5.1.3) pp. 184

Remark. If the function $f(z)$ is analytic in the annulus $R_1 < |z - a| < R_2$, then the Laurent Series representation

$$f(z) = \sum_{n=-\infty}^{\infty} A_n(z-a)^n \quad (11.7)$$

is valid for any z in the annulus.

Proof. □

Definitions 11.4.2 (Schwarzian Derivative). Schwarzian Derivative of f is defined as,

$$\{f, z\} = \frac{f'''(z)}{f'(z)} - \frac{3}{2} \left(\frac{f''(z)}{f'(z)} \right)^2$$

The leading term in the Laurent Series representation of Schwarzian derivative should be of some importance.

Remark. Laurent series representation,

$$(e^2 - 1)^{-1} = \frac{1}{z} - \frac{1}{2} + \sum_{k=1}^{\infty} (-1)^{k-1} \frac{B_k}{(2k)!} z^{2k-1}$$

where B_k are the Bernoulli numbers.

11.5 Partial Fractions and Factorisation(5.2)

11.5.1 Partial Fractions (5.2.1) pp. 187

Theorem 11.5.1 (Mittag-Leffler). Let $\{b_\nu\}$ be a sequence of complex numbers with $\lim_{\nu \rightarrow \infty} b_\nu = \infty$, and let $P_\nu(\zeta)$ be polynomials without constant term. Then there are functions which are meromorphic in the whole plane with poles at the points b_ν and the corresponding singular parts $P_\nu(\frac{1}{z-b_\nu})$. Moreover, the most general meromorphic function of this kind can be written in the form

$$f(z) = \sum_{\nu} \left(P_\nu \left(\frac{1}{z-b_\nu} \right) - p_\nu(z) \right) + g(z) \quad (11.8)$$

where $p_\nu(z)$ are suitably chosen polynomials and $g(z)$ is analytic in the whole plane.

Proof. □

Remark. 1.

$$\frac{\pi^2}{\sin^2 \pi z} = \sum_{n=-\infty}^{\infty} \frac{1}{(z-n)^2}$$

2.

$$\pi \cot \pi z = \frac{1}{z} + \sum_{n \neq 0} \left(\frac{1}{z-n} + \frac{1}{n} \right)$$

11.5.2 Infinite Products (5.2.2) pp. 191

Definitions 11.5.1 (Infinite Product). Infinite product of complex numbers $\{p_n\}$ is defined as

$$P = \lim_{n \rightarrow \infty} P_n = \lim_{n \rightarrow \infty} \prod_{k=1}^n p_k \quad (11.9)$$

if the limit exists and is nonzero. The infinite product converges if and only if at most finitely many terms are zero and infinite product of the nonzero terms converges to a finite limit.

Remark. In a convergent infinite product, the general term $p_n \rightarrow 1$. Thus, we may represent an infinite product as

$$\prod_{n=1}^{\infty} p_n = \prod_{n=1}^{\infty} (1 + a_n) \text{ where } a_n \rightarrow 0$$

Theorem 11.5.2. *The infinite product $\prod(1 + a_n)$ with $1 + a_n \neq 0$ converges simultaneously with the series $\sum \log(1 + a_n)$ whose terms represent the value of the principal branch of the logarithm.*

Proof. □

Theorem 11.5.3. *The infinite product $\prod(1 + a_n)$ converges absolutely if and only if the respective series $\sum |a_n|$ converges.*

Proof. □

11.5.3 Canonical Products (5.2.3) pp. 193

Remark (Canonical Product). An entire function $f(z)$ may be represented using another entire function $g(z)$,

$$f(z) = z^m e^{g(z)} \prod \left(1 - \frac{z}{a_n}\right) \quad (11.10)$$

if the infinite product converges absolutely on every compact subset.

Theorem 11.5.4 (Weierstrass). *There exists an entire function with arbitrarily prescribed zeroes a_n provided that, in the case of infinitely many zeroes, $a_n \rightarrow \infty$. Every entire functions with these and no other zeroes can be written in the form*

$$f(z) = z^m e^{g(z)} \prod_{n=1}^{\infty} \left(1 - \frac{z}{a_n}\right) e^{\frac{z}{a_n} + \frac{1}{2}\left(\frac{z}{a_n}\right)^2 + \dots + \frac{1}{m_n}\left(\frac{z}{a_n}\right)^{m_n}} \quad (11.11)$$

where all the product is taken over all $a_n \neq 0$, then m_n are certain integers, and $g(z)$ is an entire function.

Proof. □

Corollary 11.5.4.1. *Every function which is meromorphic in the whole plane is the quotient of two entire functions.*

Proof. □

11.5.4 The Gamma Function (5.2.4) pp 198

Definitions 11.5.2. Euler's gamma function is defined as,

$$\Gamma(z) = \frac{e^{-\gamma z} z}{z} \prod_{n=1}^{\infty} \left(1 + \frac{z}{n}\right)^{-1} e^{\frac{z}{n}} \quad (11.12)$$

where $\gamma = \lim_{n \rightarrow \infty} \left(1 + \frac{1}{2} + \frac{1}{3} + \cdots + \frac{1}{n} - \log n\right) \approx 0.57722$ is the Euler's constant.

Remark.

$$\Gamma(z+1) = z\Gamma(z) \quad (11.13)$$

Remark.

$$\Gamma(z)\Gamma(1-z) = \frac{\pi}{\sin \pi z} \quad (11.14)$$

Remark (Legendre's Duplication Formula).

$$\sqrt{\pi}\Gamma(2z) = 2^{2z-1}\Gamma\left(z + \frac{1}{2}\right) \quad (11.15)$$

Exercise

Remark (Gauss's Formula).

$$(2\pi)^{\frac{n-1}{2}} \Gamma(z) = n^{z-\frac{1}{2}} \Gamma\left(\frac{z}{n}\right) \Gamma\left(\frac{z+1}{n}\right) \Gamma\left(\frac{z+2}{n}\right) \cdots \Gamma\left(\frac{z+n-1}{n}\right) \quad (11.16)$$

Remark.

$$\Gamma\left(\frac{1}{6}\right) = 2^{-\frac{1}{2}} \left(\frac{3}{\pi}\right)^{\frac{1}{2}} \Gamma\left(\frac{1}{3}\right)^2 \quad (11.17)$$

11.6 Entire Functions(5.3)

11.6.1 Jensen's Formula (5.3.1) pp 207

Remark (Jensen's Formula). Let f be an entire function with zeroes a_1, a_2, \dots, a_n in the interior of $|z| < \rho$. Then,

$$\log |f(0)| = - \sum_{i=1}^n \log \left(\frac{\rho}{|a_i|} \right) + \frac{1}{2\pi} \int_0^{2\pi} \log |f(\rho e^{i\theta})| d\theta \quad (11.18)$$

Proof. □

Remark (Poisson-Jensen Formula). Let f be an entire function with zeroes a_1, a_2, \dots, a_n in the interior of $|z| < \rho$ and $f(z) \neq 0$. Then,

$$\log |f(0)| = - \sum_{i=1}^n \log \left| \frac{\rho^2 - \bar{a}_i z}{\rho(z - a_i)} \right| + \frac{1}{2\pi} \int_0^{2\pi} \Re \frac{\rho e^{i\theta} + z}{\rho e^{i\theta} - z} \log |f(\rho e^{i\theta})| d\theta \quad (11.19)$$

Proof. □

11.6.2 Hadamard's Theorem (5.3.2) pp 208

Definitions 11.6.1 (Order). The order of an entire function f is defined by

$$\lambda = \overline{\lim}_{r \rightarrow \infty} \frac{\log \log M(r)}{\log r} \text{ where } M(r) = \max\{f(z) : |z| = r\} \quad (11.20)$$

Remark. The order is the smallest number such that

$$M(r) \leq e^{r^{\lambda+\varepsilon}}, \quad \forall \varepsilon > 0 \text{ and sufficiently large } r \quad (11.21)$$

Theorem 11.6.1. *The genus h and the order λ of an entire function of finite order satisfies the double inequality $h \leq \lambda \leq h + 1$.*

Proof.

□

Remark. If the genus of an entire function is fractional, then the genus h and the form of the product is uniquely determined. If the genus is integral, then there is an ambiguity.

Proof.

□

Corollary 11.6.1.1. *An entire function of fractional order assumes every finite value infinitely many times.*

Proof.

□

Module 3

11.7 The Riemann Zeta Function (5.4) pp 212**Definitions 11.7.1** (Euler Zeta Function).

$$\zeta(s) = \sum_{n=1}^{\infty} n^{-s}, \quad s > 1 \quad (11.22)$$

Definitions 11.7.2 (Riemann Zeta Function). The Riemann Zeta Function given by,

$$\zeta(s) = \sum_{n=1}^{\infty} n^{-s}, \quad \Re(s) = \sigma > 1 \quad (s = \sigma + it) \quad (11.23)$$

is analytic in the half plane $\sigma > 1$.**11.7.1 The Product Development (5.4.1) pp 213****11.7.2 Extension of $\zeta(s)$ to the Whole Plane (5.4.2) pp 214****11.7.3 The Functional Equation (5.4.3) pp 218****11.7.4 The Zeroes of the Zeta Function (5.4.4) pp 218****11.8 Normal Families (5.5) pp 219****11.8.1 Equicontinuity (5.5.1) pp 219****11.8.2 Normality and Compactness (5.5.2) pp 220****11.8.3 Arzela's Theorem (5.5.3) pp 222**

Module 4

11.9 The Riemann Mapping Theorem(6.1)

11.9.1 Statement and Proof (6.1.1) pp 229

11.9.2 Boundary Behaviour (6.1.2) pp 232

Theorem 11.9.1. *Let $f : \Omega \rightarrow \Omega'$ be a surjective, conformal mapping. If Ω and Ω' are Jordan regions, then f can be extended to a topological mapping from the closure of Ω onto the closure of Ω' .*

Proof. Out of scope. □

Definitions 11.9.1. Let sequence $\{z_n\}$ be a sequence in Ω . The **sequence tends to the boundary of Ω** if the points eventually stays away from any point in Ω .

In other words, the sequence $\{z_n\}$ tends to the boundary of Ω if for any point $z_0 \in \Omega$ and any neighbourhood $|z - z_0| < \varepsilon$, there exists a natural number N such that $\forall n > N$, $|z_n - z_0| > \varepsilon$.

Definitions 11.9.2. An arc $z(t)$ **tends to the boundary of Ω** if $z(t)$ eventually stays away from any point in Ω .

In other words, the arc $z : [a, b] \rightarrow \mathbb{C}$ tends to the boundary of Ω if $\forall \varepsilon > 0$ there exists $t_0 \in [a, b]$ such that $\forall t > t_0, (t \in [a, b]) \forall z \in \Omega, |z(t) - z| > \varepsilon$.

Theorem 11.9.2. *Let f be a topological mapping of a region Ω onto Ω' . If a sequence $\{z_n\}$ or an arc $z(t)$ tends to the boundary of Ω , then the sequence $\{f(z_n)\}$ or the arc $f(z(t))$ tends to the boundary of Ω' .*

Proof. Suppose sequence $\{z_n\}$ is a sequence which tends to boundary of Ω . Let K be a compact subset of Ω' . Then $f^{-1}(K)$ is a compact subset of Ω (which is preserved by the continuous function).

The family of balls with center at $z \in K$ and radius ε is a cover of K . The radius of these balls may depend on the point $z \in K$. Since $f^{-1}(K)$ is compact, the above cover has a finite subcover. Thus, there are finite balls with center z_1, z_2, \dots, z_k and radius $\varepsilon_1, \varepsilon_2, \dots, \varepsilon_k$. Since $\{z_n\}$ stays away from each of these points, there exists natural numbers N_1, N_2, \dots, N_k such that $|z_n - z_j| > \varepsilon_j$ for every $n > N_k$ and $j = 1, 2, \dots, k$. Define $n_0 = \max \{N_1, N_2, \dots, N_k\}$ and $\varepsilon_0 = \min \{\varepsilon_1, \varepsilon_2, \dots, \varepsilon_k\}$. Then $|z_n - z_j| > \varepsilon_0$ for every $n > n_0$ and $j = 1, 2, \dots, k$.

Clearly, $z_n \notin K$ for $n > n_0$ and $f(z_n) \notin K$ for $n > n_0$. Thus, the sequence $\{f(z_n)\}$ stays away from each point in K . Since K is arbitrary, the sequence stays away from every compact subset of Ω' . Thus, the sequence stays away from any point in Ω' . Clearly, the sequence tends to the boundary of Ω' . □

11.9.3 Use of the Reflection Principle (6.1.3) pp 233

Definitions 11.9.3 (free boundary arc). Let Ω be a region and γ be a arc. Arc γ is a free boundary arc if every point of γ has a neighbourhood whose intersection with $\partial\Omega$ is same as its intersection with γ .

Definitions 11.9.4 (one-sided/two-sided boundary arc). Let γ be a free-boundary arc with respect to Ω . Then each point $z(t_0)$ on γ has a disk about $z(t_0)$ such that γ divides the disk into two half disks in such a way that at least one of the half disks is contained in Ω .

A point on γ is a one-sided boundary point if one half disk is outside Ω and two-side boundary point if both half disks are inside Ω .

A arc is one-side boundary arc if all its points are one-sided boundary points. And two-sided boundary arc if all its points are two-sided boundary points.

Theorem 11.9.3. *Let Ω be a simply connect region. Let γ be line segment in $\partial\Omega$ and is a one-sided boundary arc. Then the function which maps Ω onto the unit disk can be extended to a function analytic and one to one on $\Omega \cup \gamma$. The image of γ is an arc γ' on the unit circle.*

Proof. Let f be a conformal mapping from Ω onto unit disk, normalised by the condition $f(z_0) = 0$. Consider a disk about $x_0 \in \gamma$ such that the half disk in Ω does not contain z_0 with $f(z_0) = 0$. Then $\log f(z)$ has a single valued branch in the half disk, and its real part tends to 0 as z tends to the diameter.³ From the boundary behaviour of the conformal map f suggested by Riemann mapping theorem, we know that “as z tends to the boundary of Ω , $f(z)$ tends to the boundary of the unit disk.” Thus, $|f(z)|$ tends to 1.

By reflection principle, $\log f(z)$ has an analytic extension to the whole disk. Therefore, $\log f(z)$ is analytic at x_0 . Clearly, $f(z)$ is analytic at x_0 . Suppose $x_1 \in \gamma$ belongs to the disk. Then the extension to the neighbourhood of x_1 coincides with the extension to the neighbourhood of x_0 on their intersection and f is analytic at x_1 . Continuing like this, $f(z)$ can be extended to $\Omega \cup \gamma$ preserving its analyticity.

We claim that, $f'(z) \neq 0$ for any $z \in \gamma$. Suppose $f'(x_0) = 0$ for some $x_0 \in \gamma$. Then $f(x_0)$ is a branch point and the two subarcs of γ at x_0 form an angle π/n where $n \geq 2$. Then the unit circle should have two arcs at π/n angles which is impossible.

Without loss of generality suppose that the upper half disks are in Ω . Then

$$\frac{\partial \log |f|}{\partial y} = -\frac{\partial \arg f}{\partial x} < 0$$

on γ . Thus, γ and $\arg f$ move in the same direction. Therefore, f is injective on γ . \square

Remark. For two-side (boundary) arcs —to continue— find ‘obvious’ modifications

³If the intersection of neighbourhood of x_0 with $\partial\Omega$ is its intersection with γ , then the intersection is a diameter of the disk

11.9.4 Analytic Arcs (6.1.4) pp 234

Theorem 11.9.4. *If the boundary of Ω contains a free one-sided analytic arc γ , then the mapping function has an analytic extension to $\Omega \cup \gamma$, and γ is mapped on an arc of the unit circle.*

Proof.

□

11.10 The Weierstrass Theory(7.3)

11.10.1 The Weierstrass \wp function (7.3.1) pp 272

11.10.2 The Functions $\zeta(z)$ and $\sigma(z)$ (7.3.2) pp 273

11.10.3 The Differential Equation (7.3.3) pp 275

Subject 12

**ME010302 Partial
Differential Equations**

Subject 13

ME010303 Multivariate Calculus & Integral Transforms

13.1 Integral Transforms

13.1.1 The Weierstrass Approximation Theorem

Every continuous, real valued function on a compact interval has a polynomial approximation. [Apostol, 1973, Theorem 11.17]

Theorem 13.1.1 (Weierstrass). *Let f be a real-valued, continuous function on a compact interval $[a, b]$. Then for every $\epsilon > 0$, there is a polynomial p such that $|f(x) - p(x)| < \epsilon$ for every $x \in [a, b]$.*

Synopsis. Given a real-valued continuous function on compact interval $[a, b]$, we can construct a real-valued, continuous function g on \mathbb{R} which is periodic with period 2π . We have, if $f \in L(I)$ and f is bounded almost everywhere in I , then $f \in L^2(I)$. [Apostol, 1973, Theorem 10.52]. By Fejer's theorem ([Apostol, 1973, Theorem 11.15]), the fourier series generated by g ([Apostol, 1973, definition 11.3]) converges to the Cesaro sum ([Apostol, 1973, Definition 8.47]), which is g itself in this case. Thus for any $\epsilon > 0$, there is a finite sum of trigonometric functions. The power series expansions of trigonometric functions ([Apostol, 1973, definition 9.27]) being uniformly convergent, there exists a polynomial p_m which approximates g . And we can construct p (polynomial approximation of g) using p_m .

Proof. Define $g : \mathbb{R} \rightarrow \mathbb{R}$,

$$g(t) = \begin{cases} f(a + (b-a)t/\pi), & t \in [0, \pi) \\ f(a + (2\pi - t)(b-a)/\pi), & t \in [\pi, 2\pi] \\ g(t - 2n\pi), & t > 2\pi, n \in \mathbb{N} \\ g(t + 2n\pi), & t < 0, n \in \mathbb{N} \end{cases}$$

Thus g is a continuous, real-valued, periodic function with period 2π such that

$$f(x) = g\left(\frac{\pi(x-a)}{b-a}\right), \quad x \in [a, b] \quad (13.1)$$

The fourier series generated by g is given by,

$$g(t) \sim \frac{a_0}{2} + \sum_{k=1}^{\infty} (a_k \cos kt + b_k \sin kt)$$

$$\text{where } a_k = \frac{1}{\pi} \int_0^{2\pi} f(t) \cos kt \, dt, \quad b_k = \frac{1}{\pi} \int_0^{2\pi} f(t) \sin kt \, dt$$

Let $\{s_n(t)\}$ be the sequence of partial sums of the fourier series generated by g . And $\{\sigma_n(t)\}$ be the sequence of averages of $s_n(t)$ given by,

$$\sigma_n(t) = \frac{1}{n} \sum_{k=1}^n s_k(t), \quad \text{where } s_k(t) = \frac{a_0}{2} + \sum_{j=1}^k (a_j \cos jt + b_j \sin jt)$$

Function $f \in L(I)$ being real-valued continuous function on a compact interval, it is bounded and hence is Lebesgue square integrable. ie, $f \in L^2(I)$. Thus, $g \in L^2(I)$.

Since g is continous on \mathbb{R} , the function $s : \mathbb{R} \rightarrow \mathbb{R}$ defined by,

$$s(t) = \lim_{h \rightarrow 0^+} \frac{g(t+h) - g(t-h)}{2}$$

is well-defined on \mathbb{R} and $s(t) = g(t)$, $\forall t \in \mathbb{R}$.

Then by Fejer's Theorem, the sequence $\{\sigma_n(t)\}$ converges uniformly to $g(t)$ for every $t \in \mathbb{R}$. Thus, given $\epsilon > 0$, there exists $N \in \mathbb{N}$ such that $\forall t \in \mathbb{R}$, $|g(t) - \sigma_N(t)| < \frac{\epsilon}{2}$.

We have,

$$\sigma_N(t) = \sum_{k=0}^N (A_k \cos kt + B_k \sin kt), \quad \text{where } A_k, B_k \in \mathbb{R} \quad (13.2)$$

By the power series expansion of the trigonometric functions about origin,

$$\cos kt = \sum_{j=1}^{\infty} \left(\frac{\cos^{(j)} 0}{j!} (kt)^j \right) = \sum_{j=1}^{\infty} A'_j t^j \quad \text{where } A'_j \in \mathbb{R} \quad (13.3)$$

$$\sin kt = \sum_{j=1}^{\infty} \left(\frac{\sin^{(j)} 0}{j!} (kt)^j \right) = \sum_{j=1}^{\infty} B'_j t^j \quad \text{where } B'_j \in \mathbb{R} \quad (13.4)$$

Since the above power series expansions of trigonometric functions are uniformly convergent, their finite linear combination $\{\sigma_N(t)\}$ is also uniformly convergent. ie, Given $\epsilon > 0$ there exists $m \in \mathbb{N}$ such that for every $t \in \mathbb{R}$

$$\left| \sum_{k=0}^m C_k t^k - \sigma_N(t) \right| < \frac{\epsilon}{2} \quad \text{where } C_k \in \mathbb{R}$$

Therefore, $|p_m(t) - g(t)| \leq |p_m(t) - \sigma_N(t)| + |\sigma_N(t) - g(t)| < \epsilon$ where $p_m(t) = \sum_{k=0}^m C_k t^k$. Define $p : [a, b] \rightarrow \mathbb{R}$ by,

$$p(x) = p_m \left(\frac{\pi(x-a)}{b-a} \right) \quad (13.5)$$

By equations 13.1 and 13.5, $|p(x) - f(x)| < \epsilon$ for every $x \in [a, b]$. \square

13.1.2 Other Forms of Fourier Series

Let $f \in L([0, 2\pi])$, then the fourier series generated by f is given by,

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} (a_n \cos nx + b_n \sin nx)$$

$$\text{where } a_n = \frac{1}{\pi} \int_0^{2\pi} f(t) \cos nt \, dt, \quad b_n = \frac{1}{\pi} \int_0^{2\pi} f(t) \sin nt \, dt$$

By Euler's formula $e^{inx} = \cos nx + i \sin nx$. We have, $\cos nx = \frac{(e^{inx} + e^{-inx})}{2}$ and $\sin nx = \frac{(e^{inx} - e^{-inx})}{2i}$

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} (\alpha_n e^{inx} + \beta_n e^{-inx})$$

$$\text{where } \alpha_n = \frac{(a_n - ib_n)}{2} \quad \beta_n = \frac{(a_n + ib_n)}{2}$$

Therefore, by assigning $\alpha_0 = a_0/2$, $\alpha_{-n} = \beta_n$, we get the following exponential form of fourier series generated by f ,

$$f(x) \sim \sum_{n=-\infty}^{\infty} \alpha_n e^{inx} \text{ where } \alpha_n = \frac{1}{2\pi} \int_0^{2\pi} f(t) e^{-int} \, dt$$

Note : If f is periodic with period 2π , then the interval of integration $[0, 2\pi]$ can be replaced with any interval of length 2π . eg. $[-\pi, \pi]$

Periodic with period p

Let $f \in L([0, p])$ and f is periodic with period p . Then

$$f(x) \sim \frac{a_0}{2} + \sum_{n=1}^{\infty} \left(a_n \cos \frac{2\pi nx}{p} + b_n \sin \frac{2\pi nx}{p} \right)$$

$$\text{where } a_n = \frac{2}{p} \int_0^p f(t) \cos \frac{2\pi nt}{p} \, dt \quad b_n = \frac{2}{p} \int_0^p f(t) \sin \frac{2\pi nt}{p} \, dt$$

Therefore, we have the exponential form of the above fourier series given by,

$$f(x) \sim \sum_{n=-\infty}^{\infty} \alpha_n e^{\frac{2\pi inx}{p}}, \text{ where } \alpha_n = \frac{1}{p} \int_0^p f(t) e^{\frac{-2\pi int}{p}} \, dt$$

13.1.3 Fourier Integral Theorem

Theorem 13.1.2 (Fourier Integral Theorem). *Let $f \in L(-\infty, \infty)$. Suppose $x \in \mathbb{R}$ and an interval $[x - \delta, x + \delta]$ about x such that either*

1. *f is of bounded variation on an interval $[x - \delta, x + \delta]$ about x or*
2. *both limits $f(x+)$ and $f(x-)$ exists and both Lebesgue integrals*

$$\int_0^\delta \frac{f(x+t) - f(x+)}{t} dt \text{ and } \int_0^\delta \frac{f(x-t) - f(x-)}{t} dt$$

exists.

Then,

$$\frac{f(x+) + f(x-)}{2} = \frac{1}{\pi} \int_0^\infty \int_{-\infty}^\infty f(u) \cos v(u-x) du dv,$$

the integral \int_0^∞ being an improper Riemann integral.

Synopsis.

$$f(x+t) \frac{\sin \alpha t}{\pi t} dt \rightarrow f(u) \frac{\sin \alpha(u-x)}{\pi(u-x)} \rightarrow \frac{f(u)}{\pi} \int_0^\alpha \cos v(u-x) dv$$

By Riemann-Lebesgue lemma [Apostol, 1973, Theorem 11.6],

$$f \in L(I) \implies \lim_{\alpha \rightarrow +\infty} \int_I f(x) \sin \alpha t dt = 0$$

By Jordan's Theorem [Apostol, 1973, Theorem 10.8], if g is of bounded variation on $[0, \delta]$, then

$$\lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^\delta g(t) \frac{\sin \alpha t}{t} dt = g(0+)$$

By Dini's Theorem [Apostol, 1973, Theorem 10.9], if the limit $g(x+)$ exists and Lebesgue integral $\int_0^\delta \frac{g(t)+g(0+)}{t} dt$ exists for some $\delta > 0$, then

$$\lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^\delta g(t) \frac{\sin \alpha t}{t} dt = g(0+)$$

The order of Lebesgue integrals can be interchanged. [Apostol, 1973, Theorem 10.40]

Suppose $f \in L(X)$ and $g \in L(Y)$. Then

$$\int_X f(x) \left(\int_Y g(y) k(x, y) dy \right) dx = \int_Y g(y) \left(\int_X f(x) k(x, y) dx \right) dy$$

Proof. Consider $\int_{-\infty}^\infty f(x+t) \frac{\sin \alpha t}{\pi t} dt$. We prove that this integral is equal to the either sides.

$$\int_{-\infty}^\infty f(x+t) \frac{\sin \alpha t}{\pi t} dt = \int_{-\infty}^{-\delta} + \int_{-\delta}^0 + \int_0^{-\delta} + \int_{\delta}^\infty f(x+t) \frac{\sin \alpha t}{\pi t} dt$$

We have, function $\frac{f(x+t)}{\pi t}$ is bounded on $(-\infty, -\delta) \cup (\delta, \infty)$, hence $\frac{f(x+t)}{\pi t}$ is Lebesgue integrable on $(-\infty, -\delta) \cup (\delta, \infty)$.

By Riemann Lebesgue lemma,

$$\frac{f(x+t)}{\pi t} \in L(-\infty, -\delta) \implies \int_{-\infty}^{-\delta} f(x+t) \frac{\sin \alpha t}{\pi t} dt = 0,$$

$$\frac{f(x+t)}{\pi t} \in L(\delta, \infty) \implies \int_{\delta}^{\infty} f(x+t) \frac{\sin \alpha t}{\pi t} dt = 0$$

Case 1 Suppose f is of bounded variation on $[x-\delta, x+\delta]$, put $g(t) = f(x+t)$ then g is of bounded variation on $[-\delta, \delta]$. Thus g is of bounded variation on $[0, \delta]$. Then by Jordan's Theorem

$$\lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^{\delta} f(x+t) \frac{\sin \alpha t}{t} dt = \lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^{\delta} g(t) \frac{\sin \alpha t}{t} dt = g(0+) = f(x+)$$

Case 2 Suppose both the limits $f(x+)$ and $f(x-)$ exists and both Lebesgue integrals

$$\int_0^{\delta} \frac{f(x+t) - f(x+)}{t} dt \text{ and } \int_0^{\delta} \frac{f(x-t) - f(x-)}{t} dt$$

exists.

Thus, we have $f(x+)$ exists and the Lebesgue integral $\int_0^{\delta} \frac{f(x+t) - f(x+)}{t} dt$ exists. Put $g(t) = f(x+t)$, then $g(0+) = f(x+)$ exists and the Lebesgue integral $\int_0^{\delta} \frac{g(t) - g(0+)}{t} dt$ exists, then by Dini's Theorem,

$$\lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^{\delta} f(x+t) \frac{\sin \alpha t}{t} dt = \lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^{\delta} g(t) \frac{\sin \alpha t}{t} dt = g(0+) = f(x+)$$

Similarly, $f(x-)$ exists and the Lebesgue integral $\int_0^{\delta} \frac{f(x-t) - f(x-)}{t} dt$ exists. Put $g(t) = f(x-t)$, then $g(0+) = f(x-)$ exists and the Lebesgue integral $\int_0^{\delta} \frac{g(t) - g(0+)}{t} dt$ exists, then by Dini's Theorem,

$$\begin{aligned} \lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_{-\delta}^0 f(x+t) \frac{\sin \alpha t}{t} dt &= \lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^{\delta} f(x-\tau) \frac{\sin \alpha \tau}{\tau} d\tau \\ &= \lim_{\alpha \rightarrow +\infty} \frac{2}{\pi} \int_0^{\delta} g(\tau) \frac{\sin \alpha \tau}{\tau} d\tau = g(0+) = f(x-) \end{aligned}$$

Then by either cases,

$$\begin{aligned} \lim_{\alpha \rightarrow +\infty} \int_{-\infty}^{\infty} f(x+t) \frac{\sin \alpha t}{\pi t} dt &= \lim_{\alpha \rightarrow +\infty} \int_{-\delta}^0 + \int_0^{\delta} f(x+t) \frac{\sin \alpha t}{\pi t} dt \\ &= \frac{f(x+) + f(x-)}{2} \end{aligned}$$

We have, $\int_0^\alpha \cos v(u-x)dv = \frac{\sin v(u-x)}{u-x}$.

$$\begin{aligned} \lim_{\alpha \rightarrow +\infty} \int_{-\infty}^{\infty} f(x) \frac{\sin \alpha t}{\pi t} dt &= \lim_{\alpha \rightarrow +\infty} \int_{-\infty}^{\infty} f(u) \frac{\sin \alpha(u-x)}{u-x} du, \text{ (put } u = x+t) \\ &= \lim_{\alpha \rightarrow +\infty} \int_{-\infty}^{\infty} f(u) \left(\int_0^\alpha \cos v(u-x)dv \right) du \\ &= \lim_{\alpha \rightarrow +\infty} \int_0^\alpha \left(\int_{-\infty}^{\infty} f(u) \cos v(u-x)du \right) dv, \end{aligned}$$

since, the order of Lebesgue integrals can be reversed.

$$= \int_0^\infty \left(\int_{-\infty}^{\infty} f(u) \cos v(u-x)du \right) dv$$

where, \int_0^∞ is not a Lebesgue integral, but an improper Riemann integral

Therefore,

$$\begin{aligned} \int_0^\infty \left(\int_{-\infty}^{\infty} f(u) \cos v(u-x)du \right) dv &= \lim_{\alpha \rightarrow +\infty} \int_{-\infty}^{\infty} f(x) \frac{\sin \alpha t}{\pi t} dt \\ &= \frac{f(x+) + f(x-)}{2} \end{aligned}$$

□

Remark. If a function f on $(-\infty, \infty)$ is non-periodic, then it may not have a fourier series representation. In such cases, we have fourier intergral representation.

13.1.4 Exponential form of Fourier Integral Theorem

Let $f \in L(-\infty, \infty)$. Suppose $x \in \mathbb{R}$ and an interval $[x - \delta, x + \delta]$ about x such that either

1. f is of bounded variation on an interval $[x - \delta, x + \delta]$ about x or
2. both limits $f(x+)$ and $f(x-)$ exists and both Lebesgue intergrals

$$\int_0^\delta \frac{f(x+t) - f(x+)}{t} dt \text{ and } \int_0^\delta \frac{f(x-t) - f(x-)}{t} dt$$

exists.

Then,

$$\frac{f(x+) + f(x-)}{2} = \lim_{\alpha \rightarrow \infty} \frac{1}{2\pi} \int_{-\alpha}^{\alpha} \left(\int_{-\infty}^{\infty} f(u) e^{iv(u-x)} du \right) dv$$

Proof. Let $F(v) = \int_{-\infty}^{\infty} f(u) \cos v(u-x)du$. Then $F(v) = F(-v)$ and

$$\begin{aligned} \lim_{\alpha \rightarrow \infty} \frac{1}{2\pi} \int_{-\alpha}^{\alpha} F(v) dv &= \lim_{\alpha \rightarrow \infty} \frac{1}{\pi} \int_0^\alpha \int_{-\infty}^{\infty} f(u) \cos v(u-x) du dv \\ &= \frac{f(x+) + f(x-)}{2} \end{aligned}$$

Let $G(v) = \int_{-\infty}^{\infty} f(u) \sin v(u-x) du$. Then $G(v) = -G(-v)$ and

$$\lim_{\alpha \rightarrow \infty} \frac{1}{2\pi} \int_{-\alpha}^{\alpha} G(v) dv = 0$$

Thus

$$\lim_{\alpha \rightarrow \infty} \frac{1}{2\pi} \int_{-\alpha}^{\alpha} F(v) + iG(v) dv = \frac{f(x+) + f(x-)}{2}$$

□

13.1.5 Integral Transforms

Definitions 13.1.1. Integral transform $g(y)$ of $f(x)$ is a Lebesgue integral or Improper Riemann integral of the form

$$g(y) = \int_{-\infty}^{\infty} K(x, y) f(x) dx$$

, where K is the kernel of the transform. We write $g = \mathcal{K}(f)$.

Remark. Integral transforms(operators) are linear operators. ie, $\mathcal{K}(af_1 + bf_2) = a\mathcal{K}f_1 + b\mathcal{K}f_2$

Remark. A few commonly used integral transforms,

1. Exponential Fourier Transform \mathcal{F} ,

$$\mathcal{F}f = \int_{-\infty}^{\infty} e^{-ixy} f(x) dx$$

2. Fourier Cosine Transform \mathcal{C} ,

$$\mathcal{C}f = \int_0^{\infty} \cos xy f(x) dx$$

3. Fourier Sine Transform \mathcal{S} ,

$$\mathcal{S}f = \int_0^{\infty} \sin xy f(x) dx$$

4. Laplace Transform \mathcal{L} ,

$$\mathcal{L}f = \int_0^{\infty} e^{-xy} f(x) dx$$

5. Mellin Transform \mathcal{M} ,

$$\mathcal{M}f = \int_0^{\infty} x^{y-1} f(x) dx$$

Remark. Suppose $f(x) = 0, \forall x < 0$.

$$\int_{-\infty}^{\infty} e^{-ixy} f(x) dx = \int_0^{\infty} e^{-ixy} f(x) dx = \int_0^{\infty} \cos xy f(x) dx + i \int_0^{\infty} \sin xy f(x) dx$$

$$\mathcal{F}f = \mathcal{C}f + i\mathcal{S}f$$

Therefore Fourier Cosine \mathcal{C} and Sine \mathcal{S} transforms are special cases of fourier integral transform, \mathcal{F} provided f vanishes on negative real axis.

Remark. Let $y = u + iv$, $f(x) = 0$, $\forall x < 0$.

$$\int_0^\infty e^{-xy} f(x) dx = \int_0^\infty e^{-xu} e^{-ixv} f(x) dx = \int_0^\infty e^{-ixv} \phi_u(x) dx$$

where $\phi_u(x) = e^{-xu} f(x)$.

$$\mathcal{L}f = \mathcal{F}\phi_u$$

Therefore Laplace transform, \mathcal{L} is a special case of Fourier integral transform, \mathcal{F} .

Remark. Let $g(y) = \mathcal{F}f(x)$.

$$g(y) = \int_{-\infty}^\infty e^{-ixy} f(x) dx$$

Suppose f is continuous at x , then by fourier integral theorem,

$$\begin{aligned} f(x) &= \frac{1}{2\pi} \int_{-\infty}^\infty \left(\int_{-\infty}^\infty f(u) e^{iv(u-x)} du \right) dv \\ &= \int_{-\infty}^\infty e^{-ivx} \left(\frac{1}{2\pi} \int_{-\infty}^\infty e^{ivu} f(u) du \right) dv \\ &= \int_{-\infty}^\infty g(v) e^{-ivx} dv = \mathcal{F}g \text{ where } g(v) = \frac{1}{2\pi} \int_{-\infty}^\infty f(u) e^{ivu} du \end{aligned}$$

The above function $g(v)$ gives the **inverse fourier transformation** of f .

Let g be fourier transform of f , then f is uniquely determined by its fourier transform g by,

$$f(x) = \mathcal{F}^{-1}g(y) = \frac{1}{2\pi} \lim_{\alpha \rightarrow \infty} \int_{-\alpha}^\alpha g(y) e^{ixy} dy$$

6. Inverse Fourier Transform \mathcal{F}^{-1} ,

$$\mathcal{F}^{-1}f = \int_{-\infty}^\infty \frac{e^{ixy}}{2\pi} f(y) dy$$

13.1.6 Convolutions

Definitions 13.1.2. Let $f, g \in L(-\infty, \infty)$. Let S be the set of all points x for which the Lebesgue integral

$$h(x) = \int_{-\infty}^\infty f(t)g(x-t)dt$$

exists. Then the function $h : S \rightarrow \mathbb{R}$ is a **convolution** of f and g . And $h = f * g$.

Remark. Convolution operator is commutative. ie, $h = f * g = g * f$ (hint : take $u = x - t$)

Remark. Suppose f, g vanishes on negative real axis, then

$$h(x) = \int_{-\infty}^\infty f(t)g(x-t)dt = \int_{-\infty}^0 + \int_0^x + \int_x^\infty f(t)g(x-t)dt = \int_0^x f(t)g(x-t)dt$$

Remark. Singularity of convolution is a point at which the convolution integral fails to exist.

Theorem 13.1.3. Let $f, g \in L(\mathbb{R})$ and either f or g is bounded in \mathbb{R} . Then the convolution integral

$$h(x) = \int_{-\infty}^{\infty} f(t)g(x-t)dt$$

exists for every $x \in \mathbb{R}$ and the function h so defined is bounded in \mathbb{R} . In addition, if the bounded function is continuous on \mathbb{R} , then h is continuous and $h \in L(\mathbb{R})$.

Synopsis.

Proof. □

Remark. If f, g are both unbounded, the convolution integral may not exist.

$$\text{eg: } f(t) = \frac{1}{\sqrt{t}}, \quad g(t) = \frac{1}{\sqrt{1-t}}$$

Theorem 13.1.4. Let $f, g \in L^2(\mathbb{R})$. Then the convolution integral $f * g$ exists for each $x \in \mathbb{R}$ and the function $h : \mathbb{R} \rightarrow \mathbb{R}$ defined by $h(x) = f * g(x)$ is bounded in \mathbb{R} .

Synopsis.

Proof. □

13.1.7 The Convolution Theorem for Fourier Transforms

Theorem 13.1.5. Let $f, g \in L(\mathbb{R})$ and at least one of f or g is continuous and bounded on \mathbb{R} . Let $h = f * g$. Then for every real u ,

$$\int_{-\infty}^{\infty} h(x)e^{-ixu}dx = \left(\int_{-\infty}^{\infty} f(t)e^{-itu}dt \right) \left(\int_{-\infty}^{\infty} g(y)e^{-iyu}dy \right)$$

The integral on the left exists both as a Lebesgue integral and an improper Riemann integral.

Synopsis.

Proof. □

Remark (Application of Convolution Theorem).

$$B(p, q) = \frac{\Gamma p \Gamma q}{\Gamma p + q}, \text{ where } B(p, q) = \int_0^1 x^{p-1}(1-x)^{q-1}dx, \quad \Gamma p = \int_0^{\infty} t^{p-1}e^{-t}dt$$

13.2 Multivariate Differential Calculus

In this chapter, we deal with real functions of several variables. Instead of \mathbf{c} , we write $\bar{c} \in \mathbb{R}^n$, then $\bar{c} = (c_1, c_2, \dots, c_n)$ where $c_j \in \mathbb{R}$ for every $j = 1, 2, \dots, n$. Again, suppose $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ and $f(\bar{x}) = \bar{y}$, then $\bar{y} = (y_1, y_2, \dots, y_m)$ where each y_k is real. The unit co-ordinate vector, \bar{u}_k is given by $u_{kj} = \delta_{j,k}$

13.2.1 Directional Derivative

Motivation : The existence of all partial derivatives of a multivariate real function f at a point \bar{c} doesn't imply the continuity of f at \bar{c} . Thus, we need a suitable generalisation for the partial derivative which could characterise continuity. And directional derivative is such an attempt.

Definitions 13.2.1 (Directional Derivative). Let $S \subset \mathbb{R}^n$ and $f : S \rightarrow \mathbb{R}^m$. Let \bar{c} be an interior points of S and $\bar{u} \in \mathbb{R}^n$, then there exists an open ball $B(\bar{c}, r)$ in S . Also for some $\delta > 0$ the line segment $\alpha : [0, \delta] \rightarrow S$ given by $\alpha(t) = \bar{c} + t\bar{u}$ lie in $B(\bar{c}, r)$. Then the **directional derivative** of f at an interior point \bar{c} in the direction \bar{u} is given by

$$f'(\bar{c}, \bar{u}) = \lim_{h \rightarrow 0} \frac{f(\bar{c} + h\bar{u}) - f(\bar{c})}{h}$$

Remark. The direction derivative of f at an interior point \bar{c} in the direction \bar{u} exists only if the above limit exists.

Remark. Example, [Apostol, 1973, Exercise 12.2a]

Suppose $\bar{x}, \bar{a}, \bar{c}, \bar{u} \in \mathbb{R}^n$. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ such that $f(\bar{x}) = \bar{a} \cdot \bar{x}$. Then

$$f'(\bar{c}, \bar{u}) = \lim_{h \rightarrow 0} \frac{\bar{a} \cdot (\bar{c} + h\bar{u}) - \bar{a} \cdot \bar{c}}{h} = \bar{a} \cdot \bar{u}$$

Remark (Properties). Let $f : S \rightarrow \mathbb{R}^m$, where $S \subset \mathbb{R}^n$

1. $f'(\bar{c}, \bar{0}) = \bar{0}$

Note : The zero vectors belongs to $\mathbb{R}^n, \mathbb{R}^m$ respectively.

2. $f'(\bar{c}, \bar{u}_k) = \frac{\partial f}{\partial u_k}(\bar{c}) = D_k f(\bar{c})$, the k^{th} partial derivative of f .

3. Let $f = (f_1, f_2, \dots, f_m)$, such that $f(\bar{c}) = (f_1(\bar{c}), f_2(\bar{c}), \dots, f_m(\bar{c}))$. Then,

$$\exists f'(\bar{c}, \bar{u}) \iff \forall k, \exists f'_k(\bar{c}, \bar{u}) \text{ and } f'(\bar{c}, \bar{u}) = (f'_1(\bar{c}, \bar{u}), f'_2(\bar{c}, \bar{u}), \dots, f'_m(\bar{c}, \bar{u}))$$

ie, Directional derivative of f exists iff directional derivative of each component function f_k exists. And the components of the directional derivatives of f are the directional derivatives of the components of f .

Thus $D_k f(\bar{c}) = (D_k f_1(\bar{c}), D_k f_2(\bar{c}), \dots, D_k f_m(\bar{c}))$ holds.

4. Let $F(t) = f(\bar{c} + t\bar{u})$, then $F'(0) = f'(\bar{c}, \bar{u})$ and $F'(t) = f'(\bar{c} + t\bar{u}, \bar{u})$
5. Let $f(\bar{c}) = \bar{c} \cdot \bar{c} = \|\bar{c}\|^2$, and $F(t) = f(\bar{c} + t\bar{u})$, then $F'(t) = 2\bar{c} \cdot \bar{u} + 2t\|\bar{u}\|^2$ and $F'(0) = f'(\bar{c}, \bar{u}) = 2\bar{c} \cdot \bar{u}$
6. Let f be linear, then $f'(\bar{c}, \bar{u}) = f(\bar{u})$
7. Existence of all partial derivatives doesn't imply existence of all directional derivatives.

$$f(x, y) = \begin{cases} x + y & \text{if } x = 0 \text{ or } y = 0 \\ 1 & \text{otherwise} \end{cases}$$

For above f , directional derivatives exists only along the co-ordinates (ie, partial derivatives).

8. Existence of all directional derivatives doesn't imply continuity.

$$f(x, y) = \begin{cases} xy^2(x^2 + y^4) & x \neq 0 \\ 0 & x = 0 \end{cases}$$

Above f is discontinuous at $(0, 0)$, however all directional derivatives exists and has finite value.

13.2.2 Total Derivative

We may define a total derivative $T_c(h) = hf'(c)$ in the case of real-functions of single variable as follows :-

$$\text{Let } E_c(h) = \begin{cases} \frac{f(c+h)-f(c)}{h} - f'(c), & h \neq 0 \\ 0, & h = 0 \end{cases}$$

Then, $f(c+h) = f(c) + hf'(c) + hE_c(h)$ and as $h \rightarrow 0$, $E_c(h) \rightarrow 0$. Also $T_c(h) = f'(c)h$ is a linear function of h . ie, $T_c(ah_1 + bh_2) = aT_c(h_1) + bT_c(h_2)$. Now, we will define a total derivative of multivariate function that has these two properties.

Definitions 13.2.2 (Total Derivative). The function $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ is **differentiable** at \bar{c} if there exists a **linear** function $T_{\bar{c}} : \mathbb{R}^n \rightarrow \mathbb{R}^m$ such that $f(\bar{c} + \bar{v}) = f(\bar{c}) + T_{\bar{c}}(\bar{v}) + \|\bar{v}\|E_{\bar{c}}(\bar{v})$ where $E_{\bar{c}}(\bar{v}) \rightarrow \bar{0}$ as $\bar{v} \rightarrow \bar{0}$.

Remark. The linear function $T_{\bar{c}}$ is the total derivative of f at \bar{c} , $T_{\bar{c}}(\bar{0}) = \bar{0}$ and the condition above gives the First Order Taylor's Formula for $f(\bar{c} + \bar{v}) - f(\bar{c})$.

Remark (Properties). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ and $f'(\bar{c})(\bar{v}) = T_{\bar{c}}(\bar{v})$ be the total derivative of f at \bar{c} evaluated at \bar{v} . Then,

1. $f'(\bar{c})(\bar{v}) = f'(\bar{c}, \bar{u})$
2. If f is differentiable at \bar{c} , then f is continuous at \bar{c} .
3. $f'(\bar{c})(\bar{v}) = v_1 D_1 f(\bar{c}) + v_2 D_2 f(\bar{c}) + \cdots + v_n D_n f(\bar{c})$

Note. The above f' is a function from \mathbb{R}^n to the set of all linear functions $\mathcal{L} = \{h : \mathbb{R}^n \rightarrow \mathbb{R}^m\}$. $f'(\bar{c})$ is a linear function (in fact, total derivative $T_{\bar{c}}$) which maps \bar{v} into the directional derivatives of f at \bar{c} in the direction \bar{v} . This notation generalises f' for univariate f as well. (put $n = m = 1$)

In this subject, we use the following notations,

$D_k f(\bar{c})$ partial derivative

$f'(\bar{c}, \bar{v})$ directional derivative

$f'(\bar{c})(\bar{v})$ total derivative

$\nabla f(\bar{c})$ gradient vector

Theorem 13.2.1. If f is differentiable at \bar{c} with total derivative $T_{\bar{c}}$, then for every $\bar{u} \in \mathbb{R}^n$, $T_{\bar{c}}(\bar{u}) = f'(\bar{c}, \bar{u})$. (ie, $f'(\bar{c})(\bar{v}) = f'(\bar{c}, \bar{v})$)

Proof. For $\bar{v} = \bar{0}$, we have $T_{\bar{c}}(\bar{0}) = 0 = f'(\bar{c}, \bar{0})$.

Suppose $\bar{v} \neq \bar{0}$, then put $\bar{v} = h\bar{u}$. Since f is differentiable at \bar{c} , f has total derivative at \bar{c} . That is, there exists a linear function $T_{\bar{c}}$ such that $f(\bar{c} + h\bar{u}) = f(\bar{c}) + T_{\bar{c}}(h\bar{u}) + \|h\bar{u}\|E_{\bar{c}}(h\bar{u})$ where $E_{\bar{c}}(h\bar{u}) \rightarrow \bar{0}$ as $h\bar{u} \rightarrow \bar{0}$.

$$\begin{aligned} \implies f(\bar{c} + h\bar{u}) &= f(\bar{c}) + hT_{\bar{c}}(\bar{u}) + |h|\|\bar{u}\|E_{\bar{c}}(h\bar{u}), \quad E_{\bar{c}}(h\bar{u}) \rightarrow \bar{0} \text{ as } h\bar{u} \rightarrow \bar{0} \\ \implies \frac{f(\bar{c} + h\bar{u}) - f(\bar{c})}{h} &= T_{\bar{c}}(\bar{u}) + \frac{|h|\|\bar{u}\|E_{\bar{c}}(h\bar{u})}{h}, \quad E_{\bar{c}}(h\bar{u}) \rightarrow \bar{0} \text{ as } h \rightarrow 0 \\ \implies \lim_{h \rightarrow 0} \frac{f(\bar{c} + h\bar{u}) - f(\bar{c})}{h} &= T_{\bar{c}}(\bar{u}) + \lim_{h \rightarrow 0} \frac{|h|\|\bar{u}\|E_{\bar{c}}(h\bar{u})}{h} \\ \implies f'(\bar{c}, \bar{u}) &= T_{\bar{c}}(\bar{u}) \end{aligned}$$

□

Note. $T_{\bar{c}}$ is linear, however $E_{\bar{c}}$ is not linear. Thus $E_{\bar{c}}(h\bar{u}) \neq hE_{\bar{c}}(\bar{u})$.

As $h \rightarrow 0$, $h\bar{u} \rightarrow \bar{0}$ and $E_{\bar{c}}(h\bar{u}) \rightarrow \bar{0}$. Since the order of the function $E_{\bar{c}}(h\bar{u})$ is much smaller than that of h , the limit on the right converges to 0.

Theorem 13.2.2. *If f is differentiable at \bar{c} , then f is continuous at \bar{c} .*

Proof. Let $\bar{v} \neq \bar{0}$, then

$$\begin{aligned} \bar{v} &= v_1\bar{u}_1 + v_2\bar{u}_2 + \cdots + v_n\bar{u}_n, \\ \bar{v} \rightarrow \bar{0} &\implies \forall j, v_j \rightarrow 0 \\ T \text{ is linear} &\implies T_{\bar{c}}(\bar{v}) = v_1T_{\bar{c}}(\bar{u}_1) + v_2T_{\bar{c}}(\bar{u}_2) + \cdots + v_nT_{\bar{c}}(\bar{u}_n) \\ \text{Thus, } T_{\bar{c}}(\bar{v}) &\rightarrow \bar{0} \text{ as } \bar{v} \rightarrow \bar{0} \end{aligned}$$

Since f differentiable at \bar{c} , there exists linear function $T_{\bar{c}}$ such that

$$\begin{aligned} f(\bar{c} + \bar{v}) &= f(\bar{c}) + T_{\bar{c}}(\bar{v}) + \|\bar{v}\|E_{\bar{c}}(\bar{v}) \\ \implies \lim_{\bar{v} \rightarrow \bar{0}} f(\bar{c} + \bar{v}) &= f(\bar{c}) + \lim_{\bar{v} \rightarrow \bar{0}} T_{\bar{c}}(\bar{v}) + \lim_{\bar{v} \rightarrow \bar{0}} \|\bar{v}\|E_{\bar{c}}(\bar{v}) \\ \implies \lim_{\bar{v} \rightarrow \bar{0}} f(\bar{c} + \bar{v}) &= f(\bar{c}) \end{aligned}$$

□

Theorem 13.2.3. *Let $S \subset \mathbb{R}^n$ and $f : S \rightarrow \mathbb{R}^m$ be differentiable at an interior point \bar{c} of S , where $S \subseteq \mathbb{R}^n$. If $\bar{v} = v_1\bar{u}_1 + v_2\bar{u}_2 + \cdots + v_n\bar{u}_n$, then*

$$f'(\bar{c})(\bar{v}) = \sum_{k=1}^n v_k D_k f(\bar{c})$$

In particular, if f is real-valued ($m = 1$) we have, $f'(\bar{c})(\bar{v}) = \nabla f(\bar{c}) \cdot \bar{v}$

Proof. Suppose $f : S \rightarrow \mathbb{R}^m$ is differentiable at \bar{c} , then there exists a linear function $f'(\bar{c}) : S \rightarrow \mathbb{R}^m$ such that $f(\bar{c} + \bar{v}) = f(\bar{c}) + f'(\bar{c})(\bar{v}) + \|\bar{v}\|E_{\bar{c}}(\bar{c})$ where

$E_{\bar{c}} \rightarrow \bar{0}$ as $\bar{v} \rightarrow \bar{0}$.

$$\begin{aligned} f'(\bar{c})(\bar{v}) &= f'(\bar{c}) \left(\sum_{k=1}^n v_k \bar{u}_k \right) \\ &= \sum_{k=1}^n v_k f'(\bar{c})(\bar{u}_k), \text{ since } f'(\bar{c}) \text{ is linear} \\ &= \sum_{k=1}^n v_k D_k f(\bar{c}), \text{ since } f'(\bar{c})(\bar{u}_k) = f'(\bar{c}, \bar{u}_k) = D_k f(\bar{c}) \end{aligned}$$

Let $m = 1$, then $f : S \rightarrow \mathbb{R}$

$$\begin{aligned} f'(\bar{c})(\bar{v}) &= \sum_{k=1}^n v_k D_k f(\bar{c}) = \nabla f(\bar{c}) \cdot \bar{v} \\ &\text{since } \nabla f(\bar{c}) = (D_1 f(\bar{c}), D_2 f(\bar{c}), \dots, D_n f(\bar{c})) \end{aligned}$$

□

Remark. Let $f : S \rightarrow \mathbb{R}$, then $f(\bar{c} + \bar{v}) = f(\bar{c}) + \nabla f(\bar{c}) \cdot \bar{v} + o(\|\bar{v}\|)$ as $\bar{v} \rightarrow \bar{0}$.

Remark (Complex-valued Functions).

13.2.3 Matrix of Linear Function

Let $T : \mathbb{R}^n \rightarrow \mathbb{R}^m$ be a linear function. Let $\{\bar{u}_1, \bar{u}_2, \dots, \bar{u}_n\}$ be standard basis for \mathbb{R}^n and $\{\bar{e}_1, \bar{e}_2, \dots, \bar{e}_m\}$ be standard basis for \mathbb{R}^m . Let $\bar{v} \in \mathbb{R}^n$, then $\bar{v} = \sum_{k=1}^n v_k \bar{u}_k$ and $T(\bar{v}) = \sum_{k=1}^n v_k T(\bar{u}_k)$ and

$$\begin{aligned} T(\bar{v}) &= \begin{bmatrix} v_1 & v_2 & \vdots & v_n \end{bmatrix} \begin{bmatrix} T(\bar{u}_1) \\ T(\bar{u}_2) \\ \dots \\ T(\bar{u}_n) \end{bmatrix} \\ &= \begin{bmatrix} v_1 & v_2 & \vdots & v_n \end{bmatrix} \begin{bmatrix} t_{11}\bar{e}_1 + t_{21}\bar{e}_2 + \dots + t_{m1}\bar{e}_m \\ t_{12}\bar{e}_1 + t_{22}\bar{e}_2 + \dots + t_{m2}\bar{e}_m \\ \dots \\ t_{1n}\bar{e}_1 + t_{2n}\bar{e}_2 + \dots + t_{mn}\bar{e}_m \end{bmatrix} \\ &= \begin{bmatrix} v_1 & v_2 & \dots & v_n \end{bmatrix} \begin{bmatrix} t_{11} & t_{21} & \dots & t_{m1} \\ t_{12} & t_{22} & \dots & t_{m2} \\ \vdots & \vdots & \ddots & \vdots \\ t_{1n} & t_{2n} & \dots & t_{mn} \end{bmatrix} \begin{bmatrix} \bar{e}_1 \\ \bar{e}_2 \\ \dots \\ \bar{e}_m \end{bmatrix} \end{aligned}$$

We may take the transpose,

$$\begin{aligned} T(\bar{v}) &= \begin{bmatrix} \bar{e}_1 & \bar{e}_2 & \dots & \bar{e}_m \end{bmatrix} \begin{bmatrix} t_{11} & t_{12} & \dots & t_{1n} \\ t_{21} & t_{22} & \dots & t_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ t_{m1} & t_{m2} & \dots & t_{mn} \end{bmatrix} \begin{bmatrix} v_1 \\ v_2 \\ \dots \\ v_n \end{bmatrix} \\ T(\bar{v}) &= T \left(\sum_{k=1}^n v_k \bar{u}_k \right) = \sum_{k=1}^n v_k T(\bar{u}_k) = \sum_{k=1}^n v_k \sum_{j=1}^m t_{kj} \bar{e}_j \end{aligned}$$

Thus matrix of T is given by, $m(T) = (t_{ik})$ where $T(\bar{u}_k) = \sum_{i=1}^n t_{ik} \bar{e}_i$.

Remark (Example). Let $T : \mathbb{R}^3 \rightarrow \mathbb{R}^2$ defined by $T(x, y, z) = (2x + y, y - z)$.

$$\begin{aligned}
 T(1, 2, 3) &= T((1, 0, 0) + 2(0, 1, 0) + 3(0, 0, 1)) \\
 &= T(\bar{u}_1 + 2\bar{u}_2 + 3\bar{u}_3) \\
 &= T(\bar{u}_1) + 2T(\bar{u}_2) + 3T(\bar{u}_3) \\
 &= [1 \quad 2 \quad 3] \begin{bmatrix} T(\bar{u}_1) \\ T(\bar{u}_2) \\ T(\bar{u}_3) \end{bmatrix} \\
 &= [1 \quad 2 \quad 3] \begin{bmatrix} (2, 0) \\ (1, 1) \\ (0, -1) \end{bmatrix} \\
 &= [1 \quad 2 \quad 3] \begin{bmatrix} 2(1, 0) \\ (1, 0) + (0, 1) \\ -1(0, 1) \end{bmatrix} \\
 &= [1 \quad 2 \quad 3] \begin{bmatrix} 2\bar{e}_1 \\ \bar{e}_1 + \bar{e}_2 \\ -\bar{e}_2 \end{bmatrix} \\
 &= [1 \quad 2 \quad 3] \begin{bmatrix} 2 & 0 \\ 1 & 1 \\ 0 & -1 \end{bmatrix} \begin{bmatrix} \bar{e}_1 \\ \bar{e}_2 \end{bmatrix} \\
 &= 4\bar{e}_1 - \bar{e}_2 = 4(1, 0) - 1(0, 1) = (4, -1)
 \end{aligned}$$

$$\text{In the above case, } m(T) = \begin{bmatrix} 2 & 0 \\ 1 & 1 \\ 0 & -1 \end{bmatrix}$$

Using the matrix of linear function $m(T)$, we can compute the image of any point in \mathbb{R}^3 by matrix multiplication.

Matrix of the composition of two linear functions

Let $T : \mathbb{R}^n \rightarrow \mathbb{R}^m$ and $S : \mathbb{R}^m \rightarrow \mathbb{R}^p$ be two linear functions with domain of S containing the range of T (so that $S \circ T$ is well defined). Then $S \circ T : \mathbb{R}^n \rightarrow \mathbb{R}^p$ is defined by

$$S \circ T(\bar{x}) = S(T(\bar{x})), \quad \forall \bar{x} \in \mathbb{R}^n$$

Since S, T are linear, $S \circ T$ is also linear.

$$\begin{aligned}
 S \circ T(a\bar{x} + b\bar{y}) &= S(T(a\bar{x} + b\bar{y})) = S(aT(\bar{x}) + bT(\bar{y})) = aS(T(\bar{x})) + bS(T(\bar{y})) \\
 &= aS \circ T(\bar{x}) + bS \circ T(\bar{y}), \quad \forall a, b \in \mathbb{R}, \quad \forall \bar{x}, \bar{y} \in \mathbb{R}^n
 \end{aligned}$$

Let $\{\bar{u}_1, \bar{u}_2, \dots, \bar{u}_n\}$ be the standards basis for \mathbb{R}^n , $\{\bar{e}_1, \bar{e}_2, \dots, \bar{e}_m\}$ be the standards basis for \mathbb{R}^m and $\{\bar{w}_1, \bar{w}_2, \dots, \bar{w}_p\}$ be the standards basis for \mathbb{R}^p . Let

$\bar{v} \in \mathbb{R}^n$, then $\bar{v} = \sum_{i=1}^n v_i \bar{u}_i$, and $S \circ T(\bar{v}) = \sum_{i=1}^n v_i S \circ T(\bar{u}_i)$

$$\begin{aligned}
S \circ T(\bar{v}) &= \begin{bmatrix} v_1 & v_2 & \vdots & v_n \end{bmatrix} \begin{bmatrix} S \circ T(\bar{u}_1) \\ S \circ T(\bar{u}_2) \\ \vdots \\ S \circ T(\bar{u}_n) \end{bmatrix} \\
&= \begin{bmatrix} v_1 & v_2 & \vdots & v_n \end{bmatrix} \begin{bmatrix} S(t_{11}\bar{e}_1 + \cdots + t_{m1}\bar{e}_m) \\ S(t_{12}\bar{e}_1 + \cdots + t_{m2}\bar{e}_m) \\ \vdots \\ S(t_{1n}\bar{e}_1 + \cdots + t_{mn}\bar{e}_m) \end{bmatrix} \\
&= \begin{bmatrix} v_1 & v_2 & \vdots & v_n \end{bmatrix} \begin{bmatrix} t_{11}S(\bar{e}_1) + \cdots + t_{m1}S(\bar{e}_m) \\ t_{12}S(\bar{e}_1) + \cdots + t_{m2}S(\bar{e}_m) \\ \vdots \\ t_{1n}S(\bar{e}_1) + \cdots + t_{mn}S(\bar{e}_m) \end{bmatrix} \\
&= \begin{bmatrix} v_1 & v_2 & \cdots & v_n \end{bmatrix} \begin{bmatrix} t_{11} & t_{21} & \cdots & t_{m1} \\ t_{12} & t_{22} & \cdots & t_{m2} \\ \vdots & \vdots & \ddots & \vdots \\ t_{1n} & t_{2n} & \cdots & t_{mn} \end{bmatrix} \begin{bmatrix} S(\bar{e}_1) \\ S(\bar{e}_2) \\ \vdots \\ S(\bar{e}_m) \end{bmatrix} \\
&= \begin{bmatrix} v_1 & v_2 & \cdots & v_n \end{bmatrix} \begin{bmatrix} t_{11} & t_{21} & \cdots & t_{m1} \\ t_{12} & t_{22} & \cdots & t_{m2} \\ \vdots & \vdots & \ddots & \vdots \\ t_{1n} & t_{2n} & \cdots & t_{mn} \end{bmatrix} \begin{bmatrix} s_{11}\bar{w}_1 + s_{12}\bar{w}_2 + \cdots + s_{1p}\bar{w}_p \\ s_{12}\bar{w}_1 + s_{22}\bar{w}_2 + \cdots + s_{p2}\bar{w}_p \\ \vdots \\ s_{1m}\bar{w}_1 + s_{2m}\bar{w}_2 + \cdots + s_{pm}\bar{w}_p \end{bmatrix} \\
&= \begin{bmatrix} v_1 & v_2 & \cdots & v_n \end{bmatrix} \begin{bmatrix} t_{11} & t_{21} & \cdots & t_{m1} \\ t_{12} & t_{22} & \cdots & t_{m2} \\ \vdots & \vdots & \ddots & \vdots \\ t_{1n} & t_{2n} & \cdots & t_{mn} \end{bmatrix} \begin{bmatrix} s_{11} & s_{21} & \cdots & s_{p1} \\ s_{12} & s_{22} & \cdots & s_{p2} \\ \vdots & \vdots & \ddots & \vdots \\ s_{1m} & s_{2m} & \cdots & s_{pm} \end{bmatrix} \begin{bmatrix} \bar{w}_1 \\ \bar{w}_2 \\ \vdots \\ \bar{w}_p \end{bmatrix}
\end{aligned}$$

We may take transpose,

$$S \circ T(\bar{v}) = \begin{bmatrix} \bar{w}_1 & \bar{w}_2 & \vdots & \bar{w}_p \end{bmatrix} \begin{bmatrix} s_{11} & s_{12} & \cdots & s_{1m} \\ s_{21} & s_{22} & \cdots & s_{2m} \\ \vdots & \vdots & \ddots & \vdots \\ s_{p1} & s_{p2} & \cdots & s_{pm} \end{bmatrix} \begin{bmatrix} t_{11} & t_{12} & \cdots & t_{1n} \\ t_{21} & t_{22} & \cdots & t_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ t_{m1} & t_{m2} & \cdots & t_{mn} \end{bmatrix} \begin{bmatrix} v_1 \\ v_2 \\ \vdots \\ v_n \end{bmatrix}$$

Remember : Given $T : \mathbb{R}^n \rightarrow \mathbb{R}^m$, then we may take $m(T)$ either as $m \times n$ matrix or $n \times m$ matrix. Since, we chose $m \times n$, $m(S \circ T) = m(S)m(T)$. Otherwise, $m(S \circ T) = m(T)m(S)$. This may change for different authors.

Suppose $m(S) = (s_{ij})$ and $m(T) = (t_{ij})$ respectively. Then

$$S(e_k) = \sum_{i=1}^p s_{ik} \bar{w}_i, \quad k = 1, 2, \dots, m \text{ and}$$

$$T(u_j) = \sum_{k=1}^m t_{kj} \bar{e}_k, \quad j = 1, 2, \dots, n$$

$$\begin{aligned}
(S \circ T)(\bar{u}_j) &= S(T(\bar{u}_j)) = S\left(\sum_{k=1}^m t_{kj} \bar{e}_k\right) = \sum_{k=1}^m t_{kj} S(\bar{e}_k) \\
&= \sum_{k=1}^m t_{kj} \left(\sum_{i=1}^p s_{ik} \bar{w}_i\right) = \sum_{i=1}^p \left(\sum_{k=1}^m s_{ik} t_{kj}\right) \bar{w}_i
\end{aligned}$$

Therefore, $m(S \circ T) = \sum_{k=1}^m s_{ik} t_{kj} = (s_{ik})(t_{kj}) = m(S)m(T)$.

13.2.4 The Jacobian Matrix

Let $\bar{u}_1, \bar{u}_2, \dots, \bar{u}_n$ be the unit co-ordinate vectors in \mathbb{R}^n and $\bar{e}_1, \bar{e}_2, \dots, \bar{e}_m$ be the unit co-ordinate vectors in \mathbb{R}^m . Let function $f: \mathbb{R}^n \rightarrow \mathbb{R}^m$ be differentiable at $\bar{c} \in \mathbb{R}^n$. Then there exists a linear function $T = f'(\bar{c}): \mathbb{R}^n \rightarrow \mathbb{R}^m$ such that $f(\bar{c} + \bar{v}) = f(\bar{c}) + f'(\bar{c})(\bar{v}) + \|\bar{v}\| + E_{\bar{c}}(\bar{v})$. We have, $T(\bar{u}_k) = f'(\bar{c})(\bar{u}_k) = f'(\bar{c}, \bar{u}_k) = D_k f(\bar{c}) = D_k \sum_{i=1}^m f_i(\bar{c}) \bar{e}_i$.

Clearly, the matrix of total derivative T , $m(T) = (t_{ik}) = (D_k f_i(\bar{c}))$. This matrix is called Jacobian matrix of f at \bar{c} and is denoted by $Df(\bar{c})$.

$$Df(\bar{c}) = \begin{bmatrix} D_1 f_1(\bar{c}) & D_2 f_1(\bar{c}) & \dots & D_n f_1(\bar{c}) \\ D_1 f_2(\bar{c}) & D_2 f_2(\bar{c}) & \dots & D_n f_2(\bar{c}) \\ \vdots & \vdots & \ddots & \vdots \\ D_1 f_m(\bar{c}) & D_2 f_m(\bar{c}) & \dots & D_n f_m(\bar{c}) \end{bmatrix}$$

Properties of Jacobian matrix

1. k th row of $Df(\bar{c})$ is gradient vector of f_k

$$\nabla f_k(\bar{c}) = (D_1 f_k(\bar{c}), D_2 f_k(\bar{c}), \dots, D_n f_k(\bar{c}))$$

2. When $m = 1$, $Df(\bar{c}) = \nabla f(\bar{c})$.

3. $f'(\bar{c})(\bar{v}) = \sum_{k=1}^m (\nabla f_k(\bar{c}) \cdot \bar{v}) \bar{e}_k$

4. $\|f'(\bar{c})(\bar{v})\| \leq M\|\bar{v}\|$ where $M = \sum_{k=1}^m \|\nabla f_k(\bar{c})\|$, by property (3)

5. $f'(\bar{c})(\bar{v}) \rightarrow \bar{0}$ as $\bar{v} \rightarrow \bar{0}$, by property (4)

Chain Rule

Chain Rule for real function : $\frac{dF \circ G}{dx}(x) = \frac{d}{dy} F(y) \frac{d}{dx} G(x) = F'(y) G'(x)$

For example : $\frac{d}{dx}(ax+3)^3 = \frac{d}{dy} y^3 \frac{d}{dx}(ax+3) = 3ay^2 = 3a(ax+3)^2$

Theorem 13.2.4. Let g be differentiable at \bar{a} , with total derivative $g'(\bar{a})$ and $\bar{b} = g(\bar{a})$. Let f be differentiable at \bar{b} , with total derivative $f'(\bar{b})$. Then $h = f \circ g$ is differentiable at \bar{a} with total derivative $h'(\bar{a}) = f'(\bar{b}) \circ g'(\bar{a})$. Try to read $h'(\bar{a}) = H$, $f'(\bar{b}) = F$, $g'(\bar{a}) = G$, then $H = F \circ G \implies H(x) = F(G(x))$ In other words, $h'(\bar{a})(\bar{v}) = f'(\bar{b}) \circ g'(\bar{a})(\bar{v}) = f'(\bar{b})(g'(\bar{a})(\bar{v}))$.

Proof. Given $\epsilon > 0$, let $y \in \mathbb{R}^p$ such that $\|y\| < \epsilon$. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$ and $g : \mathbb{R}^p \rightarrow \mathbb{R}^n$, then $h = f \circ g : \mathbb{R}^p \rightarrow \mathbb{R}^m$.

We have, $h(\bar{a} + \bar{y}) - h(\bar{a}) = f(g(\bar{a} + \bar{y})) - f(g(\bar{a})) = f(\bar{b} + \bar{v}) - f(\bar{b})$ where $\bar{b} = g(\bar{a})$, and $\bar{v} = g(\bar{a} + \bar{y}) - g(\bar{a})$.

Since g is differentiable at \bar{a} , g satisfies first-order Taylor's formula.

$$\begin{aligned} g(\bar{a} + \bar{y}) &= g(\bar{a}) + g'(\bar{a})(\bar{y}) + \|\bar{y}\| E_{\bar{a}}(\bar{y}) \text{ where } E_{\bar{a}} \rightarrow \bar{0} \text{ as } \bar{y} \rightarrow \bar{0} \\ \implies \bar{v} &= g(\bar{a} + \bar{y}) - g(\bar{a}) = g'(\bar{a})(\bar{y}) + \|\bar{y}\| E_{\bar{a}}(\bar{y}) \end{aligned}$$

Clearly, as $\bar{y} \rightarrow \bar{0} \implies \bar{v} \rightarrow g'(\bar{a})(\bar{0}) = \bar{0}$. Again, we have f is differentiable at \bar{b} , thus f satisfies first-order Taylor's formula.

$$f(\bar{b} + \bar{v}) = f(\bar{b}) + f'(\bar{b})(\bar{v}) + \|\bar{v}\| E_{\bar{b}}(\bar{v}) \text{ where } E_{\bar{b}} \rightarrow \bar{0} \text{ as } \bar{v} \rightarrow \bar{0}$$

$$\begin{aligned} \implies f(\bar{b} + \bar{v}) - f(\bar{b}) &= f'(\bar{b})(\bar{v}) + \|\bar{v}\| E_{\bar{b}}(\bar{v}) \\ &= f'(\bar{b})(g'(\bar{a})(\bar{y}) + \|\bar{y}\| E_{\bar{a}}(\bar{y})) + \|\bar{v}\| E_{\bar{b}}(\bar{v}) \\ &= f'(\bar{b})(g'(\bar{a})(\bar{y})) + \|\bar{y}\| E(\bar{y}) \\ &\text{where } E(\bar{y}) = f'(\bar{b})(E_{\bar{a}}(\bar{y})) + \frac{\|\bar{v}\|}{\|\bar{y}\|} E_{\bar{b}}(\bar{v}), \bar{y} \neq \bar{0} \end{aligned}$$

$$\implies h(\bar{a} + \bar{y}) - h(\bar{a}) = f(\bar{b} + \bar{v}) - f(\bar{b}) = f'(\bar{b})(g'(\bar{a})(\bar{y})) + \|\bar{y}\| E(\bar{y})$$

Since $f'(\bar{b})$ and $g'(\bar{a})$ are linear, their composition is also linear. Therefore, h is differentiable at \bar{a} with a linear, total derivative $h'(\bar{a}) = f'(\bar{b}) \circ g'(\bar{a})$ as it satisfies first-order Taylor's formula if $E_{\bar{y}} \rightarrow \bar{0}$ as $\bar{y} \rightarrow \bar{0}$.

We have, $\|\bar{v}\| \leq \|g'(\bar{a})(\bar{y})\| + \|\bar{y}\| \|E_{\bar{a}}(\bar{y})\| \leq M\|y\| + \|E_{\bar{a}}(\bar{y})\| \|\bar{y}\|$.

$$\implies \frac{\|\bar{v}\|}{\|\bar{y}\|} \leq M + \|E_{\bar{a}}(\bar{y})\|$$

Thus, $\bar{v} \rightarrow \bar{0}$ as $\bar{y} \rightarrow \bar{0}$. Then $f'(\bar{b})(\bar{v}) \rightarrow f'(\bar{b})(\bar{0}) = \bar{0}$. And $E_{\bar{a}}(\bar{y}) \rightarrow \bar{0}$. Therefore, $E(\bar{y}) \rightarrow \bar{0} + M\bar{0} = \bar{0}$ as $\bar{y} \rightarrow \bar{0}$. \square

Matrix form of the chain rule

Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$, $g : \mathbb{R}^p \rightarrow \mathbb{R}^n$. And $h = f \circ g : \mathbb{R}^p \rightarrow \mathbb{R}^m$. Suppose g is differentiable at $\bar{a} \in \mathbb{R}^p$ and f is differentiable at $g(\bar{a}) = \bar{b} \in \mathbb{R}^n$. Then h is differentiable at \bar{a} and the Jacobian matrix of h is given by the chain rule,

$$Dh(\bar{a}) = Df(\bar{b})Dg(\bar{a}) \text{ where } h = f \circ g, \bar{b} = g(\bar{a})$$

In other words,

$$D_j h_i(\bar{a}) = \sum_{k=1}^n D_k f_i(\bar{b}) D_j g_k(\bar{a}), \quad i = 1, 2, \dots, m, \quad j = 1, 2, \dots, p$$

$$\text{For } m = 1, D_j h(\bar{a}) = \sum_{k=1}^n D_k f(\bar{b}) D_j g_k(\bar{a})$$

$$\text{For } m = 1 \text{ and } p = 1, h'(\bar{a}) = \sum_{k=1}^n Df(\bar{b})g'_k(\bar{a}) = \nabla f(\bar{b}) \cdot Dg(\bar{a})$$

Theorem 13.2.5. Let f and D_2f be continuous functions on a rectangle $[a, b] \times [c, d]$. Let p and q be differentiable on $[c, d]$, where $p(y) \in [a, b]$ and $q(y) \in [c, d]$ for each $y \in [c, d]$. Define F by the equation,

$$F(y) = \int_{p(y)}^{q(y)} f(x, y) dx, \quad y \in [c, d]$$

Then $F'(y)$ exists for each $y \in (c, d)$ and is given by,

$$F'(y) = \int_{p(y)}^{q(y)} D_2f(x, y) dx + f(q(y), y)q'(y) - f(p(y), y)p'(y)$$

The following two theorems are required for proving the theorem on differentiating an integral.

Theorem 13.2.6. Let α be of bounded variation on $[a, b]$ and assume that $f \in \mathcal{R}(\alpha)$ on $[a, b]$.

$$\text{Define } F(x) = \int_a^x f d\alpha, \quad x \in [a, b]$$

Then F is of bounded variation on $[a, b]$ and F is continuous at x if α is continuous at x . If α is increasing on $[a, b]$, then the derivative $F'(x)$ exists at each $x \in (a, b)$ where $\alpha'(x)$ exists and where f is continuous. And

$$F'(x) = f(x)\alpha'(x)$$

Theorem 13.2.7. Let $Q = \{(x, y) : a \leq x \leq b, c \leq y \leq d\}$. Assume that α is of bounded variation on $[a, b]$ and for each $y \in [c, d]$, assume that the integral

$$F(y) = \int_a^b f(x, y) d\alpha(x)$$

exists. If the partial derivative D_2f is continuous on Q , the derivative $F'(y)$ exists for each $y \in (c, d)$ and is given by

$$F'(y) = \int_a^b D_2f(x, y) d\alpha(x)$$

Proof. Let $G(x_1, x_2, x_3) = \int_{x_1}^{x_2} f(t, x_3) dt$. Then we may write $F(y)$ in terms of G . That is, $F(y) = G(p(y), q(y), y)$.

Step 1 : 1-D Chain Rule

By 1-dimensional chain rule, we have

$$\begin{aligned} F'(y) &= \frac{dF}{dy} = \frac{\partial G}{\partial p} \frac{dp}{dy} + \frac{\partial G}{\partial q} \frac{dq}{dy} + \frac{\partial G}{\partial y} \\ &= D_1G p'(y) + D_2G q'(y) + D_3G \end{aligned}$$

Step 2 : D_1G

Since the variable of differentiation is present in the limit of the integral, we use theorem 13.2.6 to compute the derivative of the integral. We may write,

$$G(p(y), q(y), y) = - \int_{q(y)}^{p(y)} f(t, y) dt \quad (13.6)$$

We are differentiating (partially) with respect to $p(y)$. Thus $q(y)$, y are constants for this differentiation.

$$\begin{aligned} G(x, a, y) &= -H(x) = - \int_a^x f(t, y) dt \\ \implies D_1 G &= -H'(x) = -f(x, y). \\ \text{Thus, } D_1 G &= -f(p(y), y) \end{aligned}$$

Step 3 : $D_2 G$

Again, variable of differentiation is present in the limit of the integral. Thus, we write,

$$G(p(y), q(y), y) = \int_{p(y)}^{q(y)} f(t, y) dt \quad (13.7)$$

Now we are differentiating (partially) with respect to the the second component of G which is $q(y)$. Clearly, $p(y)$ and y are treated as constants.

$$\begin{aligned} G(a, x, y) &= H(x) = \int_a^x f(t, y) dt \\ \implies D_2 G &= H'(x) = f(q(y), y) \end{aligned}$$

Step 4 : $D_3 G$

Now the variable of integration is not affecting the limits of the integral. Also it is given that $D_2 f$ is continuous on $[a, b] \times [c, d]$. We write

$$\begin{aligned} G(a, b, x) &= H(x) = \int_a^b f(t, x) dt \\ \implies D_3 G &= H'(x) = \int_a^b D_2 f(t, x) dt \\ \text{Thus, } D_3 G &= \int_{p(y)}^{q(y)} f(t, y) dt \end{aligned}$$

□

The mean-value theorem for differentiable functions

Theorem 13.2.8 (Mean-Value). *Let S be an open subset of \mathbb{R}^n . Assume $f : S \rightarrow \mathbb{R}^m$ is differentiable at each point of S . Let \bar{x}, \bar{y} be two points in S such that $L(\bar{x}, \bar{y}) = \{t\bar{x} + (1-t)\bar{y} : t \in [0, 1]\}$ is subset of S . Then for every $\bar{a} \in \mathbb{R}^m$, there exists a point $\bar{z} \in L(\bar{x}, \bar{y})$ such that*

$$\bar{a} \cdot (f(\bar{y}) - f(\bar{x})) = \bar{a} \cdot f'(\bar{z})(\bar{y} - \bar{x})$$

Proof. Let $\bar{u} = \bar{y} - \bar{x}$. We have S is open subset and $L(\bar{x}, \bar{y}) \subset S$, thus there exists $\delta > 0$ such that $\bar{x} + t\bar{u} \in S, \forall t \in (-\delta, 1 + \delta)$. In other words, the ‘Line

segment $L(\bar{x}, \bar{y})'$ is properly contained in S , in such a way that extending the Line from \bar{x} to \bar{y} a little bit extra one either sides is still contained in S .

Let $\bar{a} \in \mathbb{R}^m$ and $F : (-\delta, 1 + \delta) \rightarrow \mathbb{R}$ defined by $F(t) = \bar{a} \cdot f(\bar{x} + t\bar{u})$. Then F is differentiable at each $t \in (-\delta, 1 + \delta)$ and the derivative $F'(t) = \bar{a} \cdot f'(\bar{x} + t\bar{u}, \bar{u})$, the directional derivative of $f(\bar{x} + t\bar{u})$ with respect to \bar{u} .

$$f'(\bar{x} + t\bar{u}, \bar{u}) = f'(\bar{x} + t\bar{u})(\bar{u}) \implies F'(t) = \bar{a} \cdot f'(\bar{x} + t\bar{u})(\bar{u})$$

By 1-dimensional mean-value theorem, we have

$$\exists \theta \in (0, 1) \text{ such that } F(1) - F(0) = F'(\theta)$$

By definition of F , $F(1) = \bar{a} \cdot f(\bar{x} + \bar{u}) = \bar{a} \cdot f(\bar{y})$. And $F(0) = \bar{a} \cdot f(\bar{x})$. Therefore,

$$F'(\theta) = F(1) - F(0) = \bar{a} \cdot f(\bar{y}) - \bar{a} \cdot f(\bar{x}) = \bar{a} \cdot (f(\bar{y}) - f(\bar{x}))$$

We also have,

$$F'(\theta) = \bar{a} \cdot f'(\bar{x} + \theta\bar{u})(\bar{u}) = \bar{a} \cdot f'(\bar{z})(\bar{y} - \bar{x}), \text{ where } \bar{z} = \bar{x} + \theta\bar{u} \in L(\bar{x}, \bar{y})$$

□

Remark. Suppose S is convex in \mathbb{R}^m . Then for every pair of points $\bar{x}, \bar{y} \in S$, $L(\bar{x}, \bar{y}) \subset S$. Thus Mean-value theorem holds for all $\bar{x}, \bar{y} \in S$.

13.3 Multivariate Calculus

13.3.1 A sufficient condition for differentiability

Theorem 13.3.1. Suppose one of the partial derivatives D_1f, D_2f, \dots, D_nf exists at \bar{c} . And the remaining $n - 1$ partial derivatives exists in some n -ball $B(\bar{c})$ and are continuous at \bar{c} . Then f is differentiable at \bar{c} .

Proof. **Step 1 : Real-valued function**

We claim that the function $f : \mathbb{R}^n \rightarrow \mathbb{R}$ is differentiable at \bar{c} iff each component f_k is differentiable at \bar{c} .

Suppose f is differentiable at \bar{c} , then there exists a linear, total derivative function $f'(\bar{c})$ satisfying first-order Taylor's formula at \bar{c} .

ie, $f(\bar{c} + \bar{v}) = f(\bar{c}) + f'(\bar{c})(\bar{v}) + \|\bar{v}\|E_{\bar{c}}(\bar{v})$ where $E_{\bar{c}}(\bar{v}) \rightarrow 0$ as $\bar{v} \rightarrow \bar{0}$.

$$f(\bar{c} + \bar{v}) = (f_1(\bar{c} + \bar{v}), f_2(\bar{c} + \bar{v}), \dots, f_m(\bar{c} + \bar{v}))$$

$$f(\bar{c}) = (f_1(\bar{c}), f_2(\bar{c}), \dots, f_m(\bar{c}))$$

$$f'(\bar{c})(\bar{v}) = (f'_1(\bar{v}), f'_2(\bar{v}), \dots, f'_m(\bar{v}))$$

$$E_{\bar{c}}(\bar{v}) = (E_1(\bar{v}), E_2(\bar{v}), \dots, E_m(\bar{v}))$$

where each component of the error function $E_k(\bar{v}) \rightarrow 0$ as $\bar{v} \rightarrow \bar{0}$. Also since $f'(\bar{c})$ is linear, each of its components $f'_k : \mathbb{R}^n \rightarrow \mathbb{R}$ are linear.

$$\begin{aligned} f(\bar{c} + \bar{v}) &= (f_1(\bar{c} + \bar{v}), f_2(\bar{c} + \bar{v}), \dots, f_m(\bar{c} + \bar{v})) \\ &= (f_1(\bar{c}), f_2(\bar{c}), \dots, f_m(\bar{c})) + (f'_1(\bar{v}), f'_2(\bar{v}), \dots, f'_m(\bar{v})) \\ &\quad + \|\bar{v}\| (E_1(\bar{v}), E_2(\bar{v}), \dots, E_m(\bar{v})) \text{ where } E_k(\bar{v}) \rightarrow 0 \text{ as } \bar{v} \rightarrow \bar{0} \\ &= (f_1(\bar{c}), f_2(\bar{c}), \dots, f_m(\bar{c})) + (f'_1(\bar{v}), f'_2(\bar{v}), \dots, f'_m(\bar{v})) \\ &\quad + (\|\bar{v}\|E_1(\bar{v}), \|\bar{v}\|E_2(\bar{v}), \dots, \|\bar{v}\|E_m(\bar{v})) \text{ where } E_k(\bar{v}) \rightarrow 0 \text{ as } \bar{v} \rightarrow \bar{0} \\ &= (f_1(\bar{c}) + f'_1(\bar{v}) + \|\bar{v}\|E_1(\bar{v}), \dots, f_m(\bar{c}) + f'_m(\bar{v}) + \|\bar{v}\|E_m(\bar{v})) \end{aligned}$$

Thus first-order Taylor's formula for f at \bar{c} gives first-order Taylor's formula for each of its components f_k . ie, $f_k(\bar{c} + \bar{v}) = f_k(\bar{c}) + f'_k(\bar{v} + \|\bar{v}\| E_k(\bar{v}))$ where $E_k(\bar{v}) \rightarrow 0$ as $\bar{v} \rightarrow \bar{0}$. Therefore, f_k are differentiable at \bar{c} for $k = 1, 2, \dots, m$.

Suppose each component f_k of f are differentiable at \bar{c} . Then there exists linear, total derivative functions f'_k satisfying first-order Taylor's formula at \bar{c} . ie, $f_k(\bar{c} + \bar{v}) = f_k(\bar{c}) + f'_k(\bar{v}) + \|\bar{v}\| E_k(\bar{v})$ where $E_k(\bar{v}) \rightarrow 0$ as $\bar{v} \rightarrow \bar{0}$.

Define $E_{\bar{c}}(\bar{v}) = (E_1(\bar{v}), E_2(\bar{v}), \dots, E_m(\bar{v}))$. Then $E_{\bar{c}}(\bar{v}) \rightarrow \bar{0}$ as $\bar{v} \rightarrow \bar{0}$. Therefore, there exists a linear, total derivative function $f'(\bar{c}) = (f'_1, f'_2, \dots, f'_m)$ satisfying first-order Taylor's formula at \bar{c} .

Thus, if each (real-valued) component function f_k are differentiable, then f is also differentiable. Therefore, it is sufficient to prove the theorem for a real-valued function.

Step 2: Telescopic Sum

Assume (without loss of generality) that $D_1 f$ exists at \bar{c} and $D_2 f, D_3 f, \dots, D_n f$ exist and continuous in some n -ball $B(\bar{c})$. Suppose $D_r f$ exists at \bar{c} and all partial derivatives except $D_r f$ are continuous. Then $v_0 = \bar{0}$, $v_1 = y_r \bar{u}_r$, $v_2 = y_r \bar{u}_r + y_1 \bar{u}_1, \dots$. Then the following proof can be applied without any loss of generality.

Let $\bar{v} = \lambda \bar{y}$ where $\bar{y} = \frac{\bar{v}}{\|\bar{v}\|}$. Clearly, $\|\bar{y}\| = 1$ and $\lambda = \|\bar{v}\|$. Choose $\lambda > 0$ such that $\bar{c} + \bar{v} \in B(\bar{c})$ and all the partial derivatives $D_2 f, D_3 f, \dots, D_n f$ exist and are continuous in $B(\bar{c})$.

We have, $\bar{y} = (y_1, y_2, \dots, y_n) = y_1 \bar{u}_1 + y_2 \bar{u}_2 + \dots + y_n \bar{u}_n$.

Define $\bar{v}_0 = \bar{0}$, $\bar{v}_1 = y_1 \bar{u}_1, \dots$, $\bar{v}_n = y_1 \bar{u}_1 + y_2 \bar{u}_2 + \dots + y_n \bar{u}_n$.

$$\begin{aligned} f(\bar{c} + \bar{v}) - f(\bar{c}) &= (f(\bar{c} + \lambda \bar{v}_n) - f(\bar{c} + \lambda \bar{v}_{n-1})) \\ &\quad + (f(\bar{c} + \lambda \bar{v}_{n-1}) - f(\bar{c} + \lambda \bar{v}_{n-2})) \\ &\quad + \dots + (f(\bar{c} + \lambda \bar{v}_1) - f(\bar{c} + \lambda \bar{v}_0)) \\ &= \sum_{k=1}^n f(\bar{c} + \lambda \bar{v}_k) - f(\bar{c} + \lambda \bar{v}_{k-1}) \\ &= \sum_{k=1}^n f(\bar{c} + \lambda \bar{v}_{k-1} + \lambda y_k \bar{u}_k) - f(\bar{c} + \lambda \bar{v}_{k-1}) \end{aligned}$$

Step 3 : Mean-value theorem

Define $\bar{b}_k = \bar{c} + \lambda \bar{v}_{k-1}$. Then we have

$$f(\bar{c} + \bar{v}) - f(\bar{c}) = \sum_{k=1}^n f(\bar{b}_k + \lambda y_k \bar{u}_k) - f(\bar{b}_k) \quad (13.8)$$

We know that all partial derivatives exist in $B(\bar{c})$. Therefore by 1-dimensional mean-value theorem we have,

$$f(\bar{b}_k + \lambda y_k \bar{u}_k) - f(\bar{b}_k) = \lambda y_k D_k f(\bar{a}_k) \text{ where } \bar{a}_k \in L(\bar{b}_k, \bar{b}_k + \lambda y_k \bar{u}_k)$$

$$f(\bar{c} + \bar{v}) - f(\bar{c}) = \lambda \sum_{k=1}^n y_k D_k f(\bar{a}_k) \text{ where } \bar{a}_k \in L(\bar{b}_k, \bar{b}_k + \lambda y_k \bar{u}_k) \quad (13.9)$$

Step 4 : Continuity of partial derivatives in $B(\bar{c})$

As $\lambda \rightarrow 0$, $\bar{v} \rightarrow \bar{0}$. And both \bar{b}_k , $\bar{b}_k + \lambda y_k \bar{u}_k \rightarrow \bar{c}$. Clearly, \bar{a}_k in the line between \bar{b}_k and $\bar{b}_k + \lambda y_k \bar{u}_k$ also converges to \bar{c} .

For $k \geq 2$, $D_k f$ are continuous in the n -ball $B(\bar{c})$. Thus $D_k f(\bar{a}_k) \rightarrow D_k f(\bar{c})$. We may write, $D_k f(\bar{a}_k) = D_k f(\bar{c}) + E_k(\lambda)$ where $E_k(\lambda) \rightarrow \bar{0}$ as $\lambda \rightarrow 0$. Also, since $D_1 f$ exists, $D_1 f(\bar{c} + \lambda y_1 \bar{u}_1) \rightarrow D_1 f(\bar{c})$ as $\lambda \rightarrow 0$.

$$\text{Remember : } D_1 f(\bar{c}) = \lim_{h \rightarrow 0} \frac{f(\bar{c} + h \bar{u}_1) - f(\bar{c})}{h}$$

$$\begin{aligned} f(\bar{c} + \bar{v}) - f(\bar{c}) &= \lambda \sum_{k=1}^n y_k D_k f(\bar{c}) + \lambda \sum_{k=1}^n y_k E_k(\lambda) \\ &= \nabla f(\bar{c}) \cdot \bar{v} + \|\bar{v}\| E(\lambda) \\ \text{where } E(\lambda) &= \sum_{k=1}^n y_k E_k(\lambda) \rightarrow \bar{0} \text{ as } \bar{v} \rightarrow \bar{0} \end{aligned}$$

That is, we have a linear function which satisfies first-order Taylor's formula at \bar{c} . Therefore, f is differentiable. \square

13.3.2 Sufficient conditions for the equality of mixed partial derivatives

Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$. Then $D_r f$ and $D_k f$ are two partial derivatives of f . And $D_{r,k} f = D_r(D_k f)$ and $D_{k,r} f = D_k(D_r f)$.

$$D_{r,k} f = \frac{\partial^2 f}{\partial x_r \partial x_k} = \frac{\partial}{\partial x_r} \frac{\partial f}{\partial x_k} \text{ and } D_{k,r} f = \frac{\partial^2 f}{\partial x_k \partial x_r} = \frac{\partial}{\partial x_k} \frac{\partial f}{\partial x_r}$$

There are two sufficient conditions for the equality of these mixed partial derivatives in our scope. 1. differentiability of $D_k f$ or 2. continuity of $D_{r,k} f$ and $D_{k,r} f$ at \bar{c} where the mixed partial derivatives are to be equal.

Differentiability

Theorem 13.3.2. Suppose $D_r f$ and $D_k f$ exists in an n -ball about \bar{c} and are both differentiable at \bar{c} . Then $D_{r,k} f = D_{k,r} f$.

Proof. **Step 1 : Real-valued function**

It is sufficient to prove the theorem for real-valued functions. Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$, then $f(\bar{c}) = (f_1(\bar{c}), f_2(\bar{c}), \dots, f_m(\bar{c}))$. And

$$D_k f(\bar{c}) = (D_k f_1(\bar{c}), D_k f_2(\bar{c}), \dots, D_k f_m(\bar{c}))$$

Thus it is sufficient to prove that $D_{r,k} f_j(\bar{c}) = D_{k,r} f_j(\bar{c})$, $j = 1, 2, \dots, m$. That is, it is sufficient to prove equality of mixed partial derivatives of a real-valued function $f_j : \mathbb{R}^n \rightarrow \mathbb{R}$. Also, we will prove it for $n = 2$ and $\bar{c} = (0, 0)$. **Now, we will prove the theorem of a real-valued function $f : \mathbb{R}^n \rightarrow \mathbb{R}$.**

Step 2 : $\nabla(h)$

Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$. Suppose that the partial derivatives $D_k f$, $D_r f$ exist in the n -ball $B(n)$. And let $h > 0$ such that the rectangle with vertices $(0, 0)$, $(0, h)$, $(h, 0)$, (h, h) lies in $B(n)$.

Suppose $n = 3$, $c = (x, y, z)$, and we want to prove equality of $D_{2,3} f$ and $D_{3,2} f$. Then we will consider the rectangle with vertices (x, y, z) , $(x, y, z +$

$h), (x, y + h, z), (x, y + h, z + h)$. Again, we are taking $n = 2$ and $\bar{c} = (0, 0)$, only for the ease of notation as the same proof is applicable for any finite natural number, n and any vector $\bar{c} \in \mathbb{R}^n$.

$$\text{Define } \nabla(h) = f(h, h) - f(h, 0) - f(0, h) + f(0, 0) \quad (13.10)$$

$$\text{Step 3 : } D_{1,2}f = \frac{\nabla(h)}{h^2} = D_{2,1}f$$

$$\text{Define } G(x) = f(x, h) - f(x, 0) \quad (13.11)$$

Then we have, $\nabla(h) = G(h) - G(0)$ and $G'(x) = D_1f(x, h) - D_1f(x, 0)$. By 1-dimensional mean value theorem,

$$\begin{aligned} G(h) - G(0) &= hG'(x_1) \text{ where } x_1 \in (0, h) \\ &= h(D_1f(x_1, h) - D_1f(x_1, 0)) \end{aligned}$$

We have D_1f is differentiable at $(0, 0)$. There exists linear, total derivative function $(D_1f)'(0, 0)$ where $(D_1f)'(0, 0)(x, y) = \nabla D_1f(0, 0) \cdot (x, y)$ satisfying first-order Taylor's formula at $(0, 0)$.

$$\text{Remember : } f'(\bar{c})(\bar{v}) = \sum_{k=1}^n v_k D_k f(\bar{c}) = \nabla f(\bar{c}) \cdot \bar{v}$$

$$\begin{aligned} D_1f((0, 0) + (x_1, h)) &= D_1f(0, 0) + \nabla D_1f(0, 0) \cdot (x_1, h) + \|(x_1, h)\|E_1(h) \\ &\text{where } E_1(h) \rightarrow 0 \text{ as } h \rightarrow 0 \end{aligned}$$

$$D_1f(x_1, h) = D_1f(0, 0) + x_1 D_{1,1}f(0, 0) + h D_{2,1}f(0, 0) + \left| \sqrt{x_1^2 + h^2} \right| E_1(h)$$

Similarly,

$$\begin{aligned} D_1f((0, 0) + (x_1, 0)) &= D_1f(0, 0) + \nabla D_1f(0, 0) \cdot (x_1, 0) + \|(x_1, 0)\|E_2(h) \\ &\text{where } E_2(h) \rightarrow 0 \text{ as } h \rightarrow 0 \\ D_1f(x_1, 0) &= D_1f(0, 0) + x_1 D_{1,1}f(0, 0) + |x_1|E_2(h) \end{aligned}$$

Therefore $\nabla(h) = h(D_1f(x_1, h) - D_1f(x_1, 0)) = h^2 D_{2,1}f(0, 0) + E(h)$ where $E(h) = h|\sqrt{x_1^2 + h^2}|E_1(h) - h|x_1|E_2(h)$ and $E(h) \rightarrow 0$ as $h \rightarrow 0$.

Since $0 < x_1 < h$, we have

$$0 \leq E(h) \leq h^2 \left(\sqrt{2}E_1(h) - E_2(h) \right)$$

Therefore,

$$\begin{aligned} \lim_{h \rightarrow 0} \frac{\nabla(h)}{h^2} &\leq \lim_{h \rightarrow 0} \frac{h^2 D_{2,1}f(0, 0)}{h^2} + \lim_{h \rightarrow 0} \frac{h^2(\sqrt{2}E_1(h) - E_2(h))}{h^2} = D_{2,1}f(0, 0) \\ \lim_{h \rightarrow 0} \frac{\nabla(h)}{h^2} &= D_{2,1}f(0, 0) \end{aligned} \quad (13.12)$$

You may skip the following part. And conclude with the last two lines.

Similarly, define $H(y) = f(h, y) - f(0, y)$. Then we have, $\nabla(h) = H(h) - H(0)$ and $H'(y) = D_2f(h, y) - D_2f(0, y)$. By 1-dimensional mean value theorem,

$$\begin{aligned} H(h) - H(0) &= hH'(y_1) \text{ where } y_1 \in (0, h) \\ &= h(D_2f(h, y_1) - D_2f(0, y_1)) \end{aligned}$$

We have D_2f is differentiable at $(0, 0)$. Thus there exists a linear, total derivative function $(D_2f)'(0, 0)$ where $(D_2f)'(0, 0)(x, y) = \nabla D_2f(0, 0) \cdot (x, y)$ satisfying first-order Taylor's formula at $(0, 0)$. That is,

$$\begin{aligned} D_2f((0, 0) + (h, y_1)) &= D_2f(0, 0) + \nabla D_2f(0, 0) \cdot (h, y_1) + \|(h, y_1)\|E_3(h) \\ \text{where } E_3(h) &\rightarrow 0 \text{ as } h \rightarrow 0 \end{aligned}$$

$$D_2f(h, y_1) = D_2f(0, 0) + hD_{1,2}f(0, 0) + y_1D_{2,2}f(0, 0) + \left| \sqrt{h^2 + y_1^2} \right| E_3(h)$$

Again,

$$\begin{aligned} D_2f((0, 0) + (0, y_1)) &= D_2f(0, 0) + \nabla D_2f(0, 0) \cdot (0, y_1) + |y_1|E_4(h) \\ \text{where } E_4(h) &\rightarrow 0 \text{ as } h \rightarrow 0 \end{aligned}$$

$$D_2f(0, y_1) = D_2f(0, 0) + y_1D_{2,2}f(0, 0) + |y_1|E_4(h)$$

Therefore, $\nabla(h) = h(D_2f(h, y_1) - D_2f(0, y_1)) = h^2D_{1,2}f(0, 0) + E'(h)$ where $E'(h) = \left| \sqrt{h^2 + y_1^2} \right| E_3(h) - |y_1|E_4(h)$ and $E'(h) \rightarrow 0$ as $h \rightarrow 0$. And

$$\lim_{h \rightarrow 0} \frac{\nabla(h)}{h^2} = D_{1,2}f(0, 0) \quad (13.13)$$

Therefore, $D_{1,2}f(0, 0) = D_{2,1}f(0, 0)$. \square

Continuity

Theorem 13.3.3. Suppose D_rf and D_kf exists in an n -ball about \bar{c} . And $D_{r,k}f$ and $D_{k,r}f$ are continuous at \bar{c} . Then $D_{r,k}f = D_{k,r}f$.

Proof. We have $D_rf = (D_rf_1, D_rf_2, \dots, D_rf_m)$. Therefore, it is sufficient to prove the theorem for real-valued functions. Suppose $n = 2$, $\bar{c} = (0, 0)$ and the partial derivatives D_1f and D_2f exist and are continuous in some 2-ball about $(0, 0)$. Suppose (h, h) lies in that 2-ball, then $D_1f(h, h) \rightarrow D_1f(0, 0)$ as $h \rightarrow 0$. \square

Remark. A function f such that $D_{1,2}f \neq D_{2,1}f$.

$$\text{Let, } f(x, y) = \begin{cases} \frac{xy(x^2 - y^2)}{x^2 + y^2} & (x, y) \neq (0, 0) \\ 0 & (x, y) = (0, 0) \end{cases}$$

$$\begin{aligned} D_1f(x, y) &= \frac{\partial}{\partial x} \frac{x^3y - xy^3}{x^2 + y^2} \\ &= \frac{(3x^2y - y^3)(x^2 + y^2) - 2x(x^3y - xy^3)}{(x^2 + y^2)^2} \\ &= \frac{x^4y + 4x^2y^3 - y^5}{(x^2 + y^2)^2} \end{aligned}$$

$$D_1f_{(x=0)} = -y \implies D_{2,1}f_{(x=0)} = \frac{\partial}{\partial y} -y = -1$$

$$\begin{aligned}
D_2 f(x, y) &= \frac{\partial}{\partial y} \frac{x^3 y - xy^3}{x^2 + y^2} \\
&= \frac{(x^3 - 3xy^2)(x^2 + y^2) - 2y(x^3 y - xy^3)}{(x^2 + y^2)^2} \\
&= \frac{x^5 - 4x^3 y^2 - xy^4}{(x^2 + y^2)^2} \\
D_2 f_{(y=0)} = x &\implies D_{1,2} f_{(y=0)} = \frac{\partial}{\partial x} x = 1
\end{aligned}$$

Therefore, $D_{1,2}f \neq D_{2,1}f$ in the neighbourhood of $(0, 0)$. This treatment save a lot of time. After $D_1 f$, we are planning to perform $D_{2,1}f = D_2(D_1 f)$ in which the value of x is going to be treated as a constant. Therefore, we can simplify the expression by substituting $x = 0$ at this stage. If you are not confident enough to substitute that “early”. You may take partial derivative with respect to y and then substitute $x = 0$ and $y = 0$. Why don’t we substitute $y = 0$ before $D_2 f$ is something you should know already !

13.3.3 Implicit Functions and Extremum Problems

Definitions 13.3.1 (Implicit function). Let f be a function. Consider the equation, $f(\bar{x}, \bar{y}) = \bar{0}$. If there exists a function g such that $\bar{x} = g(\bar{y})$, then g is an **implicit form** of f or g is defined implicitly by f . For example, a linear system of equations $Ax - b = 0$ implicitly defines $x = A^{-1}b$ provided A has non-zero determinant.

Definitions 13.3.2 (Jacobian Determinant). Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^n$, then determinant of the Jacobian matrix $Df(\bar{x})$ is the **Jacobian determinant** of f , $J_f(\bar{x})$.

What is an implicit function ?

Consider the equation $\sin(x + y) = \cos(u + v)$. We can rewrite this equation as $\sin(x + y) - \cos(u + v) = 0$. Now, the equation is of the form $f(x, y, u, v) = 0$. Therefore, we have the function $f : \mathbb{R}^4 \rightarrow \mathbb{R}$ defined by $f(x, y, u, v) = \sin(x + y) - \cos(u + v)$.

Let us play around,

$$\begin{aligned}
\sin(x + y) = \cos(u + v) &\implies x + y = \sin^{-1} \cos(u + v) \\
&\implies (x, y) = (\sin^{-1} \cos(u + v) - k, k) \\
&\implies (x, y) = g(u, v)
\end{aligned}$$

where $g_1(u, v) = \sin^{-1} \cos(u + v) - k$, and $g_2(u, v) = k$

Now we have defined a new function $g : \mathbb{R}^2 \rightarrow \mathbb{R}^2$ from the function f . Therefore, we could say that f defines g implicitly. Now we have a few questions to ask about the nature of such implicit functions (or implicitly defined functions),

1. Does there always exists a function for any such combination ?
That is, does there exists a function h such that $(x, u) = h(y, v)$?

2. Suppose $f(x, y, u, v) = 0$. And function $h : \mathbb{R}^2 \rightarrow \mathbb{R}^2$ is implicitly defined by f . Does there always exists a neighbourhood of (u, v) in which h has continuous partial derivatives ?
3. What are the properties of f for these implicit functions to have nice properties ?

It turns out that non-zero Jacobian determinant is nice property f can have.

Theorem 13.3.4. *Let $f : \mathbb{C} \rightarrow \mathbb{C}$. Then $J_f(z) = |f'(z)|^2$.*

Proof. Suppose $f : \mathbb{C} \rightarrow \mathbb{C}$ where $f(z) = u(z) + iv(z)$ where $u : \mathbb{C} \rightarrow \mathbb{R}$ and $v : \mathbb{C} \rightarrow \mathbb{R}$. These real-valued functions u, v have respective u^*, v^* multivariate real functions such that $u^* : \mathbb{R}^2 \rightarrow \mathbb{R}$, where $u(z) = u^*(x, y)$ and $z = x + iy$. Let $f(z) = z^2 + 1$. Then $u^*(x, y) = x^2 - y^2 + 1$ and $v^*(x, y) = -2xy$. And theoretically we use derivatives of u^* when we mention derivatives of u .

Then f has a derivative at z only if the partial derivatives D_1u, D_2u, D_1v, D_2v exists at z and satisfies Cauchy-Riemann equations. ie $D_1u(z) = D_2v(z)$ and $D_1v(z) = -D_2u(z)$. [Apostol, 1973, Theorem 5.22].

Thus we have $f'(z) = D_1u + iD_1v$ [Apostol, 1973, Theorem 12.6]¹.

$$\begin{aligned} f'(z) &= D_1u(z) + iD_1v(z) \\ |f'(z)|^2 &= (D_1u(z))^2 + (D_1v(z))^2 \end{aligned}$$

For ease of representation, we write D_1u instead of $D_1u(z)$

$$|f'(z)|^2 = (D_1u)^2 + (D_1v)^2$$

We also have

$$J_f(z) = |Df(z)| = \begin{vmatrix} D_1u & D_2u \\ D_1v & D_2v \end{vmatrix} = D_1uD_2v - D_1vD_2u = (D_1u)^2 + (D_1v)^2$$

Therefore, $J_f(z) = |f'(z)|^2$. □

Functions with non-zero Jacobian determinant

That is, $f : \mathbb{R}^n \rightarrow \mathbb{R}^n$ such that $J_f \neq 0$ in an n -ball. In other words, we have an n -ball $B(\bar{x})$ such that $J_f(\bar{y}) \neq 0, \forall \bar{y} \in B(\bar{x})$.

Theorem 13.3.5. *Let B be an n -ball about \bar{a} in \mathbb{R}^n , ∂B be its boundary and $\bar{B} = B \cup \partial B$ be its closure.² Let $f : \mathbb{R}^n \rightarrow \mathbb{R}^n$ be continuous in \bar{B} and all partial derivatives, $D_j f_i(\bar{x})$ exists for every $\bar{x} \in B$. Let $f(\bar{x}) \neq f(\bar{a})$ for every $\bar{x} \in \partial B$ and $J_f(\bar{x}) \neq 0$ for every $\bar{x} \in B$. Then $f(B)$ contains an n -ball about $f(\bar{a})$.*

$$\begin{aligned} B &= \{\bar{x} : \|\bar{x} - \bar{a}\| < r\} \\ \partial B &= \{\bar{x} : \|\bar{x} - \bar{a}\| = r\} \\ \bar{B} &= \{\bar{x} : \|\bar{x} - \bar{a}\| \leq r\} \end{aligned}$$

¹Prove using first-order Taylor's formula

² \bar{B} : The line above B has a different meaning compare to \bar{a} (situations like this are an abuse of language).

Proof. Define $g : \partial B \rightarrow \mathbb{R}$ where $g(\bar{x}) = \|f(\bar{x}) - f(\bar{a})\|$. We have, $f(\bar{x}) \neq f(\bar{a})$ for every $\bar{x} \in \partial B$, thus $g(\bar{x}) > 0$ for every $\bar{x} \in \partial B$. Function f is continuous on \bar{B} , thus g is continuous on \bar{B} and thus g is continuous on its subset ∂B . Since ∂B is compact, every continuous function on ∂B attains its extrema³ and thus g attains its minimum value $m > 0$ somewhere on ∂B .

Consider n -ball T about $f(\bar{a})$ with radius $\frac{m}{2}$,

$$T = B\left(f(\bar{a}), \frac{m}{2}\right) = \left\{\bar{y} \in \mathbb{R}^n : \|f(\bar{a}) - \bar{y}\| < \frac{m}{2}\right\}$$

Therefore, it is sufficient to prove that $T \subset f(B)$.

Let $\bar{y} \in T$. Define $h : \bar{B} \rightarrow \mathbb{R}$ where $h(\bar{x}) = \|f(\bar{x}) - \bar{y}\|$. Again this continuous function h on compact set \bar{B} attains its extrema somewhere on \bar{B} . Since $\bar{y} \in T$, $h(\bar{a}) = \|f(\bar{a}) - \bar{y}\| < \frac{m}{2}$. Thus, the minimum of h on \bar{B} is less than $\frac{m}{2}$, since $\bar{a} \in \bar{B}$.

Let $\bar{x} \in \partial B$, then

$$\begin{aligned} h(\bar{x}) &= \|f(\bar{x}) - \bar{y}\| \\ &= \|f(\bar{x}) - f(\bar{a}) + f(\bar{a}) - \bar{y}\| \\ &\geq \|f(\bar{x}) - f(\bar{a})\| + \|f(\bar{a}) - \bar{y}\| \\ &= g(\bar{x}) - h(\bar{a}) \\ &> \frac{m}{2} \text{ since } g(\bar{x}) \geq m \text{ and } h(\bar{a}) < \frac{m}{2} \end{aligned}$$

Thus h doesn't attain its minimum on ∂B , but at an interior point $\bar{c} \in B$. Consider

$$h^2(\bar{x}) = \|f(\bar{x}) - \bar{y}\|^2 = \sum_{r=1}^n (f_r(\bar{x}) - y_r)^2$$

The function h^2 also has minimum at the same point \bar{c} . Thus all partial derivatives of h^2 at \bar{c} are zero. ie,

$$D_k h^2(\bar{c}) = \sum_{r=1}^n (f_r(\bar{c}) - y_r) D_k f_r(\bar{c}) = 0$$

This is a system of linear equations with non-zero determinant since $\bar{c} \in B$ and we have $J_f(\bar{c}) \neq 0$. Therefore, $f_r(\bar{c}) = y_r$. That is, $f(\bar{c}) = \bar{y} \in f(B)$. Since $\bar{y} \in T$ is arbitrary, $T \subset f(B)$. \square

Theorem 13.3.6. Let A be an open subset of \mathbb{R}^n and $f : A \rightarrow \mathbb{R}^n$ is continuous and has continuous partial derivatives $D_j f_i$ on A . If f is one-to-one on A and $J_f(\bar{x}) \neq 0$, $\forall \bar{x} \in A$, then $f(A)$ is open.

Proof. Let $\bar{b} \in f(A)$. Then $\bar{b} = f(\bar{a})$ for some $\bar{a} \in A$. We have, f is continuous, f has continuous partial derivatives on A and $J_f(\bar{x}) \neq 0$ for every $\bar{x} \in A$. Therefore, there exists an open ball $B \subset A$ containing \bar{a} such that $f(B) \subset f(A)$ contains an n -ball about $f(\bar{a})$. Since $\bar{b} \in f(A)$ is arbitrary, every point in $f(A)$ has an n -ball containing it in $f(A)$. Therefore, $f(A)$ is open.

Two assumption in above theorem are trivial. 1. f is continuous in the closed ball, \bar{B} . Set B so chosen that $\bar{B} \subset A$ and f is continuous in A .

³“Every continuous function on a compact set attains its extrema”

Thus, f is continuous in \bar{B} . 2. f has different value at boundary compared to center. ie, $f(\bar{a}) \neq f(\bar{x}), \forall x \in \partial B$. We have, f is injective on A , and $\bar{B} \subset A$. Thus f has different values for any two distinct points in it. Thus, $\forall \bar{x}, \bar{y} \in \bar{B}, \bar{x} \neq \bar{y} \implies \bar{x}, \bar{y} \in A$, and $\bar{x} \neq \bar{y} \implies f(\bar{x}) \neq f(\bar{y})$ \square

Theorem 13.3.7. Let S be an open subset of \mathbb{R}^n and $f : S \rightarrow \mathbb{R}^n$. Let components of f has continuous partial derivatives on S , $D_j f_i$ and $J_f(\bar{a}) \neq 0$ for some point $\bar{a} \in S$. Then there is an n -ball B about \bar{a} on which f is injective.

Proof. Let $\bar{z} = (\bar{z}_1, \bar{z}_2, \dots, \bar{z}_n)$ where $\bar{z}_i \in \mathbb{R}^n$. ie, $\bar{z} \in \mathbb{R}^{n^2}$. Define function $h : \mathbb{R}^{n^2} \rightarrow \mathbb{R}$ by $h(\bar{z}) = \det [D_j f_i(\bar{z}_i)]$. Since f has continuous partial derivatives on S , each component of f has continuous partial derivatives in S and thus h is continuous on S^n which is a subset of \mathbb{R}^{n^2} since S is an open subset of \mathbb{R}^n .

Let $\bar{a} \in S$ such that $J_f(\bar{a}) \neq 0$. Existence of such a point in S is assumed.

Consider, $\bar{z}_i = \bar{a}, \forall i$. Then $\bar{z} = (\bar{a}, \bar{a}, \dots, \bar{a})$. And $h(\bar{z}) = \det [D_j f_i(\bar{a})] = J_f(\bar{a}) \neq 0$.

Since h is continuous and $h(\bar{z}) \neq 0$. There exists an n -ball B about \bar{a} in S such that $h(\bar{z}) \neq 0$ for $\bar{z}_i \in B$. We claim that f is injective on B .

Suppose f is not injective. ie, There exists $\bar{x}, \bar{y} \in B(\bar{a})$ such that $\bar{x} \neq \bar{y}$ and $f(\bar{x}) = f(\bar{y})$. Open ball $B(\bar{a})$ is a convex set. And the line segment $L(\bar{x}, \bar{y}) \subset B(\bar{a})$. The function f is differentiable on S . On applying mean-value theorem to each component of f , we get

$$0 = f_i(\bar{y}) - f_i(\bar{x}) = \nabla f_i(\bar{Z}_i) \cdot (\bar{y} - \bar{x}), \quad i = 1, 2, \dots$$

where $\bar{Z}_i \in L(\bar{x}, \bar{y}) \subset B(\bar{a})$. Therefore, We have

$$\sum_{k=1}^n D_k f_i(\bar{Z}_i)(y_k - x_k) = 0$$

The determinant of this system of linear equations is nonzero, as the function f has nonzero jacobian determinant at $\bar{Z}_i \in B(\bar{a})$ for $i = 1, 2, \dots$. Thus, $y_i = x_i$ for $i = 1, 2, \dots$. This contradicts $\bar{x} \neq \bar{y}$. Hence, the function f is injective. \square

Theorem 13.3.8. Let A be an open subset of \mathbb{R}^n and assume that $f : A \rightarrow \mathbb{R}^n$ has continuous partial derivatives $D_j f_i$ on A . If $J_f(\bar{x}) \neq 0$ for all $\bar{x} \in A$, then f is an open mapping.

Proof. Let S be an open subset of A . Let $\bar{x} \in S$. Clearly, f has continuous partial derivatives on S and $J_f(\bar{x}) \neq 0$ for all $\bar{x} \in S$. Thus, there is an n -ball $B(\bar{x})$ in which f is injective. Therefore, $f(B(\bar{x}))$ is open in \mathbb{R}^n . Since $\bar{x} \in S$ is arbitrary, $S = \cup_{\bar{x} \in S} B(\bar{x})$. And $f(S) = \cup_{\bar{x} \in S} f(B(\bar{x}))$. Therefore, $f(S)$ is open. Since open set S is arbitrary, f is an open mapping. \square

Remark (Properties). Functions with non-zero Jacobian determinant has following properties :

1. If $J_f \neq 0$ in n -ball B about \bar{a} which has different values at its boundaries, then $f(B)$ has an n -ball about $f(\bar{a})$.
2. If $J_f \neq 0$, f has continuous partial derivatives in S , and f is injective in an open set A , then $f(A)$ is open.

3. Let S be an open set in \mathbb{R}^n , f has continuous partial derivatives in S , and $J_f(\bar{a}) \neq 0$ for some $\bar{a} \in S$, then f is injective on an n -ball $B(\bar{a})$ in S .
4. Let A be an open set in \mathbb{R}^n , f has continuous partial derivatives in A , and $J_f \neq 0$ in A , then f is an open mapping.

Inverse function Theorem

Theorem 13.3.9 (Inverse function). *Let S be an open subset of \mathbb{R}^n and f be a continuously differentiable function⁴ $f : S \rightarrow \mathbb{R}^n$. If $J_f(\bar{a}) \neq 0$ for some $\bar{a} \in S$, then there are two open sets $X \subset S$, and $Y \subset f(S)$ such that*

1. $\bar{a} \in X$ and $f(\bar{a}) \in Y$
2. $Y = f(X)$
3. f is injective
4. there exists another function $g : Y \rightarrow X$ such that $g(f(\bar{x})) = \bar{x}$, $\forall \bar{x} \in X$
5. g is continuously differentiable on Y

In other words, if $f \in C'$ and there exists $\bar{a} \in S$ such that $J_f(\bar{a}) \neq 0$, then f has an inverse f^{-1} in a neighbourhood of $f(\bar{a})$ and $f^{-1} \in C'$.

Proof. Step 1 : Construction of open sets X and Y .

Given that, $J_f(\bar{a}) \neq 0$ and $f \in C'$. Thus all partial derivatives of f are continuous on S . Then J_f is continuous on S . By the continuity of J_f at \bar{a} , there exists a neighbourhood of \bar{a} , say $B_1(\bar{a})$ in which $J_f \neq 0$. That is, $\forall \bar{x} \in B_1(\bar{a})$, $J_f(\bar{x}) \neq 0$. Therefore, (by theorem) there exists an n -ball $B(\bar{a})$ on which f is injective. Let B be an n -ball with center \bar{a} contained in $B(\bar{a})$. Then f is injective on B . Therefore, (by theorem) $f(B)$ contains an n -ball with center $f(\bar{a})$. Let Y be the n -ball contained in $f(B)$. And $X = f^{-1}(Y) \cap B$. That is, the inverse image of Y on B . Since f is continuous, $f^{-1}(Y)$ is open. Thus, X is an intersection of open sets. And therefore, X is open.

Step 2 : The inverse of f , say g .

Clearly $\bar{a} \in X$ and $f(\bar{a}) \in Y$. Also $Y = f(X)$ and f is injective on X (since, $X \subset B$).

The closure of B , \bar{B} is compact and f is injective and continuous on \bar{B} . Then⁵ there exists a continuous function g defined on $f(\bar{B})$ such that $g \circ f$ is the identity function on \bar{B} . That is, $\forall x \in \bar{B}$, $g(f(\bar{x})) = \bar{x}$. Thus, $g(X) = Y$ and g is unique.

Step 3 : g has continuous partial derivatives.

Define a real-valued function $h : S^n \rightarrow \mathbb{R}$ by $h(\bar{Z}) = \det[D_j f_i(\bar{Z}_i)]$ where $\bar{Z}_1, \bar{Z}_2, \dots, \bar{Z}_n \in S$ and $\bar{Z} = (\bar{Z}_1, \bar{Z}_2, \dots, \bar{Z}_n)$. Now, let $\bar{Z} = (\bar{a}, \bar{a}, \dots, \bar{a})$. Then $h(\bar{Z}) \neq 0$ and h is continuous on S^n . Therefore, \bar{Z} has a neighbourhood on which h does not vanish (that is, nonzero). Let $B_2(\bar{a})$ be the corresponding n -ball with center \bar{a} such that $\bar{Z}_i \in B_2(\bar{a}) \implies h(\bar{Z}) \neq 0$.

Let B be an n -ball with center \bar{a} contained in $B_2(\bar{a})$. Now $\bar{B} \subset B_2(\bar{a})$. And $h(\bar{Z}) \neq 0$, $\forall \bar{Z}_i \in \bar{B}$.

⁴ $f \in C'(S) : f$ is continuously differentiable on S

⁵ Existence of inverse of a continuous function on a compact set in metric spaces.

We have, $g = (g_1, g_2, \dots, g_n)$. It is enough to prove that $g_k \in C'$ for $k = 1, 2, \dots, n$. Again, it is enough to prove that $D_r g_k$ exists and is continuous for $1 \leq r \leq n$. (Fix some r and prove that $D_r g_k$ is continuous.)

Let $\bar{y} \in Y$. Define $\bar{x} = g(\bar{y})$ and $\bar{x}' = g(\bar{y} + t\bar{u}_r)$ where t is sufficiently small such that $\bar{y} + t\bar{u}_r \in Y$. Then $\bar{x}, \bar{x}' \in X$. And $f(\bar{x}') - f(\bar{x}) = t\bar{u}_r$. Therefore $f_i(\bar{x}) - f_i(\bar{x}') = 0$ when $i \neq r$. And $f_i(\bar{x}') - f_i(\bar{x}) = t$ when $i = r$. By mean-value theorem,

$$\frac{f_i(\bar{x}') - f_i(\bar{x})}{t} = \nabla f_i(\bar{Z}_i) \cdot \frac{\bar{x}' - \bar{x}}{t}$$

where $\bar{Z}_i \in L(\bar{x}, \bar{x}')$, the line segment joining \bar{x} and \bar{x}' . Since $\det[D_j f_i(\bar{Z}_i)] = h(\bar{Z}) \neq 0$, this system of linear equations in n unknowns, $\frac{x'_j - x_j}{t}$ has a unique solution. As $t \rightarrow 0$, $\bar{x}' \rightarrow \bar{x}$. And $\bar{Z}_i \rightarrow \bar{x}$. Since $J_f(\bar{x}) \neq 0$, the limit

$$\lim_{t \rightarrow 0} \frac{g_k(\bar{y} + t\bar{u}_r) - g_k(\bar{y})}{t}$$

exists. Thus, $D_r g_k(\bar{y})$ exists $\forall y \in Y$ and every r . By Cramer's rule, this limit is a quotient of two determinants of partial derivatives of f , which are all continuous since $f \in C'$. Therefore, $D_r g_k$ are all continuous and $g \in C'$. \square

Implicit function Theorem

Theorem 13.3.10. Let S be an open set in \mathbb{R}^{n+k} and f be a function $f : S \rightarrow \mathbb{R}^n$. Suppose f is continuously differentiable on S . Let $(\bar{x}_0, \bar{t}_0) \in S$ such that $\bar{x}_0 \in \mathbb{R}^n$, $\bar{t}_0 \in \mathbb{R}^k$, $f(\bar{x}_0, \bar{t}_0) = \bar{0}$ and $J_f(\bar{x}_0, \bar{t}_0) \neq 0$. Then there exists an open set T_0 containing \bar{t}_0 in \mathbb{R}^k and a unique function $g : T_0 \rightarrow \mathbb{R}^n$ such that

1. g is continuously differentiable on T_0
2. $g(\bar{t}_0) = \bar{x}_0$
3. $f(g(\bar{t}), \bar{t}) = \bar{0}$, $\forall \bar{t} \in T_0$

Proof. Given $f : S \rightarrow \mathbb{R}^n$ where $S \subset \mathbb{R}^{n+k}$. We have $f = (f_1, f_2, \dots, f_m)$. Also given that $f \in C'(S)$, $f(\bar{x}_0; \bar{t}_0) = 0$, and $J_f(\bar{x}_0; \bar{t}_0) \neq 0$. Define a function $F : S \rightarrow \mathbb{R}^{n+k}$ defined by $F = (F_1, F_2, \dots, F_{n+k})$.

$$F_m(\bar{x}; \bar{t}) = \begin{cases} f_m(\bar{x}; \bar{t}) & 1 \leq m \leq n \\ t_{m-n} & n < m \leq n+k \end{cases}$$

For example, let $n = 3$, $k = 2$. Let $\bar{x} = (1, 2, 3)$ and $\bar{t} = (4, 5)$. Suppose $f_k(1, 2, 3, 4, 5) = a_k$. Then $F(\bar{x}; \bar{t}) = (a_1, a_2, a_3, 4, 5)$.

Then the Jacobian determinant of $F'(\bar{x}; \bar{t})$ is given by

$$J_F(\bar{x}; \bar{t}) = \begin{vmatrix} D_1 f_1 & D_2 f_1 & \dots & D_m f_1 & & & \\ D_1 f_2 & D_2 f_2 & \dots & D_m f_2 & & & \\ \vdots & \vdots & \ddots & \vdots & & & \\ D_1 f_m & D_2 f_m & \dots & D_m f_m & & & \\ & & & & 0 & \dots & 0 \\ & & & & \vdots & \ddots & \vdots \\ & & & & 0 & \dots & 1 \end{vmatrix}$$

Thus $J_F(\bar{x}_0; \bar{t}_0) = J_f(\bar{x}_0; \bar{t}_0) \neq 0$. Also, $F(\bar{x}_0; \bar{t}_0) = (\bar{0}; \bar{t}_0)$. Therefore, by inverse function theorem, there exists open sets X, Y containing $(\bar{x}_0; \bar{t}_0)$ and $(\bar{0}; \bar{t}_0)$ such that F is injective on X , $X = F^{-1}(Y)$ and there exists a unique local inverse function G such that $G(F(\bar{x}; \bar{t})) = (\bar{x}; \bar{t})$ and $G \in C'(Y)$.

Let $G = (v; w)$. That is, $v_i = G_i$ for $1 \leq i \leq n$. And $w_j = G_{n+j}$ for $1 \leq j \leq k$. We have, $G(F(\bar{x}; \bar{t})) = (\bar{x}; \bar{t})$. Therefore, $v(F(\bar{x}; \bar{t})) = \bar{x}$ and $w(F(\bar{x}; \bar{t})) = \bar{t}$. Since $X \subset F^{-1}(Y)$ and F is one-to-one on X , for every $(\bar{x}; \bar{t}) \in Y$, there exists $(\bar{x}'; \bar{t}') \in X$ such that $F(\bar{x}'; \bar{t}') = (\bar{x}; \bar{t})$. By the definition of F , $\bar{t}' = \bar{t}$. Therefore, $v(\bar{x}; \bar{t}) = v(F(\bar{x}'; \bar{t}')) = \bar{x}'$ and $w(\bar{x}; \bar{t}) = w(F(\bar{x}'; \bar{t}')) = \bar{t}$. Now, we have $G : Y \rightarrow X$ defined by $G(\bar{x}; \bar{t}) = (\bar{x}'; \bar{t})$.

Let T_0 be a subset of \mathbb{R}^k defined by $T_0 = \{\bar{t} \in \mathbb{R}^k : (\bar{0}; \bar{t}) \in Y\}$. For each $\bar{t} \in T_0$ let $g : T_0 \rightarrow \mathbb{R}^n$ is defined by $g(\bar{t}) = v(\bar{0}; \bar{t})$. The set T_0 is open. And $g \in C'(T_0)$ since g is constructed from the components of G which has continuous partial derivatives on Y . (ie, $G \in C'(Y)$.)

Clearly, $g(\bar{t}_0) = v(\bar{0}; \bar{t}_0) = \bar{x}_0$. And $(\bar{0}; \bar{t}) = F(\bar{x}_0; \bar{t}_0)$. Therefore, we have $f(v(\bar{x}; \bar{t}); \bar{t}) = \bar{x}$. Let $\bar{x} = \bar{0}$, then $f(g(\bar{t}); \bar{t}) = \bar{0}$. It is enough to prove that the function g is unique. Suppose $f(g(\bar{t}); \bar{t}) = f(h(\bar{t}); \bar{t})$. Since f is one-to-one on X , $(g(\bar{t}); \bar{t}) = (h(\bar{t}); \bar{t})$ for every $\bar{t} \in T_0$. And $g(\bar{t}) = h(\bar{t})$, $\forall \bar{t} \in T_0$. \square

Extrema of function of one variable

The function $f : \mathbb{R} \rightarrow \mathbb{R}$ defined by $f(x) = x^3$ has derivative $f'(x) = 3x^2$. Thus $f'(0) = 0$. However, 0 is not a local extrema for the function f . Thus, derivative of the function vanishing at point is not sufficient for a local extrema at that point.

Theorem 13.3.11 (sufficient condition for local extrema). *Let $n \geq 1$ and function f has n th partial derivative in open interval (a, b) . Suppose for some $c \in (a, b)$,*

$$f'(c) = f''(c) = \dots = f^{(n-1)}(c) = 0 \text{ and } f^{(n)}(c) \neq 0$$

If n is even,

1. f has a local minimum at c if $f^{(n)}(c) > 0$
2. f has a local maximum at c if $f^{(n)}(c) < 0$

and If n is odd, there is neither a local minimum nor a local maximum at c .

Proof. We have $f^{(n)}(c) \neq 0$. Thus, there exists an open interval $B(c)$ such that for each $x \in B(c)$, $f^{(n)}(x)$ has the same sign as $f^{(n)}(c)$. By Taylor's theorem, we have

$$f(x) = f(c) + \sum_{k=1}^{n-1} \frac{f^{(k)}(c)}{k!} (x-c)^k + \frac{f^{(n)}(x_1)}{n!} (x-c)^n \quad (13.14)$$

where $x_1 \in L(x, c)$, the line connecting x and c . We have, $f^{(k)}(c) = 0$ for $k = 1, 2, \dots, n-1$. Thus,

$$f(x) - f(c) = \frac{f^{(n)}(x_1)}{n!} (x-c)^n$$

Case 1 : n is even.

If n is even, then $(x - c)^n > 0$. Therefore, $f(x) - f(c)$ has the same sign as $f^{(n)}(x_1)$. If $f^{(n)}(c) > 0$, then $f^{(n)}(x_1)$ has a positive value and $f(x) - f(c) > 0$ for every $x \in B(c)$. Therefore, f has a local minimum at c . Similarly, if $f^{(n)}(c) < 0$, then $f(x) - f(c) < 0$ for every $x \in B(c)$. Therefore, f has a local maximum at c .

Case 2 : n is odd.

If n is odd, then $(x - c)^n$ takes both positive and negative values. Therefore, $f(x) - f(c)$ has both positive and negative values in $B(c)$. Thus f has neither local minimum nor local maximum at c . \square

Extrema of function of several variables

Definitions 13.3.3. If function f is differentiable at \bar{a} and $\nabla f(\bar{a}) = \bar{0}$, the point \bar{a} is a **stationary point** of f .

Definitions 13.3.4. A stationary point is a **saddle point** if every n -ball $B(\bar{a})$ contains points \bar{x} such that $f(\bar{x}) > f(\bar{a})$ and other points such that $f(\bar{x}) < f(\bar{a})$.

Definitions 13.3.5. A function $Q : \mathbb{R}^n \rightarrow \mathbb{R}$ defined by $Q(\bar{x}) = \sum_{i=1}^n \sum_{j=1}^n a_{ij} x_i x_j$ where $a_{ij} \in \mathbb{R}$ is a function of the **quadratic form** or simply a quadratic form.

symmetric quadratic form $a_{ij} = a_{ji}, \forall i, j$.

positive definite quadratic form $Q(\bar{x}) > 0, \forall \bar{x} \neq \bar{0}$.

negative definite quadratic form $Q(\bar{x}) < 0, \forall \bar{x} \neq \bar{0}$.

Suppose $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$. And second derivative of f exists if the derivative of f, f' is differentiable. Thus, $f'(\bar{a} + h\bar{t}) = f'(\bar{a}) + hf''(\bar{a})(\bar{t}) + |h|\|\bar{t}\|E_{\bar{a}}(h\bar{t})$ where $E_{\bar{a}} \rightarrow \bar{0}$ as $h \rightarrow 0$. Writing the Taylor's first order formula using matrices, we can see that $f''(\bar{a})(\bar{t})$ is of the quadratic form.

Theorem 13.3.12. Let f be a function $f : \mathbb{R}^n \rightarrow \mathbb{R}^m$. Suppose that the second order partial derivatives $D_{i,j}f$ exists in an n -ball $B(\bar{a})$ and are continuous at \bar{a} where \bar{a} is a stationary point of f .

$$\text{Let } Q(\bar{t}) = \frac{1}{2}f''(\bar{a}, \bar{t}) = \frac{1}{2} \sum_{i=1}^n \sum_{j=1}^n D_{i,j}f(\bar{a})t_i t_j$$

1. If $Q(\bar{t}) > 0$ for all $\bar{t} \neq \bar{0}$, f has a relative minimum at \bar{a} .
2. If $Q(\bar{t}) < 0$ for all $\bar{t} \neq \bar{0}$, f has a relative maximum at \bar{a} .
3. If $Q(\bar{t})$ takes both positive and negative values, then f has a saddle point at \bar{a} .

Proof. Define $Q : \mathbb{R}^n \rightarrow \mathbb{R}$ by $Q(\bar{t}) = \frac{1}{2}f''(\bar{a}; \bar{t})$. Then Q is continuous for every $\bar{t} \in \mathbb{R}^n$. Let S be the boundary of the n -ball $B(\bar{0}, 1)$. That is, $S = \{\bar{t} \in \mathbb{R}^n : \|\bar{t}\| = 1\}$.

Case 1 : Suppose that $Q(\bar{t}) > 0, \forall \bar{t} \neq \bar{0}$.

We have, Q is a continuous real-valued function on a compact interval, S . Therefore, Q attains its extrema. Thus Q has a minimum value at point in S , say m . Clearly $Q(\bar{t}) > 0 \implies m > 0$.

By Taylor's formula, $f(\bar{a} + \bar{t}) - f(\bar{a}) = \nabla f(\bar{a}) \cdot \bar{t} + \frac{1}{2}f''(\bar{z}; \bar{t})$ where $\bar{z} \in L(\bar{a} + \bar{t}, \bar{a})$

For every $\bar{a} \in S$, $\nabla f(\bar{a}) = \bar{0}$. And $f(\bar{a} + \bar{t}) - f(\bar{a}) \rightarrow \frac{1}{2}f''(\bar{a}; \bar{t})$ as $\bar{t} \rightarrow \bar{0}$.

$$\begin{aligned}\text{Thus, } f(\bar{a} + \bar{t}) - f(\bar{a}) &= \frac{1}{2}f''(\bar{a}; \bar{t}) + \|\bar{t}\|^2 E(\bar{t}) \\ &= Q(\bar{t}) + \|\bar{t}\|^2 E(\bar{t}) \text{ where } E(\bar{t}) \rightarrow \bar{0} \text{ as } \bar{t} \rightarrow \bar{0}\end{aligned}$$

Let $c = \frac{1}{\|\bar{t}\|}$. Then $c\bar{t} \in S$ and $Q(c\bar{t}) = c^2 Q(\bar{t}) \geq m$. Thus, $Q(\bar{t}) \geq m\|\bar{t}\|^2$. Therefore, $f(\bar{a} + \bar{t}) - f(\bar{a}) \geq m\|\bar{t}\|^2 + \|\bar{t}\|^2 E(\bar{t})$

Choose n -ball $B(\bar{0}, r)$ such that $|E(\bar{t})| < \frac{m}{2}$ for every $\bar{t} \in B(\bar{0}, r)$. Thus,

$$-\frac{m}{2}\|\bar{t}\|^2 \leq -\|\bar{t}\|^2 |E(\bar{t})| \leq 0$$

Therefore, $f(\bar{a} + \bar{t}) - f(\bar{a}) \geq m\|\bar{t}\|^2 - \frac{m}{2}\|\bar{t}\|^2 = \frac{m}{2}\|\bar{t}\|^2$ for every $\bar{t} \in B(\bar{0}, r)$. Clearly, f has a local minimum at \bar{a} .

Case 2 : Suppose $Q(\bar{t}) < 0$, $\forall \bar{t} \neq \bar{0}$, then consider $-f$.

Clearly, function $-f$ has a local minimum at \bar{t} . Thus f has a local maximum at \bar{t} .

Case 3 : Suppose $Q(\bar{t})$ takes both positive and negative values.

$$\begin{aligned}\text{We have, } f(\bar{a} + \lambda\bar{t}) - f(\bar{a}) &= Q(\lambda\bar{t}) + \lambda^2\|\bar{t}\|^2 E(\lambda\bar{t}) \\ &= \lambda^2[Q(\bar{t}) + \|\bar{t}\|^2 E(\lambda\bar{t})]\end{aligned}$$

Choose n -ball $B(\bar{0}, r)$ such that $\|\bar{t}\|^2 E(\lambda\bar{t}) < \frac{1}{2}|Q(\bar{t})|$, $\forall \bar{t} \in B(\bar{0}, r)$. We have $\bar{t} \in B(\bar{0}, r) \implies \lambda < r$. Then, error function $E(\bar{t})$ on the RHS is small enough, not to affect the sign of the RHS. Thus $f(\bar{a} + \lambda\bar{t}) - f(\bar{a})$ has the same sign as $Q(\bar{t})$. Therefore, \bar{a} is a saddle point. \square

Theorem 13.3.13. Let f be a real-valued function $f : \mathbb{R}^2 \rightarrow \mathbb{R}$ with continuous second order partial derivatives at a stationary point $\bar{a} \in \mathbb{R}^2$. Let $A = D_{1,1}f(\bar{a})$, $B = D_{1,2}f(\bar{a}) = D_{2,1}f(\bar{a})$, and $C = D_{2,2}f(\bar{a})$. And let $\Delta = \begin{vmatrix} A & B \\ B & C \end{vmatrix} = AC - B^2$. Then we have,

1. If $\Delta > 0$ and $A > 0$, then f has a relative minimum at \bar{a} .
2. If $\Delta > 0$ and $A < 0$, then f has a relative maximum at \bar{a} .
3. if $\Delta < 0$, then f has a saddle point at \bar{a} .

Proof. We have, function $f : \mathbb{R}^2 \rightarrow \mathbb{R}$ with $\nabla f(\bar{a}) = (0, 0)$. Consider the quadratic form $Q(x, y) = \frac{1}{2}[Ax^2 + Bxy + Cy^2]$ where $A = D_{1,1}f(\bar{a})$, $B =$

$D_{1,2}f(\bar{a})$, and $C = D_{2,2}f(\bar{a})$. Therefore, $Q(x, y) = \frac{1}{2}f''(\bar{a}; \bar{t})$.

Case 1 : Suppose $A \neq 0$.

$$\begin{aligned} Q(x, y) &= \frac{1}{2A}[A^2x^2 + ABxy + ACy^2] \\ &= \frac{1}{2A}[(Ax + By)^2 - B^2y^2 + ACy^2] \\ &= \frac{1}{2A}[(Ax + By)^2 + \Delta y^2] \end{aligned}$$

If $\Delta > 0$, then $Q(x, y)$ has the same sign as A . Therefore, f has a local minimum/maximum at \bar{a} depending on the sign of A .

Case 2 : Suppose $A = 0$. Then $Q(x, y) = \frac{1}{2}[Bxy + Cy^2] = \frac{1}{2}(Bx + Cy)y$.

Now we have two lines in \mathbb{R}^2 , $y = 0$ and $Bx + Cy = 0$. These two lines divide \mathbb{R}^2 into four regions. The value of $Q(x, y)$ is positive in two of those regions and negative in the other two regions. Therefore, $f(\bar{a} + \bar{t}) - f(\bar{a})$ assumes both positive and negative values in any neighbourhood $B(\bar{a}, r)$. Therefore, \bar{a} is a saddle point. \square

13.4 Integration on Differential Forms

Definitions 13.4.1. A k -cell in \mathbb{R}^k is given by,

$$I^k = \{\bar{x} \in \mathbb{R}^k : a_i \leq x_i \leq b_i, \forall i\}$$

where $\bar{a}, \bar{b} \in \mathbb{R}^k$. Let f be a continuous, real-valued function on I^k . Then, the **integral** of f over I^k is given by,

$$\begin{aligned} \int_{I^k} f(\bar{x}) d\bar{x} &= f_0 \text{ where } f_k = f \text{ and} \\ f_{r-1} &= \int_{a_r}^{b_r} f_r(x_0, x_1, \dots, x_r) dx_r, \quad r = 1, 2, \dots, k \end{aligned}$$

In other words,

$$\int_{I^k} f(\bar{x}) d\bar{x} = \int \cdots \int_{a_k}^{b_k} [f(x_0, x_1, \dots, x_k) dx_k] dx_{k-1} \cdots dx_1$$

Theorem 13.4.1. For every $f \in \mathcal{C}(I^k)$, $L(f) = L'(f)$.

In other words, integral of a function over a k -cell is independent of the order in which those k integrations are carried out.

Proof. Step 1 : “Separable” Functions ie, $h(\bar{x}) = \prod h_i(x_i)$.

(“separable” is not standard. It is only for the purpose of understanding.)

Let $h(\bar{x}) = h_1(x_1)h_2(x_2) \dots h_k(x_k)$ where $h_j \in [a_j, b_j]$.

$$L(h) = \int_{I^k} \left(\prod_{i=1}^k h_i(x_i) \right) d\bar{x} = \prod_{i=1}^k \int_{a_k}^{b_k} h_i(x_i) dx_i = L'(h)$$

Step 2 : Algebra of “separable” functions, \mathcal{A} .

Let \mathcal{A} be all finite sums of functions such as h . Let $g \in \mathcal{A}$.

$$\begin{aligned} L(g) &= \int_{I^k} \left(\sum_j \prod_i h_{i,j}(x_i) \right) d\bar{x} \\ &= \sum_j \int_{I^k} \prod_i h_{i,j}(\bar{x}) d\bar{x} \\ &= \sum_j \prod_i \int_{a_k}^{b_k} h_{i,j}(x_i) d(x_i) \\ &= L'(g) \end{aligned}$$

Step 3 : All functions continuous on I^k .

Let $f \in \mathcal{C}(I^k)$. ie, a function which is continuous in I^k .

Stone-Weierstrass theorem - Let \mathcal{A} be an algebra of real, continuous functions on a compact set K . If \mathcal{A} separates points on K and if \mathcal{A} vanishes at no point of K , then the uniform closure \mathcal{B} of \mathcal{A} consists of all real, continuous functions on K .

Clearly, the algebra of functions \mathcal{A} separates points on I^k and vanishes nowhere on I^k . Suppose $\bar{x} \neq \bar{y}$, then there exists m such that $x_m \neq y_m$. Thus $h(\bar{x}) = x_m$ separates \bar{x} and \bar{y} . Again $h(\bar{x}) = 1$ vanishes nowhere on I^k . Thus every function which is continuous on I^k is the limit of some uniformly convergent sequence of functions in \mathcal{A} .

Let $V = \prod_{j=1}^k (b_j - a_j)$. Then by Stone-Weierstrass theorem, for any $\epsilon > 0$, there exists a function $g \in \mathcal{A}$ such that $\|f - g\| < \frac{\epsilon}{V}$ where the norm of a function f is defined by $\|f\| = \max\{f(\bar{x}) : \bar{x} \in I^k\}$.

Therefore, it is sufficient to prove that $\|L(f) - L'(f)\| < \epsilon$. Since $\|f - g\| < \frac{\epsilon}{V}$, $|L(f - g)| < \epsilon$ and $|L'(f - g)| < \epsilon$. Thus,

$$\begin{aligned} L(f) - L'(f) &= L(f) - L(g) + L'(g) - L'(f) \\ &= L(f - g) + L'(g - f) \end{aligned}$$

$$|L(f) - L'(f)| < 2\epsilon$$

Therefore, $L(f) = L'(f)$. □

Definitions 13.4.2. Let $f : \mathbb{R}^k \rightarrow \mathbb{C}$. The **support** of f is the closure of the set of all points $\bar{x} \in \mathbb{R}^k$ such that $f(\bar{x}) \neq 0$.

Remark. Let f be a continuous function with compact support. And I^k be any k -cell containing the support of f . Then, $\int_{\mathbb{R}^k} f d\bar{x} = \int_{I^k} f d\bar{x}$.

$L(f) = L'(f)$ where f is the limit function of a sequence of functions with compact support. [Apostol, 1973, §10.4 Example]

Definitions 13.4.3. Let E be an open subset in \mathbb{R}^n . Then function $G : E \rightarrow \mathbb{R}^n$ is **primitive** if it satisfies

$$G(\bar{x}) = \sum_{i \neq m} x_i \bar{e}_i + g(\bar{x}) \bar{e}_m \quad (13.15)$$

for some integer m and some function $g : E \rightarrow \mathbb{R}$ (where \bar{e}_i are the unit co-ordinate vectors).

$$G(\bar{x}) = \bar{x} + [g(\bar{x}) - x_m] \bar{e}_m \quad (13.16)$$

If g is differentiable at \bar{a} , then G is also differentiable at \bar{a} . The matrix $[\alpha_{i,j}]$ of $G'(\bar{a})$ is given by

$$\alpha_{i,j} = \begin{cases} D_j g(\bar{a}) & i = m \\ 1 & i \neq m, j = i \\ 0 & i \neq m, j \neq i \end{cases}$$

The Jacobian of G at \bar{a} is given by, $J_G(\bar{a}) = \det[G'(\bar{a})] = D_m g(\bar{a})$.

Total derivative $G'(\bar{a})$ is invertible if and only if $D_m g(\bar{a}) \neq 0$.

Definitions 13.4.4. A linear operator B on \mathbb{R}^n that interchanges some pair of members of the standard basis and leaves the others fixed is a **flip**.

Theorem 13.4.2. Suppose F is a \mathcal{C}' -mapping of an open set $E \subset \mathbb{R}^n$ into \mathbb{R}^n , $\bar{0} \in E$, $F(\bar{0}) = \bar{0}$, and $F'(\bar{0})$ is invertible. Then there is a neighbourhood of $\bar{0}$ in \mathbb{R}^n in which the representation $F(\bar{x}) = B_1 B_2 \cdots B_{n-1} G_n \circ G_{n-1} \circ \cdots \circ G_1(\bar{x})$ is valid where G_i are primitive \mathcal{C}' -mapping in some neighbourhood of $\bar{0}$, $G_i(\bar{0}) = \bar{0}$, $G'(\bar{0})$ is invertible and each B_i is either a flip or identity operator.

In other words, locally F is a composition of primitive mappings and flips.

Proof. Proof by mathematical induction on m .

$$\text{There exists a neighbourhood of } \bar{0}, V_m \text{ such that } F_m \in \mathcal{C}'(V_m) \quad (13.17)$$

$$F_m(\bar{0}) = \bar{0} \quad (13.18)$$

$$F'_m(\bar{0}) \text{ is invertible and} \quad (13.19)$$

$$P_{m-1} F_m(\bar{x}) = P_{m-1} \bar{x}, \bar{x} \in V_m \quad (13.20)$$

where the k th projection $P_k : \mathbb{R}^n \rightarrow \mathbb{R}^n$ is defined by $P_k(\bar{x}) = x_1 \bar{e}_1 + x_2 \bar{e}_2 + \cdots + x_k \bar{e}_k + 0 \bar{e}_{k+1} + \cdots + 0 \bar{e}_n$. Clearly, $P_0(\bar{x}) = \bar{0}$.

The essence of this proof lies in the fourth statement which is designed to construct a sequence of functions F_1, F_2, \dots, F_n such that the other three statements remains true for every function in this sequence.

Step 1 : Initial Case, prove that all the four statements are true for $m = 1$.

Define $F_1 = F$ and $V_1 = E$. Thus, $F \in \mathcal{C}'(E) \implies F_1 \in \mathcal{C}'(V_1)$. Also we have, $F_1(\bar{0}) = F(\bar{0}) = \bar{0}$, and $F'_1(\bar{0}) = F'(\bar{0})$ is invertible. Obviously the trivial projection, $P_0(F_1(\bar{x})) = P_0(F(\bar{x})) = \bar{0} = P_0(\bar{x})$ for every $\bar{x} \in V_1$.

Step 2 : Induction Hypothesis, suppose all the four statements are true for $m = 1, 2, \dots, n-1$.

Suppose that for each $m = 1, 2, \dots, n-1$, there exists a neighbourhood of $\bar{0}$, say V_m such that $F_m \in \mathcal{C}'(V_m)$, $F_m(\bar{0}) = \bar{0}$, $F'_m(\bar{0})$ is invertible and $P_{m-1} F_m(\bar{x}) = P_{m-1} \bar{x}$ for every $\bar{x} \in V_m$.

Step 3 : Induction Step, prove that all the four statements are true for $m = n$.

$$\begin{aligned} P_{m-1} F_m(\bar{x}) &= P_{m-1} \bar{x}, \bar{x} \in V_m \\ &= x_1 \bar{e}_1 + x_2 \bar{e}_2 + \cdots + x_{m-1} \bar{e}_{m-1} \end{aligned}$$

We may write remaining components of $F_m(\bar{x})$ using real-valued functions. That is, k th component of the function F_m , say $F_{m_k} = \alpha_k$.

$$F_m(\bar{x}) = P_{m-1}(\bar{x}) + \sum_{i=m}^n \alpha_i(\bar{x}) \bar{e}_i$$

Since $F_m \in \mathcal{C}'(V_m)$, the functions $\alpha_i \in \mathcal{C}'(V_m)$. Taking m th partial derivative on either sides, we get

$$D_m F_m(\bar{0}) = F'_m(\bar{0})\bar{e}_m = \sum_{i=m}^n D_m \alpha_i(\bar{0})\bar{e}_i$$

We have, $F'_m(\bar{0})$ is invertible. Thus $F'_m(\bar{0})\bar{e}_m \neq 0$. Thus there exists some integer k such that $m \leq k \leq n$ and $D_m \alpha_k(\bar{0})\bar{e}_k \neq 0$. Let B_m be the flip that interchanges m and k . If $m = k$, then B_m is identity map.

$$\text{Define } G_m(\bar{x}) = \bar{x} + [\alpha_k(\bar{x}) - x_m]\bar{e}_m \quad (13.21)$$

Then $G_m \in \mathcal{C}'(V_m)$, G_m is a primitive mapping, and $G'_m(\bar{0})$ is invertible since $D_m \alpha_k(\bar{0}) \neq 0$.

By inverse function theorem, there exists an open set U_m , $0 \in U_m$ and $U_m \subset V_m$ such that G_m is a bijection from U_m onto a neighbourhood of $\bar{0}$, say V_{m+1} in which G_m^{-1} is continuously differentiable.

$$\text{Define } F_{m+1}(\bar{y}) = B_m F_m \circ G_m^{-1}(\bar{y}), \bar{y} \in V_{m+1} \quad (13.22)$$

Then $F_m \in \mathcal{C}'(V_{m+1})$, $F_{m+1}(\bar{0}) = 0$ and $F'_{m+1}(\bar{0})$ is invertible by the chain rule. Also, for $\bar{x} \in U_m$, we have

$$\begin{aligned} P_m F_{m+1}(G_m(\bar{x})) &= P_m B_m F_m(\bar{x}) \\ &= P_m(P_{m-1}\bar{x} + \alpha_k(\bar{x})\bar{e}_m + \cdots) \\ &= P_{m-1}\bar{x} + \alpha_k(\bar{x})\bar{e}_m \\ &= P_m G_m(\bar{x}) \\ \implies P_m F_{m+1}(\bar{y}) &= P_m(\bar{y}) \end{aligned}$$

Thus, by mathematical induction, we have characterised a finite sequence of function F_1, F_2, \dots, F_n such that all the four statements are true.

Step 4 : Using the finite sequence of functions F_1, F_2, \dots, F_n constructed in the proof, we can represent F in terms of flips and primitive mappings.

Let $\bar{x} \in U_m$ and $\bar{y} = G_m(\bar{x})$. We have,

$$\begin{aligned} B_m F_m \circ G_m^{-1}(\bar{y}) &= F_{m+1}(\bar{y}) \\ B_m B_m F_m(\bar{x}) &= B_m F_{m+1}(G_m(\bar{x})) \\ IF_m(\bar{x}) &= B_m F_{m+1}(G_m(\bar{x})) \\ F_m(\bar{x}) &= B_m(F_{m+1}(G_m(\bar{x}))) \end{aligned}$$

$$\text{Thus, } F_m = B_m F_{m+1} \circ G_m, \quad m = 1, 2, \dots, n \quad (13.23)$$

Therefore, we have

$$\begin{aligned} F &= F_1 \\ &= B_1 F_2 \circ G_1 \\ &= B_1(B_2 F_3 \circ G_2) \circ G_1 = B_1 B_2 F_3 \circ G_2 \circ G_1 \\ &\vdots \\ F &= B_1 B_2 \cdots B_n F_n \circ G_{n-1} \circ G_{n-2} \circ \cdots \circ G_1 \end{aligned}$$

Since $P_n F_n(\bar{x}) = P_{n-1}(\bar{x})$, F_n is a primitive mapping, say G_n .

$$F = B_1 B_2 \cdots B_n G_n \circ G_{n-1} \circ G_{n-2} \circ \cdots \circ G_1$$

□

Remark. Let K be a compact subset of \mathbb{R}^n . Then a family of function $\psi_1, \psi_2, \dots, \psi_s$ where $\psi_j : K \rightarrow \mathbb{R}$ is a partition of unity if it satisfies

1. $0 \leq \psi_j(\bar{x}) \leq 1$ for every $\bar{x} \in K$, $j = 1, 2, \dots, s$.
2. $\sum_{j=1}^s \psi_j(\bar{x}) = 1$ for every $\bar{x} \in K$.

Theorem 13.4.3 (partitions of unity). *Suppose K is a compact subset of \mathbb{R}^n , and $\{V_\alpha\}$ is an open cover of K . Then there exists functions $\psi_1, \psi_2, \dots, \psi_s \in \mathcal{C}(\mathbb{R}^n)$ such that*

1. $0 \leq \psi_i \leq 1$ for $1 \leq i \leq s$
2. each ψ_i has its support in some V_α and
3. $\psi_1(\bar{x}) + \psi_2(\bar{x}) + \cdots + \psi_s(\bar{x}) = 1$ for every $\bar{x} \in K$.

In other words, every open cover of a compact subset of \mathbb{R}^n has a partition of unity with compact support in a finite subcover.

Proof. **Step 1 : Construction of ϕ_j .**

We have, $\{V_\alpha\}$ is a cover of K . Therefore, every $\bar{x} \in K$ belongs to some $V_{\alpha(\bar{x})}$. And there exists open balls $B(\bar{x})$ and $W(\bar{x})$ such that

$$\bar{x} \in B(\bar{x}) \subset \bar{B}(\bar{x}) \subset W(\bar{x}) \subset \bar{W}(\bar{x}) \subset V_{\alpha(\bar{x})} \quad (13.24)$$

The family $\{B(\bar{x})\}$ is an open cover of K . Since K is compact, the open cover $\{B(\bar{x})\}$ has a finite subcover, say $\{B(\bar{x}_1), B(\bar{x}_2), \dots, B(\bar{x}_s)\}$. That is, there are points $\bar{x}_1, \bar{x}_2, \dots, \bar{x}_s$ in K such that

$$K = B(\bar{x}_1) \cup B(\bar{x}_2) \cup \cdots \cup B(\bar{x}_s) \quad (13.25)$$

Since \mathbb{R}^n is a metric space,⁶ there exists continuous functions $\phi_j : \mathbb{R}^n \rightarrow [0, 1]$ such that $\phi_j(B(\bar{x}_j)) = \{1\}$, $\phi_j(\mathbb{R}^n - W(\bar{x}_j)) = \{0\}$ and $0 \leq \phi_j(\bar{x}) \leq 1$ for every $\bar{x} \in \mathbb{R}^n$.

Step 2 : Construction of ψ_j .

$$\begin{aligned} \text{Define, } \psi_1 &= \phi_1 \\ \psi_2 &= (1 - \phi_1)\phi_2 \\ &\vdots \\ \psi_s &= (1 - \phi_1)(1 - \phi_2) \cdots (1 - \phi_{s-1})\phi_s \end{aligned}$$

Clearly, $0 \leq \psi_j \leq 1$ for $j = 1, 2, \dots, s$. And ψ_j has a compact support in $W(\bar{x}_j) \subset V_{\alpha(\bar{x}_j)}$.

Step 3 : $\{\psi_j\}$ is a partition of unity.

⁶metric spaces have a simpler version of Urysohn's lemma or apply Urysohn's lemma

We claim that,

$$\psi_1 + \psi_2 + \cdots + \psi_j = 1 - (1 - \phi_1)(1 - \phi_2) \cdots (1 - \phi_j), \quad j = 1, 2, \dots, s \quad (13.26)$$

It is proved using mathematical induction on j . Clearly, it is true for $j = 1$. $\psi_1 = \phi_1 = 1 - 1 + \phi = 1 - (1 - \phi)$. Suppose the claim is true for $j = k$ for some $1 \leq k < s$. Then we have, $\psi_1 + \psi_2 + \cdots + \psi_k = 1 - (1 - \phi_1)(1 - \phi_2) \cdots (1 - \phi_k)$.

$$\begin{aligned} \text{Thus, } \psi_1 + \psi_2 + \cdots + \psi_k + \psi_{k+1} &= 1 - (1 - \phi_1)(1 - \phi_2) \cdots (1 - \phi_k) \\ &\quad + (1 - \phi_1)(1 - \phi_2) \cdots (1 - \phi_k)\phi_{k+1} \\ &= 1 - (1 - \phi_1)(1 - \phi_2) \cdots (1 - \phi_{k+1}) \end{aligned}$$

Thus, the claim is true for $j = 1, 2, \dots, s$. Let $\bar{x} \in K$. Then $\bar{x} \in B(\bar{x}_j)$ for some j . By definition of ϕ_j , $\phi_j(\bar{x}) = 1$. That is, $(1 - \phi_j(\bar{x})) = 0$.

$$\begin{aligned} \implies (1 - \phi_1(\bar{x}))(1 - \phi_2(\bar{x})) \cdots (1 - \phi_s(\bar{x})) &= 0 \\ \implies \psi_1(\bar{x}) + \psi_2(\bar{x}) + \cdots + \psi_s(\bar{x}) &= 1 \end{aligned}$$

Therefore, $\psi_1(\bar{x}) + \psi_2(\bar{x}) + \cdots + \psi_s(\bar{x}) = 1$ for every $\bar{x} \in K$. □

Definitions 13.4.5. By theorem, every open cover $\{V_\alpha\}$ of a compact subset K has a partition of unity $\{\psi_j\}$ with compact support in a finite subcover $\{V_{\alpha_j}\}$. Then $\{\psi_j\}$ is **subordinate** to the cover $\{V_\alpha\}$.

Corollary 13.4.3.1. If $f \in \mathcal{C}(\mathbb{R}^n)$ and the support of f lies in K , then

$$f = \sum_{i=1}^s \psi_i f. \text{ Each } \psi_i \text{ has its support in some } V_\alpha.$$

Any continuous function with support in a compact set can be represented as sum of continuous functions $\psi_j f$ with small supports.

Proof. Let K be compact subset of \mathbb{R}^n . And support of a function f lies in K . Then,

$$\begin{aligned} f(\bar{x}) &= I(\bar{x})f(\bar{x}) = \left(\sum_{j=1}^s \psi_j(\bar{x}) \right) f(\bar{x}) \\ &= \sum_{j=1}^s \psi_j(\bar{x})f(\bar{x}) = \sum_{j=1}^s (\psi_j f)(\bar{x}) \end{aligned}$$

Let $\{\psi_j\}$ be a partition of unity. Let K' be the support of f and V_{α_j} be support of each ψ_j . Then $K' \cap V_{\alpha_j}$ is the support of each $\psi_j f$. □

Theorem 13.4.4 (effect of change of variable on multiple integral). Suppose T is a one-to-one \mathcal{C}' -mapping of an open set $E \subset \mathbb{R}^k$ into \mathbb{R}^k such that $J_T(\bar{x}) \neq 0$ for all $\bar{x} \in E$. If f is a continuous function on \mathbb{R}^k whose support is compact and lies in $T(E)$, then

$$\int_{\mathbb{R}^k} f(\bar{y}) \, d\bar{y} = \int_{\mathbb{R}^k} f(T(\bar{x})) |J_T(\bar{x})| \, d\bar{x} \quad (13.27)$$

Proof. **Step 1 : ‘separable’ functions**

Let E be an open subset of \mathbb{R}^k . We claim that⁷ the theorem is true for functions $h : E \rightarrow \mathbb{R}$ of the form $h(\bar{y}) = h_1(y_1)h_2(y_2) \cdots h_k(y_k)$. We know that, every continuous function T is locally a composition of primitives and flips. Therefore it is enough to prove that the theorem is true for functions h with transformations T primitives, flips and their compositions.

Step 2 : Transformation T is a primitive

Let G be a primitive with change in m th coordinate. Then

$$G(\bar{x}) = x_1\bar{e}_1 + x_2\bar{e}_2 + \cdots + g(\bar{x})\bar{e}_m + \cdots + x_k\bar{e}_k \quad (13.28)$$

We have,

$$J_G(\bar{x}) = \begin{vmatrix} 1 & 0 & \cdots & 0 & \cdots & 0 & 0 \\ 0 & 1 & \cdots & 0 & \cdots & 0 & 0 \\ \vdots & \vdots & \ddots & \vdots & \ddots & \vdots & 0 \\ D_1g(\bar{x}) & D_2g(\bar{x}) & \cdots & D_mg(\bar{x}) & \cdots & D_{k-1}g(\bar{x}) & D_kg(\bar{x}) \\ \vdots & \vdots & \ddots & 0 & \ddots & \vdots & \vdots \\ 0 & 0 & \cdots & 0 & \cdots & 1 & 0 \\ 0 & 0 & \cdots & 0 & \cdots & 0 & 1 \end{vmatrix} \quad (13.29)$$

Therefore, $J_G(\bar{x}) = D_mg(\bar{x})$.

$$\begin{aligned} \int_{\mathbb{R}^k} h(\bar{y}) \, d\bar{y} &= \prod_{j=1}^k \int h_j(y_j) \, dy_j \\ &= \left(\prod_{j \neq m} \int h_j(y_j) \, dy_j \right) \int h_m(y_m) \, dy_m \\ G(\bar{x}) = \bar{y} &\implies (x_1, x_2, \dots, g(\bar{x}), \dots, x_k) = (y_1, y_2, \dots, y_m, \dots, y_k) \\ &\implies x_j = y_j \text{ for } j \neq m \text{ and } g(\bar{x}) = y_m \\ \int_{\mathbb{R}^k} h(\bar{y}) \, d\bar{y} &= \left(\prod_{j \neq m} \int h_j(x_j) \, 1 \, dx_j \right) \int h_m(g(\bar{x})) \, D_mg(\bar{x}) \, dx_m \\ \int_{\mathbb{R}^k} h(\bar{y}) \, d\bar{y} &= \int_{\mathbb{R}^k} h(G(\bar{x})) \, |J_G(\bar{x})| \, d\bar{x} \end{aligned}$$

Step 3 : Transformation T is a flip.

Let B be a flip that interchanges m and n coordinates. When $m = n$, B is an identity map and the theorem is trivially true.

$$B(\bar{x}) = x_1\bar{e}_1 + \cdots + x_n\bar{e}_m + \cdots + x_m\bar{e}_n + \cdots + x_k\bar{e}_k \quad (13.30)$$

The Jacobian matrix is an identity matrix with m th and n th rows interchanged.

⁷This proof is my own work. There may exist simpler proofs.

For example,

$$J_B(\bar{x}) = \begin{vmatrix} 0 & 1 & \dots & 0 \\ 1 & 0 & \vdots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \dots & 1 \end{vmatrix} \quad (13.31)$$

Therefore, $J_B(\bar{x}) = \pm 1$.

$$\begin{aligned} \int_{\mathbb{R}^k} h(\bar{y}) \, d\bar{y} &= \prod_{j=1}^k \int h_j(y_j) \, dy_j \\ &= \left(\prod_{j \neq m, n} \int h_j(y_j) \, dy_j \right) \int h_n(y_n) \, dy_n \int h_m(y_m) \, dy_m \\ B(\bar{x}) = \bar{y} &\implies (x_1, \dots, x_n, \dots, x_m, \dots, x_k) = (y_1, \dots, y_m, \dots, y_n, \dots, y_k) \\ \int_{\mathbb{R}^k} h(\bar{y}) \, d\bar{y} &= \left(\prod_{j \neq m, n} \int h_j(x_j) \, dx_j \right) \int h_n(x_m) \, dx_n \int h_m(x_n) \, dx_m \\ &= \int_{\mathbb{R}^k} h(B(\bar{x})) \, 1 \, d\bar{x} = \int_{\mathbb{R}^k} h(B(\bar{x})) \, |J_B(\bar{x})| \, d\bar{x} \end{aligned}$$

Notice that, substituting $y_n = x_m$ gives $\int h_n(x_m) \, dx_n$. This is due to the fact that $\bar{y} = B(\bar{x})$ have $x_m \bar{e}_n$ in place of $y_n \bar{e}_n$. Therefore, parameter of h_n is on the same axis \bar{e}_n .

Step 4 : If the theorem is true for two transformations, then it is true for their composition.

Suppose the theorem is true for transformations P and Q . And $S = P \circ Q$. Let $\bar{z} = P(\bar{y})$ and $\bar{y} = Q(\bar{x})$. Then $\bar{z} = S(\bar{x})$. Since $m(P)m(Q) = m(S)$, we have $J_P(\bar{y})J_Q(\bar{x}) = J_S(\bar{x})$.

$$\begin{aligned} \int_{\mathbb{R}^k} f(\bar{z}) \, d\bar{z} &= \int_{\mathbb{R}^k} f(P(\bar{y})) \, |J_P(\bar{y})| \, d\bar{y} \\ &= \int_{\mathbb{R}^k} f(P(Q(\bar{x}))) \, |J_P(Q(\bar{x}))| \, |J_Q(\bar{x})| \, d\bar{x} \\ &= \int_{\mathbb{R}^k} f(S(\bar{x})) \, |J_S(\bar{x})| \, d\bar{x} \end{aligned}$$

Therefore, the theorem is true for their composition.

Now for any function of the form $h(\bar{x}) = \prod_{j=1}^k h_j(x_j)$, the theorem is true for any continuous transformation T . Then it is true for the algebra of functions \mathcal{A} . And by Stone-Weierstrass theorem,⁸ the theorem is true for every continuous function f . \square

13.4.1 Differential Forms

Definitions 13.4.6. Suppose E is an open set in \mathbb{R}^n . A k -**surface** in E is a \mathcal{C}' -mapping Φ from a compact set $D \subset \mathbb{R}^k$ into E . D is the **parameter**

⁸I haven't checked whether the application of Stone-Weierstrass theorem causes any problem. It is the same proof technique as we have seen in $L(f) = L'(f)$.

domain of Φ .

Definitions 13.4.7. Suppose E is an open set in \mathbb{R}^n . A **differential form** of order $k \geq 1$ in E , a k -form in E is a function ω , symbolically represented by the sum

$$\omega = \sum a_{i_1 \dots i_k}(\bar{x}) dx_{i_1} \wedge dx_{i_2} \wedge \dots \wedge dx_{i_k} \quad (13.32)$$

which assigns to each k -surface Φ in E a number $\omega(\Phi) = \int_{\Phi} \omega$, according to the rule

$$\int_{\Phi} \omega = \int_D \sum a_{i_1 \dots i_k}(\Phi(\bar{u})) \frac{\partial(x_{i_1}, x_{i_2}, \dots, x_{i_k})}{\partial(u_1, u_2, \dots, u_k)} d\bar{u} \quad (13.33)$$

where D is the parameter domain of Φ and

$$\frac{\partial(x_{i_1}, x_{i_2}, \dots, x_{i_k})}{\partial(u_1, u_2, \dots, u_k)} = \begin{vmatrix} \frac{\partial x_{i_1}}{\partial u_1} & \frac{\partial x_{i_1}}{\partial u_2} & \dots & \frac{\partial x_{i_1}}{\partial u_k} \\ \frac{\partial x_{i_2}}{\partial u_1} & \frac{\partial x_{i_2}}{\partial u_2} & \dots & \frac{\partial x_{i_2}}{\partial u_k} \\ \vdots & \vdots & \ddots & \vdots \\ \frac{\partial x_{i_k}}{\partial u_1} & \frac{\partial x_{i_k}}{\partial u_2} & \dots & \frac{\partial x_{i_k}}{\partial u_k} \end{vmatrix} \quad (13.34)$$

Let $\omega = 4 dx_1 \wedge dx_3 + 3x_2^2 dx_2 \wedge dx_1$. Then ω is a 2-form with $a_{1,3}(x, y, z) = 4$, $a_{2,1}(x, y, z) = 3y^2$ and all other $a_{i_1 i_2}(x, y, z) = 0$.

Consider the upper hemi-sphere of unit radius, $S_{y \geq 0}^2$ in \mathbb{R}^3 and the closure of the unit disc $\bar{S}^1 = D \subset \mathbb{R}^2$. Then D is compact. Let $\Phi : D \rightarrow S_{y \geq 0}^2$ defined by $\Phi(x, y) = (x, +\sqrt{1-x^2-y^2}, y)$. Clearly, $\Phi \in \mathcal{C}'(D)$. Thus, Φ is a 2-surface with parameter domain D .

Clearly, $\Phi_1(x, y) = x$, $\Phi_2(x, y) = \sqrt{1-x^2-y^2}$ and $\Phi_3(x, y) = y$. We have, $\omega(\Phi) = \int_{\Phi} \omega$.

$$\omega(\Phi) = \int_{\Phi} \omega = \iint_D \omega_{\Phi}(x, y) \frac{\partial(\Phi_{i_1}, \Phi_{i_2})}{\partial(x, y)} dx dy \quad (13.35)$$

where $\omega_{\Phi}(x, y)$ is obtained by evaluating each function $a_{i_1 i_2 \dots i_k}(\Phi(x, y))$.

$$\begin{aligned} \omega(\Phi) &= \iint_D a_{1,3}(\Phi(x, y)) \begin{vmatrix} \frac{\partial \Phi_1}{\partial x} & \frac{\partial \Phi_1}{\partial y} \\ \frac{\partial \Phi_3}{\partial x} & \frac{\partial \Phi_3}{\partial y} \end{vmatrix} dx dy + \iint_D a_{2,1}(\Phi(x, y)) \begin{vmatrix} \frac{\partial \Phi_2}{\partial x} & \frac{\partial \Phi_2}{\partial y} \\ \frac{\partial \Phi_1}{\partial x} & \frac{\partial \Phi_1}{\partial y} \end{vmatrix} dx dy \\ &= \int_0^1 \int_0^{\sqrt{1-y^2}} 4 \begin{vmatrix} 1 & 0 \\ 0 & 1 \end{vmatrix} dx dy \\ &\quad + \int_0^1 \int_0^{\sqrt{1-y^2}} 3(1-x^2-y^2) \begin{vmatrix} \frac{1}{2} \frac{-2x}{\sqrt{1-x^2-y^2}} & \frac{1}{2} \frac{-2y}{\sqrt{1-x^2-y^2}} \\ 1 & 0 \end{vmatrix} dx dy \\ &= \int_0^1 \int_0^{\sqrt{1-y^2}} 4 dx dy + \int_0^1 \int_0^{\sqrt{1-y^2}} \frac{3}{\sqrt{1-x^2-y^2}} dx dy \end{aligned}$$

Remark (example 1). Integrals of 1-forms are line integrals. And $\omega(\gamma) = 0$ for every closed curve γ . [Rudin, 1976, 10.12a]

Remark (example 2). Let $\gamma : [0, 2\pi] \rightarrow \mathbb{R}^2$ defined by $\gamma(t) = (a \cos t, b \sin t)$. Then γ is a 1-surface with parameter domain $[0, 2\pi]$. [Rudin, 1976, 10.12b] Let ω be a 1-form defined by $\omega = xdy$. Then

$$\omega(\gamma) = \int_{\gamma} \omega = \int_{\gamma} x dy = \int_0^{2\pi} ab \cos^2 t dt = \pi ab$$

Similarly, $\omega = ydx$ gives

$$\omega(\gamma) = \int_{\gamma} \omega = \int_{\gamma} y dx = \int_0^{2\pi} -ab \sin^2 t dt = -\pi ab$$

Remark (example 3). Let $0 \leq r \leq 1$, $0 \leq \theta \leq \pi$ and $0 \leq \phi \leq 2\pi$. Then $D \subset \mathbb{R}^n$ defined by $\{(r, \theta, \phi)\}$ is compact. Define $\Phi : D \rightarrow \mathbb{R}^3$ by $\Phi(r, \theta, \phi) = (r \sin \theta \cos \phi, r \sin \theta \sin \phi, r \cos \theta)$. Then Φ is a 3-surface in \mathbb{R}^3 . We have

$$J_{\Phi}(r, \theta, \phi) = \begin{vmatrix} \sin \theta \cos \phi & r \cos \theta \cos \phi & -r \sin \theta \sin \phi \\ \sin \theta \sin \phi & r \cos \theta \sin \phi & r \sin \theta \cos \phi \\ \cos \theta & -r \sin \theta & 0 \end{vmatrix} = r^2 \sin \theta$$

Let $\omega = dx_1 \wedge dx_2 \wedge dx_3$. Then $\omega(\Phi) = \int_{\Phi} \omega = \int_D J_{\Phi} = \frac{4\pi}{3}$ is the volume of the unit ball $\Phi(D)$.

Remark. A k -form ω is of class \mathcal{C}' or \mathcal{C}'' if the functions $a_{i_1 \dots i_k}$ are all of class \mathcal{C}' or \mathcal{C}'' . A 0-form in E is defined to be a continuous function in E . And 0 is the only k -form in any open set $E \subset \mathbb{R}^n$.

Definitions 13.4.8. Let $\omega = a(\bar{x}) dx_{i_1} \wedge dx_{i_2} \wedge \dots \wedge dx_{i_k}$. Then $\bar{\omega}$ is the k -form obtained by interchanging a pair subscripts of ω . ie, $\bar{\omega} = a(\bar{x}) dx_{i_2} \wedge dx_{i_1} \wedge \dots \wedge dx_{i_k}$.

Elementary properties of k -forms

Definitions 13.4.9. Let E be an open set in \mathbb{R}^n . And Φ be a k -surface in E . And let ω_1, ω_2 be k -forms in E , then

1. $\omega_1 = \omega_2 \iff \omega_1(\Phi) = \omega_2(\Phi) \iff \int_{\Phi} \omega_1 = \int_{\Phi} \omega_2, \forall \Phi \in E$
2. $\omega = 0 \iff \omega(\Phi) = 0 \iff \int_{\Phi} \omega = 0, \forall \Phi \in E$
3. k -form Addition, $\omega_1 + \omega_2$
 $\omega = \omega_1 + \omega_2 \iff \omega(\Phi) = \omega_1(\Phi) + \omega_2(\Phi) \iff \int_{\Phi} \omega = \int_{\Phi} \omega_1 + \int_{\Phi} \omega_2.$
4. Scalar multiplication, $c\omega$
 $c\omega(\Phi) = c(\omega(\Phi)) \iff \int_{\Phi} c\omega = c \int_{\Phi} \omega.$
5. Inverse k -form, $-\omega$
 $-\omega(\Phi) = -(\omega(\Phi)) \iff \int_{\Phi} -\omega = - \int_{\Phi} \omega$
6. $\bar{\omega} = -\omega.$

Remark. Let $\omega = a(\bar{x}) dx_{i_1} \wedge dx_{i_2} \wedge \dots \wedge dx_{i_k}$. If $\bar{\omega} = \omega$, then $\omega = 0$. Since \wedge is anticommutative.

Thus, differential k -forms with repeated subscripts are 0. For example, $\omega = dx_i \wedge dx_j \wedge dx_i = 0$.

Definitions 13.4.10. Let $\bar{I} = (i_1, i_2, \dots, i_k)$ be an **increasing k -index**. That is, $1 \leq i_1 \leq i_2 \leq \dots \leq i_k \leq n$. Then $dx_{\bar{I}}$ of the form $dx_{\bar{I}} = dx_{i_1} \wedge dx_{i_2} \cdots dx_{i_k}$ is **basic k -form** in \mathbb{R}^n .

Remark. List of all basic k -forms

0-form 0

1-forms dx_1, dx_2, \dots, dx_n

2-forms $dx_i \wedge dx_j$ for every $1 \leq i, j \leq n$.

3-forms $dx_i \wedge dx_j \wedge dx_k$ for every $1 \leq i, j, k \leq n$.

Remark. There are $\binom{n}{k}$ basic k -forms in \mathbb{R}^n .

Every k -form can be represented in terms of basic k -forms. For every k -tuple (j_1, j_2, \dots, j_k) , $dx_{j_1} \wedge \cdots \wedge dx_{j_k} = \sigma(j_1, j_2, \dots, j_k) dx_{\bar{J}}$ where \bar{J} is the increasing k -index obtained by interchanging pairs. And σ maps odd permutations to -1 and even permutations to 1 .

Standard representation of a k -form

$$\omega = \sum_I b_I(\bar{x}) dx_I \quad (13.36)$$

For example : $x_1 dx_2 \wedge dx_1 - x_2 dx_3 \wedge dx_2 + x_3 dx_2 \wedge dx_3 + dx_1 \wedge dx_2 = (1-x) dx_1 \wedge dx_2 + (x_2+x_3) dx_2 \wedge dx_3$ is a 2-form in \mathbb{R}^3 .

$dx_1 \wedge dx_2 \wedge dx_3 = dx_2 \wedge dx_3 \wedge dx_1$, since $(1\ 2\ 3) \in A_3 \implies \sigma(1\ 2\ 3) = 1$. And, $dx_1 \wedge dx_2 \wedge dx_3 = -dx_1 \wedge dx_3 \wedge dx_2$, $(2\ 3) \notin A_3 \implies \sigma(2\ 3) = -1$. Here, A_3 is an alternating group of all even permutation on $\{1, 2, 3\}$.

Subject 14

ME010304 Functional Analysis

Module 1 - Banach Space

14.1 Metric Spaces

The following are a few metric spaces and their distance functions,

1. $\mathbb{R}, d(x, y) = |x - y|$
2. $\mathbb{R}^2, d(x, y) = (|\xi_1 - \eta_1|^2 + |\xi_2 - \eta_2|^2)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2)$
3. $\mathbb{R}^3, d(x, y) = (|\xi_1 - \eta_1|^2 + |\xi_2 - \eta_2|^2 + |\xi_3 - \eta_3|^2)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2, \xi_3)$
4. Euclidean space, $\mathbb{R}^n, d(x, y) = \left(\sum_{j=1}^n |\xi_j - \eta_j|^2 \right)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2, \dots, \xi_n)$
5. $\mathbb{C}, d(x, y) = |x - y|$ where $x = a + ib$ and $|x| = (a^2 + b^2)^{\frac{1}{2}}$
6. Unitary space, $\mathbb{C}^n, d(x, y) = \left(\sum_{j=1}^n |\xi_j - \eta_j|^2 \right)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2)$ and $\xi_j \in \mathbb{C}$.
7. $C[a, b], d(x, y) = \max_{t \in [a, b]} \{|x(t) - y(t)|\}$ where x is a continuous, complex-valued function defined on closed interval $[a, b] \subset \mathbb{R}$.
8. $B(A), d(x, y) = \sup_{t \in A} \{|x(t) - y(t)|\}$ where x is a bounded, complex-valued function defined on $A \subset \mathbb{R}$
9. $l^p, d(x, y) = \left(\sum_{j=1}^{\infty} |\xi_j - \eta_j|^p \right)^{\frac{1}{p}}$ where $x = \{\xi_j\}$ and $\sum_{j=1}^{\infty} |\xi_j|^p < \infty$. That is, set of all sequences such that the p th power series is convergent.

10. $l^\infty, d(x, y) = \sup_{j \in \mathbb{N}} \{|\xi_j - \eta_j|\}$ where $x = \{\xi_j\}$ and $|\xi_j| \leq c$. That is, l^∞ is the space of all bounded sequences of complex numbers.
11. $c, d(x, y) = \sup_{j \in \mathbb{N}} \{|\xi_j - \eta_j|\}$ where $\{\xi_j\} = \xi_1, \xi_2, \dots$ is the space of all convergent sequences of complex numbers.
12. $s, d(x, y) = \sum_{j=1}^{\infty} \frac{1}{2^j} \frac{|\xi_j - \eta_j|}{1 + |\xi_j - \eta_j|}$ where $x = \{\xi_j\} = \xi_1, \xi_2, \dots$ is the space of all sequence of complex numbers.

Clearly, $l^p \subset l^\infty \subset s$. But, $B(A), C[a, b]$ are non-comparable since there exists discontinuous, bounded functions and continuous, unbounded functions.

Definitions 14.1.1 (conjugate exponents). Real numbers p, q are conjugates, if $p > 1, q > 1$ and $\frac{1}{p} + \frac{1}{q} = 1$.

Note : if p, q are conjugates, then $p + q = pq$ and $(p - 1)(q - 1) = 1$. Also if $u = t^{p-1}$, then $t = u^{q-1}$.

Theorem 14.1.1 (Hölder¹ Inequality for sums). Let p, q be conjugates and $(\xi_j), (\eta_j)$ be two complex sequences. $(\xi_j) \in l^p$ and $(\eta_j) \in l^q$. Then

$$\sum_{j=1}^{\infty} |\xi_j \eta_j| \leq \left(\sum_{j=1}^{\infty} |\xi_j|^p \right)^{\frac{1}{p}} \left(\sum_{j=1}^{\infty} |\eta_j|^q \right)^{\frac{1}{q}} \quad (14.1)$$

Corollary 14.1.1.1 (Cauchy-Schwarz Inequality). When $p = 2$, we have its conjugate $q = 2$. Then Hölder inequality reduces to the following

$$\sum_{j=1}^{\infty} |\xi_j \eta_j| \leq \left(\sum_{j=1}^{\infty} |\xi_j|^2 \right)^{\frac{1}{2}} \left(\sum_{j=1}^{\infty} |\eta_j|^2 \right)^{\frac{1}{2}} \quad (14.2)$$

Theorem 14.1.2 (Minkowski² Inequality). Let p, q be conjugates and $(\xi_j), (\eta_j)$ be two complex sequences in l^p . Then

$$\left(\sum_{j=1}^{\infty} |\xi_j + \eta_j|^p \right)^{\frac{1}{p}} \leq \left(\sum_{j=1}^{\infty} |\xi_j|^p \right)^{\frac{1}{p}} + \left(\sum_{j=1}^{\infty} |\eta_j|^p \right)^{\frac{1}{p}} \quad (14.3)$$

A few more Concepts

product of metric spaces Let $(X_1, d_1), (X_2, d_2)$ be two metric spaces. Let $x, y \in X_1 \times X_2$ where $x = (\xi_1, \xi_2)$ and $y = (\eta_1, \eta_2)$. Then, $X_1 \times X_2$ is a metric space with following metrics,

1. $d(x, y) = d_1(\xi_1, \eta_1) + d_2(\xi_2, \eta_2)$
2. $d(x, y) = \left(d_1(\xi_1, \eta_1)^2 + d_2(\xi_2, \eta_2)^2 \right)^{\frac{1}{2}}$

¹O.Hölder

²H. Minkowski

$$3. d(x, y) = \max\{d_1(\xi_1, \eta_1), d_2(\xi_2, \eta_2)\}$$

separable A space is separable if it has a countable, dense subset.

Examples : l^∞ is not separable. But, l^p spaces are separable ($1 \leq p < \infty$).
 \mathbb{R}, \mathbb{C} are separable. $B[a, b]$ not separable.

Exercises

1. Find a sequence $\{\xi_j\}$ which converges to 0 but does not belong to any l^p space. ($1 \leq p < \infty$).
2. Find a sequence $\{\xi_j\}$ which belong to l^p space with $p > 1$, but not in l^1 .

14.1.5 Completeness

Convergent sequences in a metric space are Cauchy sequences. But, Cauchy sequences are not necessarily convergent.

Definitions 14.1.2 (complete). A metric space is complete if every Cauchy sequence in it is convergent.

Theorem 14.1.3. \mathbb{R} is complete.

Proof. Let $\{\xi_n\}$ be a Cauchy sequence in \mathbb{R} . Let $M_0 > 0$. Then, there exists $N \in \mathbb{N}$ such that $\forall n, m > N, d(x_n, x_m) < M_0$. Let $M = \max\{M_0, M_1, \dots, M_N\}$ where $\forall n \leq N, |\xi_j| < M_j$. Then $\{\xi_n\}$ is bounded by M .

By Bolzano Weierstrass³ theorem, every bounded sequence in \mathbb{R}^n has a convergent subsequence. Let $x \in \mathbb{R}$ be the limit of a convergent subsequence $\{\xi_{n_k}\}$ of $\{\xi_n\}$. Then $\xi_n \rightarrow x$ since $\{\xi_n\}$ is a Cauchy sequence. \square

Theorem 14.1.4. \mathbb{C} is complete.

Proof. Let $\{\xi_n\}$ be a Cauchy sequence in \mathbb{C} . Then the real, imaginary parts of ξ_n are a Cauchy sequences in \mathbb{R} . We know that \mathbb{R} is complete. Let $\Re(\xi_n) \rightarrow a$ and $\Im(\xi_n) \rightarrow b$. Then $\xi_n \rightarrow a + ib$ since $\{\xi_n\}$ is a Cauchy sequence. \square

Theorem 14.1.5. Finite dimensional Euclidean space, \mathbb{R}^n is complete.

Proof. Let $\{x_k\}$ be Cauchy sequence in \mathbb{R}^n . Let $\varepsilon > 0$. There exists $N \in \mathbb{N}$ such that $\forall m, r > N, d(x_m, x_r) < \varepsilon$. Let $x_m = (\xi_{1,m}, \xi_{2,m}, \dots, \xi_{n,m})$ and $x_r = (\xi_{1,r}, \xi_{2,r}, \dots, \xi_{n,r})$.

$$\begin{aligned} d(x_m, x_r) &= \left(\sum_{j=1}^n |\xi_{j,m} - \xi_{j,r}|^2 \right)^{\frac{1}{2}} < \varepsilon \\ \implies \sum_{j=1}^n |\xi_{j,m} - \xi_{j,r}|^2 &< \varepsilon^2 \\ \implies |\xi_{j,m} - \xi_{j,r}| &< \varepsilon^2, \quad j = 1, 2, \dots, n \end{aligned}$$

³Every sequence has a monotone subsequence. And by monotone converges theorem, every monotone bounded sequence has a convergent subsequence.

$$\begin{array}{cccccc}
x_1 & = & \xi_{1,1} & \xi_{1,2} & \xi_{1,3} & \dots \\
x_2 & = & \xi_{2,1} & \xi_{2,2} & \xi_{2,3} & \dots \\
x_3 & = & \xi_{3,1} & \xi_{3,2} & \xi_{3,3} & \dots \\
\vdots & & \vdots & \vdots & \vdots & \vdots \\
x & = & \xi_1 & \xi_2 & \xi_3 & \dots
\end{array}$$

Figure 14.1: Construction of limit in Sequence spaces

Therefore, $\{\xi_{j,k}\}$'s are Cauchy sequences in \mathbb{R} for $j = 1, 2, \dots, n$. Since \mathbb{R} is convergent, $\xi_{j,k} \rightarrow \xi_j$ for each j . Let $x = (\xi_1, \xi_2, \dots, \xi_n)$. Let $\varepsilon > 0$. Then there exists $N_j \in \mathbb{N}$ such that $\forall m > N_j$, $|\xi_{j,m} - \xi_j| < \frac{\varepsilon}{n}$. Let $N = \max\{N_1, N_2, \dots, N_n\}$. Then $\forall m > N$,

$$d(x_m, x) = \left(\sum_{j=1}^n |\xi_{j,m} - \xi_j|^2 \right)^{\frac{1}{2}} = \left(\sum_{j=1}^n \frac{\varepsilon^2}{n^2} \right)^{\frac{1}{2}} = \frac{\varepsilon}{\sqrt{n}} < \varepsilon$$

Therefore, $\{x_n\}$ converges to x in \mathbb{R}^n . □

Theorem 14.1.6. *Finite dimensional Unitary space, \mathbb{C}^n is complete.*

Proof. Same proof as above. □

Theorem 14.1.7. *The complex sequence space of all bounded sequences, l^∞ is complete.*

Proof. Let $\{x_k\}$ be a Cauchy sequence in l^∞ where $x_k = \xi_{k,1}, \xi_{k,2}, \dots$ are bounded sequences in \mathbb{C} for each k .

Step 1 : Construct x

Let $\varepsilon > 0$. Then there exists $N \in \mathbb{N}$ such that $\forall m, n > N$,

$$d(x_m, x_n) = \sup_{k \in \mathbb{N}} \{|\xi_{m,k} - \xi_{n,k}|\} < \varepsilon$$

From the metric, we have $\{\xi_{j,k}\}$'s are Cauchy sequences in \mathbb{C} for each k . Thus, $\xi_{j,k} \rightarrow \xi_k \in \mathbb{C}$. Let $x = \xi_1, \xi_2, \dots$. Now we need to prove that $x_k \rightarrow x$ and $x \in l^\infty$.

Step 2 : $x \in l^\infty$

We have, $x_j \in l^\infty \implies |\xi_{m,j}| < c_j, \forall m \in \mathbb{N}$. Therefore,

$$|\xi_j| \leq |\xi_j - \xi_{m,j}| + |\xi_{m,j}| \leq \varepsilon + c_j, \quad \forall j \in \mathbb{N}$$

Thus, $x \in l^\infty$.

Step 3 : $x_n \rightarrow x$

Since $|\xi_{m,j} - \xi_j| < \varepsilon$ for $m \in \mathbb{N}$,

$$d(x_m, x) = \sup_{j \in \mathbb{N}} |\xi_{m,j} - \xi_j| \leq \varepsilon$$

□

Theorem 14.1.8. *The complex sequence space of all convergent sequences, c is complete*

Proof. Since every convergent sequence is bounded, $c \subset l^\infty$. And the sequence space c is complete if c is a closed subset of the complete space l^∞ . Therefore, it is sufficient to show that $c = \bar{c}$.

Let $x = \{\xi_k\} \in \bar{c}$. Then there exists a convergent sequence $\{x_k\} = \{\xi_{k,j}\}$ in c converging to x . That is, $\xi_{k,j} \rightarrow \xi_j$.

Let $\varepsilon > 0$. Then there exists $N \in \mathbb{N}$ such that $\forall n \geq N$ and $\forall j, k \in \mathbb{N}$, we have $d(x_N, x) = \sup_{r \in \mathbb{N}} \{|\xi_{n,r} - \xi_r|\} < \frac{\varepsilon}{3}$. Therefore,

$$|\xi_{N,j} - \xi_j| < \frac{\varepsilon}{3} \text{ and } |\xi_{N,k} - \xi_k| < \frac{\varepsilon}{3} \quad (14.4)$$

Every convergent sequence in complete metric space l^∞ is also a Cauchy sequence. Since $x_N = \{\xi_{N,k}\}$ is a Cauchy sequence, there exists $N_1 \in \mathbb{N}$ and $\forall j, k \geq N_1$,

$$|\xi_{N,j} - \xi_{N,k}| < \frac{\varepsilon}{3} \quad (14.5)$$

Therefore,

$$|\xi_j - \xi_k| \leq |\xi_j - \xi_{N,j}| + |\xi_{N,j} - \xi_{N,k}| + |\xi_{N,k} - \xi_k| \leq \frac{\varepsilon}{3} + \frac{\varepsilon}{3} + \frac{\varepsilon}{3} = \varepsilon$$

Clearly, $x = \{\xi_k\}$ is a Cauchy sequence in l^∞ . And since l^∞ is complete, every Cauchy sequence in l^∞ is convergent. Therefore, x is a convergent sequence and $x \in c$. Thus, $c = \bar{c}$ and the induced^{†4} metric space c is complete. \square

Theorem 14.1.9. *For any $p \geq 1$, the sequence space l^p is complete.*

Proof. Let $\{x_k\}$ be a Cauchy sequence l^p where $x_k = \{\xi_{k,j}\}$ and

$$\sum_{j=1}^{\infty} |\xi_{k,j}|^p < \infty, \quad \forall k \in \mathbb{N}$$

Since $\{x_k\}$ is a Cauchy sequence, for $\varepsilon > 0$ there exists $N \in \mathbb{N}$ such that

$$d(x_m, x_n) = \left(\sum_{j=1}^{\infty} |\xi_{m,j} - \xi_{n,j}|^p \right)^{\frac{1}{p}} < \varepsilon, \quad \forall m, n > N \text{ since } x_k \in l^p \quad (14.6)$$

Thus, $\forall m, n > N$, $|\xi_{m,j} - \xi_{n,j}| < \varepsilon$ for each j .

Step 1 : Construction of x

Define $x = \{\xi_j\}$ where ξ_j is an accumulation point of $\{\xi_{k,j}\}$.

Step 2 : $x \in l^p$

Let $j > N$. Then, $\{\xi_{k,j}\}$ is a Cauchy sequence of complex numbers. Then,

^{†4}Let (X, d) be a complete metric space. And Y is a closed subset of X . Then the induced metric space $(Y, d|_Y)$ is complete.

$\xi_{k,j} \rightarrow \xi_j$. Applying limit $n \rightarrow \infty$ on equation 14.6, we get

$$\begin{aligned} \left(\sum_{j=1}^{\infty} |\xi_{m,j} - \xi_j|^p \right)^{\frac{1}{p}} &< \varepsilon, \quad \forall m > N \\ \implies \sum_{j=1}^{\infty} |\xi_{m,j} - \xi_j|^p &< \varepsilon^p \end{aligned}$$

Let $m > N$. Then $x_m - x \in l^p$. By Minkowski inequality, we have

$$\sum_{j=1}^{\infty} |\xi_j|^p \leq \sum_{j=1}^{\infty} |\xi_{m,j} - \xi_j|^p + \sum_{j=1}^{\infty} |\xi_{m,j}|^p < \infty \quad (14.7)$$

Thus, $x \in l^p$.

Step 3: $x_m \rightarrow x$

Since $\sum_{j=1}^{\infty} |\xi_{m,j} - \xi_j|^p < \varepsilon^p$, we have $\xi_{k,j} \rightarrow \xi_j$. Therefore, $x_m \rightarrow x$. \square

Theorem 14.1.10. *The function space of all continuous functions defined on closed interval $[a, b]$, $C[a, b]$ is complete.*

Proof. Let $\{x_k\}$ be a Cauchy sequence in $C[a, b]$.

Let $\varepsilon > 0$. Then there exists $N \in \mathbb{N}$ such that $m, n > N$,

$$d(x_m, x_n) = \max_{t \in [a, b]} \{x_m(t) - x_n(t)\} < \varepsilon$$

Thus, for each $t \in [a, b]$, sequence $\{x_k(t)\}$'s are Cauchy sequences. We know that \mathbb{C} is complete. Thus every Cauchy sequence in \mathbb{C} is convergent.

Define $x : [a, b] \rightarrow \mathbb{C}$ defined by $x(t) =$ the limit of the sequence $\{x_k(t)\}$. As $n \rightarrow \infty$, $d(x_m, x_n) \rightarrow d(x_m, x)$ and for any $m > N$, $d(x_m, x) < \varepsilon$ uniformly. It is evident from the construction that, the convergence $x_k(t) \rightarrow x(t)$ is uniform.

Since the function x_k 's are continuous and the convergence is uniform, the limit function x is also continuous. Therefore $x \in C[a, b]$. \square

Definitions 14.1.3 (uniform metric). Let (X, d) be a metric space in which every convergence $x_k \rightarrow x$ is uniform. Then the metric d is a uniform metric.

For example, the metric of $C[a, b]$ is a uniform metric.

A few examples of incomplete metric spaces

The following metric spaces are not complete,

1. The space of rational numbers, \mathbb{Q} with usual metric $d(x, y) = |x - y|$ is not complete since the rational approximations of π is sequence in \mathbb{Q} which doesn't converge in \mathbb{Q} .

$$3, 3.1, 3.14, 3.141, \dots \rightarrow \pi \notin \mathbb{Q}$$

2. The space of polynomial functions with metric $d(x, y) = \sup\{x(t) - y(t)\}$ is not complete since Taylor approximations of $\sin x$ is a sequence of polynomial functions which doesn't converge to a polynomial function.

$$x, x + \frac{-x^3}{3!}, x + \frac{-x^3}{3!} + \frac{x^5}{5!}, \dots \rightarrow \sin x \notin p$$

3. The space of continuous function on unit interval $[0, 1]$ with a different⁵ metric

$$d(x, y) = \int_0^1 |x(t) - y(t)| dt$$

is not complete since $\{x_k\}$ where $x_k : [0, 1] \rightarrow \mathbb{R}$ is defined by

$$x_k(t) = \begin{cases} 0 & t \in [0, \frac{1}{2}) \\ (t - \frac{1}{2})k & t \in [\frac{1}{2}, \frac{1}{2} + \frac{1}{k}) \\ 1 & t \in [\frac{1}{2} + \frac{1}{k}, 1] \end{cases}$$

doesn't converge to a continuous polynomial function.

14.1.6 Completion of Metric Space

Let Y be an incomplete metric space. Completion of Y is a construction of a complete metric space X in which Y is dense.

For example, \mathbb{Q} is not complete. However, \mathbb{R} is a complete metric space in which \mathbb{Q} is complete. Therefore, \mathbb{R} is a completion of \mathbb{Q} .

Definitions 14.1.4 (Isometry). Isometric functions are distance preserving functions. And two metric spaces are isometric if there exists a bijective isometry between them.

For example, function $f : X \rightarrow Y$ is an isometry if

$$\forall x, y \in X, \quad \hat{d}(f(x), f(y)) = d(x, y)$$

where d, \hat{d} are metrics in X, Y respectively. And if f is a bijection, then X, Y are isometric space.

Two metric spaces are isometric is another way of saying that the spaces are identical (same) from metric point of view.

Theorem 14.1.11 (completion). Let (X, d) be a metric space. Then, there exists a complete metric space, (\hat{X}, \hat{d}) such that it has a dense subspace W which is isometric with X . And \hat{X} is unique upto isometries.

In other words, for any metric space X there exists a unique complete metric space, \hat{X} in which X is dense.

Proof. Step 1 : Construction of (\hat{X}, \hat{d})

Let (X, d) be a metric space. Let C be the set of all Cauchy sequence in X . Define relation $\{x_k\} \sim \{y_k\}$ if and only if $\lim_{k \rightarrow \infty} d(x_k, y_k) = 0$. This is an equivalence relation. (proof not required)

⁵Area under the graph of difference function is well-defined for integrable functions.

1. Reflexive - $\hat{x} \sim \hat{x}$ since $\lim_{k \rightarrow \infty} d(x_k, x_k) = 0$.
2. Symmetric - $\hat{x} \sim \hat{y} \implies \lim_{k \rightarrow \infty} d(x_k, y_k) = \lim_{k \rightarrow \infty} d(y_k, x_k) \implies \hat{y} \sim \hat{x}$.
3. Transitive - Suppose, $\hat{x} \sim \hat{y}$ and $\hat{y} \sim \hat{z}$.
 $\hat{x} \sim \hat{y} \implies \hat{d}(\hat{x}, \hat{y}) = \lim_{k \rightarrow \infty} d(x_k, y_k) = 0$, $\hat{y} \sim \hat{z} \implies \hat{d}(\hat{y}, \hat{z}) = \lim_{k \rightarrow \infty} d(y_k, z_k) = 0$.
Then $\hat{d}(\hat{x}, \hat{z}) = \lim_{k \rightarrow \infty} d(x_k, z_k) \leq \lim_{k \rightarrow \infty} d(x_k, y_k) + d(y_k, z_k) = \hat{d}(\hat{x}, \hat{y}) + \hat{d}(\hat{y}, \hat{z}) = 0$.
Therefore $\hat{x} \sim \hat{z}$.

Let \hat{X} be the set of all equivalent classes in C . Define $\hat{d} : \hat{X} \times \hat{X} \rightarrow \mathbb{R}$ given by $\hat{d}(\hat{x}, \hat{y}) = \lim_{k \rightarrow \infty} d(x_k, y_k)$ where the Cauchy sequences $\{x_k\} \in \hat{x}$ and $\{y_k\} \in \hat{y}$.

Then \hat{d} is metric in \hat{X} .

1. \hat{d} is well-defined
Suppose $\{x_k\}, \{x'_k\} \in \hat{x}$ and $\{y_k\}, \{y'_k\} \in \hat{y}$. Then $\{x_k\} \sim \{x'_k\}$ and $\{y_k\} \sim \{y'_k\}$. In other words, $\lim_{k \rightarrow \infty} d(x_k, x'_k) = 0$ and $\lim_{k \rightarrow \infty} d(y_k, y'_k) = 0$.

By triangular inequality, we have

$$\begin{aligned} d(x_k, y_k) &\leq d(x_k, x'_k) + d(x'_k, y'_k) + d(y'_k, y_k) \\ \implies d(x_k, y_k) - d(x'_k, y'_k) &\leq d(x_k, x'_k) + d(y_k, y'_k) \end{aligned}$$

Similarly,

$$\begin{aligned} d(x'_k, y'_k) &\leq d(x'_k, x_k) + d(x_k, y_k) + d(y_k, y'_k) \\ \implies d(x'_k, y'_k) - d(x_k, y_k) &\leq d(x_k, x'_k) + d(y_k, y'_k) \end{aligned}$$

Therefore,

$$|d(x_k, y_k) - d(x'_k, y'_k)| \leq d(x_k, x'_k) + d(y_k, y'_k)$$

Apply the limit $k \rightarrow \infty$ on either sides, we get

$$\lim_{k \rightarrow \infty} |d(x_k, y_k) - d(x'_k, y'_k)| \leq \lim_{k \rightarrow \infty} d(x_k, x'_k) + \lim_{k \rightarrow \infty} d(y_k, y'_k) = 0$$

Thus, $\hat{d}(\hat{x}, \hat{y})$ depends only on the equivalent class \hat{x}, \hat{y} to which x_k, y_k belongs and is independent of the representative from these equivalent classes. Therefore, $\hat{d} : \hat{X} \times \hat{X} \rightarrow \mathbb{R}$ is well-defined.

2. $\hat{d}(\hat{x}, \hat{y}) = 0 \iff \hat{x} = \hat{y}$
 $\hat{d}(\hat{x}, \hat{y}) = 0 \iff \forall \{x_k\} \in \hat{x}, \forall \{y_k\} \in \hat{y}, \lim_{k \rightarrow \infty} d(x_k, y_k) = 0 \iff \{x_k\} \sim \{y_k\}$
3. $\hat{d}(\hat{x}, \hat{y}) = \hat{d}(\hat{y}, \hat{x})$ is trivial since $d(x_k, y_k) = d(y_k, x_k)$.
4. $\hat{d}(\hat{x}, \hat{y}) \leq \hat{d}(\hat{x}, \hat{z}) + \hat{d}(\hat{z}, \hat{y})$

Let $\{z_k\}$ be a Cauchy sequence in X . Then $\{z_k\} \in \hat{z}$ for some $\hat{z} \in \hat{X}$,

$$d(x_k, y_k) \leq d(x_k, z_k) + d(z_k, y_k)$$

Applying limit $k \rightarrow \infty$ on either sides, we get

$$\hat{d}(\hat{x}, \hat{y}) = \lim_{k \rightarrow \infty} d(x_k, y_k) \leq \lim_{k \rightarrow \infty} d(x_k, z_k) + \lim_{k \rightarrow \infty} d(z_k, y_k) = \hat{d}(\hat{x}, \hat{z}) + \hat{d}(\hat{z}, \hat{y})$$

Step 2: Construction of Isometry $T : X \rightarrow W$, $W \subset \hat{X}$.

Let $b \in X$. Then b, b, \dots is a Cauchy sequence in \hat{X} . Let $\hat{b} \in \hat{X}$ be the equivalent class of Cauchy sequences containing the constant sequence $\{b\}$. Define $T : X \rightarrow \hat{X}$ given by $T(b) = \hat{b}$.

Let $a, b \in X$, then

$$\begin{aligned} Ta = Tb &\implies \{a\} \sim \{b\} \\ &\implies \lim_{k \rightarrow \infty} d(a, b) = 0 \\ &\implies d(a, b) = 0 \implies a = b \end{aligned}$$

Thus $T : X \rightarrow T(X)$ is a bijection. Also T is an isometry since

$$\hat{d}(\hat{a}, \hat{b}) = \hat{d}(Ta, Tb) = \lim_{k \rightarrow \infty} d(a, b) = d(a, b)$$

Step 3: $W = T(X)$ is dense in \hat{X}

Let $\hat{x} \in \hat{X}$, then $\{x_k\} \in \hat{x}$. Then, $\{x_k\}$ is a Cauchy sequence.

Let $\varepsilon > 0$. Then there exists $N \in \mathbb{N}$ such that $\forall m, n \geq N$, $d(x_n, x_m) < \frac{\varepsilon}{2}$.

Thus,

$$\forall \varepsilon > 0, \exists N \in \mathbb{N} \text{ such that } d(x_n, x_N) < \frac{\varepsilon}{2}$$

Clearly, $x_N \in X$ and the constant sequence $\{x_N\} \in \hat{x}_N \in T(X) = W$.

Thus,

$$\hat{d}(\hat{x}, \hat{x}_N) = \lim_{k \rightarrow \infty} d(x_k, x_N) \leq \frac{\varepsilon}{2} < \varepsilon$$

For every $\varepsilon > 0$, there exists $\hat{x}_N \in W$ such that $\hat{d}(\hat{x}, \hat{x}_N) < \varepsilon$. Therefore, W is dense in \hat{X} .

Step 4: \hat{X} is complete

Let $\{\hat{x}_n\}$ be any Cauchy sequence in \hat{X} . Since W is dense in \hat{X} , for each \hat{x}_n there exists $\hat{z}_n \in W$ such that $\hat{d}(\hat{x}_n, \hat{z}_n) < \frac{1}{n}$. Then, $\{\hat{z}_n\}$ is a Cauchy sequence since,

$$\hat{d}(\hat{z}_m, \hat{z}_n) \leq \hat{d}(\hat{z}_m, \hat{x}_m) + \hat{d}(\hat{x}_m, \hat{x}_n) + \hat{d}(\hat{x}_n, \hat{z}_n) \leq \frac{1}{m} + \hat{d}(\hat{x}_m, \hat{x}_n) + \frac{1}{n}$$

We have, $T : X \rightarrow W$ defined by $T(z_n) = \hat{z}_n$ is a bijective, isometry. Thus, sequence $\{z_n\}$ is a Cauchy sequence in X , since $\hat{d}(\hat{z}_n, \hat{z}_m) = d(z_n, z_m)$. Let $\hat{x} \in \hat{X}$ be the equivalent class of Cauchy sequences containing $\{z_n\}$. Then,

$$\hat{d}(\hat{x}_n, \hat{x}) \leq \hat{d}(\hat{x}_n, \hat{z}_n) + \hat{d}(\hat{z}_n, \hat{x}) = \frac{1}{n} + \hat{d}(\hat{z}_n, \hat{x})$$

Apply limit $n \rightarrow \infty$ on either sides, we get

$$\lim_{n \rightarrow \infty} \hat{d}(\hat{x}_n, \hat{x}) \leq \lim_{n \rightarrow \infty} \frac{1}{n} + \lim_{n \rightarrow \infty} \hat{d}(\hat{z}_n, \hat{x}) = 0$$

Thus, $\hat{x}_n \rightarrow \hat{x}$ as $n \rightarrow \infty$. Therefore \hat{X} is complete, since every Cauchy sequence in \hat{X} converges.

Step 5: Uniqueness of \hat{X}

Suppose (\tilde{X}, \tilde{d}) is a complete metric space with dense subset \tilde{W} which is isometric with X . Let $\tilde{x}, \tilde{y} \in \tilde{X}$. Since \tilde{W} is dense in \tilde{X} , there exists sequences $\{\tilde{x}_n\}, \{\tilde{y}_n\}$ in \tilde{W} such that $\tilde{x}_n \rightarrow \tilde{x}$ and $\tilde{y}_n \rightarrow \tilde{y}$.

We have,

$$\tilde{d}(\tilde{x}, \tilde{y}) \leq \tilde{d}(\tilde{x}, \tilde{x}_n) + \tilde{d}(\tilde{x}_n, \tilde{y}_n) + \tilde{d}(\tilde{y}_n, \tilde{y})$$

$$\tilde{d}(\tilde{x}_n, \tilde{y}_n) \leq \tilde{d}(\tilde{x}_n, \tilde{x}) + \tilde{d}(\tilde{x}, \tilde{y}) + \tilde{d}(\tilde{y}, \tilde{y}_n)$$

Thus,

$$|\tilde{d}(\tilde{x}_n, \tilde{y}_n) - \tilde{d}(\tilde{x}, \tilde{y})| \leq \tilde{d}(\tilde{x}, \tilde{x}_n) + \tilde{d}(\tilde{y}, \tilde{y}_n)$$

Applying limit $n \rightarrow \infty$ on either sides, we get

$$\begin{aligned} \lim_{n \rightarrow \infty} |\tilde{d}(\tilde{x}_n, \tilde{y}_n) - \tilde{d}(\tilde{x}, \tilde{y})| &\leq \lim_{n \rightarrow \infty} \tilde{d}(\tilde{x}, \tilde{x}_n) + \lim_{n \rightarrow \infty} \tilde{d}(\tilde{y}, \tilde{y}_n) = 0 \\ \implies \tilde{d}(\tilde{x}, \tilde{y}) &= \lim_{n \rightarrow \infty} \tilde{d}(\tilde{x}_n, \tilde{y}_n) \end{aligned}$$

We have, \tilde{W} is isometric⁶ with W . Also \tilde{W}, W are dense in \tilde{X}, \hat{X} respectively. Therefore, \hat{X} and \tilde{X} are isometric. In other words, the completion of X is unique except for isometries. \square

14.2 Banach Spaces

Definitions 14.2.1 (vector space). A set of vectors X , a field of scalars K together with vector addition $+: X \times X \rightarrow X$ and scalar multiplication $\times: K \times X \rightarrow X$ is a vector(linear) space if it satisfies

1. $x + y = y + x$
2. $x + (y + z) = (x + y) + z$
3. $\exists 0 \in X$ such that $x + 0 = x$
4. $\forall x \in X, \exists -x \in X$ such that $x + (-x) = 0$
5. $\alpha(\beta x) = (\alpha\beta)x$
6. $1x = x$
7. $\alpha(x + y) = \alpha x + \alpha y$
8. $(\alpha + \beta)x = \alpha x + \beta x$

Note : $0x = 0, \alpha 0 = 0$ and $(-1)x = -x$.

For example, $\mathbb{R}^n, \mathbb{C}^n, C[a, b], l^p, l^\infty$ are vector spaces.

⁶Suppose \tilde{W} is isometric with X . Then there exists an isometry $S: \tilde{W} \rightarrow X$ such that $\tilde{d}(\tilde{x}_n, \tilde{y}_n) = d(S(\tilde{x}_n), S(\tilde{y}_n))$. Since $T: X \rightarrow W$ is also an isometry, $T^{-1} \circ S: \tilde{W} \rightarrow W$ is an isometry where $\tilde{d}(\tilde{x}_n, \tilde{y}_n) = \hat{d}(T^{-1} \circ S(\tilde{x}_n), T^{-1} \circ S(\tilde{y}_n))$.

14.2.1 A few definitions

Unlike other texts^{†7}. And the **subspace** Y **generated a subset** M of a vector space X is the set of all linear combinations of vectors in M . That is,

$$Y = \text{span } M$$

A set M is **linear independent** if there exists a non-zero r -tuple $(\alpha_1, \alpha_2, \dots, \alpha_r)$ such that $\alpha_1 x_1 + \alpha_2 x_2 + \dots + \alpha_r x_r = 0$ where $x_1, x_2, \dots, x_r \in M$.

A vector space X is **finite dimensional** if there exists an integer n which is the maximum cardinality for any linearly independent subset of X , say **dimension** of X , $\dim X$.

Let X be a finite dimensional vector space of dimension n . Then **basis** of X is a linearly independent subset of cardinality n .

Note : Let B be a basis of vector space X . Then $X = \text{span } B$. And every element $x \in X$ has a unique representation $x = \alpha_1 x_1 + \alpha_2 x_2 + \dots + \alpha_n x_n$ where $B = \{x_1, x_2, \dots, x_n\}$.

Let X be an infinite dimensional vector space. Let B be a linearly independent subset of X which span X , then B is a **Hamel basis** for X .

By Zorn's lemma, every vector space has a basis. And for any vector space, all the bases are of same cardinality (**dimension**).

Let X be a finite dimensional vector space of dimension n . Then any proper subspace Y of X has dimension less than n .

14.2.2 Banach Space

Definitions 14.2.2 (norm). A real-valued function $\|\cdot\|$ on a vector space X , $\|\cdot\| : X \rightarrow \mathbb{R}$ is a norm if it satisfies

1. $\|x\| \geq 0, \quad \forall x \in X$
2. $\|x\| = 0 \iff x = 0$
3. $\|\alpha x\| = |\alpha| \|x\|, \quad \forall x \in X, \quad \forall \alpha \in K$
4. $\|x + y\| \leq \|x\| + \|y\|, \quad \forall x, y \in X$

Note : Let $\|\cdot\|$ be a norm on a vector space X . Then $d(x, y) = \|x - y\|$ is a metric induced by the norm.

Proof. Let $d(x, y) = \|x - y\|$.

1. $d(x, y) = \|x - y\| \geq 0$ since $x - y \in X, \quad \|x - y\| \geq 0$
2. $d(x, y) = \|x - y\| = 0 \iff x - y = 0 \iff x = y$
3. $d(x, y) = \|x - y\| = \|-1(y - x)\| = |-1| \|y - x\| = d(y, x)$

^{†7}Usually, we define **space generated by a set** as the intersection of all spaces containing that set.

$$4. d(x, y) = \|x - z + z - y\| \leq \|x - z\| + \|z - y\| = d(x, z) + d(z, y)$$

□

normed space is a vector space with a norm defined on it.

Banach space is a complete, normed space.

Note : Norm is continuous.

Proof. Let X be a normed space. We have,

$$\lim_{h \rightarrow 0} \|x + hy\| \leq \|x\| + \lim_{h \rightarrow 0} |h| \|y\| = \|x\|$$

$$\|x\| \leq \lim_{h \rightarrow 0} \|x + hy\| + \lim_{h \rightarrow 0} |-h| \|y\| = \lim_{h \rightarrow 0} \|x + hy\|$$

Therefore, $\lim_{h \rightarrow 0} \|x + hy\| = \|x\|$. And the norm is continuous. □

Challenge 8. Characterise metric spaces, which are normed space ?

By defining $\|x\| = d(x, 0)$, we may construct a norm on various metric spaces so that the induced metric is the same. We observe the following conditions are necessary and sufficient for this function to be a norm.

Proof. Let (X, d) be a metric space. Let $d(x+z, y+z) = d(x, y)$ and $d(\alpha x, \alpha y) = |\alpha|d(x, y)$ for any $x, y, z \in X$ and for any $\alpha \in K$. Then,

1. $\|x\| = d(x, 0) \geq 0$
2. $\|x\| = d(x, 0) = 0 \iff x = 0$
3. $\|\alpha x\| = d(\alpha x, 0) = |\alpha|d(x, 0)$
4. $\|x + y\| \leq d(x + y, 0) \leq d(x + y, y) + d(y, 0) = d(x, 0) + d(y, 0) = \|x\| + \|y\|$

□

Lemma 14.2.1 (translational invariance). *Let d be a metric induced by a norm on a normed space X . Then*

1. $d(x + z, y + z) = d(x, y), \quad \forall x, y, z \in X$
2. $d(\alpha x, \alpha y) = |\alpha|d(x, y), \quad \forall x, y \in X, \forall \alpha \in K$

Proof. Let $(X, \|\cdot\|)$ be a normed space. Define $d(x, y) = \|x - y\|$. Then,

1. $d(x + z, y + z) = \|x + z - (y + z)\| = \|x - y\| = d(x, y)$
2. $d(\alpha x, \alpha y) = \|\alpha x - \alpha y\| = \|\alpha(x - y)\| = |\alpha|\|x - y\| = |\alpha|d(x, y)$

□

Normed spaces

The following normed spaces can be easily constructed using the above construction.

1. $\mathbb{R}, \|x\| = |x|$
2. $\mathbb{R}^2, \|x\| = (|\xi_1|^2 + |\xi_2|^2)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2)$
3. $\mathbb{R}^3, \|x\| = (|\xi_1|^2 + |\xi_2|^2 + |\xi_3|^2)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2, \xi_3)$
4. Euclidean space, $\mathbb{R}^n, \|x\| = \left(\sum_{j=1}^n |\xi_j|^2 \right)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2, \dots, \xi_n)$
5. $\mathbb{C}, \|x\| = |x|$ where $x = a + ib$ and $|x| = (a^2 + b^2)^{\frac{1}{2}}$
6. Unitary space, $\mathbb{C}^n, \|x\| = \left(\sum_{j=1}^n |\xi_j|^2 \right)^{\frac{1}{2}}$ where $x = (\xi_1, \xi_2)$ and $\xi_j \in \mathbb{C}$.
7. $C[a, b], \|x\| = \max_{t \in [a, b]} \{|x(t)|\}$ where x is a continuous, complex-valued function defined on closed interval $[a, b] \subset \mathbb{R}$.
8. $B(A), \|x\| = \sup_{t \in A} \{|x(t)|\}$ where x is a bounded, complex-valued function defined on $A \subset \mathbb{R}$
9. $l^p, \|x\| = \left(\sum_{j=1}^{\infty} |\xi_j|^p \right)^{\frac{1}{p}}$ where $x = \{\xi_j\}$ and $\sum_{j=1}^{\infty} |\xi_j|^p < \infty$. That is, set of all sequences such that the p th power series is convergent.
10. $l^{\infty}, \|x\| = \sup_{j \in \mathbb{N}} \{|\xi_j|\}$ where $x = \{\xi_j\}$ and $|\xi_j| \leq c$. That is, l^{∞} is the space of all bounded sequences of complex numbers.
11. $c, \|x\| = \sup_{j \in \mathbb{N}} \{|\xi_j|\}$ where $\{\xi_j\} = \xi_1, \xi_2, \dots$ is the space of all convergent sequences of complex numbers.

Test for norm which can be induced from a metric

There exists metric which cannot be induced by any norm. Suppose (X, d) is a metric space which doesn't feature translational invariance. Suppose there exists a norm on X which induces d . Then d must have translational invariance which is a contradiction.

For example, metric space of all sequence of complex numbers with metric

$$d(x, y) = \sum_{j=1}^{\infty} \frac{1}{2^j} \frac{|\xi_j - \eta_j|}{1 + |\xi_j - \eta_j|}$$

where $x = \{\xi_j\} = \xi_1, \xi_2, \dots$ is not translationally invariant since,

$$d(\alpha x, \alpha y) = \sum_{j=1}^{\infty} \frac{1}{2^j} \frac{|\alpha \xi_j - \alpha \eta_j|}{1 + |\alpha \xi_j - \alpha \eta_j|} = \sum_{j=1}^{\infty} \frac{1}{2^j} \frac{|\xi_j - \eta_j|}{\left|\frac{1}{\alpha}\right| + |\xi_j - \eta_j|} \neq |\alpha| d(x, y)$$

Therefore, above metric cannot be induced from any norm defined on the sequence space s .

Completion of incomplete normed space

Consider the set of all continuous functions defined on closed interval $[a, b]$ together with the norm $\|\cdot\| : C[a, b] \rightarrow \mathbb{R}$ defined by

$$\|x\| = \left(\int_a^b |x(t)|^2 dt \right)^{\frac{1}{2}}$$

is an incomplete normed space. The completion of this space is $L^2[a, b]$, the collection of all Lebesgue measurable function x defined on closed interval $[a, b]$ such that $|x|^2$ is Lebesgue integrable.

Consider the collection of continuous functions defined on $[a, b]$ with norm,

$$\|x\| = \left(\int_a^b |x(t)|^p dt \right)^{\frac{1}{p}}$$

We have its completion $L^p[a, b]$, the collection of all Lebesgue measurable function x defined on closed $[a, b]$ such that $|x|^p$ are Lebesgue integrable.

14.2.3 Further properties of normed spaces

1. Let X be a Banach space. And Y is a subset of X . The subspace Y is Banach if and only if Y is closed in X .
2. Every normed space X has a Banach space \hat{X} in which X is dense.

Theorem 14.2.2. *Let X be a Banach space. And Y is a subset of X . The subspace Y is complete if and only if Y is closed in X .*

Proof. Let $(X, \|\cdot\|)$ be a Banach space. Then (X, d) is a metric space with induced metric $d : X \times X \rightarrow \mathbb{R}$ defined by $d(x, y) = \|x - y\|$. Let $Y \subset X$. Clearly, Y is a normed space with induced norm. And if Y is closed then every Cauchy sequence in Y does converge to a point in Y . Therefore, Y is a Banach space.

Suppose Y is a Banach space with induced norm. Since Y is complete, every convergent sequence in Y converges to a point in Y itself. Therefore, Y is a closed subset of X . \square

Convergence in normed spaces

We have the notion of convergence defined for metric spaces. Now, we may extend that notion to normed spaces.

Definitions 14.2.3 (convergence). A sequence $\{x_n\}$ in a normed X converges to $x \in X$ if

$$\lim_{n \rightarrow \infty} \|x_n - x\| = 0$$

A series in a normed space $(X, \|\cdot\|)$ is convergent if the sequence of partial sums converges. If it converges, the limit of the sequence of partial sums is the **sum** of the series. A series is **absolutely convergent** if the series of norms converges.

Schauder basis

Definitions 14.2.4 (Schauder basis). A sequence $\{e_n\}$ of normed space $(X, \|\cdot\|)$ is a Schauder basis of X if for any $x \in X$, there exists a unique sequence of scalars $\{\alpha_n\}$ such that the series $\sum_{n=1}^{\infty} \alpha_n e_n$ converges to x .

$$\lim_{n \rightarrow \infty} \left\| x - \sum_{k=1}^n \alpha_k e_k \right\| = 0$$

Note : Every normed space with a Schauder basis is separable. But there exists separable, Banach spaces without a Schauder basis.

Completion of normed spaces

Theorem 14.2.3 (completion). Let $(X, \|\cdot\|)$ be a normed space. Then there exists a Banach space \hat{X} with a dense subset W and an isometry A from W onto X . And \hat{X} is unique except for isometries.

Proof. Let $(X, \|\cdot\|)$ be a normed space. Then (X, d) where $d : X \times X \rightarrow \mathbb{R}$ defined by $d(x, y) = \|x - y\|$ is a metric space. By the completion of metric spaces, there exists a unique, complete metric space \hat{X} with a dense subset W and an isometry from W onto X .

Step 1 : \hat{X} is a vector space

Let $\hat{x}, \hat{y} \in \hat{X}$. Let sequence $\{x_n\}$ be some Cauchy sequence in \hat{x} and sequence $\{y_n\}$ be some Cauchy sequence in \hat{y} .

1. The vector set, \hat{X} is the equivalent classes of Cauchy sequence in X .
2. The scalar field is real/complex field K of numbers.
3. Vector addition $+$: $\hat{X} \times \hat{X} \rightarrow \hat{X}$ is defined by $\hat{x} + \hat{y} = \hat{z}$. Let $\{x_n\} \in \hat{x}$, $\{y_n\} \in \hat{y}$. Then $\{x_n + y_n\} = \{z_n\} \in \hat{z}$.

Suppose $\hat{x}, \hat{y} \in \hat{X}$. We need prove that vector addition is well-defined as well as closed. In other words, $\hat{x} + \hat{y}$ is independent of the choice of

$\{x_n\} \in \hat{x}$, $\{y_n\} \in \hat{y}$ and is an equivalent class of Cauchy sequences.

Step 1.3a : Vector Sum is well-defined

Let $\{x_n\}, \{x'_n\} \in \hat{x}$ and $\{y_n\}, \{y'_n\} \in \hat{y}$. Then, $\{x_n\} \sim \{x'_n\}$ and $\{y_n\} \sim \{y'_n\}$. By triangular inequality,

$$\|x_n + y_n - (x'_n + y'_n)\| \leq \|x_n - x'_n\| + \|y_n - y'_n\|$$

Applying limit $n \rightarrow \infty$ on either sides we get,

$$\lim_{n \rightarrow \infty} \|x_n + y_n - (x'_n + y'_n)\| \leq \lim_{n \rightarrow \infty} \|x_n - x'_n\| + \lim_{n \rightarrow \infty} \|y_n - y'_n\| = 0$$

Clearly, $\{x_n + y_n\} \sim \{x'_n + y'_n\}$ and the vector sum $\hat{x} + \hat{y}$ is consistent, independent of the choice of the sequences.

Step 1.3b: Vector addition is closed

Let $\hat{x}, \hat{y} \in \hat{X}$. Then every sequence $\{x_n\} \in \hat{x}$ and every sequence $\{y_n\} \in \hat{y}$ are Cauchy sequences. Then,

$$\forall \varepsilon > 0, \exists N_1 \in \mathbb{N} \text{ such that } \forall n, m \geq N_1, \|x_n - x_m\| < \frac{\varepsilon}{2}$$

$$\forall \varepsilon > 0, \exists N_2 \in \mathbb{N} \text{ such that } \forall n, m \geq N_2, \|y_n - y_m\| < \frac{\varepsilon}{2}$$

We need to prove that $\hat{x} + \hat{y}$ is also an equivalent class of Cauchy sequences. Let $N = \max\{N_1, N_2\}$. Then $\forall \varepsilon > 0, \forall n, m \geq N$ we have,

$$\|x_n + y_n - (x_m + y_m)\| \leq \|x_n - x_m\| + \|y_n - y_m\| < \frac{\varepsilon}{2} + \frac{\varepsilon}{2} = \varepsilon$$

4. Scalar multiplication $\times : K \times \hat{X} \rightarrow \hat{X}$ defined by $\alpha \hat{x}$ is the collection of Cauchy sequences equivalent to $\{\alpha x_n\}$ where $\{x_n\} \in \hat{x}$ and $\alpha \in K$.

We need to prove that scalar product is well-defined and is closed. Let $\hat{x} \in \hat{X}$ and $\alpha \in K$.

Step 1.4a : Scalar product is well-defined

Let $\{x_n\}, \{x'_n\} \in \hat{x}$. Then $\{x_n\} \sim \{x'_n\}$. And both sequences are Cauchy sequences.

$$\lim_{n \rightarrow \infty} \|\alpha x_n - \alpha x'_n\| = |\alpha| \lim_{n \rightarrow \infty} \|x_n - x'_n\| = 0$$

Therefore, $\alpha \hat{x}$ is the equivalent class of sequences which are equivalent to sequence $\{\alpha x_n\}$ where sequence $\{x_n\} \in \hat{x}$. Clearly, $\alpha \hat{x}$ is independent of the choice of sequence $\{x_n\} \in \hat{x}$.

Step 1.4b: Scalar product is closed

We have, $\{x_n\}$ is a Cauchy sequence. Then,

$$\forall \varepsilon > 0, \exists N \in \mathbb{N} \text{ such that } \forall n, m \geq N, \|x_n - x_m\| < \frac{\varepsilon}{|\alpha|}$$

Clearly, $\forall n, m > N$

$$\|\alpha x_n - \alpha x_m\| = |\alpha| \|x_n - x_m\| < \varepsilon$$

Therefore, $\alpha \hat{x}$ is an equivalent class of Cauchy sequences.

Therefore, \hat{X} is vector space.

Step 2 : Construction of norm

We introduce a norm on \hat{X} such that the metric \hat{d} coincides with the induced metric.

Define $\|\cdot\| : W \rightarrow \mathbb{R}$ as $\|\hat{x}\| = \hat{d}(\hat{x}, \hat{0})$ for every $\hat{x} \in W$. We know that $A : W \rightarrow X$ is an isometry. Therefore, $\hat{d}(\hat{x}, \hat{y}) = d(x, y)$ for every $\hat{x}, \hat{y} \in W$ where $A(\hat{x}) = x$ and the constant sequence $\{x\} \in \hat{x}$. Clearly, for each $\hat{x} \in W$, $\|\hat{x}\| = \hat{d}(\hat{x}, \hat{0}) = d(x, 0) = \|x\|$ where $x \in X$.

Let $\hat{x} \in \hat{X}$. We know that, W is dense in \hat{X} , there exists a Cauchy sequence $\{\hat{x}_n\}$ in W such that $\hat{d}(\hat{x}_n, \hat{x}) < \frac{1}{n}$. Thus, there exists a sequence $\{\hat{x}_n\}$ in W such that $\hat{x}_n \rightarrow \hat{x}$. Now we may extend the norm function to \hat{X}

$$\|\cdot\| : \hat{X} \rightarrow \mathbb{R} \text{ defined by } \|\hat{x}\| = \lim_{n \rightarrow \infty} \|\hat{x}_n\| \quad (14.8)$$

where $\{\hat{x}_n\} \in W$ and $\lim_{n \rightarrow \infty} \hat{d}(\hat{x}_n, \hat{x}) = 0$.

Now, we show that $\|\cdot\|$ defined on \hat{X} is a norm

1. We have, \hat{d} is metric on \hat{X} . Thus, $\hat{d}(\hat{x}, \hat{y}) \geq 0$, $\forall \hat{x}, \hat{y} \in \hat{X}$.

$$\|\hat{x}\| = \lim_{n \rightarrow \infty} \|\hat{x}_n\| = \lim_{n \rightarrow \infty} \hat{d}(\hat{x}_n, \hat{0}) \geq 0$$

2. Suppose $\hat{x} = \hat{0}$. Then $\hat{x} \in W$ and $\|\hat{x}\| = \hat{d}(\hat{x}, \hat{0}) = 0$.
Suppose $\|\hat{x}\| = 0$. Then,

$$\|\hat{x}\| = \lim_{n \rightarrow \infty} \|\hat{x}_n\| = \lim_{n \rightarrow \infty} \hat{d}(\hat{x}_n, \hat{0}) = \lim_{n \rightarrow \infty} d(x_n, 0) = 0$$

Clearly, sequence $\{x_n\}$ is a Cauchy sequence in X converging to $0 \in X$. Therefore, $\{x_n\} \in \hat{0}$ and $\hat{x}_n \rightarrow \hat{0}$. But, we know that $\hat{x}_n \rightarrow \hat{x}$. Since \hat{X} is a metric space, $\hat{x} = \hat{0}$.

3. We have,

$$\|\alpha \hat{x}\| = \lim_{n \rightarrow \infty} \|\alpha \hat{x}_n\| = |\alpha| \lim_{n \rightarrow \infty} \|\hat{x}_n\| = |\alpha| \|\hat{x}\|$$

4. Suppose $\hat{y} \in \hat{X}$. Since W is dense in \hat{X} , there exists a Cauchy sequence $\{\hat{y}_n\}$ in W such that $\hat{y}_n \rightarrow \hat{y}$. Thus,

$$\|\hat{x} + \hat{y}\| = \lim_{n \rightarrow \infty} \|\hat{x}_n + \hat{y}_n\| \leq \lim_{n \rightarrow \infty} \|\hat{x}_n\| + \lim_{n \rightarrow \infty} \|\hat{y}_n\| = \|\hat{x}\| + \|\hat{y}\|$$

Thus, $(\hat{X}, \|\cdot\|)$ is complete, normed space in which W is dense.

Step 3 : \hat{X} is unique

Suppose \tilde{X} is another Banach space with dense subset \tilde{W} which has a isometry \tilde{A} from \tilde{W} onto X . Then, the norm $\|\cdot\|$ defined on \tilde{W} ,

$$\|\tilde{x}\| = \tilde{d}(\tilde{x}, \tilde{0}) = d(x, 0), \quad \forall \tilde{x} \in \tilde{W}$$

And we may extend this norm to \tilde{X} ,

$$\|\tilde{x}\| = \lim_{n \rightarrow \infty} \|\tilde{x}_n\|$$

where sequence $\{\tilde{x}_n\}$ is a Cauchy sequence in \tilde{W} and $\tilde{x}_n \rightarrow \tilde{x}$.

Clearly, \tilde{W} is isometric with W and the topological closure of \tilde{W} is isometric with the topological closure of W . That is, \tilde{X} is isometric with \hat{X} . \square

Exercise

Definitions 14.2.5 (seminorm). A real-valued function p on a vector space X , $p : X \rightarrow \mathbb{R}$ is a seminorm if it satisfies

1. $p(x) \geq 0, \quad \forall x \in X$
2. $x = 0 \implies p(x) = 0$
3. $p(\alpha x) = |\alpha|p(x), \quad \forall x \in X, \quad \forall \alpha \in K$
4. $p(x + y) \leq p(x) + p(y), \quad \forall x, y \in X$

Lemma 14.2.4. Let $(X, \|\cdot\|)$ be a normed space. Then, the vector addition $+(x, y) = x + y$ and scalar multiplication $\times(\alpha, x) = \alpha x$ are continuous with respect to norm.

Proof. Let $x, y \in X$ and $\alpha \in K$. Let sequences $\{x_n\}, \{y_n\}$ converges to x, y respectively. Let sequence $\{\alpha_n\}$ converges to $\alpha \in K$. Then,

$$\lim_{n \rightarrow \infty} \|x_n + y_n - (x + y)\| \leq \lim_{n \rightarrow \infty} \|x_n - x\| + \lim_{n \rightarrow \infty} \|y_n - y\| = 0$$

And,

$$\begin{aligned} \lim_{n \rightarrow \infty} \|\alpha_n x_n - \alpha x\| &= \lim_{n \rightarrow \infty} \|\alpha_n x_n - \alpha x_n + \alpha x_n - \alpha x\| \\ &\leq \lim_{n \rightarrow \infty} \|(\alpha_n - \alpha)x_n\| + \lim_{n \rightarrow \infty} \|\alpha(x_n - x)\| \\ &= \lim_{n \rightarrow \infty} |\alpha_n - \alpha| \lim_{n \rightarrow \infty} \|x_n\| + |\alpha| \lim_{n \rightarrow \infty} \|x_n - x\| \\ &= 0 \end{aligned}$$

\square

Lemma 14.2.5. Let $(X, \|\cdot\|)$ be a normed space. Let sequence $\{x_n\}$ converges to $x \in X$ and sequences $\{y_n\}$ converges to $y \in X$. Then, sequence $\{x_n + y_n\}$ converges to $x + y$. Let sequence $\{\alpha_n\}$ converges to α . Then sequence $\{\alpha_n x_n\}$ converges to αx .

Proof. Suppose $x_n \rightarrow x$, $y_n \rightarrow y$ and $\alpha_n \rightarrow \alpha$. We have, vector addition and scalar multiplication are continuous with respect to norm.

$$\lim_{n \rightarrow \infty} \|x_n + y_n - (x + y)\| = 0 \text{ and } \lim_{n \rightarrow \infty} \|\alpha_n x_n - \alpha x\| = 0$$

Therefore, $x_n + y_n \rightarrow x + y$ and $\alpha_n x_n \rightarrow \alpha x$. \square

Lemma 14.2.6 (separable). *Let $(X, \|\cdot\|)$ be a normed space. Suppose X has a Schauder basis. Then, X is separable.*

Proof. refer : Exercise 2.3.10

hint : Let $B = \{e_1, e_2, \dots\}$ be a Schauder basis of normed space $(X, \|\cdot\|)$. Then B is a dense subset of X . \square

Lemma 14.2.7 (quotient space). *Let $(X, \|\cdot\|)$ be a normed space. Let Y be a closed subset of X . Then $(X/Y, \|\cdot\|_0)$ is a normed space with norm defined by*

$$\|\hat{x}\|_0 = \inf_{x \in \hat{x}} \|x\|, \quad \text{for any coset } \hat{x} \in X/Y$$

Proof. refer : Exercise 2.3.14

hint : You can skip algebraic details. N1, N2 are trivial. N3 - use $\alpha\hat{x} = \alpha x + Y$ where $\hat{x} = x + Y$. \square

Lemma 14.2.8 (product of normed spaces). *Let $(X_1, \|\cdot\|_1)$, $(X_2, \|\cdot\|_2)$ be normed spaces. Let $X = X_1 \times X_2$. Then $(X, \|\cdot\|)$ is a normed space with norm defined by,*

$$\|x\| = \max\{\|x_1\|_1, \|x_2\|_2\} \text{ where } x = (x_1, x_2)$$

Proof. refer : Exercise 2.3.15

hint : Let $x = (x_1, x_2)$. WLOG suppose $\|x_1\|_1 \geq \|x_2\|_2$. Then $\|x\| = \|x_1\|_1$ and the rest is obvious. \square

14.2.4 Finite dimensional normed spaces and subspaces

Lemma 14.2.9. *Let $(X, \|\cdot\|)$ be a normed space. Let $\{x_1, x_2, \dots, x_n\}$ be any linearly independent subset of X . Then there exists a real-number $c > 0$ such that for any subset $\{\alpha_1, \alpha_2, \dots, \alpha_n\}$ from field K*

$$\left\| \sum_{j=1}^n \alpha_j x_j \right\| \geq c \sum_{j=1}^n |\alpha_j| \quad (14.9)$$

Proof. Let $\{x_1, x_2, \dots, x_n\}$ be a linearly independent subset of a normed space X .

$$\left\| \sum_{j=1}^n \alpha_j x_j \right\| \geq c \sum_{j=1}^n |\alpha_j|$$

Let $s = \sum_{j=1}^n |\alpha_j|$. Suppose $s \neq 0$. Consider $\beta_j = \alpha_j/s$. Then,

$$\left\| \sum_{j=1}^n \beta_j x_j \right\| \geq c \text{ where } \sum_{j=1}^n |\beta_j| = 1$$

It is sufficient to prove that for any real-number $c > 0$, there does not exist an n -tuple of scalars violating the above inequality.

On the contrary, suppose that for every positive real number $c \in \mathbb{R}$ there exists a subset $\{\beta_1, \beta_2, \dots, \beta_n\}$ such that $\|\beta_1 x_1 + \beta_2 x_2 + \dots + \beta_n x_n\| < c$. Then, for each $m \in \mathbb{N}$, there exist a subset of n scalars $\{\beta_{m,1}, \beta_{m,2}, \dots, \beta_{m,n}\}$ such that

$$\|\beta_{m,1}x_1 + \beta_{m,2}x_2 + \dots + \beta_{m,n}x_n\| < \frac{1}{m}$$

Define, $y_m = \beta_{m,1}x_1 + \beta_{m,2}x_2 + \dots + \beta_{m,n}x_n$. Then, $\|y_m\| \rightarrow 0$ as $m \rightarrow \infty$.

Now, every $\beta_{j,k}$ is bounded since for each m , $|\beta_{m,1}| + |\beta_{m,2}| + \dots + |\beta_{m,n}| = 1$. The sequence $\{\beta_{m,1}\}$ of first terms of the sequence $\{y_m\}$ is a bounded sequence. By Bolzano-Weierstrass theorem, sequence $\{\beta_{m,1}\}$ has a convergent subsequence, say $\beta_{m,1} \rightarrow 1$ as $m \rightarrow \infty$. Let sequence $\{y_{1,m}\}$ be subsequence of sequence $\{y_m\}$ such that the sequence of first terms is convergent.

Again, sequence $\{\beta_{m,2}\}$ of second terms of the sequence $\{y_{1,m}\}$ is a bounded sequence. And $\{y_{1,m}\}$ has a subsequence $\{y_{2,m}\}$ such that both the sequence of first terms $\{\beta_{m,1}\}$ and $\{\beta_{m,2}\}$ are convergent, say $\beta_{2,m} \rightarrow \beta_2$ as $m \rightarrow \infty$.

Continuing like this n times, we get sequence $\{y_{n,m}\}$ such that sequence corresponding each term is convergent. That is, $\beta_{m,j} \rightarrow \beta_j$ for $j = 1, 2, \dots, n$.

Define $y = \beta_1 x_1 + \beta_2 x_2 + \dots + \beta_n x_n$. Clearly, the sequence $\{y_{n,m}\}$ is a convergent subsequence of $\{y_m\}$. And $y_{n,m} \rightarrow y$ as $m \rightarrow \infty$. Suppose $\{y_{n,m}\}$ is convergent, then $y_{n,m} \rightarrow y$ as the limit is unique. By the continuity of norm, we have $\|y_m\| \rightarrow \|y\|$.

Since $\|y\| = 0$, we have $y = \beta_1 x_1 + \beta_2 x_2 + \dots + \beta_n x_n = 0$. We have $\{x_1, x_2, \dots, x_n\}$ is linearly independent. Then, $\beta_j = 0$ for each j . And, for each m , sequence $\{|\beta_{m,n}|\}$ converges to 0 and $|\beta_{m,1}| + |\beta_{m,2}| + \dots + |\beta_{m,n}| = 1$. This is a contradiction. Therefore, there exists such a positive real-number c satisfying the lemma. \square

Theorem 14.2.10. *Every finite dimensional subspace of a normed space is complete.*

Proof. Let Y be a finite dimensional subspace of a normed space X . Let $\dim Y = n$ and $B = \{y_1, y_2, \dots, y_n\}$ be a basis for Y . Let sequence $\{y_m\}$ be a Cauchy sequence in Y . Then,

$$y_m = \alpha_{m,1}y_1 + \alpha_{m,2}y_2 + \dots + \alpha_{m,n}y_n$$

Since $\{y_m\}$ is a Cauchy sequence, we have $N \in \mathbb{N}$ such that $\forall s, t > N$

$$\|y_s - y_t\| = \left\| \sum_{j=1}^n \alpha_{s,j}y_j - \sum_{j=1}^n \alpha_{t,j}y_j \right\| < \varepsilon$$

Since basis of Y is a linearly independent subset of X , applying lemma we get

$$c \sum_{j=1}^n |\alpha_{s,j} - \alpha_{t,j}| \leq \left\| \sum_{j=1}^n (\alpha_{s,j} - \alpha_{t,j}) y_j \right\| < \varepsilon$$

Therefore, for each $j \in \mathbb{N}$, sequence $\{\alpha_{m,j}\}$ are Cauchy sequences in K . Suppose $\alpha_{m,j} \rightarrow \alpha_j$. Define $y = \alpha_1 y_1 + \alpha_2 y_2 + \alpha_n y_n$. Then, $y_m \rightarrow y$ as $m \rightarrow \infty$. Therefore, Y is complete. \square

Theorem 14.2.11. *Every finite dimensional subspace Y of a normed space X is closed in X .*

Proof. Let Y be a finite dimensional subspace of a normed space X . Then Y is complete. Let \hat{X} be the completion of X . Now, Y is complete subspace of a Banach space \hat{X} . Then, Y is closed \hat{X} . And $Y \subset X \subset \hat{X}$. Therefore, Y is closed in X . \square

Equivalent Norm

Consider normed space $(X, \|\cdot\|)$. Then (X, d) is a metric space with induced metric $d(x, y) = \|x - y\|$. Since every metric space is a topological space, there exists a unique topology (X, \mathcal{T}) with respect to its norm. Let $\|\cdot\|_0, \|\cdot\|_1$ be two different norms defined a vector space X . Then, the norms are equivalent if their topologies coincide.

Alternatively, we have the following inequality to characterise the equivalent of norms,

Definitions 14.2.6. Let X be a vector space. Let $\|\cdot\|_0, \|\cdot\|_1$ be two normed defined on X . These norms are equivalent if there exists positive, real numbers a, b such that for any $x \in X$,

$$a\|x\|_0 \leq \|x\|_1 \leq b\|x\|_0 \quad (14.10)$$

For a topologist, it will interesting to prove that the topologies coincide if the above inequality holds.

Proof. Let G be an open set in the induced metric space $(X; d_0)$ where $d_0(x, y) = \|x - y\|_0$. Let $x \in G$. Then there exists an n -Ball $B_0(x, r)$ such that $x \in B_0(x, r)$ and $B_0(x, r) \subset G$, since G is open. Then $B_1(x, r/a)$ is an n -Ball in (X, d_1) contained in G . Therefore, G is open in the induced metric space (X, d_1) where $d_1(x, y) = \|x - y\|_1$. Similarly, if G is an open set in (X, d_1) , then G is an open set in (X, d_0) . \square

Remark. Two norms that induces same topological space are not necessarily equivalent.

Proof. Let (R, d_1) be a metric space. Define $d_2 = d_1/(d_1 + 1)$. Then $(R, d_2) \cong (R, d_1)$. However, d_1, d_2 are not equivalent since $d_1(x, y) > c d_2(x, y)$ for every $c \in \mathbb{R}$ when $d_1(x, y) = n$. Refer <https://math.stackexchange.com/questions/109733> \square

Theorem 14.2.12. *Every norm on any finite dimensional normed spaces are equivalent.*

Let $(X, \|\cdot\|)$ be a finite dimensional normed space. Then there exists a unique norm on X upto equivalence. That is, the topology induced by norm is independent of the norm for finite dimensional normed spaces.

Proof. Let X be a vector space with $\dim X = n$ and basis $B = \{x_1, x_2, \dots, x_n\}$. Let $\|\cdot\|_1, \|\cdot\|_2$ be two norms defined on X . Let $x \in X$. Then,

$$x = \sum_{j=1}^n \alpha_j x_j$$

We know that, basis is a linearly independent subset of X itself. By the lemma for finite dimensional, normed spaces, there exists a positive, real number c such that

$$\|x\|_1 = \left\| \sum_{j=1}^n \alpha_j x_j \right\|_1 \geq c \sum_{j=1}^n |\alpha_j|$$

By triangular inequality, we have

$$\|x\|_1 = \left\| \sum_{j=1}^n \alpha_j x_j \right\| \leq \sum_{j=1}^n \|\alpha_j x_j\| = \sum_{j=1}^n |\alpha_j| \|x_j\| \leq k \sum_{j=1}^n |\alpha_j|$$

where $k = \max\{\|x_j\|\}$.

Let $a = \frac{c}{k}$. Then $a\|x\|_1 \leq \|x\|_2$. Similarly, we can obtain a positive, real number b such that $b\|x\|_2 \leq \|x\|_1$. Therefore, the norms are equivalent. \square

14.2.5 Compactness and Finite dimension

Definitions 14.2.7 (compact). Let X be a metric space. The space X is **compact** if every sequences in X has a convergent subsequence. A subset M is a **compact subset** if every sequence in M has a convergent subsequence.

Lemma 14.2.13 (compact). *Compact subsets of metric spaces are closed and bounded.*

Proof. Let C be a compact subset of metric space X . Let \tilde{X} be the completion of X . Let $\{x_n\}$ be a Cauchy sequence in C . Since, C is compact $\{x_n\}$ has a convergent subsequence which converges to $x \in C$. Clearly, $\{x_n\}$ is a Cauchy sequence in \tilde{X} . Then, $\{x_n\} \rightarrow x \in \tilde{X}$. Thus, every Cauchy sequence in C converges. Therefore, C is closed, since every complete space is closed.

Let $y \in C$. Suppose C is unbounded. Then, there exists a sequence $\{y_n\}$ in C such that $B(y_n, y) > n$. Clearly, $\{y_n\}$ doesn't have any convergent subsequence which is a contradiction. Therefore, C is bounded. \square

Theorem 14.2.14. *In a finite dimensional metric space, a subset M is compact if and only if M is closed and bounded.*

Proof. Let X be a metric space with $\dim X = n$. Let $C \subset X$. Suppose C is compact. Then, C is closed and bounded.

Suppose C is closed and bounded. Let $\{y_n\}$ be a sequence in C . Let $\{x_1, x_2, \dots, x_n\}$ be a basis for X . Then $y_j = \sum_{k=1}^n \xi_{j,k} x_k$. By lemma of linear combinations, we have

$$\|y_j\| = \left\| \sum_{k=1}^n \xi_{j,k} x_k \right\| \geq c \sum_{k=1}^n |\xi_{j,k}|$$

Clearly, for each j , sequences $\{\xi_{j,k}\}$'s are bounded. By Bolzano-Weierstrass theorem, there exists a convergent subsequence converging to ξ_j . Define, $y = \sum_{k=1}^n \xi_k x_k$. Then $\{y_n\}$ has a subsequence which converges to y (obtained by successive construction of subsequences.) Since C is closed, $y \in C$. Let \tilde{X} be the completion of X . Then $\{y_n\}$ has a subsequence converging to $y \in \tilde{X}$. Thus, every sequence in C has a convergent subsequence. Therefore, C is compact. \square

Lemma 14.2.15 (Riesz). *Let Y, Z be subspaces of a normed space X . Suppose Y is closed and is a proper subset of Z . Then for any real number $\theta \in (0, 1)$, there exists $z \in Z$ such that*

$$\|z\| = 1 \text{ and } \|z - y\| \geq \theta, \forall y \in Y$$

Proof. Let X be a normed space. Let Y be a closed subspace of X . And Z is a subspace of X properly containing Y . Thus, $Z - Y$ is non-empty. Let $v \in Z - Y$. Define $a = \inf\{\|v - y\| : y \in Y\}$. ^{†8} Let $\theta \in (0, 1)$. Then, $a < \frac{a}{\theta}$. And there exists $y_0 \in Y$ such that $a < \|v - y_0\| < \frac{a}{\theta}$. Define $z = \frac{v - y_0}{\|v - y_0\|}$. Then $\|z\| = 1$. And $\|z - y\| = c\|v - y_0 - c^{-1}y\|$ where $c^{-1} = \|v - y_0\|$. Clearly, $y_0 + c^{-1}y \in Y$ since Y is subspace and $y, y_0 \in Y$. Thus, $\|z - y\| = c\|v - (y_0 + c^{-1}y)\| \geq ca \geq \theta$ since $c^{-1} = \|v - y_0\| \leq \frac{a}{\theta}$. \square

Theorem 14.2.16 (compact set characterisation of finiteness). *Let X be normed space in which closed unit ball is compact. Then X is a finite dimensional normed space.*

Proof. Let X be normed space. Let $M = \{x : \|x\| = 1\}$ be a compact subset of X . Suppose $\dim X = \infty$. Choose $x_1 \in X$ such that $\|x_1\| = 1$. For any $x \in X$, if $x \neq 0$, then there exists $x/\|x\| \in X$ has norm 1.

The subspace X_1 spanned by x_1 is a proper, closed subspace of X . Since X is infinite dimensional, from Riesz's lemma there exists $x_2 \in X$ such that $\|x_2\| = 1$ and $\|x_2 - x_1\| \geq \frac{1}{2}$. Again, the subspace X_2 spanned by $\{x_1, x_2\}$ is a proper, closed subspace of X . Continuing like this, we get a sequence $\{x_n\}$ in M such that $\|x_n - x_m\| \geq \frac{1}{2}$ for any $n \neq m$. Since $\|x_n - x_m\| \geq \frac{1}{2}$, the sequence doesn't have a convergent subsequence. This is a contradiction since M is compact. Therefore, X is finite dimensional. \square

Theorem 14.2.17. *Let X, Y be metric spaces. Let $T : X \rightarrow Y$ be a continuous function. Then, the image of compact subset of X under T is a compact subset of Y .*

⁸Simply, $a = \|v - Y\|$ where $v \in Z - Y$

Proof. Let $T : X \rightarrow Y$ be a continuous function where X, Y are metric spaces. Let M be a compact subset of X . Let sequence $\{y_n\}$ be a sequence in $T(M)$. Then $y_n = Tx_n$ for some $x_n \in M$. Thus, we have a sequence $\{x_n\}$ in M corresponding to each sequence $\{y_n\}$ in $T(M)$. Since M is compact, every sequence in M has a convergent subsequence, say $\{x_{n_k}\} \rightarrow x$. Then, the corresponding sequence $\{Tx_{n_k}\}$ is subsequence of sequence $\{y_n\}$.

Since T is continuous, $\{Tx_{n_k}\}$ converges to Tx in $T(M)$. Thus, every sequence in $T(M)$ has a convergent subsequence. That is, M is compact. \square

Corollary 14.2.17.1. *A continuous mapping T of a compact subset M of a metric space X into \mathbb{R} assumes a maximum and a minimum value at some points of M .*

Proof. Let X be a metric space and M be a compact subset of X . Let $T : X \rightarrow \mathbb{R}$ be a continuous function. Then, $T(M)$ is compact. Since $T(M)$ is a compact subset of \mathbb{R} , $T(M)$ is a closed and bounded subset of \mathbb{R} . Thus, $\inf T(M), \sup T(M) \in T(M)$. Therefore, there exists $x, y \in M$ such that $Tx = \inf T(M) = \min T(M)$ and $Ty = \sup T(M) = \max T(M)$. \square

Module 2 - Linear Operators

14.2.6 Linear Operators

Definitions 14.2.8 (linear operator). Let X, Y be vector spaces with the same field K . A function $T : X \rightarrow Y$ is a linear operator if it satisfies

1. $T(x + y) = Tx + Ty, \forall x, y \in X$
2. $T(\alpha x) = \alpha Tx, \forall x \in X, \forall \alpha \in K$

Usually, we address a linear, function from a vector space into another as linear transformation. And linear operators are functions from a vector space into itself. Kreyszig uses linear operator for both.

1. $T(\alpha x + \beta y) = T(\alpha x) + T(\beta y) = \alpha Tx + \beta Ty$
2. $T0 = T(0x) = 0Tx = 0$
3. $T : X \rightarrow T(X)$ is a homomorphism of vector(linear) spaces

Examples of linear operators

1. Identity operator, $I : X \rightarrow X, Ix = x, \forall x \in X$

$$I(\alpha x + \beta y) = \alpha Ix + \beta Iy = \alpha x + \beta y$$

2. Zero operator, $O : X \rightarrow X, Ox = 0, \forall x \in X$

$$0(\alpha x + \beta y) = \alpha 0(x) + \beta 0(y) = 0$$

3. Differentiation operator, $T : X \rightarrow X, Tx(t) = x'(t), \forall t \in [a, b]$ where X is the set of all polynomials over $[a, b]$.

$$T(\alpha x + \beta y) = \alpha T(x) + \beta T(y) = \alpha x' + \beta y'$$

4. Integration operator, $T : X \rightarrow X, Tx(t) = \int_a^b k(t, \tau)x(\tau)d\tau$ where $X = C[a, b]$.

$$T(\alpha x + \beta y) = \alpha \left(\int_a^b k(t, \tau)x(\tau)d\tau \right) + \beta \left(\int_a^b k(t, \tau)y(\tau)d\tau \right)$$

5. Multiplication by parameter, $T : X \rightarrow X, Tx(t) = tx(t)$ where $X = C[a, b]$.

$$T(\alpha x + \beta y) = \alpha T(x) + \beta T(y) = \alpha tx + \beta ty$$

6. $T : \mathbb{R}^3 \rightarrow \mathbb{R}^3, T(x) = x \times a$ where $a \in \mathbb{R}^3$.

$$T(\alpha x + \beta y) = \alpha T(x) + \beta T(y) = \alpha(x \times a) + \beta(y \times a)$$

7. $T : \mathbb{R}^3 \rightarrow \mathbb{R}, T(x) = x \cdot a$ where $a \in \mathbb{R}^3$.

$$T(\alpha x + \beta y) = \alpha T(x) + \beta T(y) = \alpha(x \cdot a) + \beta(y \cdot a)$$

8. $T : \mathbb{R}^n \rightarrow \mathbb{R}^r$, $T(x) = Ax$ where $A \in \mathbb{R}^{n \times r}$ is an $n \times r$ matrix of real-numbers.

$$T(\alpha x + \beta y) = \alpha T(x) + \beta T(y) = \alpha Ax + \beta Ay$$

Theorem 14.2.18 (Range, Null space). *Let T be a linear operator. Then*

1. *Range, $\mathcal{R}(T)$ is a vector space.*
2. *If $\dim X = n < \infty$, then $\dim \mathcal{R}(T) \leq n$.*
3. *Null space, $\mathcal{N}(T)$ is a vector space.*

Proof. Let $T : X \rightarrow Y$ be a linear operator where X, Y are vector spaces over the same field K . Let $y_1, y_2 \in \mathcal{R}(T)$. Then $y_1 = Tx_1$ and $y_2 = Tx_2$ for some $x_1, x_2 \in X$. Let $\alpha, \beta \in K$. Since X is a vector space, $\alpha x_1 + \beta x_2 \in X$. We have, T is a linear operator. Thus,

$$\alpha y_1 + \beta y_2 = \alpha Tx_1 + \beta Tx_2 = T(\alpha x_1 + \beta x_2)$$

Therefore, $\alpha y_1 + \beta y_2 \in \mathcal{R}(T)$.

Let $\dim X = n$. Let $\{y_1, y_2, \dots, y_{n+1}\}$ be elements from $\mathcal{R}(T)$. Then for each j , we have $y_j = Tx_j$. Since $\dim X = n$, the set $\{x_1, x_2, \dots, x_{n+1}\}$ is linearly dependent. That is, there exist α_j 's in K such that

$$\alpha_1 x_1 + \alpha_2 x_2 + \dots + \alpha_{n+1} x_{n+1} = 0$$

Therefore,

$$T(\alpha_1 x_1 + \alpha_2 x_2 + \dots + \alpha_{n+1} x_{n+1}) = \alpha_1 y_1 + \alpha_2 y_2 + \dots + \alpha_{n+1} y_{n+1} = 0$$

Thus, any subset of $\mathcal{R}(T)$ having $n+1$ or more elements is linearly dependent. Clearly, the dimension of $\mathcal{R}(T) \leq n = \dim X$.

In other words, linear operators preserves linear dependence.

Let $x_1, x_2 \in \mathcal{N}(T)$. Then $Tx_1 = 0$ and $Tx_2 = 0$. Let $\alpha, \beta \in K$. Since T is linear,

$$T(\alpha x_1 + \beta x_2) = \alpha T(x_1) + \beta T(x_2) = 0$$

Therefore, $\alpha x_1 + \beta x_2 \in \mathcal{N}(T)$ for any $\alpha, \beta \in K$. □

Definitions 14.2.9 (Inverse). Let X, Y be vector spaces over the same field K . Let $T : X \rightarrow Y$ be an injective, linear operator. Then, $T^{-1} : \mathcal{R}(T) \rightarrow X$ defined by $y = Tx \iff T^{-1}y = x$ is the image operator of T .

Theorem 14.2.19 (inverse). *Let X, Y be vector spaces over the same field K . Let $T : X \rightarrow Y$ be a linear operator. Then,*

1. *T^{-1} exists if and only if $(Tx = 0 \implies x = 0)$.*
2. *If T^{-1} exists, then it is a linear operator.*
3. *If $\dim \mathcal{D}(T) = n < \infty$, and T^{-1} exists, then $\dim \mathcal{R}(T) = \dim \mathcal{D}(T)$.*

Proof. Suppose $Tx = 0 \implies x = 0$. Let $Tx_1 = Tx_2$. Then $Tx_1 - Tx_2 = T(x_1 - x_2) = 0$. That is, $x_1 - x_2 = 0$. Therefore, T is injective. Suppose T^{-1} exists, then T is injective and $Tx_1 = Tx_2 \implies x_1 = x_2$. Thus, $Tx = Tx_1 - Tx_2 = 0 \implies x = x_1 - x_2 = 0$.

Let $T : X \rightarrow Y$ be a linear operator where X, Y are vector spaces over the same field K . Suppose T^{-1} exists. Let $y_1 = Tx_1$ and $y_2 = Tx_2$. Then, $T^{-1}y_1 = x_1$ and $T^{-1}y_2 = x_2$. Let $\alpha, \beta \in K$. We have,

$$\alpha x_1 + \beta x_2 = \alpha T^{-1}(y_1) + \beta T^{-1}(y_2)$$

Also we have,

$$\begin{aligned} T(\alpha x_1 + \beta x_2) &= \alpha Tx_1 + \beta Tx_2 = \alpha y_1 + \beta y_2 \\ \implies \alpha x_1 + \beta x_2 &= T^{-1}(\alpha y_1 + \beta y_2) = \alpha T^{-1}y_1 + \beta T^{-1}y_2 \end{aligned}$$

Therefore, T^{-1} is a linear operator.

Since $T : X \rightarrow Y$ is a linear transformation. If $\dim X < \infty$, then $\mathcal{R}(T) \leq \dim \mathcal{D}(T)$. Again, $T^{-1} : \mathcal{R}(T) \rightarrow X$ is a linear transformation. Thus, $\dim \mathcal{D}(T) \leq \dim \mathcal{R}(T)$. Therefore, $\dim \mathcal{D}(T) = \dim \mathcal{R}(T)$. \square

Theorem 14.2.20 (Inverse of Product). *Let $T : X \rightarrow Y$ and $S : Y \rightarrow Z$ be bijective linear operators, where X, Y, Z are vector spaces. Then the inverse $(ST)^{-1} : Z \rightarrow X$ exists and $(ST)^{-1} = T^{-1}S^{-1}$.*

Proof. The operator $ST : X \rightarrow Z$ is bijective since S, T are bijective. Therefore, $(ST)^{-1} : Z \rightarrow X$ exists and $ST(ST)^{-1} = I_Z$.

$$\begin{aligned} S^{-1}ST(ST)^{-1} &= T(ST)^{-1} = S^{-1}I_Z = S^{-1} \text{ and} \\ T^{-1}S^{-1}ST(ST)^{-1} &= T^{-1}T(ST)^{-1} = (ST)^{-1} = T^{-1}S^{-1} \end{aligned}$$

\square

14.2.7 Bounded and Continuous Linear Operators

Definitions 14.2.10 (bounded operator). Let X, Y be normed spaces and $T : X \rightarrow Y$ be a linear operator. Then T is bounded if there exists a real number c such that $\|Tx\| \leq c\|x\|$.

Suppose $T : X \rightarrow Y$ be a bounded linear operator. Then $\|T\| : B(X, Y) \rightarrow \mathbb{R}$ is well-defined.

$$\|T\| = \sup_{\substack{x \in \mathcal{D}(T) \\ x \neq 0}} \frac{\|Tx\|}{\|x\|}$$

Lemma 14.2.21. *Let $T : X \rightarrow Y$ be a bounded linear operator. Then,*

1. *We have an alternate formula,*

$$\|T\| = \sup_{\substack{x \in \mathcal{D}(T) \\ \|x\|=1}} \|Tx\|$$

2. $\|T\|$ satisfies norm axioms

Proof. Suppose $x \neq 0$ and $\|x\| = a$. Then $a \neq 0$ since X is a normed space. Define $y = \frac{1}{a}x$. Then $\|y\| = 1$. Also we have,

$$\|T\| = \sup_{\substack{x \in \mathcal{D}(T) \\ x \neq 0}} \frac{1}{a} \|Tx\| = \sup_{\substack{x \in \mathcal{D}(T) \\ x \neq 0}} \left\| T \left(\frac{1}{a}x \right) \right\| = \sup_{\substack{y \in \mathcal{D}(T) \\ \|y\|=1}} \|Ty\|$$

1. We have

$$\|T\| = \sup_{\substack{x \in \mathcal{D}(T) \\ x \neq 0}} \frac{\|Tx\|}{\|x\|}$$

Since X, Y are normed spaces $\|Tx\|, \|x\| \geq 0$ and $\|T\| \geq 0$.

2. Also we have, Y is normed space and $Tx \in Y$. Thus, $\|Tx\| = 0$ if and only if $Tx = 0$. And, $\|T\| = 0$ if and only if $Tx = 0, \forall x \in \mathcal{D}(T)$. Therefore, $T = 0$. That is, $Tx = 0x = 0, \forall x \in \mathcal{D}(T)$.

3. Let X, Y be normed spaces over the field K . Let $\alpha \in K$ and T be a linear operator from X into Y . Then,

$$\begin{aligned} \|(\alpha T)\| &= \sup_{\substack{x \in \mathcal{D}(\alpha T) \\ \|x\|=1}} \|(\alpha T)(x)\| \\ &= \sup_{\substack{x \in \mathcal{D}(T) \\ \|x\|=1}} \|\alpha T(x)\| \\ &= |\alpha| \sup_{\substack{x \in \mathcal{D}(T) \\ \|x\|=1}} \|Tx\| \\ &= |\alpha| \|T\| \end{aligned}$$

4. Let X, Y be normed spaces and T_1, T_2 are linear operators from X to Y . Then,

$$\begin{aligned} \|T_1 + T_2\| &= \sup_{\substack{x \in \mathcal{D}(T_1+T_2) \\ \|x\|=1}} \|(T_1 + T_2)(x)\| \\ &= \sup_{\substack{x \in \mathcal{D}(T_1+T_2) \\ \|x\|=1}} \|T_1(x) + T_2(x)\| \\ &\leq \sup_{\substack{x \in \mathcal{D}(T_1) \\ \|x\|=1}} \|T_1(x)\| + \sup_{\substack{x \in \mathcal{D}(T_2) \\ \|x\|=1}} \|T_2(x)\| \\ &\leq \|T_1\| + \|T_2\| \end{aligned}$$

□

Norm of a few Linear operators

1. Identity Operator $I_X : X \rightarrow X$ defined by $I(x) = x$.

$$\|I\| = \sup_{\substack{x \in \mathcal{D}(I) \\ \|x\|=1}} \|Ix\| = 1$$

2. Zero Operator $0 : X \rightarrow Y$ defined by $0x = 0$.

$$\|0\| = \sup_{\substack{x \in \mathcal{D}(0) \\ \|x\|=1}} \|0x\| = 0$$

3. Differential Operator $D : K[t] \rightarrow K[t]$ defined by $D(x(t)) = x'(t)$ where $K[t]$ is the vector space of all polynomial functions defined on $[0, 1]$ together with norm $\|x\| = \sup_{t \in [0,1]} |x(t)|$.

The differential operator is not bounded. Consider $x(t) = t^n$ where $n \in \mathbb{N}$. Then $D(x) = x'(t) = nt^{n-1}$. Thus, $\|x\| = 1$ and $\|D(x)\| = n$. Clearly, differential operator D is not bounded. Therefore, $\|D\|$ is not defined.

4. Integral Operator $T : C[0, 1] \rightarrow C[0, 1]$ defined by

$$T(x) = \int_0^1 k(t, \tau) x(\tau) d\tau$$

where $k(t, \tau)$ is the kernel of the integral operator. Clearly, the integral of any continuous function is zero if only if $k(t, \tau) = 0$ for any $\tau \in [0, 1]$.

$$\begin{aligned} \|T(x)\| &= \left\| \int_0^1 k(t, \tau) x(\tau) d\tau \right\| \\ &\leq \int_0^1 \|k(t, \tau)\| \|x(\tau)\| d\tau \end{aligned}$$

Let $k_0 = \max_{\tau \in [0,1]} \{\|k(t, \tau)\|\}$. Replacing $\|k(t, \tau)\|$ with k_0 , we get

$$\|T(x)\| \leq k_0 \|x\| \int_0^1 d\tau = k_0 \|x\|$$

Therefore, integral operator is bounded and $\|T\| \leq k_0$.

5. Matrix multiplication $T : \mathbb{R}^n \rightarrow \mathbb{R}^r$ defined by $y = Ax$ where $A = (\alpha_{jk})$ where $\alpha_{jk} \in K$. Let $x = (\xi_k)$ and $y = (\eta_j)$ where $\xi_k, \eta_j \in K$. Then

$$\eta_j = \sum_{k=1}^n \alpha_{jk} \xi_k$$

Therefore,

$$\begin{aligned}\|Tx\|^2 &= \sum_{j=1}^r \eta_j^2 \\ &= \sum_{j=1}^r \left(\sum_{k=1}^n \alpha_{jk} \xi_k \right)^2\end{aligned}$$

By Cauchy-Schwarz inequality

$$\begin{aligned}\|T(x)\|^2 &\leq \sum_{j=1}^r \left(\sum_{k=1}^n \alpha_{jk}^2 \sum_{k=1}^n \xi_k^2 \right) \\ &\leq \sum_{j=1}^r \sum_{k=1}^n \alpha_{jk}^2 \sum_{j=1}^r \sum_{k=1}^n \xi_k^2 \\ &\leq c^2 \|x\|^2 \\ \|Tx\| &\leq c \|x\|\end{aligned}$$

where $c = \left(\sum_{j=1}^r \sum_{k=1}^n \alpha_{jk}^2 \right)^{\frac{1}{2}}$. Clearly, matrix multiplication is bounded.

Theorem 14.2.22 (finite dimension). *Every linear operator on finite dimensional normed space is bounded.*

Proof. Let X be normed space with $\dim(X) = n < \infty$. Let $T : X \rightarrow X$ be a linear operator on X . Then,

$$\begin{aligned}\|Tx\| &= \left\| T \left(\sum_{j=1}^n \xi_j e_j \right) \right\| \\ &= \left\| \sum_{j=1}^n \xi_j T(e_j) \right\| \text{ since } T \text{ is linear} \\ &\leq \sum_{j=1}^n |\xi_j| \|T(e_j)\|\end{aligned}$$

Let e_k is an element of the standard basis of X such that $\|T(e_k)\|$ is maximum. Then,

$$\|T(x)\| \leq \|T(e_k)\| \sum_{j=1}^n |\xi_j|$$

We have, the standard basis of X , $\{e_1, e_2, \dots, e_n\}$ is a linearly independent set of vectors. We have $x = \sum_{j=1}^n \xi_j e_j$. Thus, by linear combination lemma, there

exists a positive real number c such that $\sum_{j=1}^n |\xi_j| \leq \frac{1}{c} \left\| \sum_{j=1}^n \xi_j e_j \right\| = \frac{1}{c} \|x\|$. Thus,

$$\|T(x)\| \leq \frac{\|T(e_k)\|}{c} \|x\|$$

Therefore, linear operators on finite dimensional spaces are bounded. \square

Theorem 14.2.23. *Let X, Y be normed spaces and $T : X \rightarrow Y$ be a linear operator. Then*

1. *T is continuous if and only if T is bounded.*

2. *T is continuous at a point, then it is continuous.*

Proof. Suppose $T \neq 0$. If $T = 0$, then the results are trivial as zero operator is always continuous and bounded. Since $T \neq 0$, $\|T\| \neq 0$.

Suppose T is bounded. Let $\varepsilon > 0$. Define $\delta = \frac{\varepsilon}{\|T\|}$. Let $x_0 \in \mathcal{D}(T)$. Let $x \in \mathcal{D}(T)$ such that $\|x - x_0\| < \delta = \frac{\varepsilon}{\|T\|}$. Now, we have

$$\|Tx - Tx_0\| = \|T(x - x_0)\| \leq \|T\| \|x - x_0\| \leq \delta \|T\| \leq \varepsilon$$

Therefore, T is continuous.

Let T be continuous at $x_0 \in \mathcal{D}(T)$. Then, for any $\varepsilon > 0$, there exists $\delta > 0$ such that if $\|x - x_0\| \leq \delta$, then $\|T(x) - T(x_0)\| \leq \varepsilon$.

Let $y \neq 0$. Define $x \in X$ such that $x = x_0 + \frac{\delta}{\|y\|} y$. Then $\|x - x_0\| = \delta$. Therefore,

$$\varepsilon \geq \|Tx - Tx_0\| = \|T(x - x_0)\| = \frac{\delta}{\|y\|} \|Ty\|$$

And $\|Ty\| \leq \frac{\varepsilon}{\delta} \|y\|$. Therefore, T is bounded.

Suppose T is continuous at a point x_0 . Then T is bounded and therefore T is continuous on $\mathcal{D}(T)$. \square

Corollary 14.2.23.1. *Let T be a bounded linear operator.*

1. *$x_n \rightarrow x \implies Tx_n \rightarrow Tx$*

2. *Null space $\mathcal{N}(T)$ is closed.*

Proof. Let $T : X \rightarrow Y$ be a bounded linear operator. Suppose $x_n \rightarrow x$.

$$\lim_{n \rightarrow \infty} \|Tx_n - Tx\| = \lim_{n \rightarrow \infty} \|T(x_n - x)\| \leq \|T\| \|x_n - x\| = 0$$

Clearly, $Tx_n \rightarrow Tx$.

Suppose $x \in \overline{\mathcal{N}(T)}$.^{†9} Then there exists a sequence $\{x_n\} \in \mathcal{N}(T)$ such that $x_n \rightarrow x$. Then, $Tx_n \rightarrow Tx$. However, $Tx_n = 0$ since $x_n \in \mathcal{N}(T)$. Therefore, $Tx = 0$ and $x \in \mathcal{N}(T)$. Clearly, $\mathcal{N}(T)$ is closed. \square

Remark. Range of a bounded linear operator is not necessarily closed.

Proof. Consider $T : l^\infty \rightarrow l^\infty$ defined by $Tx = y$ where $x = (\xi_j)$, $y = (\eta_j)$ and $\eta_j = \xi_j/j$. Since l^∞ is not complete, there exists a sequence $\{x_n\}$ in l^∞ which converges to an unbounded sequence $x \notin l^\infty$.

For example, consider $x_n = (\xi_{n,j})$ where^{†10}

$$\xi_{n,j} = \begin{cases} j - \frac{1}{n} & j \leq n \\ 0 & j > n \end{cases}$$

Clearly, $x_n \in l^\infty$ since x_n is bounded by n . And $\{Tx_n\} = (\eta_{n,j})$ defined by $\eta_{n,j} = \frac{\xi_{n,j}}{j}$ is a sequence of bounded functions which converges to $y = (\eta_j)$ where $\eta_j = 1/j$. However, there doesn't exist a bounded sequence x such that $Tx = y$. Therefore, the range of T is not closed. \square

Remark. Let $T_2 : X \rightarrow Y$ and $T_1 : Y \rightarrow Z$ be bounded linear operators where X, Y, Z are normed spaces. Then $\|T_1 T_2\| = \|T_1\| \|T_2\|$.

Proof. Suppose T_1, T_2 are bounded, linear operators. Then $T_1 T_2$ is linear since $T_1(T_2(\alpha x + \beta y)) = T_1(\alpha T_2 x + \beta T_2 y) = \alpha T_1 T_2 x + \beta T_1 T_2 y$.

And $T_1 T_2$ is bounded since

$$\begin{aligned} \|T_1 T_2(x)\| &\leq \|T_1\| \|T_2(x)\| \text{ since } T_1 \text{ is bounded} \\ &\leq \|T_1\| \|T_2\| \|x\| \text{ since } T_2 \text{ is bounded} \\ &\leq c\|x\| \text{ where } c = \|T_1\| \|T_2\| \end{aligned}$$

Now we have,

$$\begin{aligned} \|T_1 T_2\| &= \sup_{\substack{x \in \mathcal{D}(T_2) \\ \|x\|=1}} \|T_1 T_2 x\| \\ &\leq \sup_{\substack{x \in \mathcal{D}(T_2) \\ \|x\|=1}} \|T_1\| \|T_2\| \|x\| \\ &\leq \|T_1\| \|T_2\| \end{aligned}$$

\square

⁹To prove X is closed, we often choose $x \in \overline{X}$ which is limit of an arbitrary sequence x_n in X . Then it is enough to show that the limit $x \in X$.

¹⁰We know that l^∞ is the set of all bounded sequences. We may use simpler sequence $x_n = (\xi_{n,j})$ where $\xi_{n,j} = j$, $j \leq n$ and vanishes for $j > n$. That is, $x_1 = 1, 0, 0, \dots$, $x_2 = 1, 2, 0, 0, \dots$ and $x_n = 1, 2, \dots, n, 0, 0, \dots$. Then $Tx_n = 1, 1, \dots, 1, 0, 0, \dots$. Clearly, $x_n \rightarrow y$ where $y = 1, 2, 3, \dots$ is unbounded and $Tx_n \rightarrow Ty$ where $Ty = 1, 1, 1, \dots$ is bounded.

Remark. Let $T : X \rightarrow X$ be a bounded, linear operator. Then $\|T^2\| \leq \|T\|^2$. And by mathematical induction, $\forall n \in \mathbb{N}$, $\|T^n\| \leq \|T\|^n$.

Definitions 14.2.11 (equal operator). Two linear operators T_1, T_2 are equal if $\mathcal{D}(T_1) = \mathcal{D}(T_2)$ and $\forall x \in \mathcal{D}(T_1)$, $T_1x = T_2x$.

Theorem 14.2.24 (bounded linear extension). *Let $T : X \rightarrow Y$ be a bounded, linear operator from a normed space into a Banach space. Then T has a bounded, linear extension $\tilde{T} : \overline{\mathcal{D}(T)} \rightarrow Y$ such that $\|\tilde{T}\| = \|T\|$.*

Proof. Suppose $x \in \overline{\mathcal{D}(T)}$. Then there exists a sequence $\{x_n\}$ in $\mathcal{D}(T)$ such that $x_n \rightarrow x$. By Cauchy-criterion, there exists $N \in \mathbb{N}$ such that $\forall n, m > N$, $\|x_n - x_m\| \leq \varepsilon$. Since T is a bounded linear operator, for every $n, m > N$ we have

$$\|Tx_n - Tx_m\| = \|T(x_n - x_m)\| \leq \|T\| \|x_n - x_m\| \leq \varepsilon \|T\|$$

Therefore, sequence $\{Tx_n\}$ in Y converges to a point, say $y \in Y$ since Y is complete. Define $\tilde{T} : \overline{\mathcal{D}(T)} \rightarrow Y$ such that $\tilde{T}(x) = y$.

Now, we show that the function \tilde{T} is defined independent of the choice of the sequence $\{x_n\}$. Let $x_n \rightarrow x$ and $z_n \rightarrow x$. Consider $\{v_n\} : x_1, z_1, x_2, z_2, \dots$. Then $v_n \rightarrow x$. And $\{Tv_n\}$ converges. Then, the subsequences $\{Tx_n\}$ and $\{Tz_n\}$ converges and they must have the same limit. Thus, \tilde{T} is uniquely defined (well defined).

Suppose $x_n \rightarrow x$ and $z_n \rightarrow z$. Then \tilde{T} is linear since

$$T(\alpha x_n + \beta z_n) = \alpha Tx_n + \beta Tz_n \rightarrow \alpha \tilde{T}x + \beta \tilde{T}z = \tilde{T}(\alpha x + \beta z)$$

Since \tilde{T} is an extension of T , $\tilde{T}x = Tx$ in the domain of T . Also we have, $\|Tx_n\| \leq \|T\| \|x_n\|$. Suppose $Tx_n \rightarrow y = \tilde{T}x$. We have, the norm function which maps $x \rightarrow \|x\|$ is continuous. As $n \rightarrow \infty$ we have, $\|\tilde{T}x\| \leq \|T\| \|x\|$. Thus, \tilde{T} is bounded and $\|\tilde{T}\| \leq \|T\|$. Since \tilde{T} is an extension of T , we also have $\|T\| \leq \|\tilde{T}\|$. Therefore, $\|\tilde{T}\| = \|T\|$. \square

Challenge 9. Any linear operator on X has a linear extension into \overline{X} preserving norm. Whether this extension is unique on \overline{X} ?

14.2.8 Linear Functionals

Definitions 14.2.12 (linear functional). Let X be a vector space over field K . Then linear functional on X is a linear operator $f : X \rightarrow K$.

Definitions 14.2.13 (bounded linear functional). A linear functional on X is bounded if there exists a real number c such that

$$|f(x)| \leq c\|x\|, \quad \forall x \in X$$

Theorem 14.2.25. *Linear functional on a normed space is continuous if and only if bounded.*

Proof. Let f be a linear functional on a normed space X . Then f can be viewed as a linear operator from normed space X into normed space K . We know that, linear operators are continuous if and only if bounded. Since f is a linear operator, f is continuous if and only if f is bounded. \square

Norm

Let X be a normed space. Then norm is a non-linear functional on X .

Let X be a normed space. Then $\|\cdot\| : X \rightarrow \mathbb{R}$ is a non-linear functional on X . The functional $\|\cdot\|$ is non-linear due to triangular inequality.

Dot Product

The dot product is a bounded linear functional on any finite dimensional Euclidean space.

Let $a \in \mathbb{R}^3$. Then $f : \mathbb{R}^3 \rightarrow \mathbb{R}$ defined by $f(x) = x \cdot a$ is a bounded, linear functional on \mathbb{R}^3 . We know that dot product is linear.

$$(\alpha x + y) \cdot a = \alpha x \cdot a + y \cdot a$$

By Cauchy-Schwarz's inequality, we have

$$\|x \cdot a\| = \left\| \sum \xi_j a_j \right\| \leq \left\| \left(\sum \xi_j^2 \right)^{\frac{1}{2}} \left(\sum a_j^2 \right)^{\frac{1}{2}} \right\| \leq \|x\| \|a\|$$

Thus f is bounded and for any $x \in X$, we have

$$|f(x)| = \|x \cdot a\| \leq \|x\| \|a\|$$

This inequality is true for any $x \in X$ and the supremum as well.

$$\|f\| = \sup_{\|x\|=1} |f(x)| \leq \sup_{\|x\|=1} \|x\| \|a\| = \|a\|$$

When $x = a$ we have,

$$\|f\| = \sup_{x \neq 0} \frac{|f(x)|}{\|x\|} \geq \frac{|f(a)|}{\|a\|} = \frac{|a \cdot a|}{\|a\|} \leq \frac{\|a\|^2}{\|a\|} = \|a\|$$

Thus norm of the dot product, $\|f\| = \|a\|$ where $f(x) = x \cdot a$.

Definite Integral

Let $X = C[a, b]$, the set of all continuous functions on closed interval $[a, b]$ with norm $\|x\| = \sup_{x \in [a, b]} |x(t)|$. Let $f(x) = \int_a^b x(t) dt$.

We know that, f is linear, since

$$\begin{aligned} f(\alpha x + \beta y) &= \int_a^b (\alpha x + \beta y)(t) dt \\ &= \int_a^b (\alpha x)(t) + (\beta y)(t) dt \\ &= \int_a^b \alpha x(t) dt + \int_a^b \beta y(t) dt \\ &= \alpha \int_a^b x(t) dt + \beta \int_a^b y(t) dt \\ &= \alpha f(x) + \beta f(y) \end{aligned}$$

And f is bounded.

$$|f(x)| = \left| \int_a^b x(t) dt \right| \leq \int_a^b |x(t)| dt \leq \|x\| \int_a^b dt \leq (b-a)\|x\|$$

since $\|x\| = \sup |x(t)|$, we have $|x(t)| \leq \|x\|$ for any $t \in [a, b]$.

Thus, $\|f\| \leq (b-a)$. And for constant function x_0 on $[a, b]$ where $x_0(t) = 1$, $\forall t \in [a, b]$, we know that $\|x_0\| = \sup |x_0(t)| = 1$. And $f(x_0) = \int_a^b dt = (b-a)$.

$$\|f\| = \sup_{x \neq 0} \frac{|f(x)|}{\|x\|} \leq \frac{|f(x_0)|}{\|x_0\|} = (b-a)$$

Therefore, $\|f\| = (b-a)$ where $f(x) = \int_a^b f(x) dt$.

A linear functional on l^2 space

Let $a \in l^2$. That is, $a = (\alpha_j)$ is a sequence such that $\sum |\alpha_j|^2 < \infty$. Consider $f(x) = \sum_{j=1}^{\infty} \xi_j \alpha_j$ where $x = (\xi_j) \in l^2$.

We have $x, a \in l^2$. Thus, $\sum \xi_j^2 < \infty$ and $\sum \alpha_j^2 < \infty$.

We have, $f(x)$ is absolutely convergent since

$$\sum |\xi_j \alpha_j| \leq \left(\sum \xi_j^2 \right)^{\frac{1}{2}} \left(\sum \alpha_j^2 \right)^{\frac{1}{2}} < \infty$$

By Cauchy-Schwarz's inequality,

$$\sum (\xi_j \alpha_j)^2 \leq \sum |\xi_j \alpha_j|^2 \leq \sum \xi_j^2 \sum \alpha_j^2 < \infty$$

Therefore, $f(x) \in l^2$. Clearly, $f(x)$ is well-defined.

Let $x, y \in l^2$, $x = (\xi_j)$ and $y = (\eta_j)$. Then, f is linear since,

$$f(\alpha x + \beta y) = \sum_{j=1}^{\infty} (\alpha \xi_j + \beta \eta_j) \alpha_j = \alpha \sum_{j=1}^{\infty} \xi_j \alpha_j + \beta \sum_{j=1}^{\infty} \eta_j \alpha_j = \alpha f(x) + \beta f(y)$$

By Cauchy-Schwarz's inequality, we have

$$|f(x)| = \left| \sum_{j=1}^{\infty} \xi_j \alpha_j \right| \leq \sum_{j=1}^{\infty} |\xi_j \alpha_j| \leq \left(\sum_{j=1}^{\infty} \xi_j^2 \right)^{\frac{1}{2}} \left(\sum_{j=1}^{\infty} \alpha_j^2 \right)^{\frac{1}{2}} = \|x\| \|a\|$$

Thus,

$$\|f\| = \sup_{\|x\|=1} |f(x)| \leq \sup_{\|x\|=1} \|x\| \|a\| = \|a\|$$

And when $x = a$, we get $\|f\| = \sup_{x \neq 0} |f(x)| \geq |f(a)| = \|a\|$. Therefore, $\|f\| = \|a\|$.

Remark. Let X be a normed space over field K . Then the set X^* of all linear functional on X is a vector space over the same field K .

Proof. Let X^* be the set of all linear functional on X . Then X^* is a vector space with vector addition $f + g$ defined as $(f + g)(x) = f(x) + g(x)$ and scalar multiplication (αf) defined as $(\alpha f)(x) = \alpha f(x)$.

Let $f, g \in X^*$.

$$\begin{aligned} (f + g)(\alpha x + \beta y) &= f(\alpha x + \beta y) + g(\alpha x + \beta y) \\ &= \alpha f(x) + \beta f(y) + \alpha g(x) + \beta g(y) \\ &= \alpha(f + g)(x) + \beta(f + g)(y) \end{aligned}$$

$$\begin{aligned} (\gamma f)(\alpha x + \beta y) &= \gamma f(\alpha x + \beta y) \\ &= \gamma \alpha f(x) + \gamma \beta f(y) \\ &= \alpha(\gamma f)(x) + \beta(\gamma f)(y) \end{aligned}$$

Then $\alpha f + \beta g \in X^*$. Therefore, vector addition is closed.

Let $f, g \in X^*$. Then $(f + g)(x) = f(x) + g(x) = g(x) + f(x) = (g + f)(x)$. Thus, vector addition is commutative.

Consider $0 : X \rightarrow K$ defined by $0(x) = 0, \forall x \in X$. Clearly $0 \in X^*$ since $0(\alpha x + \beta y) = 0 = \alpha 0(x) + \beta 0(y)$. And $(f + 0)(x) = f(x) + 0(x) = f(x)$. Thus, $0 \in X^*$ is the additive identity.

Let $f \in X^*$. Consider $-f : X \rightarrow K$ defined by $(-f)(x) = -f(x), \forall x \in X$. Clearly, $-f \in X^*$ since $-f(\alpha x + \beta y) = \alpha(-f(x)) + \beta(-f(y))$. Then, $(f + (-f))(x) = f(x) + (-f)(x) = 0 = 0(x)$. Thus, every linear functional f in X^* has a unique additive inverse in X^* . \square

Definitions 14.2.14 (Algebraic Dual Space). Let X be a normed space. The set X^* of all linear functionals on X is the algebraic dual space of X .

Definitions 14.2.15 (Second Algebraic Dual Space). Let X be a normed space and X^* be its algebraic dual space of all linear functionals on X . Then the algebraic dual of X^* , $(X^*)^*$ is the second algebraic dual X^{**} of X .

Remark. Let $g \in X^{**}$ and $f \in X^*$. Then $g(f) = f(x)$ for some fixed $x \in X$.

Remark. Let $g_x : X^* \rightarrow K$ defined by $g_x(f) = f(x)$. Then g_x is linear since

$$\begin{aligned} g_x(\alpha f_1 + \beta f_2) &= (\alpha f_1 + \beta f_2)(x) \\ &= (\alpha f_1)(x) + (\beta f_2)(x) \\ &= \alpha f_1(x) + \beta f_2(x) \\ &= \alpha g_x(f_1) + \beta g_x(f_2) \end{aligned}$$

Definitions 14.2.16 (canonical mapping). Let $C : X \rightarrow X^{**}$ defined by $x \rightarrow g_x$ where $C(x) = g_x$ and $g_x(f) = f(x), \forall f \in X^*$. Then C is a canonical mapping on X .

Remark. Canonical mapping on X is an embedding of X into X^{**} . In other words, it is an isomorphism of X onto the subspace $\mathcal{R}(C)$ of X^{**} .

Definitions 14.2.17 (algebraic reflexivity). Let X be a normed space. If the canonical mapping $C : X \rightarrow X^{**}$ is surjective, then X is isomorphic to X^{**} . Then, X is **algebraically reflexive**.

Remark. If X is algebraically reflexive, then X^{**} is its algebraic reflection.

Definitions 14.2.18 (isomorphism). Isometry/Isomorphism of a metric space is a bijection that preserves metric/distance.

$$T : X \rightarrow Y \text{ such that } \forall x, y \in X, \tilde{d}(Tx, Ty) = d(x, y)$$

Isomorphism of a vector space is a bijection that preserves both the binary operations, say vector addition and scalar multiplication.

$$T : X \rightarrow Y \text{ such that } \forall x, y \in X, \forall \alpha, \beta \in K, T(\alpha x + \beta y) = \alpha Tx + \beta Ty$$

Isomorphism of a normed space is a vector space isomorphism that preserves norm.

$$T : X \rightarrow Y \text{ such that } \forall x \in X, \|Tx\| = \|x\|$$

Exercise §2.8

Definitions 14.2.19 (Null space of a set). Let X be a normed space and X^* be its algebraic dual space. Let $M^* \subset X^*$. Then null space of M^* , $\mathcal{N}(M^*)$ is the set of all $x \in X$ such that $f(x) = 0, \forall f \in M^*$.

$$\mathcal{N}(M^*) = \{x \in X : f(x) = 0, \forall f \in M^*\}$$

Definitions 14.2.20 (quotient space). Let Y be a subspace of a vector space X . The quotient space X/Y is set $\{x + Y : x \in X\}$ together with vector addition $(x_1 + Y) + (x_2 + Y) = (x_1 + x_2) + Y$ and $\alpha(x + Y) = (\alpha x + Y)$ where $x + Y = \{x + y \in X : y \in Y\}$.

Definitions 14.2.21 (codimension). Let Y be a subspace of vector space X . The codimension of Y is the dimension of the quotient space X/Y .

Definitions 14.2.22 (hyperplane). Let Y be a subspace of vector space X with codimension 1. Then the quotient space is a hyperplane parallel to Y .

Remark ([Kreyszig, 2014, §2.8 Exercise 12]). Let $f \in X^*$ and $f \neq 0$. Then, constant functionals on X forms a hyperplane parallel to the null space of f .

Definitions 14.2.23 (half space). Let $f \in X^*$ and $f \neq 0$ such that $f(x) = c, \forall x \in X$. Then the hyperplane divides the vector space X into two half spaces X_1, X_2 such that

$$X_1 = \{x \in X : f(x) \leq c\} \quad \text{and} \quad X_2 = \{x \in X : f(x) \geq c\}$$

14.2.9 Linear Operators and Functionals on finite dimensional spaces

Matrix of Linear Operator

Let X, Y be finite dimensional vector spaces over the same field K . Let $E = \{e_1, e_2, \dots, e_n\}$ be a basis for X and $B = \{b_1, b_2, \dots, b_r\}$ be a basis for Y . Let $T : X \rightarrow Y$ be a linear operator uniquely defined by $y_k = Te_k$. Then, $y_k = \sum \tau_{j,k} b_k = Te_k$. And the matrix $T_{EB} = (\tau_{j,k})$ is the matrix of the linear operator T with respect to the bases E and B .

For example¹¹, Let $T : \mathbb{R}^3 \rightarrow \mathbb{R}^3$ defined by $(\xi_1, \xi_2, \xi_3) \rightarrow (\xi_1, \xi_2, -\xi_1 - \xi_2)$.

$$y_1 = Te_1 = T(1, 0, 0) = (1, 0 - 1) = 1(1, 0, 0) + 0(0, 1, 0) - 1(0, 0, 1)$$

$$y_2 = Te_2 = T(0, 1, 0) = (0, 1, -1) = 0(1, 0, 0) + 1(0, 1, 0) - 1(0, 0, 1)$$

$$y_3 = Te_3 = T(0, 0, 1) = (0, 0, 0) = 0(1, 0, 0) + 0(0, 1, 0) + 0(0, 0, 1)$$

Thus,

$$T_{E,B} = \begin{pmatrix} 1 & 0 & 0 \\ 0 & 1 & 0 \\ -1 & -1 & 0 \end{pmatrix}$$

Linear operator T determines a unique matrix representing T with respect to the given bases. Similarly, any $r \times n$ matrix determines a unique linear operator which it represents.

Definitions 14.2.24 (Dual basis). Let X be a finite dimensional vector space over K with basis $E = \{e_1, e_2, \dots, e_n\}$. Let f_1, f_2, \dots, f_n be linear functionals $f_j : X \rightarrow K$ such that $f_j(e_k) = \delta_{j,k}$. Then $\{f_1, f_2, \dots, f_n\}$ is the dual basis of the basis E .

Theorem 14.2.26. Let X be a finite dimensional vector space. Let $E = \{e_1, e_2, \dots, e_n\}$ be a basis for X and $F = \{f_1, f_2, \dots, f_n\}$ be the dual basis of E . Then F is a basis for the algebraic dual of X , X^* . And $\dim X^* = \dim X$.

Proof. Let $\sum \beta_k f_k = 0$. Then, for each $j = 1, 2, \dots, n$

$$0 = 0(e_j) = \left(\sum_{k=1}^n \beta_k f_k \right) (e_j) = \sum_{k=1}^n \beta_k f_k(e_j) = \sum_{k=1}^n \beta_k \delta_{j,k} = \beta_j$$

Clearly, $\beta_1 f_1 + \beta_2 f_2 + \dots + \beta_n f_n = 0$ if and only if $\beta_j = 0$ for all j . Therefore, F is linearly independent.

Let $f : X \rightarrow K$ such that $f(e_j) = \alpha_j$. Let $x = \sum \xi_j e_j$ and $f_j(x) = \xi_j$. Since f is linear,

$$f(x) = \sum_{j=1}^n \xi_j f(e_j) = \sum_{j=1}^n \xi_j \alpha_j = \sum_{j=1}^n \alpha_j f_j(x) = \left(\sum_{j=1}^n \alpha_j f_j \right) (x)$$

Therefore, every linear functional on X is a linear combination of F .

Clearly, F is a basis for X^* and $\dim X^* = n = \dim X$ □

¹¹[Kreyszig, 2014, §2.10 Exercise 2]

Lemma 14.2.27. *Let X be a finite dimensional vector space. If $x_0 \in X$ such that $f(x_0) = 0$ for any $f \in X^*$, then $x_0 = 0$.*

Proof. Let $x_0 = \sum \xi_j e_j$. Then,

$$f(x_0) = 0 = \sum_{j=1}^n \xi_j f(e_j) = \sum_{j=1}^n \xi_j \alpha_j$$

Thus, $\sum \xi_j \alpha_j = 0$ for any choice of $\alpha_1, \alpha_2, \dots, \alpha_n \in K$. Therefore, for each j , we have $\xi_j = 0$. Clearly, $x_0 = 0$. \square

Theorem 14.2.28 (algebraic reflexive). *A finite dimensional vector space is algebraically reflexive.*

Proof. Let $C : X \rightarrow X^{**}$ be the canonical mapping such that

$$Cx(f) = g_x(f) = f(x)$$

where $x \in X$, $f \in X^*$ and $g \in X^{**}$.

Isomorphism C is injective, since $C_{x_0} = 0$ only if $x_0 = 0$.

$$\begin{aligned} C_{x_0}(f) &= g_{x_0}(f) = f(x_0) = 0, \quad \forall f \in X^* \\ \implies x_0 &= 0 \text{ by lemma} \end{aligned}$$

Thus, there exists an inverse function $C^{-1} : \mathcal{R}(C) \rightarrow X$ and $\dim X \leq \dim \mathcal{R}(C)$. Thus, $\dim X^{**} = \dim X^* = \dim X$ and $\dim \mathcal{R}(C) = \dim X^{**}$. Therefore, C is surjective and the vector space X is algebraically reflexive. \square

14.2.10 Normed spaces of Operators, Dual space

Definitions 14.2.25 ($B(X, Y)$). Let X, Y be normed spaces. The set $B(X, Y)$ of all bounded, linear functionals from X into Y is a vector space together with vector addition, $T_1 + T_2$ defined by,

$$T_1 + T_2(x) = T_1(x) + T_2(x)$$

and scalar multiplication αT defined by,

$$(\alpha T)(x) = \alpha T(x)$$

In addition, $B(X, Y)$ is a normed space with norm

$$\|T\| = \sup_{\substack{x \in X \\ x \neq 0}} \frac{\|Tx\|}{\|x\|} = \sup_{\substack{x \in X \\ \|x\|=1}} \|Tx\|$$

Proof. Clearly, vector addition is closed, associative, and commutative. The identity vector 0 is defined by $0x = 0$, $\forall x \in X$. And additive inverse $(-T)$ is defined by $(-T)(x) = -T(x)$ for each $T \in B(X, Y)$. Also we have, $1T = T$, $\alpha(\beta T) = \alpha\beta T$, $\alpha(T_1 + T_2) = \alpha T_1 + \alpha T_2$ and $(\alpha_1 + \alpha_2)T = \alpha_1 T + \alpha_2 T$. Thus, $B(X, Y)$ is a vector space. \square

Theorem 14.2.29. *If Y is a Banach space, then $B(X, Y)$ is a Banach space.*

Proof. Let $\{T_n\}$ be a Cauchy sequence in $B(X, Y)$. That is, $\forall \varepsilon > 0, \exists N \in \mathbb{N}$ such that $\forall n, m > N$ we have $\|T_n - T_m\| \leq \varepsilon$. Thus, for any $x \in X$ we have

$$\|T_n(x) - T_m(x)\| \leq \|T_n - T_m\| \|x\| \leq \varepsilon \|x\|$$

Fix $x \in X$ and $\varepsilon > 0$, then we have $\varepsilon_x > 0$ such that $\varepsilon_x \|x\| < \varepsilon$. Then, there exists $N \in \mathbb{N}$ such that $\forall n, m > N$, $\|T_n(x) - T_m(x)\| < \varepsilon$. Thus, $\{T_n(x)\}$ is Cauchy in Y . Since, Y is complete, $\{T_n(x)\}$ converges, say $T_n(x) \rightarrow y$.

Define $T : X \rightarrow Y$ such that $Tx = y$ where $T_n(x) \rightarrow y$ as $n \rightarrow \infty$. Operator T is linear since

$$T(\alpha x + \beta z) = \lim_{n \rightarrow \infty} T_n(\alpha x + \beta z) = \lim_{n \rightarrow \infty} \alpha T_n(x) + \lim_{n \rightarrow \infty} \beta T_n(z) = \alpha T(x) + \beta T(z)$$

We have,

$$\|T_n(x) - T(x)\| = \left\| T_n(x) - \lim_{m \rightarrow \infty} T_m(x) \right\| \leq \|T_n - T_m\| \|x\| \leq \varepsilon \|x\|$$

Thus, $T_n - T$ is bounded and

$$\sup_{\substack{x \in X \\ \|x\|=1}} \|(T_n - T)x\| = \sup_{\substack{x \in X \\ \|x\|=1}} \|T_n(x) - T(x)\| \leq \varepsilon$$

Thus, $\|T_n - T\| \rightarrow 0$ as $n \rightarrow \infty$. We have, $T = T_n - (T_n - T)$ and

$$\|T(x)\| \leq \|T_n(x)\| + \|(T_n - T)(x)\| = \varepsilon \|x\| + \varepsilon \|x\|$$

Therefore, T is bounded. □

Definitions 14.2.26 (Dual space X'). Let X be a normed space. The set of X' of all bounded, linear functionals $B(X, K)$ is the dual space of X . The dual space X' of X is a normed space with norm

$$\|f\| = \sup_{\substack{x \in X \\ x \neq 0}} \frac{|f(x)|}{\|x\|} = \sup_{\substack{x \in X \\ \|x\|=1}} |f(x)|$$

Theorem 14.2.30. *The dual space X' of any normed space X is a Banach space.*

Proof. We know that, $B(X, Y)$ is Banach space if Y is a Banach space. We have $K = \mathbb{R}$ or $K = \mathbb{C}$. And \mathbb{R} as well as \mathbb{C} are complete normed spaces. Therefore, $X' = B(X, K)$ is a Banach space. □

14.2.11 Dual Spaces

Remark. Dual space of \mathbb{R}^n is \mathbb{R}^n

Proof. Let $f \in (\mathbb{R}^n)'$. Let $E = \{e_1, e_2, \dots, e_n\}$ be a basis for \mathbb{R}^n . Then f is of the form, $f = \sum \xi_j \alpha_j$ where $\alpha_j = f(e_j)$.

By Cauchy-Schwarz's inequality, we have

$$|f(x)| \leq \sum |\xi_j \alpha_j| \leq \left(\sum \xi_j^2 \right)^{\frac{1}{2}} \left(\sum \alpha_j^2 \right)^{\frac{1}{2}} = \|x\| \left(\sum \alpha_j^2 \right)^{\frac{1}{2}}$$

Thus,

$$\|f\| = \sup_{\substack{x \in \mathbb{R}^n \\ \|x\|=1}} |f(x)| \leq \left(\sum \alpha_j^2 \right)^{\frac{1}{2}}$$

Also for $x = (\alpha_j)$ we get,

$$\|f\| = \left(\sum \alpha_j^2 \right)^{\frac{1}{2}}$$

And for any $c = (\alpha_1, \alpha_2, \dots, \alpha_n) \in \mathbb{R}^n$, we have a linear functional f such that $f(x) = \sum \xi_k \alpha_k$ where $x = (\xi_1, \xi_2, \dots, \xi_n)$. We know that, for finite dimensional normed space X , every linear functional on X is bounded. Therefore, f is bounded and $f \in \mathbb{R}^{n'}$. Thus, the map from the dual space of \mathbb{R}^n into \mathbb{R}^n given by $f \rightarrow c$ is an isomorphism preserving the Euclidean norm on \mathbb{R}^n . Therefore, the dual space of \mathbb{R}^n is \mathbb{R}^n itself. \square

Remark. Dual space of l^1 is l^∞

Proof. The Schauder basis for l^1 is (e_k) where $e_k = (\delta_{k,j})$. That is, each sequence e_k has one as its k th term and all other terms are zeroes. For example, $(\frac{1}{2}, \frac{1}{4}, \dots) = \sum \frac{1}{2^k} e_k$. Then any $x \in l^1$ is of the form

$$x = (\xi_j) = \sum_{k=1}^{\infty} \xi_k \delta_{k,j} = \sum_{k=1}^{\infty} \xi_k e_k$$

Let $f \in l^{1'}$. Then,

$$f(x) = f\left(\sum \xi_k e_k\right) = \sum \xi_k f(e_k) = \sum \xi_k \gamma_k$$

where $\gamma_k = f(e_k)$. Since f is bounded, we have $|f(x)| \leq \|f\| \|x\|$. Also we have, $\|e_k\| = 1$ since $e_k = (\delta_{k,j})$. Thus,

$$|\gamma_k| = |f(e_k)| \leq \|f\| \|e_k\| = \|f\|$$

We have, $|\gamma_k| \leq \|f\|$ for any k . Therefore,

$$\sup_k |\gamma_k| \leq \|f\|$$

For any $b = (\beta_k) \in l^\infty$, there exists a functional $g \in l^{1'}$ such that

$$g(x) = \sum \xi_k \beta_k$$

where $x = (\xi_k) \in l^1$. Clearly, g is linear. And,

$$|g(x)| \leq \sum |\xi_k \beta_k| \leq \sup_j |\beta_j| \sum |\xi_j| = \|x\| \sup_j |\beta_j|$$

Thus, g is bounded and $g \in l^{1'}$. And,

$$|f(x)| = \left| \sum \xi_k \gamma_k \right| \leq \sup_j |\gamma_j| \sum |\xi_k| = \|x\| \sup_j |\gamma_j|$$

Thus,

$$\|f\| = \sup_{\substack{x \in l^1 \\ \|x\|=1}} |f(x)| \leq \sup_j |\gamma_j|$$

Therefore, $\|f\| = \sup_j |\gamma_j|$.

Clearly, the mapping from the dual space of l^1 into l^∞ given by $f \rightarrow c$ where $c = \sum \gamma_j$ is an isomorphism preserving the norm on l^∞ . Thus, the dual space of l^1 is isomorphic to l^∞ . \square

Remark. Dual space of l^p is l^q where p, q are conjugate exponents.

Proof. Sequence (e_k) is a Schauder basis for l^p where $e_k = \delta_{k,j}$. And every $x \in l^p$ is of the form

$$x = (\xi_k) = \sum_{k=1}^{\infty} \xi_k e_k$$

Let $f \in l^{p'}$. Then f is linear and

$$f(x) = f\left(\sum_{k=1}^{\infty} \xi_k e_k\right) = \sum_{k=1}^{\infty} \xi_k f(e_k) = \sum_{k=1}^{\infty} \xi_k \gamma_k$$

where $\gamma_k = f(e_k)$.

Define sequence $\{x_n\}$ in l^p such that $x_n = (\xi_{k,n})$ where

$$\xi_{k,n} = \begin{cases} \frac{|\gamma_k|^q}{\gamma_k} & k \leq n \text{ and } \gamma_k \neq 0 \\ 0 & k > n \text{ or } \gamma_k = 0 \end{cases}$$

Then,

$$f(x_n) = \sum \xi_{k,n} \gamma_k = \sum |\gamma_k|^q$$

Let p, q be conjugate exponents. Then $pq = p + q$. And $|\xi_{k,n}| = |\gamma_k|^{(q-1)}$. Thus,

$$\begin{aligned} |f(x_n)| &\leq \|f\| \|x_n\| \\ &= \|f\| \left(\sum |\xi_{k,n}|^p \right)^{\frac{1}{p}} \\ &= \|f\| \left(\sum |\gamma_k|^{(q-1)p} \right)^{\frac{1}{p}} \\ &= \|f\| \left(\sum |\gamma_k|^q \right)^{\frac{1}{p}} \text{ since } pq = p + q \end{aligned}$$

Therefore,

$$|f(x_n)| = \sum |\gamma_k|^q \leq \|f\| \left(\sum |\gamma_k|^q \right)^{\frac{1}{p}}$$

$$\Rightarrow \left(\sum |\gamma_k|^q \right)^{\left(1 - \frac{1}{p}\right)} = \left(\sum |\gamma_k|^q \right)^{\frac{1}{q}} \leq \|f\|$$

Let $b = (\beta_k) \in l^q$. And there exists a functional g on l^p such that

$$g(x) = \sum \xi_k \beta_k$$

where $x = (\xi_k) \in l^p$. Clearly, g is linear. And,

$$|g(x)| \leq \left| \sum \xi_k \beta_k \right| \leq \left(\sum |\xi_k|^p \right)^{\frac{1}{p}} \left(\sum |\beta_k|^q \right)^{\frac{1}{q}} = \|x\| \left(\sum |\beta_k|^q \right)^{\frac{1}{q}}$$

Therefore, g is a bounded, linear functional on l^p . That is, $g \in l^{p'}$.

And we have,

$$\begin{aligned} |f(x)| &\leq \left| \sum \xi_k \gamma_k \right| \\ &\leq \left(\sum |\xi_k|^p \right)^{\frac{1}{p}} \left(\sum |\gamma_k|^q \right)^{\frac{1}{q}} \\ &\leq \|x\| \left(\sum |\gamma_k|^q \right)^{\frac{1}{q}} \\ \|f\| &\leq \sup_{\substack{x \in l^p \\ \|x\|=1}} \left(\sum |\gamma_k|^q \right)^{\frac{1}{q}} \end{aligned}$$

Then,

$$\|f\| = \left(\sum |\gamma_k|^q \right)^{\frac{1}{q}}$$

Clearly, the mapping from the dual space of l^p into l^q given by $f \rightarrow c$ is an isomorphism preserving norm. Therefore, the dual space of l^p is isomorphic to l^q where p, q are conjugate exponents. That is, $\frac{1}{p} + \frac{1}{q} = 1$. \square

Module 3 - Hilbert Spaces

14.3 Hilbert Spaces

14.3.1 Inner Product space, Hilbert space

Definitions 14.3.1 (inner product). Let X be a vector space over the field K . Inner product is a function $\langle \cdot, \cdot \rangle : X \times X \rightarrow K$ which satisfies

1. $\langle x + y, z \rangle = \langle x, z \rangle + \langle y, z \rangle$
2. $\langle \alpha x, y \rangle = \alpha \langle x, y \rangle$
3. $\langle x, y \rangle = \overline{\langle y, x \rangle}$
4. $\langle x, x \rangle \geq 0$ and
 $\langle x, x \rangle = 0 \iff x = 0$

Definitions 14.3.2. A vector space X on field K together with an inner product function $\langle \cdot, \cdot \rangle : X \times X \rightarrow K$ is an inner product space.

Definitions 14.3.3 (Hilbert). Hilbert spaces are complete, inner product spaces.

Remark. Inner product spaces are normed spaces with norm

$$\|x\| = \sqrt{\langle x, x \rangle}$$

Definitions 14.3.4 (sesquilinear). An operator $T : X \rightarrow Y$ is sesquilinear if

$$T(\alpha x + \beta y) = \alpha T(x) + \bar{\beta} T(y)$$

Inner product is a sesquilinear operator.

Remark. Inner product spaces are metric spaces with metric

$$d(x, y) = \|x - y\| = \sqrt{\langle x - y, x - y \rangle}$$

Remark. Let X be a real, inner product space. Then,

$$\langle x, y \rangle = \langle y, x \rangle$$

Remark. Let X be a complex, inner product space. Then, $\langle x, y \rangle = \overline{\langle y, x \rangle}$.

Proof. Let $x = a + ib$ and $y = c + id$ where $a, b, c, d \in \mathbb{R}$. Then,

$$\langle x, y \rangle = \langle a + ib, c + id \rangle = \langle a, c \rangle - i \langle a, d \rangle + i \langle b, c \rangle + \langle b, d \rangle$$

$$\begin{aligned} \langle y, x \rangle &= \langle c + id, a + ib \rangle \\ &= \langle c, a \rangle - i \langle c, b \rangle + i \langle d, a \rangle + \langle d, b \rangle \\ &= \langle a, c \rangle - i \langle b, c \rangle + i \langle a, d \rangle + \langle b, d \rangle \\ &= \overline{\langle x, y \rangle} \end{aligned}$$

Therefore,

$$\langle x, y \rangle = \overline{\langle y, x \rangle}$$

□

Remark. From the axiom of inner product, we have

1. $\langle \alpha x + \beta y, z \rangle = \alpha \langle x, z \rangle + \beta \langle y, z \rangle$
2. $\langle x, \alpha y \rangle = \overline{\alpha} \langle x, y \rangle$
3. $\langle x, \alpha y + \beta z \rangle = \overline{\alpha} \langle x, y \rangle + \overline{\beta} \langle x, z \rangle$

Lemma 14.3.1 (parallelogram equality). *Let X be an inner product space. Let $\|\cdot\|$ be a the norm induced by the inner product. That is, $\|x\| = \sqrt{\langle x, x \rangle}$. Then,*

$$\|x + y\|^2 + \|x - y\|^2 = 2(\|x\|^2 + \|y\|^2)$$

Proof. Let $\|\cdot\|$ be a norm induced by inner product. Then $\|x\|^2 = \langle x, x \rangle$ and $\|y\|^2 = \langle y, y \rangle$.

$$\begin{aligned} \|x + y\|^2 + \|x - y\|^2 &= \langle x + y, x + y \rangle + \langle x - y, x - y \rangle \\ &= \langle x, x \rangle + \langle x, y \rangle + \langle y, x \rangle + \langle y, y \rangle \\ &\quad + \langle x, x \rangle + \langle x, -y \rangle + \langle -y, x \rangle + \langle -y, -y \rangle \\ &= \langle x, x \rangle + \langle x, y \rangle + \langle y, x \rangle + \langle y, y \rangle \\ &\quad + \langle x, x \rangle - \langle x, y \rangle - \langle y, x \rangle + \langle y, y \rangle \\ &= 2 \langle x, x \rangle + 2 \langle y, y \rangle = 2(\|x\|^2 + \|y\|^2) \end{aligned}$$

□

Not all normed spaces are inner product spaces.

For example, $C[a, b]$, l^p are not inner product spaces.

Definitions 14.3.5 (orthogonality). Let X be an inner product space and $x, y \in X$. Then x, y are orthogonal if $\langle x, y \rangle = 0$, say $x \perp y$.

Let A, B be subsets of X . Then A and B are orthogonal if $\forall a \in A, \forall b \in B, \langle a, b \rangle = 0$.

$$A \perp B \iff \forall a \in A, \forall b \in B, a \perp b = 0 \iff \forall a \in A, \forall b \in B, \langle a, b \rangle = 0$$

Examples

Euclidean Space

Banach space \mathbb{R}^n is a Hilbert space with inner product,

$$\langle x, y \rangle = \sum_{j=1}^n \xi_j \eta_j$$

Unitary Space

Banach space \mathbb{C}^n is a Hilbert space with inner product,

$$\langle x, y \rangle = \sum_{j=1}^n \xi_j \overline{\eta_j}$$

L^2 Space

Banach space of Lebesgue square integrable functions is a Hilbert space with inner product,

$$\langle x, y \rangle = \left(\int_a^b x(t) \overline{y(t)} dt \right)^{\frac{1}{2}}$$

L^2 space is the completion of the incomplete normed space of all continuous complex-valued functions on $[a, b]$ with norm $\|x\| = \sqrt{\langle x, x \rangle}$.

Remember that, the space of all continuous functions on $[a, b]$ is incomplete since the Cauchy sequence of continuous functions $x_n : [a, b] \rightarrow [0, 1]$ defined by

$$x_n(t) = \begin{cases} 0 & a \leq t \leq \frac{b+a}{2} \\ \frac{2t-b-a}{2n} & \frac{b+a}{2} < t < \frac{b+a}{2} + \frac{1}{n} \\ 1 & \frac{b+a}{2} + \frac{1}{n} \leq t \leq b \end{cases}$$

doesn't converge to continuous function.

Also note that, the above sequence is not a Cauchy sequence in $C[a, b]$ with norm $\|x\| = \max |x(t)|$ since

$$\|x_n(t) - x_m(t)\| = 1 - \frac{1}{mn}$$

 l^2 Space

Banach space l^2 is a Hilbert space with inner product,

$$\langle x, y \rangle = \sum_{j=1}^{\infty} \xi_j \overline{\eta_j}$$

Counter-examples

Remark. Banach space $C[a, b]$ with norm $\|x\| = \max |x(t)|$ is not an inner product space.

Proof. We have $C[a, b]$, the space of all continuous real-valued function on compact interval $[a, b]$ is a normed space with

$$\|x\| = \max_{t \in [a, b]} |x(t)|$$

Let $x(t) = 1$ and $y(t) = \frac{t-a}{b-a}$. Then $x + y(t) = 1 + \frac{t-a}{b-a}$ and $x - y(t) = \frac{b-t}{b-a}$. And $\|x\| = 1$, $\|y\| = 1$, $\|x + y\| = 2$ ¹² and $\|x - y\| = 1$. Clearly, $\|x + y\|^2 + \|x - y\|^2 = 5 \neq 4 = 2(\|x\|^2 + \|y\|^2)$. Therefore, $\|\cdot\|$ can't be induced from any inner product on $C[a, b]$. Thus, $C[a, b]$ (with above norm) is not an inner product space. \square

¹²Since $x+y(b) = 2$ is the maximum value of $x+y$ in $[a, b]$, we have $\|x+y\| = \max |x+y| = 2$.

Remark. Banach space l^p is not an inner product space, if $p \neq 2$.

Proof. Suppose $p \neq 2$. We have, $e_j = \delta_{j,k}$ is a Schauder basis for l^p space. Let $x = e_1 + e_2$ and $y = e_1 - e_2$ where $e_j = \delta_{j,k}$. Then $x + y = 2e_1$ and $x - y = 2e_2$. And $\|x\| = 2^{\frac{1}{p}}$, $\|y\| = 2^{\frac{1}{p}}$, $\|x + y\| = 2$ and $\|x - y\| = 2$. Then,

$$\|x + y\|^2 + \|x - y\|^2 = 8 \neq 4 \times 2^{\frac{2}{p}} = 2(\|x\|^2 + \|y\|^2), \quad \because p \neq 2$$

Thus, the norm on l^p can't be induced from an inner product when $p \neq 2$. Therefore, l^p is not an inner product space, if $p \neq 2$. \square

Remark (Hilbert). The sequence space l^2 was introduced by David Hilbert and is known as **Hilbert's sequence space**. It is a separable, complete, inner product space. Later, the condition of separability was dropped by mathematicians as most of the theory didn't make use of it.

Remark. Let X be a real inner product space. Then,

$$\langle x, y \rangle = \frac{1}{4} (\|x + y\|^2 - \|x - y\|^2)$$

Theorem 14.3.2 (Polarization identity). *Let X be a complex inner product space. Then, the real and imaginary parts of $\langle x, y \rangle$ are*

$$\Re(\langle x, y \rangle) = \frac{1}{4} (\|x + y\|^2 - \|x - y\|^2)$$

$$\Im(\langle x, y \rangle) = \frac{1}{4} (\|x + iy\|^2 - \|x - iy\|^2)$$

Proof. Let X be a complex, inner product space. Let $x, y \in X$. Then,

$$\begin{aligned} \|x + y\|^2 - \|x - y\|^2 &= \langle x + y, x + y \rangle - \langle x - y, x - y \rangle \\ &= \langle x, x \rangle + \langle x, y \rangle + \langle y, x \rangle + \langle y, y \rangle \\ &\quad - \langle x, x \rangle - \langle x, -y \rangle - \langle -y, x \rangle - \langle -y, -y \rangle \\ &= \langle x, x \rangle + \langle x, y \rangle + \langle y, x \rangle + \langle y, y \rangle \\ &\quad - \langle x, x \rangle + \langle x, y \rangle + \langle y, x \rangle - \langle y, y \rangle \\ &= 2(\langle x, y \rangle + \langle y, x \rangle) \end{aligned}$$

We know that, $\langle x, y \rangle = \overline{\langle y, x \rangle}$. Thus,

$$\|x + y\|^2 - \|x - y\|^2 = 2(\langle x, y \rangle + \overline{\langle x, y \rangle}) = 4\Re \langle x, y \rangle$$

Therefore,

$$\Re \langle x, y \rangle = \frac{1}{4} (\|x + y\|^2 - \|x - y\|^2)$$

Let $z = iy$.

$$\begin{aligned} \|x + iy\|^2 - \|x - iy\|^2 &= \|x + z\|^2 - \|x - z\|^2 \\ &= 2\langle x, z \rangle + 2\langle z, x \rangle \\ &= 2\langle x, iy \rangle + 2\langle iy, x \rangle \\ &= -2i\langle x, y \rangle + 2i\langle y, x \rangle \\ &= -2i(\langle x, y \rangle - \overline{\langle x, y \rangle}) \\ &= -2i(\langle x, y \rangle - \overline{\langle x, y \rangle}) \\ &= 4\Im \langle x, y \rangle \end{aligned}$$

$$\Im \langle x, y \rangle = \frac{1}{4} (\|x + iy\|^2 - \|x - iy\|^2)$$

□

Exercise §3.1

Theorem 14.3.3 (Pythagorean Identity). *Let X be an inner product space. Let $x, y \in X$. If $x \perp y$, then*

$$\|x + y\|^2 = \|x\|^2 + \|y\|^2$$

Proof. Suppose $x \perp y$. Then $\langle x, y \rangle = \langle y, x \rangle = 0$. And,

$$\begin{aligned} \|x + y\|^2 &= \langle x + y, x + y \rangle \\ &= \langle x, x \rangle + \langle x, y \rangle + \langle y, x \rangle + \langle y, y \rangle \\ &= \|x\|^2 + \|y\|^2 \end{aligned}$$

□

Remark.

$$\|x + y\|^2 = \|x\|^2 + \|y\|^2 \not\Rightarrow x \perp y$$

Proof.

$$\begin{aligned} \|x + y\|^2 &= \langle x + y, x + y \rangle \\ &= \langle x, x \rangle + \langle x, y \rangle + \langle y, x \rangle + \langle y, y \rangle \\ &= \|x\|^2 + \|y\|^2 + \langle x, y \rangle + \langle y, x \rangle \end{aligned}$$

Thus, if $\|x + y\|^2 = \|x\|^2 + \|y\|^2$, then $\langle x, y \rangle + \langle y, x \rangle = 0$.

We know that $\langle x, y \rangle = \overline{\langle y, x \rangle}$. And, $\langle x, y \rangle + \overline{\langle x, y \rangle} = 0$ only if $\langle x, y \rangle$ is purely imaginary. Thus, the Pythagorean theorem is true if $\langle x, y \rangle$ is purely imaginary. □

Theorem 14.3.4 (Apollonius Identity). *Let X be an inner product space. Let $x, y, z \in X$. Then,*

$$\|z - x\|^2 + \|z - y\|^2 = \frac{1}{2}\|x - y\|^2 + 2\left\|z - \frac{1}{2}(x + y)\right\|^2$$

Proof. Let $s = z - \frac{1}{2}(x + y)$ and $t = \frac{1}{2}(x - y)$. Then $s + t = z - y$ and $s - t = z - x$.

$$\begin{aligned} \|z - y\|^2 + \|z - x\|^2 &= \|s + t\|^2 + \|s - t\|^2 \\ &= 2(\|s\|^2 + \|t\|^2) \\ &= 2\left(\left\|z - \frac{1}{2}(x + y)\right\|^2 + \left\|\frac{1}{2}(x - y)\right\|^2\right) \\ &= 2\left\|z - \frac{1}{2}(x + y)\right\|^2 + \frac{1}{2}\|x - y\|^2 \end{aligned}$$

□

14.3.2 Further properties of inner product spaces

Remark.

$$\|\alpha x\|^2 = \langle \alpha x, \alpha x \rangle = \alpha \bar{\alpha} \langle x, x \rangle = |\alpha|^2 \|x\|^2$$

Lemma 14.3.5 (Schwarz Inequality). *Let X be an inner product space. Let $x, y \in X$.*

$$|\langle x, y \rangle| \leq \|x\| \|y\|$$

*And equality holds if $\{x, y\}$ are linearly independent. Also,*¹³

$$\|x + y\| \leq \|x\| + \|y\|$$

And equality holds if $y = 0$ or $x = cy$.

Proof. If $y = 0$, then $\langle x, y \rangle = 0$ and $\|y\| = 0$. And $|\langle x, y \rangle| = \|x\| \|y\| = 0$. Similarly, if $x = 0$, then $\|\langle x, y \rangle\| = \|x\| \|y\| = 0$.

Suppose $x \neq 0$ and $y \neq 0$. Then

$$\begin{aligned} \|x - \alpha y\|^2 &= \langle x - \alpha y, x - \alpha y \rangle \\ &= \langle x, x \rangle - \bar{\alpha} \langle x, y \rangle - \alpha \langle y, x \rangle + \alpha \bar{\alpha} \langle y, y \rangle \end{aligned}$$

Let $\bar{\alpha} = \frac{\langle y, x \rangle}{\langle y, y \rangle}$. Then

$$\begin{aligned} \|x - \alpha y\|^2 &= \langle x, x \rangle - \frac{\langle y, x \rangle \langle x, y \rangle}{\langle y, y \rangle} \\ &= \|x\|^2 - \frac{\|\langle x, y \rangle\|^2}{\|y\|^2} \end{aligned}$$

We have

$$0 \leq \|x - \alpha y\|^2 \leq \|x\|^2 - \frac{\|\langle x, y \rangle\|^2}{\|y\|^2}$$

Thus,

$$\|\langle x, y \rangle\| \leq \|x\| \|y\|$$

Clearly, equality holds if and only if $\|x - \alpha y\|^2 = 0$ or $x = 0$ or $y = 0$. In other words, equality holds if and only if $x = \alpha y$ or $x = 0$ or $y = 0$. Therefore, equality holds if and only if x, y are linearly independent.

We have,

$$\begin{aligned} \|x + y\|^2 &= \langle x + y, x + y \rangle \\ &= \|x\|^2 + \langle x, y \rangle + \langle y, x \rangle + \|y\|^2 \end{aligned}$$

Applying modulus on either sides, we get

$$\|x + y\|^2 \leq \|x\|^2 + |\langle x, y \rangle| + |\langle y, x \rangle| + \|y\|^2$$

¹³Every inner product space is also a normed space. Then the induced norm should possess the triangular inequality property. I think it is too late to prove this !

By Schwarz inequality, we have

$$\|x + y\|^2 \leq \|x\|^2 + 2\|x\| \|y\| + \|y\|^2 = (\|x\| + \|y\|)^2$$

Therefore,

$$\|x + y\| \leq \|x\| + \|y\|$$

And equality holds if and only if

$$\langle x, y \rangle + \langle y, x \rangle = 2\|x\| \|y\|$$

That is,

$$\Re \langle x, y \rangle = \|x\| \|y\| \geq |\langle x, y \rangle|$$

We know that the real part of a complex number can never exceed its absolute value. Therefore, $\|x\| \|y\| = |\langle x, y \rangle|$.

From Schwarz inequality, the equality holds if and only if x, y are linearly independent. Therefore, $\|x + y\| = \|x\| + \|y\|$ if and only if x, y are linearly independent. \square

Lemma 14.3.6 (Continuity). *Let X be an inner product space. Let $\{x_n\}, \{y_n\}$ be sequences in X . Suppose $x_n \rightarrow x$ and $y_n \rightarrow y$. Then $\langle x_n, y_n \rangle \rightarrow \langle x, y \rangle$.*

In other words, Inner product is continuous.

Proof. We have $x_n \rightarrow x$ and $y_n \rightarrow y$ as $n \rightarrow \infty$. Then, $\|x_n - x\| \rightarrow 0$ and $\|y_n - y\| \rightarrow 0$ as $n \rightarrow \infty$.

$$\begin{aligned} |\langle x_n, y_n \rangle - \langle x, y \rangle| &= |\langle x_n, y_n \rangle - \langle x_n, y \rangle + \langle x_n, y \rangle - \langle x, y \rangle| \\ &\leq |\langle x_n, y_n - y \rangle| + |\langle x_n - x, y \rangle| \end{aligned}$$

By Schwarz inequality

$$\leq \|x_n\| \|y_n - y\| + \|x_n - x\| \|y\|$$

Therefore, $|\langle x_n, y_n \rangle - \langle x, y \rangle| \rightarrow 0$ as $n \rightarrow \infty$. In other words, $\langle x_n, y_n \rangle \rightarrow \langle x, y \rangle$ as $n \rightarrow \infty$. \square

Definitions 14.3.6 (isomorphism). Let X, Z be inner product spaces. If there exists a bijection $T : X \rightarrow Z$ such that $\langle Tx, Ty \rangle = \langle x, y \rangle$ for any $x, y \in X$. Then T is an isomorphism preserving inner product.

Theorem 14.3.7. *Every inner product space X has a unique completion H in which it is dense.*

Proof. Let X be an inner product space. Then X is a normed space with inner product induced norm $\|x\| = \sqrt{\langle x, x \rangle}$. By completion theorem for normed spaces, there exists a unique¹⁴ Banach space H which contains a dense subset W which is isomorphic to X .

¹⁴Unique upto isomorphism - If there exists two Banach spaces H_1, H_2 which contain a dense subset which is isomorphic to X , then H_1, H_2 are isomorphic normed spaces.

We have, W is dense in H and H is the set of all equivalent classes of Cauchy sequences on X . And $\hat{x}_n \sim \hat{w}$ if and only if $\|x_n - w\| \rightarrow 0$ as $n \rightarrow \infty$ for some $w \in W$.

Defining an inner product on H ,

$$\langle \hat{x}, \hat{y} \rangle = \lim_{n \rightarrow \infty} \langle x_n, y_n \rangle$$

By polarisation identity,

$$\begin{aligned} \langle x_n, y_n \rangle &= \frac{1}{4} (\|x_n + y_n\|^2 - \|x_n - y_n\|^2) + \frac{i}{4} (\|x_n + iy_n\|^2 - \|x_n - iy_n\|^2) \\ &= \frac{1}{4} (\langle x_n + y_n, x_n + y_n \rangle - \langle x_n - y_n, x_n - y_n \rangle) \\ &\quad + \frac{i}{4} (\langle x_n + iy_n, x_n + iy_n \rangle - \langle x_n - iy_n, x_n - iy_n \rangle) \end{aligned}$$

And,

$$\begin{aligned} \langle Tx, Ty \rangle &= \langle \hat{x}, \hat{y} \rangle \\ &= \lim_{n \rightarrow \infty} \langle x_n, y_n \rangle \\ &= \frac{1}{4} (\langle x + y, x + y \rangle - \langle x - y, x - y \rangle) \\ &\quad + \frac{i}{4} (\langle x + iy, x + iy \rangle - \langle x - iy, x - iy \rangle) \\ &= \langle x, y \rangle \end{aligned}$$

Thus, $T : X \rightarrow W$ is an isomorphism preserving norms.

Let H_1, H_2 be two completions of X . Then, we have normed space isomorphism $T : H_1 \rightarrow H_2$ such that $\|Tx\| = \|x\|$. Clearly, $\langle Tx, Ty \rangle = \langle x, y \rangle$. Therefore, completion of X is unique upto an inner product isomorphism. \square

Theorem 14.3.8. *Let Y be a subspace of a Hilbert space. Then,*

1. *Y is complete if and only if Y is closed in H .*
2. *If Y is finite dimensional, then Y is complete.*
3. *If H is separable, then Y is separable.*

Proof. Let H is a Hilbert space. Then H is a Banach space with induced norm $\|x\| = \sqrt{\langle x, x \rangle}$. Then, any subset Y of H is complete if and only if Y is closed in H .

Let Y be a finite dimensional subspace of Hilbert space H . Then, H is a Banach space with induced norm and any subspace of H is a normed space. We know that any finite dimensional normed space is complete. Thus, Y is complete.

Let H be a separable, Hilbert space. Then H has a countable, dense subset W . Let Y be a subspace of H . Then $W \cap Y$ is countable, dense subset of Y . Thus, subspace Y is separable. \square

Every subset of separable inner product space is separable.

14.3.3 Orthogonal complements and direct sums

Definitions 14.3.7 (distance from a set). Let X be metric space. Let $x \in X$ and $M \subset X$. The distance of x from M is defined as

$$d(x, M) = \inf_{y \in M} d(x, y)$$

If X is normed space, then

$$d(x, M) = \inf_{y \in M} \|x - y\|$$

Definitions 14.3.8 (convex). Let M be a subset of a vector space X . Then M is convex if for any two points $x, y \in M$, the segment $z = \alpha x + (1 - \alpha)y$ is contained in M . That is,

$$\forall x, y \in M, \forall \alpha \in [0, 1], \alpha x + (1 - \alpha)y \in M$$

Theorem 14.3.9 (minimising vector). Let X be an inner product space and M be a nonempty, convex, complete subset of X . Then for any $x \in X$, there exists a unique $y \in M$ such that $d(x, M) = d(x, y)$.

Proof. Let X be an inner product space with metric $d(x, y) = \sqrt{\langle x - y, x - y \rangle}$.

Step 1 : Existence of minimising vector

Let $\delta = d(x, M)$. Let $\{\delta_n\}$ be a sequence converging to δ . Then, there exists a sequence $\{y_n\}$ such that $y_n \in M$ and $\delta < d(x, y_n) < \delta_n$.

We know that

$$\delta = d(x, M) = \inf_{y \in M} d(x, y) = \inf_{y \in M} \|x - y\|$$

$$\left\| \frac{y_n + y_m}{2} - x \right\| \leq \inf_{y \in M} \|x - y\| = \delta, \text{ since } \frac{y_n + y_m}{2} \in M$$

Define $v_n = y_n - x$ and $v_m = y_m - x$. Then $v_n - v_m = y_n - y_m$ and $v_n + v_m = (y_n + y_m) - 2x$. From parallelogram equality, we have

$$\begin{aligned} \|v_n - v_m\|^2 &= 2(\|v_n\|^2 + \|v_m\|^2) - \|v_n + v_m\|^2 \\ \implies \|y_n - y_m\|^2 &= 2(\|y_n - x\|^2 + \|y_m - x\|^2) - \|y_n + y_m - 2x\|^2 \\ &\leq 2(\delta_n^2 + \delta_m^2) - (2\delta)^2 \end{aligned}$$

Given $\varepsilon > 0$, there exists sufficiently large natural number N such that $4(\delta_N^2 - \delta^2) < \varepsilon$. Clearly, $\forall n > N$, $\delta < \delta_n < \delta_N$ and $\delta_n^2 - \delta^2 < \delta_N^2 - \delta^2$. Thus, the sequence $\{y_n\}$ is Cauchy. Since M is complete and $\{y_n\}$ is Cauchy, $y_n \rightarrow y \in M$. Therefore, there do exist $y \in M$ such that $d(x, M) = d(x, y) = \|x - y\| = \delta$.

Step 2 : Uniqueness of minimising vector

Let $y, y_0 \in M$ such that $\|x - y\| = \|x - y_0\| = \delta$.

$$\|y - y_0\|^2 = \|y - x - y_0 + x\|^2$$

Let $v = y - x$ and $w = y_0 - x$

$$\|y - y_0\|^2 = \|v - w\|^2$$

By parallelogram equality, we have

$$\begin{aligned} \|v - w\|^2 &= 2\|v\|^2 + 2\|w\|^2 - \|v + w\|^2 \\ \Rightarrow \|y - y_0\|^2 &= 2\|y - x\|^2 + 2\|y_0 - x\|^2 - \|y + y_0 - 2x\|^2 \\ &\leq 2\delta^2 + 2\delta^2 - 4\delta^2 = 0 \quad \text{since } d\left(x, \frac{y + y_0}{2}\right) \geq \delta \end{aligned}$$

Thus, $y = y_0$. Therefore, the (distance) minimising vector (in M) is unique for any $x \in X$. \square

In Banach spaces, there may not exist such a point or there may exist many points at minimal distance !

Definitions 14.3.9 (minimising vector). Let X be a Hilbert space. Let $x \in X$ and $M \subset X$ be a complete, convex subset. Then minimising vector of M is a unique vector $y \in M$ such that distance from x to y is minimal among other points in M .

Lemma 14.3.10 (Orthonality of minimising vector). Let X be a Hilbert space. Let Y be a complete, convex subspace of X . Let $x \in X$ and $y \in Y$ be its minimising vector. Then $z = x - y$ is orthogonal to Y .

Proof. Suppose $z \perp Y$ is false.

Then there exists $y_1 \in Y$ such that $\langle y_1, z \rangle = \beta \neq 0$. Let $\alpha \in K$. Then,

$$\begin{aligned} \|z - \alpha y_1\|^2 &= \langle z - \alpha y_1, z - \alpha y_1 \rangle \\ &= \langle z, z \rangle - \bar{\alpha} \langle z, y_1 \rangle - \alpha \langle y_1, z \rangle + \alpha \bar{\alpha} \langle y_1, y_1 \rangle \end{aligned}$$

When $\alpha = \frac{\beta}{\langle y_1, y_1 \rangle} = \frac{\langle z, y_1 \rangle}{\langle y_1, y_1 \rangle}$. Then $\bar{\alpha} \beta = \frac{\bar{\beta} \beta}{\langle y_1, y_1 \rangle} = \frac{|\beta|^2}{\|y_1\|^2}$. Then,

$$\|z - \alpha y_1\|^2 = \|z\|^2 - \frac{|\beta|^2}{\|y_1\|^2} < \delta^2$$

This is a contradiction, since $\|z - \alpha y_1\| = \|x - y - \alpha y_1\| = \|x - y_2\| \geq \delta$ where $y_2 = y + \alpha y_1 \in Y$. Therefore, $z \perp Y$ is true. \square

Definitions 14.3.10 (direct sum). A vector space X is a direct sum of two subspaces Y, Z if each $x \in X$ has a unique representation of the form $x = y + z$ where $y \in Y$ and $z \in Z$. And we write, $X = Y \oplus Z$.

Definitions 14.3.11 (algebraic complement). Let $X = Y \oplus Z$. Then, Y, Z are algebraic complements.

Definitions 14.3.12 (orthogonal complement). Let $X = Y \oplus Z$ where $Z = Y^\perp = \{x \in X : x \perp Y\}$. Then Y, Z are orthogonal complements.

Theorem 14.3.11 (direct sum). Let Y be any closed subspace of a Hilbert space H . Then $H = Y \oplus Z$ where $Z = Y^\perp$.

Proof. Let H be a Hilbert space. Let Y be a closed subspace of H . Then Y is complete. And Y is convex, since Y is a subspace.

Since Y is a complete, convex subset of an inner product space H , for each $x \in H$ there exists a unique minimising vector $y \in Y$ such that $x - y \perp Y$. Define $Z = \{x \in H : (x - y) \perp Y\}$. If $z \in Z$, then there exists $x \in H$ and $y \in Y$ such that $z = x - y \in Y^\perp$. Thus each $x \in H$ has a representation, $x = y + z$ where $y \in Y$ and $z = x - y \in Y^\perp$.

We know that $Y \cap Y^\perp = \{0\}$, the zero subspace of H . Suppose $x = y_1 + z_1$ and $x = y_2 + z_2$ where $y_1, y_2 \in Y$ and $z_1, z_2 \in Y^\perp$. Then $y_1 - y_2 = z_2 - z_1$. Since Y is a subspace, $y_1 - y_2 \in Y$ and $z_2 - z_1 \in Y^\perp$. Thus, $y_1 - y_2 = z_2 - z_1 \in Y \cap Y^\perp = \{0\}$. That is, $y_1 - y_2 = 0$. Therefore, the representation $x = y + z$ is unique. \square

Definitions 14.3.13 (Projection). Let H be an inner product space. Let Y be a closed subspace of H . We have, $H = Y \oplus Y^\perp$. Then, the function $P : H \rightarrow Y$ defined by $P(x) = y$ is an orthogonal projection of H onto Y .

Properties of Projection

1. P is idempotent. That is, $P^2 = P$
Let $x \in H$ such that $x = y + z$. We have, $P(x) = y$. Then, $P^2(x) = P(y) = P(y + 0) = y = P(x)$.
2. $P|_Y$ is the identity operator on Y .
Let $y \in Y \subset H$. Then $y = y + 0$ and $P(y) = y$.

Lemma 14.3.12 (Null Space). Let $H = Y \oplus Y^\perp$. Let $P : H \rightarrow Y$ be the projection defined by $P(x) = y$. Then $Y^\perp = \mathcal{N}(P)$.

Proof. Let $H = Y \oplus Y^\perp$. Let $x \in H$. Then $x = y + z$ where $y \in Y$ and $z \in Y^\perp$. Let $P : H \rightarrow Y$ be the projection defined by $P(x) = y$.

Let $z \in Y^\perp$. Then $z = 0 + z$ and $P(z) = 0$. Thus, $Y^\perp \subset \mathcal{N}(P)$.

Let $z \in \mathcal{N}(P)$. Then, $P(z) = 0$. That is, $z = 0 + z$ where $z \in Y^\perp$. Thus, $\mathcal{N}(P) \subset Y^\perp$. \square

Definitions 14.3.14 (Annihilator). Let M be a nonempty subset of an inner product space H . Annihilator of M is given by, $M^\perp = \{x \in H : x \perp M\}$.

Remark. Suppose $M \subset V$. Then $V^\perp \subset M^\perp$.

Remark. Let M be a nonempty subset of an inner product space H . Then, annihilator M^\perp is a subspace of H .

Proof. Let $v \in M$. Let $x, y \in M^\perp$. Then $\langle x, v \rangle = 0$ and $\langle y, v \rangle = 0$. Let $\alpha, \beta \in K$. Thus, $\langle \alpha x + \beta y, v \rangle = \alpha \langle x, v \rangle + \beta \langle y, v \rangle = 0$. Clearly, $\alpha x + \beta y \in M^\perp$.

Also, we know that $\langle 0, v \rangle = 0$. Thus, $0 \in M^\perp$. Therefore M^\perp is a subspace of H . \square

Remark. Let M be a nonempty subset of an inner product space H . Then,

$$M \subset M^{\perp\perp}$$

Proof. Let $v \in M$. Then for any $x \in M^\perp$, we have $\langle x, v \rangle = 0$. Thus, $v \in (M^\perp)^\perp = M^{\perp\perp}$.

$$v \in M \implies v \perp M^\perp \implies v \in M^{\perp\perp}$$

Therefore, $M \subset M^{\perp\perp}$. □

Lemma 14.3.13 (closed subspace). *Let Y be a closed subspace of a Hilbert space H . Then, $Y = Y^{\perp\perp}$.*

Proof. We know that $Y \subset Y^{\perp\perp}$. It is enough to prove that $Y^{\perp\perp} \subset Y$. Let $x \in Y^{\perp\perp}$. Then $x \perp Y^\perp$. If $x = y + z$ where $y \in Y$ and $z \in Y^\perp$. Clearly, $y \in Y^{\perp\perp}$ and $z = x - y \in Y^{\perp\perp}$. Then $z = 0$ since $z \in Y^\perp \cap Y^{\perp\perp} = \{0\}$. Therefore, $x = y + 0 = y \in Y$. Thus, $Y = Y^{\perp\perp}$. □

Remark. Let Y be a closed subspace of Hilbert space H . Then, $H = Y \oplus Z$ where $Z = Y^\perp$. And, $H = Z \oplus Z^\perp$ where $Z^\perp = Y^{\perp\perp} = Y$. And $P_Z : H \rightarrow Z$ where $P_Z(x) = z$ where $x = y + z$ is another projection from H onto Z .

Lemma 14.3.14 (dense). *Let M be a nonempty subset of a Hilbert space H . The span of M is dense in H if and only if $M^\perp = \{0\}$.*

Proof. Suppose $\text{span } M$ is dense in H . Suppose M^\perp is nonempty. Let $x \in M^\perp$, ($x \neq 0$). Then [there exists a sequence](#) $\{x_n\}$ in $\text{span } M$ such that $x_n \rightarrow x$. Since $x \in M^\perp$ and $x_n \in \text{span } M$, we have $\langle x_n, x \rangle = 0$. We know that, [inner product is continuous](#). Thus, $\langle x_n, x \rangle \rightarrow \langle x, x \rangle = 0$. Clearly, $x = 0$.

Suppose $M^\perp = \{0\}$. Define $V = \text{span } M$. Let $x \in V^\perp$, then $x \in M^\perp$ since $M \subset V$. Clearly, $x = 0$. Therefore, $V^\perp = \{0\}$. Consider the closed subspace \overline{V} . We know that $V \subset \overline{V}$. Thus, $\overline{V}^\perp \subset V^\perp = \{0\}$. Thus, $H = \overline{V} \oplus \{0\} = \overline{V}$. Therefore, $\text{span } M$ is dense in H . □

14.3.4 Orthonormal sets and sequences

Definitions 14.3.15 (orthogonal). Let H be an inner product space. Subset M of H is an orthogonal set if its elements are pairwise orthogonal.

$$\forall x, y \in M, x \perp y \quad (x \neq y)$$

Definitions 14.3.16 (orthonormal). Let H be an inner product space. Subset M of H is orthonormal set if it is an orthogonal set with elements of unit norm.

$$\langle x, y \rangle = \begin{cases} 0 & x \neq y \\ 1 & x = y \end{cases}$$

Remark. Let $M = \{x_1, x_2, \dots, x_n\}$ be an orthogonal set. Then by Pythagorean theorem,

$$\|x_1 + x_2\|^2 = \|x_1\|^2 + \|x_2\|^2 \quad \text{since } x_1 \perp x_2$$

By finite mathematical induction,

$$\|x_1 + x_2 + \dots + x_n\|^2 = \|x_1\|^2 + \|x_2\|^2 + \dots + \|x_n\|^2$$

Lemma 14.3.15. *Orthonormal sets are linearly independent.*

Proof. Let $E = \{e_k : k \in I\}$ be an orthonormal set. Let S be a finite subset of the index set I . Suppose $\sum_{k \in S} \alpha_k e_k = 0$ where $\alpha_k \in K$. Fix $j \in S$. Then,

$$\langle 0, e_j \rangle = \left\langle \sum_{k \in S} \alpha_k e_k, e_j \right\rangle = \sum_{k \in S} \alpha_k \langle e_k, e_j \rangle = \sum_{k \in S} \alpha_k \delta_{k,j} = \alpha_j$$

Clearly, $\alpha_j = 0, \forall j \in S$. Thus, zero vector can be expressed as a finite linear combination of vectors only if all the scalars are zero. Therefore, M is linearly independent. \square

Examples of Orthonormal Sets in different Hilbert Spaces

1. \mathbb{R}^3 , $M = \{(1, 0, 0), (0, 1, 0), (0, 0, 1)\}$
2. l^2 , $M = \{e_j\}$ where $e_j = \delta_{n,j}$
3. $L^2[0, 2\pi]$, $M = \left\{ \frac{\cos nt}{\sqrt{\pi}} : n = 0, 1, 2, \dots \right\}$
4. $L^2[0, 2\pi]$, $M = \left\{ \frac{\sin nt}{\sqrt{\pi}} : n \in \mathbb{N} \right\}$

Theorem 14.3.16 (Bessel Inequality). *Let sequence $\{e_k\}$ be an orthonormal sequence in an inner product space X . Let $x \in X$. Then,*

$$\sum_{k=1}^{\infty} |\langle x, e_k \rangle|^2 \leq \|x\|^2$$

Proof. Let $\{e_k\}$ be an orthonormal sequence in an inner product space X . Let $x \in X$. Let $Y_n = \text{span} \{e_1, e_2, \dots, e_n\}$. Then, there exists $y \in Y_n$ such that

$$y = \sum_{k=1}^n \langle x, e_k \rangle e_k$$

Define $z = x - y$. Then,

$$\begin{aligned} \langle z, y \rangle &= \langle x - y, y \rangle \\ &= \langle x, y \rangle - \langle y, y \rangle \\ &= \left\langle x, \sum_{k=1}^n \langle x, e_k \rangle e_k \right\rangle - \|y\|^2 \\ &= \sum_{k=1}^n \overline{\langle x, e_k \rangle} \langle x, e_k \rangle - \|y\|^2 \\ &= \sum_{k=1}^n |\langle x, e_k \rangle|^2 - \sum_{k=1}^n |\langle x, e_k \rangle|^2 = 0 \end{aligned}$$

Thus, $z \perp y$. Therefore, we have $x = y + z$ where $y \in Y_n$ and $z \perp y$.

By Pythagorean theorem,

$$\|x\|^2 = \|y\|^2 + \|z\|^2$$

Therefore,

$$0 \leq \|z\|^2 = \|x\|^2 - \|y\|^2 = \|x\|^2 - \sum_{k=1}^n |\langle x, e_k \rangle|^2$$

Thus,

$$\sum_{k=1}^n |\langle x, e_k \rangle|^2 \leq \|x\|^2$$

The series is monotonic increasing and bounded by $\|x\|^2$. Therefore, converges and the sum is bounded by $\|x\|^2$.

$$\lim_{n \rightarrow \infty} \sum_{k=1}^n |\langle x, e_k \rangle|^2 = \sum_{k=1}^{\infty} |\langle x, e_k \rangle|^2 \leq \|x\|^2$$

□

Remark (Gram-Schmidt Process). Let X be a finite dimensional inner product space. Let set $\{x_1, x_2, \dots, x_n\}$ be a set of n linearly independent vectors in X . Then an orthonormal set of vectors can be obtained by the following process.

Define $e_1 = x_1/\|x_1\|$. Then, $x_2 = \langle x_2, e_1 \rangle e_1 + v_2$. Thus, $v_2 = x_2 - \langle x_2, e_1 \rangle e_1$ and $v_2 \perp e_1$. Define $e_2 = v_2/\|v_2\|$. Continuing like this, we get

$$v_n = x_n - \sum_{k=1}^{n-1} \langle x_n, e_k \rangle e_k$$

such that $v_n \perp \text{span} \{e_1, e_2, \dots, e_{n-1}\}$. And $e_n = v_n/\|v_n\|$.

14.3.5 Series related to Orthonormal sequences and sets

Remark. Define $u_n(t) = \cos nt$ and $v_n(t) = \sin nt$. Then

$$\langle u_j, u_k \rangle = \int_0^{2\pi} \cos jt \cos kt \, dt = \begin{cases} 0 & j \neq k \\ \pi & j = k \neq 0 \\ 2\pi & j = k = 0 \end{cases}$$

$$\langle v_j, v_k \rangle = \int_0^{2\pi} \sin jt \sin kt \, dt = \begin{cases} 0 & j \neq k \\ \pi & j = k \neq 0 \\ 2\pi & j = k = 0 \end{cases}$$

$$\langle u_j, v_k \rangle = \int_0^{2\pi} \cos jt \sin kt \, dt = \frac{1}{2} \int_0^{2\pi} \sin(k+j)t + \sin(k-j)t \, dt = 0$$

Clearly, $u_j \perp u_k$ ($j \neq k$), $v_j \perp v_k$ ($j \neq k$) and $\forall j, k, u_j \perp v_k$.

Theorem 14.3.17 (Fourier). *Any periodic, (reasonably) continuous function may be expressed as series of sine and cosine terms.*

Remark. We may write $x \in C[0, 2\pi]$ as $x = y + z$ where $y, z \in C[0, 2\pi]$ such that

$$y(t) = \sum_{k=0}^{\infty} \frac{\langle x, u_k \rangle}{\pi} u_k = \sum_{k=0}^{\infty} a_k \cos kt \text{ where } a_k = \frac{1}{\pi} \int_0^{2\pi} x(t) \cos kt \, dt$$

$$z(t) = \sum_{k=0}^{\infty} \frac{\langle x, v_k \rangle}{\pi} v_k = \sum_{k=0}^{\infty} b_k \sin kt \text{ where } b_k = \frac{1}{\pi} \int_0^{2\pi} x(t) \sin kt \, dt$$

The coefficients a_k, b_k 's are the **Fourier Coefficients** of $x(t)$.

Challenge 10. Find the orthonormal series that gives the Fourier representation of $x(t)$? [Apostol, 1973]

Theorem 14.3.18 (convergence). *Let $\{e_k\}$ be an orthonormal sequence in a Hilbert space H . Then,*

1. *The series $\sum \alpha_k e_k$ converges if and only if $\sum |\alpha_k|^2$ converges.*
2. *If $\sum \alpha_k e_k$ converges, then coefficients α_k are the Fourier coefficients $\langle x, e_k \rangle$ of the sum x .*

$$x = \sum_{k=1}^{\infty} \langle x, e_k \rangle e_k$$

3. *For any $x \in H$, the series $\sum \langle x, e_k \rangle e_k$ converges.*

Proof. Consider the series $\sum \alpha_k e_k$ and the sequence of partial sums $\{s_n\}$.

$$s_n = \alpha_1 e_1 + \alpha_2 e_2 + \cdots + \alpha_n e_n$$

Similarly, consider the series $\sum |\alpha_k|^2$ and the sequence of partial sums $\{\sigma_n\}$.

$$\sigma_n = |\alpha_1|^2 + |\alpha_2|^2 + \cdots + |\alpha_n|^2$$

$$\begin{aligned} \|s_n - s_m\|^2 &= \|\alpha_{m+1} e_{m+1} + \alpha_{m+2} e_{m+2} + \cdots + \alpha_n e_n\|^2 \\ &= |\alpha_{m+1}|^2 + |\alpha_{m+2}|^2 + \cdots + |\alpha_n|^2 \text{ since } \langle e_n, e_m \rangle = \delta_{n,m} \\ &= |\sigma_n - \sigma_m| \end{aligned}$$

Clearly, $\{s_n\}$ is Cauchy if and only if $\{\sigma_n\}$ is Cauchy. We know that, \mathbb{R}, H are complete spaces and every Cauchy sequence in these spaces converges. Therefore, $\sum \alpha_k e_k$ converges if and only if $\sum |\alpha_k|^2$ converges.

Suppose $\sum \alpha_k e_k$ converges to sum x . Consider the sequence of partial sum $\{s_n\}$. Then, $s_n \rightarrow x$. Fix e_k ,

$$\langle s_n, e_k \rangle = \left\langle \sum_{j=1}^n \alpha_j e_j, e_k \right\rangle = \sum_{j=1}^n \langle \alpha_j e_j, e_k \rangle = \sum_{j=1}^n \alpha_j \delta_{j,k} = \alpha_k$$

Since inner product is continuous $\langle s_n, e_j \rangle \rightarrow \langle x, e_j \rangle$ as $n \rightarrow \infty$. Thus, $\alpha_j = \langle x, e_j \rangle$. Therefore,

$$x = \sum_{k=1}^{\infty} \alpha_k e_k = \sum_{k=1}^{\infty} \langle x, e_k \rangle e_k$$

Let H be a Hilbert space. Let $x \in H$ and $\{e_k\}$ be an orthonormal sequence in H . Define $\alpha_k = \langle x, e_k \rangle$. By Bessel Inequality,

$$\sum_{k=1}^{\infty} |\alpha_k|^2 = \sum_{k=1}^{\infty} |\langle x, e_k \rangle|^2 \leq \|x\|^2$$

Therefore, $\sum |\alpha_k|^2$ converges. We know that, if $\sum |\alpha_k|^2$ converges then $\sum \alpha_k e_k$ converges and the sum is given by

$$x = \sum_{k=1}^{\infty} \alpha_k e_k$$

□

Lemma 14.3.19 (Fourier coefficients). *Any x in an inner product space has at most countably many non-zero Fourier coefficients with respect to an orthonormal family $\{e_k : k \in I\}$ in X .*

Proof. Step 1 :

Let $x \in X$. Suppose $x = \sum \alpha_k e_k$. Without loss of generality, consider the rearrange of $\{e_k\}$ such that $x = \sum \langle x, e_k \rangle e_k$. Then,

$$\sum_{k=1}^n |\langle x, e_k \rangle|^2 \leq \|x\|^2, \forall n$$

Let $m \in \mathbb{N}$. Then the number of e_k 's such that $|\langle x, e_k \rangle| > \frac{1}{m}$ is less than $m\|x\|^2$. Thus the number of e_k 's such that $|\langle x, e_k \rangle| > 0$ is countable since it is a countable union of finite sets.

Step 2 : Independent of rearrangements

It remains to prove that, The series representation is independent of rearrangement of elements in the orthonormal sequence $\{e_k\}$.

Let $\{w_k\}$ be a rearrangement of $\{e_k\}$. Then, there exists a bijection $\phi : w \rightarrow e$ given by $\phi(w_j) = e_k$ where $j \rightarrow k$ is a bijection on \mathbb{N} .

Define $\alpha_k = \langle x, e_k \rangle$, $\beta_j = \langle x, w_j \rangle$,

$$x_1 = \sum_{n=1}^{\infty} \alpha_n e_n \quad \text{and} \quad x_2 = \sum_{n=1}^{\infty} \beta_n w_n$$

Then,

$$\alpha_j = \langle x, e_j \rangle = \langle x_1, e_j \rangle \quad \text{and} \quad \beta_k = \langle x, w_k \rangle = \langle x_2, w_k \rangle$$

$$\begin{aligned} \langle x_1 - x_2, e_j \rangle &= \langle x_1, e_j \rangle - \langle x_2, e_j \rangle \\ &= \langle x, e_j \rangle - \langle x_2, e_j \rangle = 0 \end{aligned}$$

Similarly, $\langle x_1 - x_2, w_k \rangle = 0$.

$$\begin{aligned} \|x_1 - x_2\|^2 &= \left\langle x_1 - x_2, \sum \alpha_n e_n - \sum \beta_k w_k \right\rangle \\ &= \sum \bar{\alpha}_j \langle x_1 - x_2, e_j \rangle - \sum \bar{\beta}_k \langle x_1 - x_2, w_k \rangle = 0 \end{aligned}$$

Clearly, $x_1 = x_2$. Thus, the sum is the same for representation of x with respect to any rearrangement of $\{e_k\}$. \square

14.3.6 Total Orthonormal sets and sequences

Definitions 14.3.17 (total orthonormal). A subset M in a vector X is total if the span of M is dense in X . Total orthonormal is a subset which is total and orthonormal in inner product space X .

Every non-empty Hilbert space has a total orthonormal subset.

Proof. It is included in module IV. \square

In a Hilbert space H , total orthonormal subsets have same cardinality.

Proof. It is out of our scope. This cardinality is the Hilbert dimension or orthonormal dimension of H . \square

Theorem 14.3.20 (duality). Let M a subset of an inner product space X . Then

1. If M total in X , then there does not exists a non-zero $x \in X$ which is orthogonal to every element of M .

$$x \perp M \implies x = 0$$

2. If X is complete and X has no non-zero element orthogonal to M , then M is total in X .

If M is total in X , then $M^\perp = \{0\}$. Conversely, if X is complete and $M^\perp = \{0\}$, then M is total in X .

Proof. Let H be the completion of inner product space X . Then, X is dense in H . And $\text{span } M$ is dense in X since M is total. Thus, $\text{span } M$ is dense in H .

Thus, orthogonal complement of M in H is $\{0\}$. Thus, if $x \in X$ and $x \perp M = 0$. Then, $x \in H$ and $x = y + 0 \in M \oplus M^\perp$ and $x \perp M = y = 0$. Therefore, $x = 0$.

Let X be a Hilbert space. Suppose X has no non-zero element orthogonal to M . Then, $M^\perp = \{0\}$. By lemma, $\text{span } M$ is dense in X since $M^\perp = \{0\}$. Therefore, M is total in X . \square

Theorem 14.3.21. An orthonormal subset M is total in a Hilbert space if and only if Parsevals relation is true for any $x \in H$.

$$\sum_k |\langle x, e_k \rangle|^2 = \|x\|^2$$

Proof. Suppose Parsevals relation is true for each $x \in H$. Suppose M is not total in Hilbert space H . Then, there exist $x \in H$ such that $x \perp M$ and $x \neq 0$. Since $x \in M^\perp$, we have $\forall k, \langle x, e_k \rangle = 0$, then $\|x\|^2 = 0$ by Parsevals relation. And $x = 0$ since $\|x\| = 0$. This is a contradiction. Therefore, M is total in H .

Suppose M is total in H . Then $M^\perp = \{0\}$. Let $x \in H$. We know that, there are at most countably many non-zero Fourier coefficients for any $x \in H$. Arrange non-zero Fourier coefficients of x in a (finite) sequence $\{\langle x, e_k \rangle\}$.

Define

$$y = \sum_k \langle x, e_k \rangle e_k$$

Case 1 : $e_k \in M$ such that $\langle x, e_k \rangle \neq 0$

Fix an e_j such that $\langle x, e_j \rangle \neq 0$. Then,

$$\begin{aligned} \langle x - y, e_j \rangle &= \langle x, e_j \rangle - \langle y, e_j \rangle \\ &= \langle x, e_j \rangle - \left\langle \sum_k \langle x, e_k \rangle e_k, e_j \right\rangle \\ &= \langle x, e_j \rangle - \sum_k \langle x, e_k \rangle \langle e_k, e_j \rangle \\ &= \langle x, e_j \rangle - \sum_k \langle x, e_k \rangle \delta_{j,k} \\ &= \langle x, e_j \rangle - \langle x, e_j \rangle = 0 \end{aligned}$$

Thus, if $\langle x, e_j \rangle \neq 0$, then $x - y \perp e_j$.

Case 2 : $v \in M$ such that $\langle x, v \rangle = 0$

Fix $v \in M$ such that $\langle x, v \rangle = 0$. Then,

$$\langle x - y, v \rangle = \langle x, v \rangle - \langle y, v \rangle = \langle x, v \rangle - \sum_k \langle x, e_k \rangle \langle e_k, v \rangle = 0$$

since $\langle x, v \rangle = 0$ and $\forall k, \langle e_k, v \rangle = 0$.

Clearly, from cases 1 & 2, we have $x - y \perp M$. Thus, $x - y \in M^\perp = \{0\}$. Therefore, $x = y$.

$$\begin{aligned} \|x\|^2 &= \left\langle \sum_k \langle x, e_k \rangle e_k, \sum_m \langle x, e_m \rangle e_m \right\rangle \\ &= \sum_k \langle x, e_k \rangle \overline{\langle x, e_k \rangle} \\ &= \sum_k |\langle x, e_k \rangle|^2 \end{aligned}$$

Thus, Parsevals relation is true for any $x \in H$. □

Definitions 14.3.18 (Orthogonal Projection). The orthogonal projection of x is the best approximation of x with respect to an orthonormal set $\{e_k\}$.

Theorem 14.3.22. *Let H be a Hilbert space. Then,*

1. *If H is separable, then every orthonormal subset in H is countable.*
2. *If H contains an orthonormal sequence which is total in H , then H is separable.*

Proof. Let H be a Hilbert space. Suppose H is separable. Then, H has a dense subset, say B . And H has an orthonormal subset which may be obtained from a Hamel basis by Gram-Schmidt process, say M .

Let $x, y \in M$. Then,

$$\|x - y\|^2 = \langle x - y, x - y \rangle = \langle x, x \rangle + \langle y, y \rangle = 2$$

Thus, $d(x, y) = \sqrt{2}$ for any $x, y \in M$. Thus, x, y have disjoint neighbourhoods N_x, N_y respectively. We know that, the dense set B intersects with any neighbourhood in H . Let $b \in N_x \cap B$ and $\tilde{b} \in N_y \cap B$. Clearly, $b \neq \tilde{b}$ since $N_x \cap N_y = \emptyset$.

Suppose M is uncountable. Then there exists uncountably many such pair (b, \tilde{b}) of elements in B for each pair (x, y) of elements in M . This is a contradiction since B is countable. Therefore, M is countable.

Let sequence $\{e_k\}$ be a total orthonormal sequence in H . Let A be the set of all linear combinations of the form,

$$\gamma_1^{(n)} e_1 + \cdots + \gamma_n^{(n)} e_n$$

where $\gamma_k^{(n)} = a_k^{(n)} + i b_k^{(n)}$ and $a_k^{(n)}, b_k^{(n)}$ are rationals. Then, A is countable.

It is enough to show that A is dense in H . Let $x \in H$ and $\varepsilon > 0$. Since the sequence $\{e_k\}$ is total in H , there exists a natural number n such that $Y_n = \text{span} \{e_1, e_2, \dots, e_n\}$ contains a point y whose distance from x is less than $\varepsilon/2$. Then,

$$y = \sum_{k=1}^n \langle x, e_k \rangle e_k \quad \text{and} \quad \|x - y\| < \frac{\varepsilon}{2}$$

Therefore,

$$\left\| x - \sum_{k=1}^n \langle x, e_k \rangle e_k \right\| < \frac{\varepsilon}{2}$$

□

Definitions 14.3.19 (isomorphism). Let H, \tilde{H} be Hilbert spaces. A bijection $T : H \rightarrow \tilde{H}$ preserving inner product is an isomorphism. That is,

$$\langle Tx, Ty \rangle = \langle x, y \rangle$$

Theorem 14.3.23. *Two Hilbert space over the same field are isomorphic if and only if they have the same Hilbert dimension.*

Proof. Let $T : H \rightarrow \tilde{H}$ be an isomorphism. Let M be a total, orthonormal subset of H . Then $T(M)$ is a total, orthonormal subset of \tilde{H} . Thus, these sets have same cardinality. That is, they have same Hilbert dimension.

Suppose H, \tilde{H} are two Hilbert spaces over the same field K . Suppose they have same Hilbert dimension. Then they have total, orthonormal subsets $M_1 = \{e_k\}$ and $M_2 = \{w_j\}$ respectively. Let $x \in H$. Then,

$$x = \sum_{k \in I} \langle x, e_k \rangle e_k$$

where $\langle x, e_k \rangle \neq 0$ for at most countably many e_k 's. And from Bessel inequality, we know that this series converges. Define $T : H \rightarrow \tilde{H}$ such that $e_k \rightarrow w_k$ by rearranging $\{w_j\}$ if necessary.^{†15} Consider,

$$\begin{aligned} T(x) &= T\left(\sum_{k \in I} \langle x, e_k \rangle e_k\right) \\ &= \sum_{k \in I} \langle x, e_k \rangle T(e_k) \\ &= \sum_{k \in I} \langle x, e_k \rangle w_k \end{aligned}$$

This series converges, thus $T(x) = \tilde{x} \in \tilde{H}$.

And T preserves norms since,

$$\begin{aligned} \|\tilde{x}\|^2 &= \|T(x)\|^2 \\ &= \sum_{k \in I} \langle x, e_k \rangle w_k \\ &= \sum_{k \in I} |\langle x, e_k \rangle|^2 = \|x\|^2 \end{aligned}$$

Thus, $\|x - y\| = \|T(x - y)\| = \|Tx - Ty\| = 0$. Therefore, T is injective.

Suppose $\tilde{x} = \sum_{k \in I} \alpha_k w_k$. Then $\sum |\alpha_k|^2 < \infty$ by Bessel inequality. Thus, $x = \sum_{k \in I} \alpha_k e_k \in H$. Therefore, T is surjective.

And T preserves inner product since,

$$\begin{aligned} \langle \tilde{x}, \tilde{y} \rangle &= \langle Tx, Ty \rangle \\ &= \left\langle \sum_{k \in I} \langle x, e_k \rangle w_k, \sum_{j \in I} \langle y, e_j \rangle w_j \right\rangle \\ &= \sum_{k \in I} \langle x, e_k \rangle \overline{\langle y, e_k \rangle} \end{aligned}$$

¹⁵We know that, for Hilbert spaces the representation exists against any rearrangement of orthonormal sequence.

$$\begin{aligned}
&= \left\langle \sum_{k \in I} \langle x, e_k \rangle e_k, \sum_{j \in I} \langle y, e_j \rangle e_j \right\rangle \\
&= \langle x, y \rangle
\end{aligned}$$

□

14.3.8 Representation of Functionals on Hilbert spaces

Theorem 14.3.24 (Riesz). *Every bounded linear functional f on a Hilbert space H can be represented in terms of inner product,*

$$f(x) = \langle x, z \rangle$$

where z is uniquely determined by f and $\|z\| = \|f\|$.

Proof. Let f be a bounded, linear functional on H . If $f = 0$ or $x = 0$, then the result is trivial.

Step 1 : f has a representation

Suppose $f \neq 0$. If $\mathcal{N}(f) = H$. Then $f(x) = \langle x, z \rangle = 0$ for any $x \in H$ which is a contradiction since $f \neq 0$. Thus, $\mathcal{N}(f) \neq H$ and $\mathcal{N}(f)^\perp \neq \emptyset$.

Let $z_0 \in \mathcal{N}(f)^\perp$. Let $x \in H$. Define $v = f(x)z_0 - f(z_0)x$. Then $f(v) = f(x)f(z_0) - f(z_0)f(x) = 0$. Thus, $v \in \mathcal{N}(f)$.

$$\begin{aligned}
0 &= \langle v, z_0 \rangle \\
&= \langle f(x)z_0 - f(z_0)x, z_0 \rangle \\
&= f(x) \langle z_0, z_0 \rangle - f(z_0) \langle x, z_0 \rangle
\end{aligned}$$

Then,

$$f(x) = \frac{f(z_0)}{\langle z_0, z_0 \rangle} \langle x, z_0 \rangle = \langle x, z \rangle$$

Therefore, f has a representation, $f(x) = \langle x, z \rangle$ for any $x \in H$ where z_0 is a non-vanishing point of f and

$$z = \frac{f(z_0)}{\langle z_0, z_0 \rangle} z_0$$

Step 2 : z is unique

Suppose $f(x) = \langle x, z_1 \rangle = \langle x, z_2 \rangle$ for all $x \in H$. Then $\langle x, z_1 - z_2 \rangle = \langle x, z_1 \rangle - \langle x, z_2 \rangle = 0$. Put $x = z_1 - z_2$. Then $\langle z_1 - z_2, z_1 - z_2 \rangle = \|z_1 - z_2\|^2 = 0$. Thus, $z_1 = z_2$. And the representation $f(x) = \langle x, z \rangle$ is unique.

Step 3 : norm of f

Since f is bounded,

$$\|z\|^2 = \langle z, z \rangle = f(z) \leq \|f\| \|z\|$$

Thus, $\|z\| \leq \|f\|$.

Also we have,

$$|f(x)| \leq |\langle x, z \rangle| \leq \|x\| \|z\|$$

The inequality is true for any $x \in H$. Thus,

$$\|f\| = \sup_{x \in H} \frac{|f(x)|}{\|x\|} \leq \|z\|$$

Therefore, $\|f\| = \|z\|$ □

Lemma 14.3.25 (equality). *Let $\langle v_1, w \rangle = \langle v_2, w \rangle$ for each w in an inner product space X . Then $v_1 = v_2$. In particular, if $\langle v_1, w \rangle = 0$ for every $w \in X$, then $v_1 = 0$.*

Proof. Let v_1, v_2 be vectors in an inner product space X . Suppose $\langle v_1, w \rangle = \langle v_2, w \rangle$ for each $w \in X$. Then,

$$\langle v_1 - v_2, w \rangle = \langle v_1, w \rangle - \langle v_2, w \rangle = 0$$

Let $w = v_1 - v_2$. Then $\|v_1 - v_2\|^2 = \langle v_1 - v_2, v_1 - v_2 \rangle = 0$. Thus, $v_1 = v_2$.

Suppose $\langle v_1, w \rangle = 0$ for each $w \in X$. Then $\|v_1\|^2 = \langle v_1, v_1 \rangle = 0$. Therefore, $v_1 = 0$. □

Definitions 14.3.20 (sesquilinear form). Let X, Y be vector spaces over the same field K . Then a sesquilinear form is a mapping $h : X \times Y \rightarrow K$ such that

1. $h(x_1 + x_2, y) = h(x_1, y) + h(x_2, y)$
2. $h(x, y_1 + y_2) = h(x, y_1) + h(x, y_2)$
3. $h(\alpha x, y) = \alpha h(x, y)$
4. $h(x, \beta y) = \bar{\beta} h(x, y)$

Definitions 14.3.21 (bilinear form). Let X, Y be vector spaces over the same field K . Then a bilinear form is a mapping $h : X \times Y \rightarrow K$ such that

1. $h(x_1 + x_2, y) = h(x_1, y) + h(x_2, y)$
2. $h(x, y_1 + y_2) = h(x, y_1) + h(x, y_2)$
3. $h(\alpha x, y) = \alpha h(x, y)$
4. $h(x, \beta y) = \beta h(x, y)$

Definitions 14.3.22. Let X, Y be inner product spaces over the same field K . A function $h : X \times Y \rightarrow K$ is bounded if

$$|h(x, y)| \leq c \|x\| \|y\|$$

Definitions 14.3.23. Suppose $h : X \times Y \rightarrow K$ is bounded. Then,

$$\|h\| = \sup_{\substack{x \neq 0 \\ y \neq 0}} \frac{|h(x, y)|}{\|x\| \|y\|}$$

Theorem 14.3.26 (Riesz representation). *Let H_1, H_2 be Hilbert spaces over K . And function $h : H_1 \times H_2 \rightarrow K$ be a bounded, sesquilinear form. Then h has representation,*

$$h(x, y) = \langle Sx, y \rangle$$

where $S : H_1 \rightarrow H_2$ is a bounded, linear operator uniquely determined by h and $\|h\| = \|S\|$.

Proof. Step 1 : Constructon of S

Let $h : H_1 \times H_2 \rightarrow K$ be a bounded, sesquilinear form. Then $f(y) = \overline{h(x, y)}$ is a bounded, linear functional on H_2 . By Riesz's theorem for bounded functionals, f has an inner product representation, $f(y) = \langle y, z \rangle$. Thus, $h(x, y) = \langle z, y \rangle$. Define an operator $S : H_1 \rightarrow H_2$ such that $S(x) = z$. Then $h(x, y) = \langle Sx, y \rangle$.

Step 2 : S is linear

Let $x_1, x_2 \in H_1$, $y \in H_2$ and $\alpha, \beta \in K$. We have $\langle Sx, y \rangle = h(x, y)$. Thus,

$$\begin{aligned} \langle S(\alpha x_1 + \beta x_2), y \rangle &= h(\alpha x_1 + \beta x_2, y) \\ &= \alpha h(x_1, y) + \beta h(x_2, y) \\ &= \alpha \langle S(x_1), y \rangle + \beta \langle S(x_2), y \rangle \\ &= \langle \alpha S(x_1) + \beta S(x_2), y \rangle \end{aligned}$$

Clearly, $S(\alpha x_1 + \beta x_2) = \alpha S(x_1) + \beta S(x_2)$. Therefore, S is linear.

Step 3 : S is bounded

Since h is a bounded sesquilinear form. We have,

$$\begin{aligned} \|h\| &= \sup_{\substack{x \neq 0 \\ y \neq 0}} \frac{|h(x, y)|}{\|x\| \|y\|} \\ &= \sup_{\substack{x \neq 0 \\ Sx \neq 0}} \frac{|\langle Sx, y \rangle|}{\|x\| \|y\|} \end{aligned}$$

Suppose $y = Sx$. Then,

$$\begin{aligned} \|h\| &\geq \sup_{\substack{x \neq 0 \\ Sx \neq 0}} \frac{|\langle Sx, Sx \rangle|}{\|x\| \|Sx\|} \\ &\geq \sup_{\substack{x \neq 0 \\ Sx \neq 0}} \frac{\|Sx\|^2}{\|x\| \|Sx\|} = \|S\| \end{aligned}$$

Thus, S is bounded and $\|S\| \leq \|h\|$.

By Schwarz inequality, $|\langle Sx, y \rangle| \leq \|Sx\| \|y\|$. Thus,

$$\|h\| = \sup_{\substack{x \neq 0 \\ y \neq 0}} \frac{|\langle Sx, y \rangle|}{\|x\| \|y\|} \leq \sup_{\substack{x \neq 0 \\ y \neq 0}} \frac{\|Sx\| \|y\|}{\|x\| \|y\|} = \sup_{x \neq 0} \frac{\|Sx\|}{\|x\|} = \|S\|$$

Therefore $\|S\| = \|h\|$.

Suppose $T : H_1 \rightarrow H_2$ be a bounded, linear operator. Then

$$h(x, y) = \langle Sx, y \rangle = \langle Tx, y \rangle$$

Clearly, $Sx = Tx$ for each $x \in H_1$. Therefore, S is unique.

□

Module 4 - Hilbert-Adjoint Operator & Hahn-Banach Theorem

14.4 Hilbert Spaces

14.4.1 Hilbert-Adjoint Operator

Theorem 14.4.1 (Hilbert-Adjoint operator). *Let H_1, H_2 be Hilbert spaces. Let $T : H_1 \rightarrow H_2$ be any bounded, linear operator. Then there exists a unique, bounded, linear operator $T^* : H_2 \rightarrow H_1$ such that $\langle Tx, y \rangle = \langle x, T^*y \rangle$ and $\|T^*\| = \|T\|$.*

Proof. Let $T : H_1 \rightarrow H_2$ be a bounded, linear operator. Then, we have $h(y, x) = \langle y, Tx \rangle$ is a sesquilinear form on $H_2 \times H_1$, by the linearity of T and conjugate linearity of inner product.

By Schwarz inequality,

$$|h(y, x)| = |\langle y, Tx \rangle| \leq \|y\| \|Tx\| \leq \|T\| \|x\| \|y\|$$

Therefore, h is bounded and $\|h\| \leq \|T\|$.

Also we have,

$$\|h\| = \sup_{\substack{x \neq 0 \\ y \neq 0}} \frac{|\langle y, Tx \rangle|}{\|y\| \|x\|} \geq \sup_{\substack{x \neq 0 \\ Tx \neq 0}} \frac{|\langle Tx, Tx \rangle|}{\|Tx\| \|x\|} = \sup_{x \neq 0} \frac{\|Tx\|}{\|x\|} = \|T\|$$

Therefore, $\|h\| = \|T\|$.

By Riesz theorem, h has an inner product representation using a unique linear operator $S = T^*$ given by,

$$h(y, x) = \langle T^*y, x \rangle$$

where $\|T^*\| = \|h\|$. Therefore, $\langle y, Tx \rangle = \langle T^*y, x \rangle$ and $\|T\| = \|T^*\|$. \square

Definitions 14.4.1 (Hilbert-Adjoint operator). Let H_1, H_2 be Hilbert spaces. Let $T : H_1 \rightarrow H_2$ be a bounded, linear operators. Then Hilbert-Adjoint operator T^* of T is given by

$$T^* : H_2 \rightarrow H_1 \text{ such that } \forall x \in H_1, y \in H_2, \langle Tx, y \rangle = \langle x, T^*y \rangle$$

Lemma 14.4.2 (zero vector). *Let X, Y be inner product spaces. Let $Q : X \rightarrow Y$ be a bounded, linear operator. Then*

1. $Q = 0$ if and only if $\langle Qx, y \rangle = 0$ for all $x \in X$ and $y \in Y$.
2. If $Q : X \rightarrow X$, X is complex, and $\forall x \in X, \langle Qx, x \rangle = 0$, then $Q = 0$.

Proof. If $Q = 0$, then $Qx = 0$ and $\langle Qx, y \rangle = \langle 0, y \rangle = 0$.

Suppose $\forall x \in X, \forall y \in Y, \langle Qx, y \rangle = 0$. Since $\langle y, y \rangle = 0$ if and only if $y = 0$. Thus, zero is the only vector in Y which is orthogonal to entire Y . Then

$\forall x \in X, Qx = 0$. That is, $Q = 0$.

Suppose $Q : X \rightarrow X$ such that for any $x \in X, \langle Qx, x \rangle = 0$. Let $\alpha x + y \in X$. Then

$$\begin{aligned} 0 &= \langle Q(\alpha x + y), \alpha x + y \rangle \\ &= |\alpha|^2 \langle Qx, x \rangle + \langle Qy, y \rangle + \alpha \langle Qx, y \rangle + \bar{\alpha} \langle Qy, x \rangle \end{aligned}$$

Put $\alpha = 1$, we get $\langle Qx, y \rangle + \langle Qy, x \rangle = 0$.

Put $\alpha = i$, we get $\langle Qx, y \rangle - \langle Qy, x \rangle = 0$. Thus, $\langle Qx, y \rangle = 0$ for any $y \in X$. Therefore, $Qx = 0$ for any $x \in X$ and Thus $Q = 0$. \square

Theorem 14.4.3 (properties of Hilbert-Adjoint operators). *Let H_1, H_2 be Hilbert spaces. Let $\alpha \in K$. Let $S : H_1 \rightarrow H_2$ and $T : H_1 \rightarrow H_2$ be bounded linear operators. Then,*

1. $\langle T^*y, x \rangle = \langle y, Tx \rangle$
2. $(S + T)^* = S^* + T^*$
3. $(\alpha T)^* = \bar{\alpha}T^*$
4. $(T^*)^* = T$
5. $\|T^*T\| = \|TT^*\| = \|T\|^2$
6. $T^*T = 0 \iff T = 0$
7. $(ST)^* = T^*S^*$

Proof. 1. $\langle T^*y, x \rangle = \langle y, Tx \rangle$

$$\begin{aligned} \langle T^*y, x \rangle &= \overline{\langle x, T^*y \rangle} \\ &= \overline{\langle Tx, y \rangle} \\ &= \langle y, Tx \rangle \end{aligned}$$

2. $(S + T)^* = S^* + T^*$

$$\begin{aligned} \langle x, (S + T)^*y \rangle &= \langle (S + T)x, y \rangle \\ &= \langle Sx, y \rangle + \langle Tx, y \rangle \\ &= \langle x, S^*y \rangle + \langle x, T^*y \rangle \\ &= \langle x, (S^* + T^*)y \rangle \end{aligned}$$

3. $(\alpha T)^* = \bar{\alpha}T^*$

$$\begin{aligned} \langle x, (\alpha T)^*y \rangle &= \bar{\alpha} \langle x, T^*y \rangle \\ &= \bar{\alpha} \langle Tx, y \rangle \\ &= \langle (\bar{\alpha}T)x, y \rangle \\ &= \langle x, (\bar{\alpha}T)^*y \rangle \end{aligned}$$

4. $(T^*)^* = T$

$$\begin{aligned}\langle x, (T^*)^* y \rangle &= \langle T^* x, y \rangle \\ &= \overline{\langle y, T^* x \rangle} \\ &= \overline{\langle Ty, x \rangle} \\ &= \langle x, Ty \rangle\end{aligned}$$

5. $\|T^*T\| = \|TT^*\| = \|T\|^2$

$$\begin{aligned}\|Tx\|^2 &= \langle Tx, Tx \rangle \\ &= \langle Tx, (T^*)^* x \rangle \\ &= \langle T^*Tx, x \rangle\end{aligned}$$

By Schwarz inequality, $\langle T^*Tx, x \rangle \leq \|T^*Tx\| \|x\|$.

$$\begin{aligned}\|Tx\|^2 &\leq \|T^*Tx\| \|x\| \\ &\leq \|T^*T\| \|x\|^2\end{aligned}$$

Taking supremum of $\|x\| = 1$, we get $\|T\|^2 \leq \|T^*T\|$. Since T^*, T are bounded, linear operators, we have $\|T^*T\| \leq \|T^*\| \|T\|$. Also we have, $\|T^*\| = \|T\|$. Thus,

$$\|T\|^2 \leq \|T^*T\| \leq \|T^*\| \|T\| \leq \|T\|^2$$

Therefore, $\|T^*T\| = \|T\|^2$.

Replace T by T^* we get,

$$\|(T^*)^*T^*\| = \|T^*\|^2 = \|T\|^2$$

6. $T^*T = 0 \iff T = 0$

$$\|T\|^2 = \|T^*T\| = 0 \iff \|T\| = 0$$

7. $(ST)^* = T^*S^*$

$$\begin{aligned}\langle (ST)x, y \rangle &= \langle Tx, S^*y \rangle \\ \implies \langle x, (ST)^*y \rangle &= \langle x, (T^*S^*)y \rangle\end{aligned}$$

□

14.4.2 Self-Adjoint, Unitary and Normal Operators

Definitions 14.4.2. Let H be a Hilbert space. Let $T : H \rightarrow H$ be a bounded, linear operator.

- Then T is self-adjoint/Hermitian if $T^* = T$.

- Then T is unitary, if $T^* = T^{-1}$.
- Then T is normal, if $TT^* = T^*T$.

Remark. If T is unitary, then $T^*T = T^{-1}T = TT^{-1} = TT^*$. If T is self-adjoint, then $T^*T = T^2$. Therefore, if T is self-adjoint or unitary, then T is normal. And $2iI$ is a normal operator which is neither self-adjoint or unitary.

Remark. Let $T : \mathbb{C}^n \rightarrow \mathbb{C}^n$ be a linear operator. Let A, B be the matrices of transformations T, T^* with respect to the standard basis on \mathbb{C}^n . Then

$$\langle Tx, y \rangle = x^t A^t \bar{y} = \langle x, T^*y \rangle = x^t \bar{B}y$$

Thus, $B = \bar{A}^t$.

Remark. Suppose T is self-adjoint/Hermitian, then $A = \bar{A}^t$. ie, A is Hermitian. If K is real then A is symmetric since $A = A^t$.

Suppose T is unitary, then $A^{-1} = \bar{A}^t$. ie, A is unitary. If K is real, then A is orthogonal since $A^{-1} = A^t$.

Suppose T is normal, then $BA = AB$ where $B = \bar{A}^t$. ie, A is normal since A commutes with its conjugate transpose.

Theorem 14.4.4 (self-adjoint). *Let T be a bounded, linear operator on a Hilbert space H .*

1. *If T is self-adjoint, $\langle Tx, x \rangle$ is real for any $x \in H$.*
2. *If H is complex and $\langle Tx, x \rangle$ is real for all $x \in E$. Then T is self-adjoint.*

Proof. Suppose T is self-adjoint. Then $\langle Tx, y \rangle = \langle x, Ty \rangle$.

$$\langle Tx, x \rangle = \langle x, Tx \rangle = \overline{\langle Tx, x \rangle}$$

Therefore, $\langle Tx, x \rangle$ is real.

Suppose H is complex, Hilbert space. Suppose $\langle Tx, x \rangle$ is real for any $x \in H$. Then,

$$\langle Tx, x \rangle = \overline{\langle x, Tx \rangle} = \langle x, Tx \rangle$$

And,

$$\langle Tx, x \rangle = \langle x, T^*x \rangle$$

Thus,

$$\langle x, (T - T^*)x \rangle = \langle x, 0 \rangle = 0, \forall x \in H$$

Thus, $T - T^* = 0$ (by zero vector lemma). Therefore, $T = T^*$. \square

Theorem 14.4.5 (self-adjoint product). *The product of two self-adjoint, bounded operators S, T is self-adjoint if and only if S, T commute.*

Proof. We have, $(ST)^* = T^*S^* = TS$. Thus,

$$(ST)^* = ST \iff TS = ST$$

\square

Theorem 14.4.6. *Let T_n be a sequence of self-adjoint, bounded linear operators $T_n : H \rightarrow H$. Suppose $\{T_n\}$ converges to T with respect to the norm on $B(H, H)$. Then the limit operator T is a self-adjoint, bounded, linear operator on H .*

Proof. Let $T_n : H \rightarrow H$ be self-adjoint, bounded, linear operator for each $n \in \mathbb{N}$. Suppose $T_n \rightarrow T$. Since T_n are bounded and linear, T is also bounded and linear.

It remains to prove that T is self-adjoint. Let $(T_n - T)^*$ be the adjoint operator of $T_n - T$. Then $\|(T_n - T)^*\| = \|T_n - T\|$. Therefore,

$$\|T_n^* - T^*\| = \|(T_n - T)^*\| = \|T_n - T\|$$

We have, $T_n^* = T_n$ and $\|T_n^* - T_n\| = 0$. Thus,

$$\begin{aligned} \|T - T^*\| &\leq \|T - T_n\| + \|T_n - T_n^*\| + \|T_n^* - T^*\| \\ &\leq 2\|T_n - T\| \end{aligned}$$

As $n \rightarrow \infty$, $\|T_n - T\| \rightarrow 0$ since $T_n \rightarrow T$. Clearly the left side is independent of the limit, and $\|T - T^*\| = 0$. Thus, $T - T^* = 0$. In other words, $T = T^*$. \square

Theorem 14.4.7. *Let H be a Hilbert space. Let $U : H \rightarrow H$ and $V : H \rightarrow H$ are unitary. Then*

1. U is isometric. ie, $\forall x \in H, \|Ux\| = \|x\|$
2. $\|U\| = 1$
3. U^{-1} is unitary
4. UV is unitary
5. U is normal
6. A bounded, linear operator T on a complex, Hilbert space H is unitary if and only if T is isometric and surjective.

Proof. Let H be a Hilbert space. Let U, V be unitary operators. Then $U^{-1} = U^*$ and $V^{-1} = V^*$.

1. U is isometric.

$$\|Ux\|^2 = \langle Ux, Ux \rangle = \langle x, U^*Ux \rangle = \langle x, Ix \rangle = \|x\|^2$$

Therefore, $\|Ux\| = \|x\|$. In other words, U is isometric.

2. $\|U\| = 1$

We have, U is isometric. That is, $\|Ux\| = \|x\|$. Thus,

$$\|U\| = \sup_{\substack{x \in H \\ x \neq 0}} \frac{\|Ux\|}{\|x\|} = 1$$

3. U^{-1} is unitary

$$(U^{-1})^* = (U^*)^* = U = (U^{-1})^{-1}$$

If $U^{-1} = V$, then $V^* = V^{-1}$. Therefore, U^{-1} is unitary.

4. UV is unitary since

$$(UV)^* = V^*U^* = V^{-1}U^{-1} = (UV)^{-1}$$

5. U is normal since

$$UU^* = UU^{-1} = I = U^{-1}U = U^*U$$

6. T unitary iff T is isometric and surjective. (H complex)

Suppose T is isometric and surjective. Since T is isometric, we have $\|Tx\| = \|x\|$. Suppose $Tx = Ty$. Then,

$$\|Tx - Ty\| = \|T(x - y)\| = \|x - y\| = 0$$

Thus, $Tx = Ty \implies x = y$. Therefore, T is injective. Thus, T^{-1} exists.

$$\langle T^*Tx, x \rangle = \langle Tx, Tx \rangle = \|Tx\|^2 = \|x\|^2 = \langle x, x \rangle = \langle T^{-1}Tx, x \rangle$$

Therefore, $T^* = T^{-1}$. In other words, T is unitary.

Suppose T is unitary. Then T is isometric by part (a). We have, $T^* = T^{-1}$. Thus T is bijective, otherwise T^{-1} won't even exist.

□

Exercise §3.10

Definitions 14.4.3 (unitarily equivalence). Let H be a Hilbert space. Let $S : H \rightarrow H$ and $T : H \rightarrow H$ be linear operators. Then S is unitarily equivalent to T if there exists a unitary operator $U : H \rightarrow H$ such that $S = UTU^{-1}$.

14.5 Fundamental Theorems for Banach Spaces

14.5.1 Zorn's Lemma

Definitions 14.5.1 (POSET). A set M is a POSET if it satisfies

1. $a \leq a, \forall a \in M$
2. $a \leq b, b \leq a \implies a = b$
3. $a \leq b, b \leq c \implies a \leq c$

Lemma 14.5.1 (Zorn). Let M be a non-empty POSET. Suppose every chain in M has an upper bound. Then M has a maximal element.

Applications : Zorn's Lemma

Remark. Every non-trivial vector space has a Hamel basis.

Proof. Let X be a nontrivial vector space. Let M be the set of all linearly independent subsets of X . Since X has nonzero elements, M is nonempty. We know that, M together with set inclusion is a POSET since subsets of linearly independent sets are also linearly independent.

Let C be a chain in M . Then the union of all subsets in C is a bound for C . Thus, every chain in M is bounded above. By Zorn's lemma, M has a maximal element, say B .

Let $Y = \text{span } B$. Suppose $Y \neq X$. Then there exists $z \in X$ such that $z \notin Y$. Then $B \cup \{z\}$ is a linearly independent subset of X . This is a contradiction since B is maximal. Therefore, $Y = X$. Thus, any element $x \in X$ belongs to $\text{span } B$. In other words, B is a Hamel basis for X . \square

Remark. Every nontrivial Hilbert space has a total, orthonormal set.

Proof. Let M be the set of all orthonormal subset of Hilbert space H . Set M is nonempty since H is nontrivial and for any $x \in H$, we have $\frac{x}{\|x\|} \in M$. Set M together with set inclusion is a POSET. And any chain C in M is bounded above by the union of all subsets in C . By Zorn's lemma, M has a maximal element, say F .

Suppose F is not total in H . Then there exists $z \in H$ such that $z \perp F$. Define $e = \frac{z}{\|z\|}$. Then $F \cup e$ is an orthonormal subset of H contradicting maximality of F . Therefore, F is total. Clearly, every Hilbert space has a total, orthonormal subset. \square

14.5.2 Hahn-Banach Theorem

Definitions 14.5.2 (extension problem). Let X be any set. Let Z be any subset of X and G be an entity defined on Z . The extension problem is the question whether there exists an extension of G from Z into X such that basic properties of G are preserved.

Definitions 14.5.3 (subadditive). A functional p on a vector space X is subadditive if $p(x + y) \leq p(x) + p(y)$, $\forall x, y \in X$.

Definitions 14.5.4 (positive homogenous). A function p on a vector space X is positive homogenous if $p(\alpha x) = \alpha p(x)$, $\forall \alpha \geq 0, \alpha \in \mathbb{R}$.

For example, Let X be a normed space. Then the norm on X is a positive homogenous, subadditive functional on X . In other words, norm is a sublinear functional on X .

Theorem 14.5.2 (Hahn-Banach - Linear Functional). Let X be a real vector space. Let p be a sublinear functional on X . Let f be a linear functional defined on a subspace Z of X such that $f(x) \leq p(x)$, $\forall x \in Z$. Then, f has a linear extension \tilde{f} from Z into X satisfying $\tilde{f}(x) \leq p(x)$, $\forall x \in X$.

Proof. Let X be a real vector space and Z be a subspace of X . Let p be a sublinear functional on X . Let f be linear functional on Z such that $f(x) \leq p(x) \forall x \in Z$.

Step 1 : Construction of \tilde{f} , by Zorn's lemma

Let E be the set all linear extensions g of f such that $g(x) = f(x)$, $\forall x \in Z$ and $g(x) \leq p(x)$, $\forall x \in \mathcal{D}(g)$. Then E is nonempty, since $f \in E$. Define a partial order \leq on E such that $g \leq h$ if $\mathcal{D}(g) \subset \mathcal{D}(h)$ and $g(x) = h(x)$, $\forall x \in \mathcal{D}(g)$. Clearly E is a POSET.

Let C be a chain in E . Now we can construct, an upper bound for any such chain in E by considering the extension on the union of domains of all extension in C . Define $\hat{Y} = \bigcup_{g \in C} \mathcal{D}(g)$. Define $\hat{g} : \hat{Y} \rightarrow \mathbb{R}$ such that $\hat{g}(x) = g(x)$, $x \in \mathcal{D}(g)$. Clearly, every chain C in E is bounded above by such an extension \hat{g} . By Zorn's lemma, E has a maximal element, say \tilde{f} . Since $\tilde{f} \in E$, we have $\tilde{f}(x) \leq p(x)$, $\forall x \in \mathcal{D}(\tilde{f})$.

Step 2 : $\mathcal{D}(\tilde{f}) = X$

Suppose $\mathcal{D}(\tilde{f}) \neq X$. Then there exists $y_1 \in X$ such that $y_1 \notin \mathcal{D}(\tilde{f})$. Consider the subspace Y_1 of X spanned by $\mathcal{D}(\tilde{f}) \cup \{y_1\}$. Then any $x \in Y_1$ is of the form $x = y + \alpha y_1$ where $y \in \mathcal{D}(\tilde{f})$.

This representation is unique since $x = y + \alpha y_1 = \tilde{y} + \beta y_1 \implies y - \tilde{y} = (\beta - \alpha)y_1 \in \mathcal{D}(\tilde{f})$. We know that $y_1 \notin \mathcal{D}(\tilde{f})$, thus $\alpha = \beta$.

Define $g_1 : Y_1 \rightarrow \mathbb{R}$ such that $g_1(x) = \tilde{f}(x) + \alpha c$ where $c \in \mathbb{R}$ is fixed. Clearly g_1 is linear since \tilde{f} is linear and

$$g_1(x + \tilde{x}) = g_1(y + \alpha y_1 + \tilde{y} + \beta y_1) = \tilde{f}(y + \tilde{y}) + c(\alpha + \beta) = g_1(y + \alpha y_1) + g_1(\tilde{x} + \beta y_1)$$

And $g_1 \leq \tilde{f}$ since $\mathcal{D}(\tilde{f}) \subset \mathcal{D}(g_1)$ and $g_1(x) = \tilde{f}(x)$, $\forall x \in \mathcal{D}(\tilde{f})$. However, it remains to prove that $g_1(x) \leq p(x)$. We claim that $g_1(x) \leq p(x)$, $\forall x \in \mathcal{D}(g_1)$.

Step 3 : Claim $g_1(x) \leq p(x)$, $\forall x \in \mathcal{D}(g_1)$

Suppose $y, z \in \mathcal{D}(\tilde{f})$.

$$\begin{aligned} \tilde{f}(y) - \tilde{f}(z) &= \tilde{f}(y - z) \leq p(y - z) \\ &\leq p(y + y_1 - z - y_1) \\ &\leq p(y + y_1) + p(-z - y_1) \end{aligned}$$

Rearranging terms, we get

$$-\tilde{f}(z) - p(-z - y_1) \leq p(y + y_1) - \tilde{f}(y)$$

Applying supremum over z on the left and infimum over y on the right, we get $m_0 \leq m_1$ where $m_0 = \sup_{z \in \mathcal{D}(\tilde{f})} -\tilde{f}(z) - p(-z - y_1)$ and

$$m_1 = \inf_{y \in \mathcal{D}(\tilde{f})} p(y + y_1) - \tilde{f}(y).$$

Let $m_0 \leq c \leq m_1$. Then,

$$-p(-z - y_1) - \tilde{f}(z) \leq c \leq p(y + y_1) - \tilde{f}(y), \quad \forall y, z \in \mathcal{D}(\tilde{f})$$

Case 1 : $\alpha < 0$

Consider $z = \alpha^{-1}y$,

$$-p(-y_1 - \frac{y}{\alpha}) - \tilde{f}(\frac{y}{\alpha}) \leq m_0 \leq c$$

Multiplying with $-\alpha$

$$\begin{aligned} \alpha p(-y_1 - \frac{y}{\alpha}) + \tilde{f}(y) &\leq -\alpha c \\ g_1(x) = \tilde{f}(y) + \alpha c &\leq -\alpha p(-y_1 - \frac{y}{\alpha}) \\ &\leq p(\alpha y_1 + y) = p(x) \end{aligned}$$

Case 2 : $\alpha > 0$

Consider $y = \alpha^{-1}y$,

$$c \leq m_1 \leq p(\frac{y}{\alpha} + y_1) - \tilde{f}(\frac{y}{\alpha})$$

Multiplying with α

$$\begin{aligned} \alpha c &\leq \alpha p(\frac{y}{\alpha} + y_1) - \alpha \tilde{f}(\frac{y}{\alpha}) \\ &\leq p(y + \alpha y_1) - \tilde{f}(y) \\ g_1(x) = \tilde{f}(y) + \alpha c &\leq p(y + \alpha y_1) = p(x) \end{aligned}$$

Case 3 : $\alpha = 0$

Suppose $\alpha = 0$, then $x = y + 0y_1 \in \mathcal{D}(g_1)$. Thus, $g_1(x) = \tilde{f}(x) \leq p(x)$.

Thus, the claim is true. And we have a proper extension g_1 of \tilde{f} which is a contradiction as \tilde{f} is maximal in E . Thus, there does not exist such a nonzero vector $y_1 \in X - \mathcal{D}(\tilde{f})$. Therefore, $\mathcal{D}(\tilde{f}) = X$. \square

14.5.3 Hahn-Banach Theorem for complex vector spaces and normed spaces

Theorem 14.5.3 (Hahn-Banach : generalised). *Let X be a real/complex vector space. Let p be a subadditive, real-valued functional on X and $p(\alpha x) = |\alpha|p(x)$, $\forall \alpha \in K$. Let f be a linear functional defined on a subspace Z of X such that $|f(x)| \leq p(x)$. Then, f has a linear extension \tilde{f} from Z into X satisfying $|\tilde{f}(x)| \leq p(x)$, $\forall x \in X$.*

Proof. **Case 1 : Real Vector Space**

By Hahn-Banach theorem, we know that there exists an extension \tilde{f} such that $\tilde{f} \leq p$, in the case of real vector space X .

Case 2 : Complex Vector Space

Let X be a complex vector space. Then any functional f on Z is of the form $f = f_1 + if_2$ where both f_1, f_2 are real valued functionals. And f_1, f_2 are

linear functionals since f is a linear functional on Z . Clearly, $f_1(x) \leq p(x)$ and $f_2(x) \leq p(x)$ since the real/imaginary parts are bounded above by the absolute value of f and f itself is bounded. Therefore, f_1, f_2 are bounded functionals on Z .

By Hahn-Banach theorem $f_1(x)$ has an extension $\tilde{f}_1(x)$. We have,

$$f(ix) = f_1(ix) + if_2(ix) = -f_2(x) + if_1(x)$$

Thus, $f_2(x) = -f_1(ix)$. Define

$$\tilde{f}(x) = \tilde{f}_1(x) - i\tilde{f}_1(ix)$$

Then, \tilde{f} is an extension of f . It remains to prove that \tilde{f} is linear and $\tilde{f}(x) \leq p(x)$, $\forall x \in X$.

Clearly, \tilde{f} is linear, since

$$\begin{aligned} \tilde{f}((a+ib)x) &= \tilde{f}_1(ax+ibx) - i\tilde{f}_1(iax-bx) \\ &= a\tilde{f}_1(x) + b\tilde{f}_1(ix) - ia\tilde{f}_1(ix) + ib\tilde{f}_1(x) \\ &= a(\tilde{f}_1(x) - i\tilde{f}_1(ix)) + ib(\tilde{f}_1(x) - i\tilde{f}_1(ix)) \\ &= (a+ib)(\tilde{f}_1(x) - i\tilde{f}_1(ix)) \\ &= (a+ib)\tilde{f}(x) \end{aligned}$$

Let $x \in X$ such that $\tilde{f}(x) = 0$. Suppose $p(x) < 0$, then $p(-x) < 0$ since $p(-x) = |-1|p(x) = p(x)$. We know that p is subadditive. That is, $p(x+y) \leq p(x) + p(y)$. Thus, $p(0) = p(x-x) \leq p(x) + p(-x) = 2p(x) < 0$. But $p(0) = p(0x) = 0p(x) = 0$. This is a contradiction thus $p(x) \geq 0$. Therefore, if $\tilde{f}(x) = 0$, then $\tilde{f}(x) \leq p(x)$.

Let $x \in X$ such that $\tilde{f}(x) \neq 0$. Then $\tilde{f}(x) = |\tilde{f}(x)|e^{i\theta}$. Thus, $|\tilde{f}(x)| = e^{-i\theta}\tilde{f}(x) = \tilde{f}(xe^{-i\theta})$. Since $|\tilde{f}(x)|$ is real, $\tilde{f}(xe^{-i\theta})$ is also real. By Hahn-Banach theorem on linear functionals on real vector spaces, $\tilde{f}_1(x) \leq p(x)$, $\forall x \in X$. Therefore,

$$|\tilde{f}(x)| = \tilde{f}(xe^{-i\theta}) = \tilde{f}_1(xe^{-i\theta}) \leq p(xe^{-i\theta}) = |e^{-i\theta}|p(x) = p(x)$$

Thus, if $\tilde{f}(x) \neq 0$, then $\tilde{f}(x) \leq p(x)$. Clearly, \tilde{f} is a bounded functional on X . Therefore, \tilde{f} is a linear extension of f such that $\tilde{f}(x) \leq p(x)$, $\forall x \in X$. \square

Theorem 14.5.4 (Hahn-Banach : Normed space). *Let f be a bounded, linear functional on a subspace Z of normed space X . Then there exists a bounded, linear functional \tilde{f} on X which is an extension to f on Z and has the same norm, $\|\tilde{f}\| = \|f\|$ where*

$$\|\tilde{f}\| = \sup_{\substack{x \in X \\ \|x\|=1}} |\tilde{f}(x)| \quad \|f\| = \sup_{\substack{x \in Z \\ \|x\|=1}} |f(x)|$$

Proof. Without loss of generality suppose that Z is non-trivial. If $Z = \{0\}$, then $f(x) = f(0) = 0$ since f is linear. Thus, f has an extension \tilde{f} defined by

$\tilde{f}(x) = 0, \forall x \in X$ and $\|f\| = \|\tilde{f}\| = 0$.

Suppose Z is non-trivial. Then $|f(x)| \leq \|f\| \|x\|$ since f is bounded^{†16}. Define functional p where $p(x) = \|f\| \|x\|, \forall x \in X$. Then,

$$p(x+y) = \|f\| \|x+y\| \leq \|f\|(\|x\| + \|y\|) = p(x) + p(y)$$

And,

$$p(\alpha x) = \|f\| \|\alpha x\| = |\alpha| \|f\| \|x\| = |\alpha| p(x)$$

Clearly, p is a sublinear functional on X such that $f(x) \leq p(x), \forall x \in Z$.

Given any functional f on Z , there exists a sublinear functional p such that $f(x) \leq p(x)$. Then, by Hahn-Banach theorem, there exists a bounded, linear functional $\tilde{f}: X \rightarrow K$ such that $\tilde{f}(x) \leq p(x), \forall x \in X$ and \tilde{f} is an extension of f . That is, $\tilde{f}(x) = f(x), \forall x \in Z$.

Since $\tilde{f}(x) \leq p(x), \forall x \in X$. We have,

$$\|\tilde{f}\| = \sup_{x \neq 0} \frac{|\tilde{f}(x)|}{\|x\|} \leq \sup_{x \neq 0} \frac{p(x)}{\|x\|} = \|f\|$$

Therefore, any bounded, linear functional on any subspace of any normed space has a bounded, linear extension preserving norm. \square

Theorem 14.5.5. *Let X be a normed space. Let $x_0 \in X$. Then there exists a bounded, linear functional \tilde{f} on X such that $\|\tilde{f}\| = 1$ and $\tilde{f}(x_0) = \|x_0\|$.*

Proof. Let $x_0 \in X$ and $x_0 \neq 0$. Let Z be the subspace spanned by x_0 . That is, $Z = \{\alpha x_0 : \alpha \in K\}$. Define $f: Z \rightarrow K$ where $f(x) = f(\alpha x_0) = \alpha \|x_0\|$. Then $|f(x)| = |\alpha| \|x_0\| = \|x\|$. Clearly, f is bounded, linear functional with $\|f\| = 1$. By Hahn-Banach theorem for normed spaces f has an extension \tilde{f} on X such that $\|\tilde{f}\| = \|f\|$ and $\tilde{f}(x_0) = f(x_0) = \|x_0\|$. \square

Corollary 14.5.5.1. *Let X be a normed space. Let $x \in X$.*

$$\|x\| = \sup_{\substack{f \in X' \\ f \neq 0}} \frac{|f(x)|}{\|f\|}$$

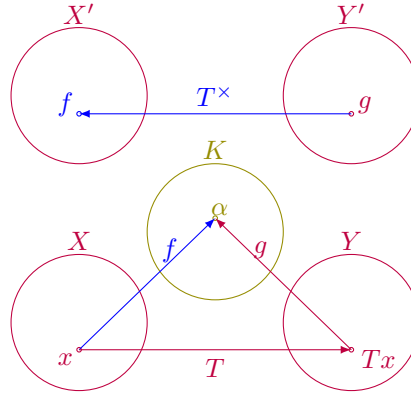
If $f(x_0) = 0, \forall f \in X'$, then $x_0 = 0$.

Proof. Suppose $x \in X$ such that $f(x) = 0$ for every bounded, linear functional f on X . Considering each f as bounded, linear functional on the subspaces spanned by x , we get bounded linear extensions \tilde{f} on X by Hahn-Banach theorem for normed spaces.

Then,

$$\sup_{f \neq 0} \frac{|f(x)|}{\|f\|} \geq \sup_{f \neq 0} \frac{|\tilde{f}(x)|}{\|\tilde{f}\|} = \|x\|$$

^{†16} Author uses $\|f\|_Z$ to emphasis that the functional is defined only in Z . Proper notation would be $\|f\|_{B(Z,K)}$ and $\|\tilde{f}\|_{B(X,K)}$ instead of $\|f\|_Z$ and $\|\tilde{f}\|_X$



$$\forall g \in Y', \exists f \in X', \text{ such that } g(Tx) = (T^\times g)(x) = f(x) = \alpha \in K$$

Figure 14.2: Adjoint-Operator

Also,

$$\sup_{f \neq 0} \frac{|f(x)|}{\|f\|} \leq \sup_{f \neq 0} \frac{\|f\| \|x\|}{\|f\|} = \|x\|$$

Therefore,

$$\sup_{f \neq 0} \frac{|f(x)|}{\|f\|} = \|x\|$$

□

14.5.5 Adjoint Operator

Definitions 14.5.5 (adjoint-operator). Let X, Y be normed spaces over the same field K . Let X', Y' be the dual spaces of X, Y . Let $T : X \rightarrow Y$ be a bounded, linear operator. Then the adjoint-operator is $T^\times : Y' \rightarrow X'$ defined by $f(x) = (T^\times g)(x) = g(Tx)$.

Theorem 14.5.6 (norm). Let T^\times be the adjoint operator of a bounded, linear operator T . Then T^\times is a bounded, linear operator. And $\|T^\times\| = \|T\|$.

Proof. Let $g_1, g_2 \in Y'$ and $\alpha, \beta \in K$. Then,

$$\begin{aligned} (T^\times(\alpha g_1 + \beta g_2))(x) &= (\alpha g_1 + \beta g_2)(Tx) \\ &= \alpha g_1(Tx) + \beta g_2(Tx) \\ &= \alpha(T^\times g_1)(x) + \beta(T^\times g_2)(x) \end{aligned}$$

Clearly, T^\times is linear.

We have, $f = T^\times g$.

$$\begin{aligned}
 f(\alpha x_1 + \beta x_2) &= (T^\times g)(\alpha x_1 + \beta x_2) \\
 &= g(T(\alpha x_1 + \beta x_2)) \\
 &= g(\alpha T x_1 + \beta T x_2) \\
 &= \alpha g(T x_1) + \beta g(T x_2) \\
 &= \alpha (T^\times g) x_1 + \beta (T^\times g) x_2 \\
 &= \alpha f(x_1) + \beta f(x_2)
 \end{aligned}$$

Thus, f is linear. We have $|g(Tx)| \leq \|g\| \|T\| \|x\|$, since functional g and operator T are bounded. Thus,

$$\|T^\times g\| = \|f\| = \sup_{x \neq 0} \frac{|f(x)|}{\|x\|} = \sup_{x \neq 0} \frac{|g(Tx)|}{\|x\|} \leq \sup_{x \neq 0} \frac{\|g\| \|T\| \|x\|}{\|x\|} \leq \|g\| \|T\|$$

Taking supremum over all $g \in Y'$, we get

$$\|T^\times\| \leq \sup_{g \neq 0} \frac{\|T^\times g\|}{\|g\|} = \sup_{g \neq 0} \frac{\|g\| \|T\|}{\|g\|} = \|T\|$$

We know that for any non-zero $x_0 \in X$ there exists a $g_0 \in Y'$ such that $\|g_0\| = 1$ and $g_0(Tx_0) = \|Tx_0\|$. Let $f_0 = T^\times g_0$. Then,

$$\|Tx_0\| = g_0(Tx_0) = f_0(x_0) \leq \|f_0\| \|x_0\| = \|T^\times g_0\| \|x_0\| \leq \|T^\times\| \|g_0\| \|x_0\|$$

Thus, $\|Tx_0\| \leq \|T^\times\| \|x_0\|$. Taking supremum over x_0 we get, $\|T\| \leq \|T^\times\|$.

We have $\|Tx_0\| \leq \|T\| \|x_0\|$. And $\|T\| = c$ is the smallest real number satisfying $\|Tx_0\| \leq c\|x_0\|$, $\forall x_0 \in X$. Thus, $\|T^\times\| \geq c$. Therefore,

$$\|T^\times\| \geq \|T\|$$

Therefore, $\|T^\times\| = \|T\|$. □

Remark (Properties of adjoint-operator). Let $S : Y \rightarrow Z$ and $T : X \rightarrow Y$ be bounded, linear operators. Let $\alpha \in K$. Then,

1. If T_E is the matrix representation of T , then transpose of T_E is the matrix representation of T^\times .

Proof. Let T_E be the matrix representation of linear operator $T : \mathbb{R}^n \rightarrow \mathbb{R}^n$ with respect to standard basis $E = \{e_1, e_2, \dots, e_n\}$. Let $x = (\xi_j)$ and $y = (\eta_j)$. Then $y = Tx$ implies $\eta_j = \sum_{k=1}^n \tau_{j,k} \xi_k$. Define $T_E = (\tau_{j,k})$. Then $y = T_E x$. We know that $(\mathbb{R}^n)' = \mathbb{R}^n$. Let g be a bounded, linear operator on \mathbb{R}^n . Then $g \in \mathbb{R}^n$ and $g = \sum_{k=1}^n \alpha_k f_k$ where $F = \{f_1, f_2, \dots, f_n\}$ is the dual basis of E . That is, $f_k(e_j) = \delta_{k,j}$.

$$g(y) = g(T_E x) = \sum_{j=1}^n \alpha_j \eta_j = \sum_{j=1}^n \sum_{k=1}^n \alpha_j \tau_{j,k} \xi_k$$

Thus,

$$g(T_E x) = \sum_{k=1}^n \beta_k \xi_k \text{ where } \beta_k = \sum_{j=1}^n \tau_{j,k} \alpha_j$$

Define $f : X \rightarrow K$ where $f(x) = g(Tx)$. Then,

$$f(x) = g(T_E x) = \sum_{k=1}^n \beta_k \xi_k$$

Suppose $f = T_E^\times g$. Then matrix of T_E^\times is given by $(\tau_{j,k})$ since $\beta_k = \sum_{j=1}^n \tau_{j,k} \alpha_j$. Thus, matrix of T^\times is the transpose of that of T_E . \square

$$2. (S + T)^\times = S^\times + T^\times$$

Proof.

$$\begin{aligned} ((S + T)^\times g)(x) &= g((S + T)x) \\ &= g(Sx + Tx) \\ &= g(Sx) + g(Tx) \\ &= (S^\times g)(x) + (T^\times g)(x) \end{aligned}$$

\square

$$3. (\alpha T)^\times = \alpha T^\times$$

Proof.

$$\begin{aligned} ((\alpha T)^\times g)(x) &= g((\alpha T)x) \\ &= g(\alpha Tx) \\ &= \alpha g(Tx) \\ &= \alpha (T^\times g)(x) \end{aligned}$$

\square

$$4. (T^\times)^{-1} = (T^{-1})^\times$$

Suppose $T : X \rightarrow Y$ is invertible and $T^{-1} : Y \rightarrow X$. Then $(T^\times)^{-1}$ exists and $(T^\times)^{-1} = (T^{-1})^\times$.

Proof. We have $T \in B(X, Y)$ and $T^\times \in B(Y', X')$. Define $S = T^{-1}$. Then $S \in B(Y, X)$. Thus S^\times exists, since S is a bounded, linear operator. And $S^\times \in B(X', Y')$.

Let $g \in Y'$. Then,

$$g(y) = g(TSy) = (T^\times g)(Sy) = (S^\times T^\times g)(y), \quad \forall g \in Y'$$

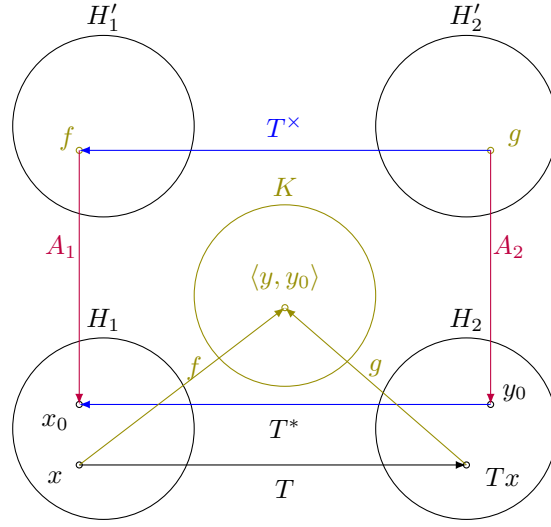


Figure 14.3: Adjoint-Operator vs Hilbert-Adjoint-Operator

Clearly, $T^\times S^\times$ is an identity operator on Y' .

Let $f \in X'$. Then,

$$f(x) = f(STx) = (S^\times f)(Tx) = (T^\times S^\times f)(x), \quad \forall f \in X'$$

Clearly, $S^\times T^\times$ is an identity operator on X' .

Therefore, $S^\times = (T^{-1})^\times = (T^\times)^{-1}$. □

Remark (T^\times vs T^).* Let $T : H_1 \rightarrow H_2$. Let T^\times be the adjoint-operator of T . That is, $T^\times : H'_2 \rightarrow H'_1$. Let $T^* : H_2 \rightarrow H_1$ be the Hilbert-adjoint operator of T . That is, $\langle Tx, y \rangle = \langle x, T^*y \rangle$. Then,

$$T^* = A_1 T^\times A_2^{-1}$$

where $A_1 : H'_1 \rightarrow H_1$ defined by $A_1(f) = x_0$ such that $f(x) = \langle x, x_0 \rangle$ and $A_2 : H'_2 \rightarrow H_2$ defined by $A_2(g) = y_0$ such that $g(y) = \langle y, y_0 \rangle$.

Proof. Let H_1, H_2 be normed spaces. Let $T : H_1 \rightarrow H_2$ be a bounded, linear operator. Then there exists unique adjoint operator $T^\times : H'_2 \rightarrow H'_1$ defined by $(T^\times g)(x) = g(Tx)$, $\forall g \in Y'$, $\forall x \in H_1$. And there exists unique Hilbert adjoint operator $T^* : H_2 \rightarrow H_1$ defined by $\langle Tx, y \rangle = \langle x, T^*y \rangle$, $\forall x \in H_1$, $\forall y \in H_2$.

Let $f \in H'_1$. By Riesz theorem, $\forall f \in H'_1$ there exists unique $x_0 \in H_1$ such that $f(x) = \langle x, x_0 \rangle$. Define $A_1 : H'_1 \rightarrow H_1$ defined by $A_1(f) = x_0$ where $f(x) = \langle x, x_0 \rangle$, $\forall x \in H_1$. Clearly, A_1 is bijective by its construction. Let $g \in H'_2$. By Riesz theorem, $\forall g \in H'_2$, there exists unique $y_0 \in H_2$ such that $g(y) = \langle y, y_0 \rangle$. Define $A_2 : H'_2 \rightarrow H_2$ defined by $A_2(g) = y_0$ where

$g(y) = \langle y, y_0 \rangle$, $\forall y \in H_2$. Clearly, A_2 is bijective by its construction.

Let $A_1(f_1) = x_1$ and $A_1(f_2) = x_2$. Then, we have $f_1(x) = \langle x, x_1 \rangle$ and $f_2(x) = \langle x, x_2 \rangle$. Thus,

$$\begin{aligned} (\alpha f_1 + \beta f_2)(x) &= \alpha f_1(x) + \beta f_2(x) \\ &= \alpha \langle x, x_1 \rangle + \beta \langle x, x_2 \rangle \\ &= \langle x, \bar{\alpha}x_1 \rangle + \langle x, \bar{\beta}x_2 \rangle \\ &= \langle x, \bar{\alpha}x_1 + \bar{\beta}x_2 \rangle \\ A_1(\alpha f_1 + \beta f_2) &= \bar{\alpha}x_1 + \bar{\beta}x_2 \\ &= \bar{\alpha}A_1(f_1) + \bar{\beta}A_1(f_2) \end{aligned}$$

Thus, A_1 is conjugate linear. Similarly, A_2 is also conjugate linear.

Let $S = A_1 T^\times A_2^{-1}$. Let $A_1(f) = x_0$ and $A_2(g) = y_0$. Then $S : H_2 \rightarrow H_1$ defined by $S(y_0) = (A_1 T^\times A_2^{-1})(y_0) = (A_1 T^\times)(g) = A_1(f) = x_0$. And S is linear since A_1 and A_2^{-1} are conjugate linear. Then,

$$\langle Tx, y_0 \rangle = g(Tx) = f(x) = \langle x, x_0 \rangle = \langle x, Sy_0 \rangle$$

But, we have, $\langle Tx, y_0 \rangle = \langle x, T^* y_0 \rangle$. And we know that, $w_1 = w_2$ if $\langle v, w_1 \rangle = \langle v, w_2 \rangle, \forall v$. Thus, $S = T^* = A_1 T^\times A_2^{-1}$. \square

Remark. Comparison

1. Matrix of adjoint operator is transpose. But, matrix of Hilbert adjoint operator is complex conjugate transpose.
2. Adjoint operator is linear, but Hilbert-adjoint operator is conjugate linear.

Exercises

[Kreyszig, 2014, §4.5 Exercise 6]

Remark.

$$(T^\times)^n = (T^n)^\times$$

Proof. Let $T : X \rightarrow X$ be a bounded, linear operator.

$$g(T^2 x) = g(T(Tx)) = (T^\times g)(Tx) = T^\times T^\times g(x) = (T^\times)^2 g(x)$$

Suppose $(T^\times)^{k-1} = (T^{k-1})^\times$ for some $k \in \mathbb{N}$. Then,

$$g(T^k x) = g(T(T^{k-1} x)) = (T^\times g)(T^{k-1} x) = T^\times (T^\times)^{k-1} g(x) = (T^\times)^k g(x)$$

Therefore, by finite mathematical induction the result is true. \square

Definitions 14.5.6 (Annihilator in a normed space). Let X be a normed space and M be a subset of X . Then the annihilator of M is given by,

$$M^a = \{f \in X' : f(M) = \{0\}\}$$

Definitions 14.5.7 (Annihilator in a dual space). Let X be a normed space and X' be its dual space. Let B be a subset of X' . Then the annihilator of B is given by,

$${}^aB = \{x \in X : f(x) = 0, \forall f \in B\}$$

Remark.

$$\mathcal{R}(T) \subset {}^a\mathcal{N}(T^\times)$$

Proof. Let $T : X \rightarrow Y$ be a bounded, linear operator. Let $f \in \mathcal{N}(T^\times)$. Then $T^\times f = 0$. Let $y = Tx \in \mathcal{R}(T)$. Then,

$$f(y) = f(Tx) = (T^\times f)(x) = 0, \forall y \in \mathcal{R}(T)$$

Thus, $y \in \mathcal{R}(T) \implies y \in {}^a\mathcal{N}(T^\times)$. □

Subject 15

ME010305 Optimization Technique

15.1 Linear Programming

Definitions 15.1.1 (LPP). Linear Programming Problem (LPP) is the optimization of linear object function under certain linear equality/inequalities constraints.

General LPP A general LPP is of the form,

$$\text{minimise : } f(X) = CX \text{ (object function)} \quad (15.1)$$

$$\text{subject to : } AX = B \text{ (linear constraints)} \quad (15.2)$$

$$\text{such that } X \geq 0 \text{ (non-negativity conditions)} \quad (15.3)$$

That is,

$$\text{minimise : } f = c_1x_1 + c_2x_2 + \cdots + c_nx_n \quad (15.4)$$

$$\text{subject to : } \begin{pmatrix} a_{11} & a_{12} & \cdots & a_{1m} \\ a_{21} & a_{22} & \cdots & a_{2m} \\ \vdots & \vdots & \ddots & \vdots \\ a_{m1} & a_{m2} & \cdots & a_{mm} \end{pmatrix} \begin{pmatrix} x_1 \\ x_2 \\ \cdots \\ x_m \end{pmatrix} = \begin{pmatrix} b_1 \\ b_2 \\ \vdots \\ b_m \end{pmatrix} \quad (15.5)$$

$$\text{such that } X \geq 0 \quad (15.6)$$

The coefficients of the object function, c_i are the **cost coefficients**.

Feasible Solution is a vector X satisfying both constraints and non-negativity constraints. The set of all feasible solutions is denoted by S_F . This is a convex set and is often represented graphically by a polytope.

Basic Solution In an LPP with m constraints and n variables, a **basic solution** is a vector X in which $n - m$ variables are zero. These variables are the non-basic variables and the remaining m variables are the basic variable. The set of all basic variables is a **basis**.

Basic Feasible Solution is a basic solution which is feasible. The **b.f.s.** are vertices of the polytope S_F (if nonempty).

Optimal Solution is a vector X which optimizes the object function f . For a bounded S_F , always there exists a b.f.s which is optimal.

Solving LPP Due to the presence of inequalities, there is no analytic tool for solving an LPP. There are a few numerical methods for solving an LPP. Simplex is popular among them.

Simplex Algorithm First, we compute a b.f.s. (which is a vertex of the polytope S_F). If there is a desirable b.f.s. (an adjacent vertex) which improves the object function, then we jump to the corresponding b.f.s. (vertex). We continue this process until there is no scope of further improvement.

Canonical Form Consider the LPP (in matrix form)

$$\begin{aligned} & \text{minimise } f(X) = CX \\ & \text{subject to } AX = B \\ & \text{such that } X \geq 0 \end{aligned}$$

Rearranging the non-basic variable, we get

$$\begin{pmatrix} a_{11} & a_{12} & \cdots & a_{1m} \\ a_{21} & a_{22} & \cdots & a_{2m} \\ \vdots & \vdots & \ddots & \vdots \\ a_{m1} & a_{m2} & \cdots & a_{mm} \end{pmatrix} \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_m \end{pmatrix} = \begin{pmatrix} b_1 - a_{1,(m+1)}x_{m+1} - a_{1,(m+2)}x_{m+2} - \cdots - a_{1n}x_n \\ b_2 - a_{2,(m+1)}x_{m+1} - a_{2,(m+2)}x_{m+2} - \cdots - a_{2n}x_n \\ \vdots \\ b_m - a_{m,(m+1)}x_{m+1} - a_{m,(m+2)}x_{m+2} - \cdots - a_{m,n}x_n \end{pmatrix}$$

Left multiplying with A^{-1} on either sides, we get

$$\begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_m \end{pmatrix} = \begin{pmatrix} \bar{b}_1 - \bar{a}_{1,(m+1)}x_{m+1} - \bar{a}_{1,(m+2)}x_{m+2} - \cdots - \bar{a}_{1n}x_n \\ \bar{b}_2 - \bar{a}_{2,(m+1)}x_{m+1} - \bar{a}_{2,(m+2)}x_{m+2} - \cdots - \bar{a}_{2n}x_n \\ \vdots \\ \bar{b}_m - \bar{a}_{m,(m+1)}x_{m+1} - \bar{a}_{m,(m+2)}x_{m+2} - \cdots - \bar{a}_{m,n}x_n \end{pmatrix}$$

where $\begin{pmatrix} A^{-1} \end{pmatrix} \begin{pmatrix} b_1 \\ b_2 \\ \vdots \\ b_m \end{pmatrix} = \begin{pmatrix} \bar{b}_1 \\ \bar{b}_2 \\ \vdots \\ \bar{b}_m \end{pmatrix}$ and $\begin{pmatrix} A^{-1} \end{pmatrix} \begin{pmatrix} a_{1,(m+1)} \\ a_{2,(m+1)} \\ \vdots \\ a_{m,(m+1)} \end{pmatrix} = \begin{pmatrix} \bar{a}_{1,(m+1)} \\ \bar{a}_{2,(m+1)} \\ \vdots \\ \bar{a}_{m,(m+1)} \end{pmatrix}$

The constraints

$$\sum_{j=1}^n a_{ij}x_j = b_i, \quad j = 1, 2, \dots, n$$

in the form,

$$x_i + \sum_{j=m+1}^n \bar{a}_{i,j} = \bar{b}_i, \quad i = 1, 2, \dots, m$$

is said to be canonical. Also we can replace x_i 's in object function to get,

$$f(X) = \sum_{i=1}^m c_i \left(\bar{b}_i - \sum_{j=m+1}^n \bar{a}_{ij} x_j \right) + \sum_{j=m+1}^n c_j x_j = \sum_{i=1}^m c_i \bar{b}_i + \sum_{j=m+1}^n \bar{c}_j x_j$$

where $\bar{c}_j = c_j - \sum_{i=1}^m c_i \bar{a}_{ij}$, $j = m+1, m_2, \dots, n$. From the canonical form, we can compute the corresponding b.f.s., $X = (\bar{b}_1, \bar{b}_2, \dots, \bar{b}_m, 0, 0, \dots, 0)$ and object function $f(X) = \sum_{i=1}^m \bar{b}_i c_i$ since \bar{c}_j are zero for $j > m$. These \bar{c}_j 's are called the **relative cost coefficients**.

15.1.1 Simplex Method

$$\begin{aligned} &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_3 - 5x_4 \\ &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 \leq 1 \\ &\quad \quad \quad x_2 + 4x_3 - 2x_4 \leq 2 \\ &\quad \quad \quad x_1, x_2, x_3, x_4 \geq 0 \end{aligned}$$

Phase	Basis	B	P_1	P_2	P_3	P_4	P_5	P_6
1	x_5	1	4	-1	2	3	1	0
	x_6	2	0	1	0	4	0	1
	f	0	6	-2	4	-5	0	0
2	x_4	$1/3$	$4/3$	$-1/3$	$2/3$	1	$1/3$	0
	x_6	$2/3$	$-16/3$	$7/3$	$-8/3$	0	$-4/3$	1
	f	$5/3$	$38/3$	$-11/3$	$22/3$	0	$5/3$	0
3	x_4	$3/7$	$4/7$	0	$2/7$	1	$1/7$	$1/7$
	x_2	$2/7$	$-16/7$	1	$-8/7$	0	$-4/7$	$3/7$
	f	$19/7$	$30/7$	0	$22/7$	0	$-3/7$	$11/7$
4	x_5	3	4	0	2	7	1	1
	x_2	2	0	1	0	4	0	1
	f	4	6	0	4	3	0	2

Figure 15.1: Simplex Steps

15.1.2 Unrestricted Variable

$$\begin{aligned} &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_3 - 5x_4 \\ &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 \leq 1 \\ &\quad \quad \quad x_2 + 4x_3 - 2x_4 \leq 2 \\ &\quad \quad \quad x_1, x_2, x_4 \geq 0 \text{ and } x_3 \text{ unrestricted} \end{aligned}$$

In above LPP, the variable x_3 is unrestricted. Replace x_3 with two variables x_{31} and x_{32} such that $x_3 = x_{31} - x_{32}$, $x_{31} \geq 0$, $x_{32} \geq 0$. We get,

$$\begin{aligned} &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_{31} - 4x_{32} - 5x_4 \\ &\text{subject to : } 4x_1 - x_2 + 2x_{31} - 2x_{32} - 3x_4 \leq 1 \\ &\quad \quad \quad x_2 + 4x_{31} - 4x_{32} - 2x_4 \leq 2 \\ &\quad \quad \quad x_1, x_2, x_{31}, x_{32}, x_4 \geq 0 \end{aligned}$$

15.1.3 Artificial Variable

$$\begin{aligned} &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_3 - 5x_4 \\ &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 \leq 1 \\ &\quad \quad \quad x_2 - 4x_3 + 2x_4 \leq -2 \\ &\quad \quad \quad x_1 - 2x_2 + x_3 \leq -5 \\ &\quad \quad \quad x_1, x_2, x_3, x_4 \geq 0 \end{aligned}$$

First, we need to change all right-hand side constants into on-negative.

$$\begin{aligned} &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_3 - 5x_4 \\ &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 \leq 1 \\ &\quad \quad \quad -x_2 + 4x_3 - 2x_4 \geq 2 \\ &\quad \quad \quad -x_1 + 2x_2 - x_3 \geq 5 \\ &\quad \quad \quad x_1, x_2, x_3, x_4 \geq 0 \end{aligned}$$

Now we need to introduce three slack variables x_5, x_6, x_7 ,

$$\begin{aligned} &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_3 - 5x_4 \\ &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 + x_5 = 1 \\ &\quad \quad \quad -x_2 + 4x_3 - 2x_4 - x_6 = 2 \\ &\quad \quad \quad -x_1 + 2x_2 - x_3 - x_7 = 5 \\ &\quad \quad \quad x_1, x_2, x_3, x_4, x_5, x_6, x_7 \geq 0 \end{aligned}$$

Two Phase Method

First phase we compute a basic feasible solution (b.f.s.) for the first iteration the simplex method from an auxiliary problem with the artificial variables x_8, x_9 each introduced into ' \geq ' constraints. In second phase, original problem is optimized using simplex method.

If the auxiliary problem does not have a feasible solution in which every artificial variables are zero, then original LPP has no feasible solution.

Phase I: Auxiliary Problem

$$\begin{aligned}
 &\text{minimise : } g(x) = x_8 + x_9 \\
 &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 + x_5 = 1 \\
 &\quad -x_2 + 4x_3 - 2x_4 - x_6 + x_8 = 2 \\
 &\quad -x_1 + 2x_2 - x_3 - x_7 + x_9 = 5 \\
 &\quad x_1, x_2, x_3, x_4, x_5, x_6, x_7, x_8, x_9 \geq 0
 \end{aligned}$$

Phase II: Original Problem with an initial b.f.s. —to be corrected—

$$\begin{aligned}
 &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_3 - 5x_4 \\
 &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 + x_5 = 1 \\
 &\quad -x_2 + 4x_3 - 2x_4 - x_6 + x_8 = 2 \\
 &\quad -x_1 + 2x_2 - x_3 - x_7 + x_9 = 5 \\
 &\quad x_1, x_2, x_3, x_4, x_5, x_6, x_7, x_8, x_9 \geq 0
 \end{aligned}$$

Big M Method

In Big M method, arbitrarily large multiples of artificial variables are added to the object function. If the optimal solution of the modified problem has any artificial variables with nonzero value, then it has no feasible solution.

$$\begin{aligned}
 &\text{minimise : } f(x) = 6x_1 - 2x_2 + 4x_3 - 5x_4 + 100x_8 + 100x_9 \\
 &\text{subject to : } 4x_1 - x_2 + 2x_3 - 3x_4 + x_5 = 1 \\
 &\quad -x_2 + 4x_3 - 2x_4 - x_6 + x_8 = 2 \\
 &\quad -x_1 + 2x_2 - x_3 - x_7 + x_9 = 5 \\
 &\quad x_1, x_2, x_3, x_4, x_5, x_6, x_7, x_8, x_9 \geq 0
 \end{aligned}$$

15.2 Integer Programming

Definitions 15.2.1. Integer Linear Programming Problem(ILPP) is the integral conditioning of an optimization of a linear object function under certain linear equality/inequality constraints.

General ILPP A general ILPP is of the form,

$$\text{minimise : } f(X) = CX \text{ (object function)} \quad (15.7)$$

$$\text{subject to : } AX = B \text{ (constraints)} \quad (15.8)$$

$$\text{such that } X \geq 0 \text{ (non-negativity conditions)} \quad (15.9)$$

$$\text{and } X \text{ is an integer vector. (integer condition)} \quad (15.10)$$

Remark. Let T_F be the set of all integer solutions. Let $[T_F]$ be the convex hull of T_F . Let S_F be the set of all real solutions. Then, $T_F \subset [T_F] \subset S_F$.

Finding optimal solution from T_F is a much more difficult than finding optimal solution from $[T_F]$. Using Simplex method on $[T_F]$, we can find the optimal integer solutions.

15.3 Networks

15.4 Non-linear Programming

Semester IV

Semester 4

Subject 16

ME010401 Spectral Theory

Module 1

16.4 Normed and Banach Spaces

Algebraically Reflexive A vector space X is algebraically reflexive if the canonical mapping $C : X \rightarrow X^{**}$ defined by $C(x) = g_x \in X^{**}$ is surjective where X^*, X^{**} are the first and second algebraic dual spaces of X and $\forall f \in X^*, g_x(f) = f(x)$. [?, §2.8]

Finite dimensional vector spaces are algebraically reflexive. [?, §2.9.3]

16.4.6 Reflexive Spaces

§4.6 Summary

1. reflexive \implies completeness, However completeness $\not\Rightarrow$ reflexive
2. finite dimensional \implies reflexive
3. Hilbert spaces are reflexive
4. Separable space with non-separable dual space is non-reflexive

Lemma 16.4.1. For every fixed $x \in X$, the functional $g_x : X' \rightarrow \mathbb{R}$ defined by $g_x(f) = f(x)$ is a bounded linear functional on X' so that $g_x \in X''$, and has the norm

$$\|g_x\| = \|x\|$$

Proof. From Hahn-Banach theorem for normed spaces, we know that for any $x \in X$

$$\|x\| = \sup_{\substack{f \in X' \\ f \neq 0}} \frac{|f(x)|}{\|f\|} = \sup_{\substack{f \in X' \\ f \neq 0}} \frac{|g_x(f)|}{\|f\|} = \|g_x\| \quad (16.1)$$

Clearly, g_x is bounded, $g_x \in X''$ and $\|g_x\| = \|x\|$. □

Lemma 16.4.2. *The Canonical mapping $C : X \rightarrow X''$ given by $C(x) = g_x$ such that $\forall f \in X'$, $g_x(f) = f(x)$ is an isomorphism of X onto $\mathcal{R}(C)$.*

Proof. **Step 1 : C is linear**

$$g_{\alpha x + \beta y} = f(\alpha x + \beta y) = \alpha f(x) + \beta f(y) = \alpha g_x(f) + \beta g_y(f) \quad (16.2)$$

Therefore, C is linear.

Step 2 : C is injective

We have, C is linear. Thus, $g_{x-y} = g_x - g_y$. Then,

$$\|C(x) - C(y)\| = \|g_x - g_y\| = \|g_{x-y}\| = \|x - y\| \quad (16.3)$$

Thus, C is an isometry. If $C(x) = C(y)$, then $\|C(x) - C(y)\| = 0$. Thus, $\|x - y\| = 0$ implies $x = y$. Therefore, C is injective. Therefore, C is an isomorphism from X onto $\mathcal{R}(C)$. \square

Definitions 16.4.1 (Reflexive). A normed space X is reflexive if the canonical mapping $C : X \rightarrow X''$ defined by $C(x) = g_x$ such that $\forall f \in X'$, $g_x(f) = f(x)$ is surjective where X', X'' are the first and second dual spaces of X .

Example, \mathbb{R}^n is reflexive since $\mathbb{R}^{n''} = \mathbb{R}^{n'} = \mathbb{R}^n$.

Theorem 16.4.3. *Reflexive \implies Completeness*

Proof. Let X be a reflexive normed space. Then $C : X \rightarrow X''$ is surjective and $X \simeq X''$. We know that, X'' is a dual space of X' . Thus, X'' is complete since $B(X', K)$ is a Banach space. Therefore, X is complete. \square

Theorem 16.4.4. *Finite dimensional \implies reflexive*

Proof. Let X be a finite dimensional normed space. Then every linear operator $T : X \rightarrow K$ is bounded. Thus, $X' = X^*$ and $\dim X = \dim X^*$. Thus, $X'' = X^{**}$. We know that, finite dimensional normed spaces are algebraically reflexive. Thus, $\mathcal{R}(C) = X^{**} = X''$. Therefore, X is reflexive. \square

Theorem 16.4.5. *Every Hilbert space is reflexive.*

Proof. Let H be a Hilbert space. Define $A : H' \rightarrow H$, $A(f) = z$ such that $f(x) = \langle x, z \rangle$. By Riesz representation theorem, A is well-defined, bijective and conjugate linear. That is, $A(\alpha f) = \bar{\alpha} A f$.

Claim : H' is a Hilbert space with inner product,

$$\langle f_1, f_2 \rangle = \langle A f_2, A f_1 \rangle = \langle z_2, z_1 \rangle \quad (16.4)$$

IP1

$$\begin{aligned} \langle f_1 + f_2, f_3 \rangle &= \langle A f_3, A(f_1 + f_2) \rangle = \langle A f_3, A f_1 + A f_2 \rangle \\ &= \langle A f_3, A f_1 \rangle + \langle A f_3, A f_2 \rangle = \langle f_1, f_3 \rangle + \langle f_2, f_3 \rangle \end{aligned}$$

$$\text{IP2 } \langle \alpha f_1, f_2 \rangle = \langle A f_2, A(\alpha f_1) \rangle = \langle A f_2, \bar{\alpha} A f_1 \rangle = \alpha \langle A f_2, A f_1 \rangle$$

$$\text{IP3 } \langle f_1, f_2 \rangle = \langle A f_2, A f_1 \rangle = \overline{\langle A f_1, A f_2 \rangle} = \overline{\langle f_2, f_1 \rangle}$$

$$\begin{aligned} \text{IP4 } \langle f_1, f_1 \rangle = \langle Af_1, Af_1 \rangle \geq 0 \text{ and} \\ \langle f_1, f_1 \rangle = 0 \iff \langle Af_1, Af_1 \rangle = 0 \iff Af_1 = 0 \iff f_1 = 0 \end{aligned}$$

Similarly, let $g \in H''$ and $g(f) = \langle f, f_0 \rangle$ be its Riesz representation. Then,

$$g(f) = \langle f, f_0 \rangle = \langle Af_0, Af \rangle$$

We have, $f(x) = \langle x, z \rangle$. Let $Af_0 = x$. Then,

$$f(x) = \langle x, z \rangle = \langle Af_0, Af \rangle$$

Thus, there exists $x \in H$ such that $g(f) = f(x)$. Therefore, $C : H \rightarrow H''$ is surjective. Thus, $\mathcal{R}(C) = H''$ and H is reflexive. \square

Lemma 16.4.6. *Let Y be a proper, closed subspace of a normed space X . Let $x_0 \in X - Y$ and $\delta = \inf_{y \in Y} \|y - x_0\|$. Then, there exists $f \in X'$ such that*

$$\|f\| = 1, \quad f(y) = 0, \quad \forall y \in Y, \quad f(x_0) = \delta \quad (16.5)$$

Proof. Let Y be a proper, closed subspace of X . Let $x_0 \in X - Y$ and $\delta = d(x_0, Y) = \inf_{y \in Y} \|x_0 - y\|$. Consider the subspace Z spanned by Y and x_0 . Then, every element $z \in Z$ is of the form $z = y + \alpha x_0$ where $\alpha \in K$.

Consider $f : Z \rightarrow K$ be defined by $f(y + \alpha x_0) = \alpha \delta$. Clearly, we have $f(y) = f(y + 0x_0) = 0$, $\forall y \in Y$ and $f(x_0) = f(0 + 1x_0) = \delta$.

Step 1 : f is linear.

Let $z_1 = y_1 + \alpha_1 x_0$ and $z_2 = y_2 + \alpha_2 x_0$. Then, $f(z_1) = \alpha_1 \delta$ and $f(z_2) = \alpha_2 \delta$.

$$\begin{aligned} f(\beta_1 z_1 + \beta_2 z_2) &= f(\beta_1 y + \beta_1 \alpha_1 x_0 + \beta_2 y + \beta_2 \alpha_2 x_0) \\ &= f((\beta_1 + \beta_2)y + (\beta_1 \alpha_1 + \beta_2 \alpha_2)x_0) \\ &= (\beta_1 \alpha_1 + \beta_2 \alpha_2)\delta = \beta_1 \alpha_1 \delta + \beta_2 \alpha_2 \delta = \beta_1 f(z_1) + \beta_2 f(z_2) \end{aligned}$$

Thus, f is linear.

Step 2 : f is bounded.

Suppose $z = y + \alpha x_0$.

$$|f(z)| = |\alpha| \delta = |\alpha| \inf_{y' \in Y} \|y' - x_0\|$$

By the definition of infimum, we have $\inf_{y' \in Y} \|y' - x_0\| \leq \|y - x_0\|$, $\forall y \in Y$.

Consider $-y/\alpha \in Y$.

$$|f(z)| \leq |\alpha| \left\| \frac{-y}{\alpha} - x_0 \right\| \leq \| -y - \alpha x_0 \| \leq \|z\|$$

Thus, f is a bounded and $\|f\| \leq 1$.

By the definition of infimum, there exists a sequence $\{y_n\}$ such that $z_n = y_n - x_0$ and $\|z_n\| = \|y_n - x_0\| \rightarrow \delta$ as $n \rightarrow \infty$. Clearly, $f(z_n) \rightarrow -\delta$ as $n \rightarrow \infty$. Then,

$$\|f\| = \sup_{z \in Z} \frac{f(z)}{\|z\|} \geq \frac{f(z_n)}{\|z_n\|}$$

Thus, $\|f\| \geq 1$ as $n \rightarrow \infty$. Therefore, we have $\|f\| = 1$. By Hahn-Banach theorem, we can extend f to X without increasing the norm. \square

Theorem 16.4.7. *If X' is separable, then X is separable.*

Proof. Suppose X' is separable. Consider the countable dense subset of the unit sphere, $\{f_n\}$. Clearly, $\|f_n\| = \sup_{\|x\|=1} |f_n(x)| = 1$. By the definition of supremum, there exists a sequence $\{x_n\}$ in X such that $\|x_n\| = 1$ and $\|f_n(x_n)\| \geq \frac{1}{2}$.

Suppose X is not separable. Let Y be the span of $\{x_n\}$. Then Y is a separable subspace of the non-separable space X . Since Y is closed, there exists bounded, linear functional $f \in X'$ such that $\|f\| = 1$, $f(y) = 0, \forall y \in Y$ and $f(x_n) = 0, \forall n$.

$$\begin{aligned} \frac{1}{2} &\leq \|f_n(x_n)\| \leq \|f_n(x_n) - f(x_n)\| \\ &\leq \|(f_n - f)(x_n)\| \leq \|f_n - f\| \|x_n\| \leq \|f_n - f\| \end{aligned}$$

Then, $\{f_n\}$ is not dense in unit sphere of X' which is a contradiction. Therefore, X is separable. \square

Remark. Let X be separable space with a non-separable dual space X' . Then X is non-reflexive.

Proof. Let X be a separable space with non-separable dual space X' . Suppose X is reflexive. Then $X \simeq X''$ and X'' is separable. However, X' is separable since X'' is separable, which is a contradiction. Therefore, X is not reflexive. \square

l^1 is non-reflexive.

Proof. l^1 is separable and $l^{1'} = l^\infty$ is non-separable. Thus, l^1 is non-reflexive. \square

Problems

1. What the functionals f, g_x if $X = \mathbb{R}^n$.

Answer. Every linear functional on finite dimensional space is bounded. Let $x = (x_1, x_2, \dots, x_n)$. A linear functional $f : \mathbb{R}^n \rightarrow \mathbb{R}$ is of the form,

$$f(x) = a_0 + a_1x_1 + \dots + a_nx_n, \quad a_k \in \mathbb{R}$$

Let $F = \{f : \mathbb{R}^n \rightarrow \mathbb{R} : f \text{ is linear}\}$. A linear functional $g_x : F \rightarrow \mathbb{R}$ is of the form

$$g_x(f) = a_0 + a_1f_1(x) + a_2f_2(x) + \dots + a_nf_n(x) \text{ where } f_j(e_i) = \delta_{ij},$$

\square

2. Prove lemma 4.6-7 for Hilbert space X

Answer. □

3. If a normed space X is reflexive, show that X' is reflexive

Answer. □

4. Show that a Banach space X is reflexive if and only if its dual space X' is reflexive.

Answer. □

5. Let Y be a proper, closed subspace of a normed space X . Let $x_0 \in X - Y$. Then there exists a bounded, linear functional h on X such that $\|h\| = \frac{1}{\delta}$, $h(y) = 0$, $\forall y \in Y$ and $h(x_0) = 1$.

Answer. □

6. Show that different closed subspaces Y_1, Y_2 of a normed space X have different annihilators.

Answer. □

7. Let Y be a closed subspace of a normed space X such that every $f \in X'$ which is zero everywhere on Y is zero everywhere on the whole space X . Show that then $Y = X$.

Answer. □

8. Let M be any subset of a normed space X . Show that an $x_0 \in X$ is an element of $A = \underline{\text{span } M}$ if and only if $f(x_0) = 0$, $\forall f \in X'$ such that $f|_M = 0$.

Answer. □

9. Show that a subset M of a normed space X is total in X if and only if every $f \in X'$ which is zero everywhere on M is zero everywhere in X .

Answer. □

10. Show that if a normed space X has a linearly independent subset of n elements, so does the dual space X' .

Answer. □

16.4.7 Category Theorem

rare A subset M in X is rare if its closure has no interior points.

meager A subset M is meager in X if M is union of countable family of rare subsets of X .

non-meager A subset M is non-meager in X if it is not meager in X .

Theorem 16.4.8 (Baire's Category Theorem). *Every complete, non-empty metric space is nonmeager in itself.*

Proof. Let X be a complete, non-empty metric space. Suppose X is meager in itself. Then, there exists a countable family of rare sets M_k such that

$$X = \bigcup_{k=1}^{\infty} M_k$$

We have, M_1 is rare in X and $\bar{M}_1 \neq X$. Thus, $\bar{M}_1^c \neq \phi$ and \bar{M}_1^c is an open set. Let $p_1 \in \bar{M}_1^c$ such that

$$p_1 \in B_1 = B(p_1, \varepsilon_1) \subset \bar{M}_1^c \text{ where } \varepsilon_1 < \frac{1}{2}$$

We have, M_2 is rare in X . Then $\bar{M}_2^c \cap B(p_1, \frac{\varepsilon_1}{2}) \neq \phi$. Suppose $\bar{M}_2^c \cap B(p_1, \frac{\varepsilon_1}{2}) = \phi$. Then $B(p_1, \frac{\varepsilon_1}{2}) \subset \bar{M}_2$ which is a contradiction.

Let $p_2 \in \bar{M}_2^c \cap B(p_1, \frac{\varepsilon_1}{2})$ such that

$$p_2 \in B_2 = B(p_2, \varepsilon_2) \subset \bar{M}_2^c \cap B(p_1, \frac{\varepsilon_1}{2}) \text{ where } \varepsilon_2 < \frac{\varepsilon_1}{2}$$

Then, $\bar{M}_3^c \cap B(p_2, \frac{\varepsilon_2}{2}) \neq \phi$. Let $p_3 \in \bar{M}_3^c \cap B(p_2, \frac{\varepsilon_2}{2})$ such that

$$p_3 \in B_3 = B(p_3, \varepsilon_3) \subset \bar{M}_3^c \cap B(p_2, \frac{\varepsilon_2}{2}) \text{ where } \varepsilon_3 < \frac{\varepsilon_2}{2}$$

Continuing like this, we get a sequence $\{p_n\}$ such that

$$\forall k \in \mathbb{N}, p_k \in B_k = B(p_k, \varepsilon_k) \subset \bar{M}_k^c \cap B(p_{k-1}, \frac{\varepsilon_{k-1}}{2}) \text{ where } \varepsilon_k < \frac{\varepsilon_{k-1}}{2}$$

We have,

$$B_{k+1} \subset B(p_k, \frac{\varepsilon_k}{2}) \subset B_k, \quad \varepsilon_k < \frac{1}{2^k}$$

This sequence is Cauchy since the open ball $B_{k+1} \subset B_k \subset \dots \subset B_3 \subset B_2 \subset B_1$ and the radius of the open balls, $\varepsilon_k < 2^{-k}$ reduces at least by half in each successive step.

Since X is complete, the sequence $\{p_n\}$ converges to $p \in X$. □

Theorem 16.4.9 (Uniform Boundedness). *Let X be a Banach space and Y be a normed space. Let $\{T_n\}$ be sequence of bounded, linear operators $T_n : X \rightarrow Y$ such that $\{\|T_n(x)\|\}$ is bounded for every $x \in X$. Then the sequence of norms $\{\|T_n\|\}$ is bounded.*

Proof. Let X be a Banach space. Let X

□

Remark. Space of polynomials $x(t) = \sum_j \alpha_j t^j$ with norm $\|x\| = \max_j |\alpha_j|$ is not complete.

Proof.

□

Remark. There exist real-valued continuous functions whose Fourier series diverge at a given point t_0 .

Proof.

□

16.4.8 Strong and Weak Convergence

16.4.9 Convergence of Sequences of Operators and Functionals

16.4.12 Open Mapping Theorem

16.4.13 Closed Linear Operators

Module 2

16.5 Linear Operators**16.5.1 Banach Fixed Point Theorem****16.7 Spectral Theory of Linear Operators in Normed Spaces****16.7.1 Spectral Theory in Finite Dimensional Normed Spaces****16.7.2 Basic Concepts****16.7.3 Spectral Properties of Bounded Linear Operators****16.7.4 Further Properties of Resolvent and Spectrum****Resolvent Equation**

Theorem 16.7.1. *Let X be a complex Banach space. Let $T \in B(X, X)$ and $\lambda, \mu \in \rho(T)$. Then*

1. $R_\mu - R_\lambda = (\mu - \lambda)R_\mu R_\lambda$
2. R_λ commutes with any $S \in B(X, X)$ which commutes with T
3. $R_\mu R_\lambda = R_\lambda R_\mu$

Proof. **Part 1 :** Hilbert Relation/Resolvent Equation

$$\begin{aligned}
 R_\mu - R_\lambda &= R_\mu(T_\lambda R_\lambda) - (R_\mu T_\mu)R_\lambda \\
 &= R_\mu(T_\lambda R_\lambda - T_\mu R_\lambda) \\
 &= R_\mu(T_\lambda - T_\mu)R_\lambda \\
 &= R_\mu(T - \lambda I - T + \mu I)R_\lambda \\
 &= R_\mu(\mu - \lambda)R_\lambda
 \end{aligned}$$

Part 2 : Commutativity Property 1

Suppose T commutes with S , that is $TS = ST$. Let $\lambda \in \rho(T)$. Then

$$T_\lambda S = (T - \lambda I)S = TS - \lambda S = ST - \lambda S = S(T - \lambda I) = ST_\lambda$$

$$\begin{aligned}
 R_\lambda S &= R_\lambda S(T_\lambda R_\lambda) && \because T_\lambda R_\lambda = I \\
 &= R_\lambda T_\lambda S R_\lambda && \because ST_\lambda = T_\lambda S \\
 &= S R_\lambda
 \end{aligned}$$

That is, T commutes with R_λ .

Part 3 : Commutativity Property 2

T commutes with $T \implies T$ commutes with R_μ by part 2.

T commutes with $R_\mu \implies R_\mu$ commutes with R_λ by part 2. □

Spectral Mapping Theorem - Polynomials

Theorem 16.7.2. *Let X be a complex Banach space. Let $T \in B(X, X)$ and p be a complex polynomial given by*

$$p(\lambda) = \alpha_n \lambda^n + \alpha_{n-1} \lambda^{n-1} + \cdots + \alpha_0, \quad \alpha_j \in \mathbb{C}, \quad \alpha_n \neq 0$$

Then,

$$p(\sigma(T)) = \sigma(p(T)) \quad (16.6)$$

Proof. The result is trivial when $\sigma(T) = \phi$ or $n = 0$. Suppose $\sigma(T) \neq \phi$ and $n > 0$.

Part 1 : $\sigma(p(T)) \subset p(\sigma(T))$

Let $\mu \in \sigma(p(T))$. Define $S = p(T) - \mu I$. Then $\mu \in \sigma(S)$ and $S_\mu = p(T) - \mu I$. Define $s_\mu(\lambda) = p(\lambda) - \mu$. Being a complex polynomial, $s_\mu(\lambda)$ must factorise into linear factors. Thus,

$$s_\mu(\lambda) = \alpha_n (\lambda - \gamma_1)(\lambda - \gamma_2) \cdots (\lambda - \gamma_n)$$

And,

$$S_\mu = \alpha_n (T - \gamma_1 I)(T - \gamma_2 I) \cdots (T - \gamma_n I)$$

Suppose $\gamma_k \in \rho(T)$ for every k . Then $(T - \gamma_k I)$ has bounded inverse and consequently S_μ will have a bounded inverse,

$$S_\mu^{-1} = \frac{1}{\alpha_n} (T - \gamma_1 I)^{-1} (T - \gamma_2 I)^{-1} \cdots (T - \gamma_n I)^{-1}$$

This is a contradiction, since $\mu \in \sigma(S)$. Thus there exists some $\gamma_k \in \sigma(T)$.

$$s_\mu(\gamma_k) = p(\gamma_k) - \mu = 0$$

Thus, $\mu = p(\gamma_k)$ where $\gamma_k \in \sigma(T)$. Therefore, $\mu \in p(\sigma(T))$.

Part 2 : $p(\sigma(T)) \subset \sigma(p(T))$

Let $\kappa \in p(\beta)$ where $\beta \in \sigma(T)$. Now, there are two alternate cases.

Part 2A : $(T - \beta I)$ has no inverse

Suppose $(T - \beta I)$ has no inverse. Define $s_\kappa(\lambda) = p(\lambda) - \kappa = (\lambda - \beta)g(\lambda)$. And $S_\kappa = (T - \beta I)g(T)$. Clearly, $g(T)$ commutes with $(T - \beta I)$. Thus, $S_\kappa = g(T)(T - \beta I)$. If S_κ has an inverse, then

$$I = (T - \beta I)g(T)S_\kappa^{-1} = S_\kappa^{-1}g(T)(T - \beta I)$$

Then, $(T - \beta I)$ has an inverse which is a contradiction. Therefore S_κ does not have an inverse. Thus, $\kappa \in \sigma(p(T))$.

Part 2B : $(T - \beta I)$ has an inverse

Suppose $(T - \beta I)$ has an inverse. We have, $\mathcal{R}(T - \beta I) \neq X$. If not, the map is surjective and by bounded inverse theorem, $(T - \beta I)^{-1}$ is bounded. Then, β is a regular point of T . That is, $\beta \notin \sigma(T)$. This is a contradiction.

We have $S_\kappa = (T - \beta I)g(T)$. Thus $\mathcal{R}(S_\kappa) \neq X$. And $\kappa \in \sigma(p(T))$. □

Linear Independence

Theorem 16.7.3. *Let T be a linear operators and $\lambda_1, \lambda_2, \dots$ be its eigenvalues. Then the eigenvectors x_1, x_2, \dots, x_n corresponding to different eigenvalues $\lambda_1, \lambda_2, \dots, \lambda_n$ form a linearly independent set.*

Proof. Suppose $\{x_1, x_2, \dots, x_n\}$ is linearly dependent. Let x_m be the first vector such that $\{x_1, x_2, \dots, x_{m-1}, x_m\}$ is linearly dependent.

$$x_m = \alpha_1 x_1 + \alpha_2 x_2 + \dots + \alpha_{m-1} x_{m-1}$$

Then,

$$\begin{aligned} (T - \lambda_m I)x_m &= \sum_{j=1}^{m-1} \alpha_j (T - \lambda_m I)x_j \\ 0 &= \sum_{j=1}^{m-1} \alpha_j (\lambda_j - \lambda_m)x_j \end{aligned}$$

This is a contradiction to the assumption that $\{x_1, x_2, \dots, x_{m-1}\}$ is linearly independent. Thus, no such m exists. Therefore, $\{x_1, x_2, \dots, x_n\}$ is linearly independent. \square

16.7.5 Use of Complex Analysis in Spectral Theory

Module 3

16.7.6 Banach Algebra

16.7.7 Further Properties of Banach Algebra

16.8 Compact Linear Operators

16.8.1 Compact Linear Operators on Normed Spaces

16.8.2 Further Properties of Compact Linear Operators

**16.8.3 Spectral Properties of Compact Linear Operators
on Normed Spaces**

**16.8.4 Further Spectral Properties of Compact Linear Op-
erators**

Module 4

16.9 Bounded, Self-Adjoint Linear Operators

16.9.1 Spectral Properties of Bounded Self-Adjoint Linear Operators

Note : A bounded, self-adjoint operator may not have eigenvalues.

Theorem 16.9.1. *Let H be a complex, Hilbert space. Let $T : H \rightarrow H$ be a bounded, self-adjoint linear operator. Then*

1. *All the eigenvalues(if they exist) are real.*
2. *Eigenvectors corresponding to different eigenvalues are orthogonal.*

Proof. Part 1 : Eigenvalues are real.

Let λ be an eigenvalue of T and $x \in H$ such that $Tx = \lambda x$, ($x \neq 0$). Then,

$$\lambda \langle x, x \rangle = \langle \lambda x, x \rangle = \langle Tx, x \rangle = \langle x, Tx \rangle = \langle x, \lambda x \rangle = \bar{\lambda} \langle x, x \rangle$$

Since $x \neq 0$, we have $\langle x, x \rangle \neq 0$ and $\lambda \in \mathbb{R}$.

Part 2 : Eigenvectors are orthogonal.

Let λ, μ be eigenvalues of T and x, y be corresponding eigenvectors. That is, $Tx = \lambda x$ and $Ty = \mu y$.

$$\lambda \langle x, y \rangle = \langle \lambda x, y \rangle = \langle Tx, y \rangle = \langle x, Ty \rangle = \langle x, \mu y \rangle = \bar{\mu} \langle x, y \rangle = \mu \langle x, y \rangle$$

Since $\lambda \neq \mu$, we have $\langle x, y \rangle = 0$. That is, $x \perp y$. \square

Theorem 16.9.2. *Let $T : H \rightarrow H$ be a bounded, self-adjoint linear operator. Then $\lambda \in \rho(T)$ if and only if there exists $c > 0$ such that $\|T_\lambda x\| \geq c\|x\|$.*

Proof. Let $\lambda \in \rho(T)$. Then $R_\lambda = T_\lambda^{-1} : H \rightarrow H$ exists and is bounded. Since $R_\lambda \neq 0$, $\|R_\lambda\| = k > 0$.

$$\|x\| = \|R_\lambda T_\lambda x\| \leq \|R_\lambda\| \|T_\lambda x\| = k \|T_\lambda x\|$$

Thus,

$$\|T_\lambda x\| \geq c\|x\| \text{ where } c = \frac{1}{k} \quad (16.7)$$

Converse Part

Suppose $T_\lambda x_1 = T_\lambda x_2$. Then,

$$0 = \|T_\lambda x_1 - T_\lambda x_2\| = \|T_\lambda(x_1 - x_2)\| \geq c\|x_1 - x_2\|$$

Since $c > 0$, we have $\|x_1 - x_2\| = 0$ and $x_1 = x_2$. Therefore, $T_\lambda : H \rightarrow T_\lambda(H)$ is bijective.

Let $x_0 \perp \overline{T_\lambda(H)}$. Then $x_0 \perp T_\lambda(H)$.

$$0 = \langle T_\lambda x, x_0 \rangle = \langle Tx, x_0 \rangle - \langle \lambda x, x_0 \rangle = \langle Tx, x_0 \rangle - \lambda \langle x, x_0 \rangle$$

$$\langle x, \bar{\lambda}x_0 \rangle = \lambda \langle x, x_0 \rangle = \langle Tx, x_0 \rangle = \langle x, Tx_0 \rangle$$

Thus, $Tx_0 = \bar{\lambda}x_0$.

Suppose $x_0 \neq 0$. Then, $\bar{\lambda}$ is an eigenvalue of T . We know that, the eigenvalues of bounded, self-adjoint linear operators are real. Thus, $\lambda \in \mathbb{R}$. That is, $Tx_0 = \lambda x_0$. Then, $T_\lambda x_0 = Tx_0 - \lambda x_0 = 0$ which is a contradiction since $0 = \|T_\lambda x_0\| \geq c\|x_0\| > 0$. Therefore, $x_0 = 0$.

By projection theorem, $H = Y \oplus Z$ where $Y = \overline{T_\lambda(H)}$, $Z = Y^\perp = \{0\}$. Thus, $\overline{T_\lambda(H)} = H$.

Suppose $y \in \overline{T_\lambda(H)}$. Then there exists a sequence $\{y_n\}$ in $T_\lambda(H)$ such that $y_n \rightarrow y$. Since $y_n \in T_\lambda(H)$, there exists $x_n \in H$ such that $y_n = T_\lambda x_n$. The sequence $\{x_n\}$ is Cauchy since

$$\|x_n - x_m\| \leq \frac{1}{c} \|T_\lambda(x_n - x_m)\| = \frac{1}{c} \|y_n - y_m\|$$

We know that, $\{y_n\}$ is convergent and thus Cauchy. Thus, sequence $\{x_n\}$ is Cauchy in H . Since H is complete, $x_n \rightarrow x \in H$. And T is continuous. Therefore, $y_n = T_\lambda x_n \rightarrow T_\lambda x$. Since the limit is unique, $T_\lambda x = y$. And $y \in T_\lambda(H)$. Therefore, $T_\lambda(H)$ is closed. \square

Theorem 16.9.3 (Spectrum). *Let $T : H \rightarrow H$ be a bounded, self-adjoint linear operator on a complex Hilbert space H . Then the spectrum $\sigma(T)$ of T is real.*

Proof. Let $\lambda \in \sigma(T)$ where $\lambda = \alpha + i\beta$.

$$\langle T_\lambda x, x \rangle = \langle Tx, x \rangle - \lambda \langle x, x \rangle$$

We know that $\langle Tx, x \rangle \in \mathbb{R}$ since T is self-adjoint. Thus,

$$\overline{\langle T_\lambda x, x \rangle} = \overline{\langle Tx, x \rangle} - \bar{\lambda} \overline{\langle x, x \rangle} = \langle Tx, x \rangle - \bar{\lambda} \langle x, x \rangle$$

Then, the imaginary part of $\langle T_\lambda x, x \rangle$ can be obtained as,

$$-2i \operatorname{Im} \langle T_\lambda x, x \rangle = \overline{\langle T_\lambda x, x \rangle} - \langle T_\lambda x, x \rangle = (\lambda - \bar{\lambda}) \langle x, x \rangle = 2i\beta \|x\|^2$$

The absolute value of imaginary part is always less than or equal to the absolute of the complex number. And by Schwarz inequality,

$$|\beta| \|x\|^2 = |\operatorname{Im} \langle T_\lambda x, x \rangle| \leq |\langle T_\lambda x, x \rangle| \leq \|T_\lambda x\| \|x\|$$

Then, $|\beta| \|x\| \leq \|T_\lambda x\|$. Suppose $\beta \neq 0$. Then we have $|\beta| = c > 0$ such that $\langle T_\lambda x, x \rangle \geq c\|x\|$. That is, $\lambda \in \rho(T)$ which is a contradiction. Therefore, $\beta = 0$ and $\lambda \in \mathbb{R}$. \square

Exercise

- 1.
- 2.

3. R_λ is bounded since

$$\|x\| = \|T_\lambda R_\lambda x\| \geq c\|R_\lambda x\|$$

4. $T : H \rightarrow H$ defined by $Tx = \lambda_0 x$. Then $c \leq |\lambda - \lambda_0|$.
5. $\langle W^*TWx, x \rangle = \langle TWx, Wx \rangle = \langle Wx, T^*Wx \rangle = \langle Wx, TWx \rangle = \langle x, W^*TWx \rangle$
6. T is bounded, but not self-adjoint. And $S : l^2 \rightarrow l^2$ is defined by $(\xi_1, \xi_2, \xi_3, \dots) \rightarrow (0, \xi_1, \xi_2, \xi_3, \dots)$.
- 7.
8. closure or exterior ? (I don't get it)
9. Let $x, x' \in L^2[0, 1]$ be real-valued functions which are Lebesgue square-integrable. Then $\bar{x}(t) = x(t)$ and $\bar{x}'(t) = x'(t)$.

$$\|Tx(t)\| = \|y(t)\| = \|tx(t)\| \leq \|x(t)\| \quad \because t \in [0, 1]$$

$$\langle Tx(t), x'(t) \rangle = \int_0^1 tx(t)\bar{x}'(t) dt = \langle \bar{x}(t), T\bar{x}'(t) \rangle = \langle x(t), Tx'(t) \rangle$$

10.

- 16.9.2 Further Spectral Properties of Bounded Self-Adjoint Linear Operators**
- 16.9.3 Positive Operators**
- 16.9.5 Projection Operators**
- 16.9.6 Further Properties of Projections**

Subject 17

**ME010402 Analytic
Number Theory**

Subject 18

ME800401 Differential Geometry

18.1 Graphs and Level Set

Definitions 18.1.1. Let function $f : U \rightarrow \mathbb{R}$ where $U \subset \mathbb{R}^{n+1}$. Let c be a real number. Then the **Level set** of f at height c is the set of all points in U with image c .

$$f^{-1}(c) = \{(x_1, x_2, \dots, x_{n+1}) \in U : f(x_1, x_2, \dots, x_{n+1}) = c\} \quad (18.1)$$

Definitions 18.1.2. Let function $f : U \rightarrow \mathbb{R}$ where $U \subset \mathbb{R}^{n+1}$. Then,

$$\text{graph}(f) = \{(x_1, x_2, \dots, x_{n+2}) \in \mathbb{R}^{n+2} : f(x_1, x_2, \dots, x_{n+1}) = x_{n+2}\} \quad (18.2)$$

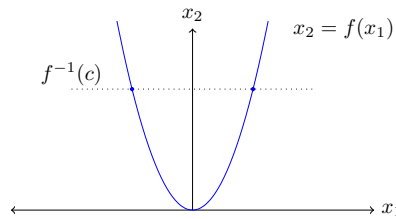
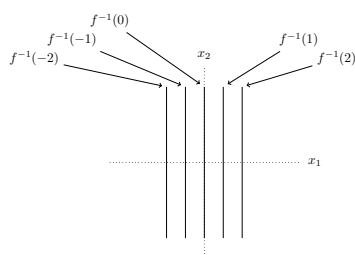
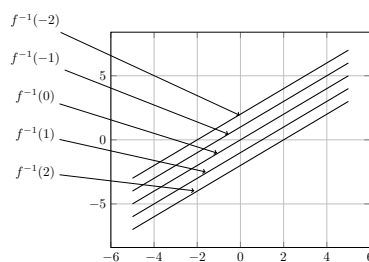
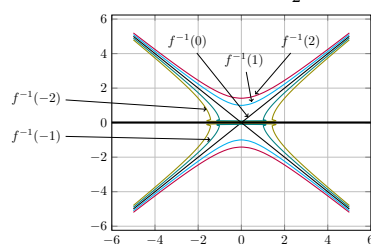


Figure 18.1: Graph of $f(x_1) = x_1^2$ and Level set $f^{-1}(c)$

Exercise

In exercise 1.1-1.4, sketch typical level curves and the graph of each function

1. $f(x_1, x_2) = x_1$
2. $f(x_1, x_2) = x_1 - x_2$
3. $f(x_1, x_2) = x_1^2 - x_2^2$

(a) Level Sets for $f(x_1, x_2) = x_1$ (b) Level Set for $f(x_1, x_2) = x_1 - x_2$ (c) Level Set for $f(x_1, x_2) = x_1^2 - x_2^2$

4. $f(x_1, x_2) = 3r^8 - 8r^6 + 6r^4$ where $r^2 = x_1^2 + x_2^2$.

In exercise 1.5-1.9, sketch the level sets of $f^{-1}(c)$ for $n = 0, 1$ and 2 of each function at the heights indicated

5. $f(x_1, x_2, \dots, x_{n+1}) = x_{n+1}$; $c = -1, 0, 1, 2$
6. $f(x_1, x_2, \dots, x_{n+1}) = 0x_1^2 + x_2^2 + \dots + x_{n+1}^2$; $c = 0, 1, 4$
7. $f(x_1, x_2, \dots, x_{n+1}) = x_1 - x_2^2 - \dots - x_{n+1}^2$; $c = -1, 0, 1, 2$
8. $f(x_1, x_2, \dots, x_{n+1}) = x_1^2 - x_2^2 - \dots - x_{n+1}^2$; $c = -1, 0, 1$
9. $f(x_1, x_2, \dots, x_{n+1}) = x_1^2 + x_2^2/4 + \dots + x_{n+1}^2/(n+1)^2$; $c = 1$
10. Show that graph of any function $f : \mathbb{R} \rightarrow \mathbb{R}$ is a level set for some function $F : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$.

Solution : Let $f : \mathbb{R}^n \rightarrow \mathbb{R}$ and $F : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$ such that

$$F(x_1, x_2, \dots, x_{n+1}) = x_{n+1} - f(x_1, x_2, \dots, x_n)$$

Then, $F^{-1}(0) = \text{graph}(f)$.

18.2 Vector Fields

Definitions 18.2.1. A vector \mathbf{v} at a point $p \in \mathbb{R}^{n+1}$ is a pair $\mathbf{v} = (p, v)$ where $v \in \mathbb{R}^{n+1}$.

vector addition $\mathbf{v} + \mathbf{w} = (p, v) + (p, w) = (p, v + w)$.

scalar multiplication Let $c \in \mathbb{R}$, then $c\mathbf{v} = c(p, v) = (p, cv)$.

dot product $\mathbf{v} \cdot \mathbf{w} = (p, v) \cdot (p, w) = v \cdot w$

cross product $\mathbf{v} \times \mathbf{w} = (p, v) \times (p, w) = (p, v \times w)$

Remark. Angle θ between \mathbf{v} and \mathbf{w} is given by,

$$\cos \theta = \mathbf{v} \cdot \mathbf{w} = (p, v) \cdot (p, w) = v \cdot w \quad (18.3)$$

And the length of a vector \mathbf{v} is given by,

$$\|\mathbf{v}\| = \mathbf{v} \cdot \mathbf{v} = (p, v) \cdot (p, v) = v \cdot v = \|v\| \quad (18.4)$$

Remark. Let $c \in \mathbb{R}$ and $p \in \mathbb{R}^{n+1}$. Let \mathbf{v}, \mathbf{w} be two vectors at p . That is, $\mathbf{v} = (p, v)$ and $\mathbf{w} = (p, w)$ for some $v, w \in \mathbb{R}^{n+1}$. Then **the set of all vectors at p is a vector space** with vector addition $\mathbf{v} + \mathbf{w} = (p, v + w)$ and scalar multiplication $c\mathbf{v} = (p, cv)$. This vector space is denoted by \mathbb{R}_p^{n+1} .

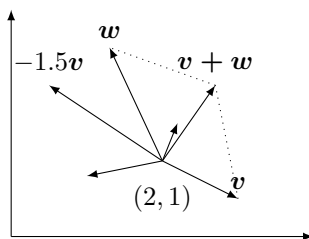


Figure 18.3: The vector space of all vectors at $(2, 1)$, $\mathbb{R}_{(2,1)}^2$

Definitions 18.2.2. The vector field \mathbf{X} on \mathbb{R}^{n+1} is a function which assigns to each point of \mathbb{R}^{n+1} a vector at that point. That is, $\mathbf{X}(p) = (p, X(p))$.

For example, $\mathbf{X}(p) = (p, X(p))$ where the associated function of the vector field, $X : \mathbb{R}^2 \rightarrow \mathbb{R}^2$ defined by $X(p) = (1, 2)$ assigns a constant vector $(1, 2)$ at every vector in \mathbb{R}^2 .

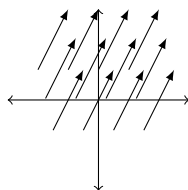


Figure 18.4: Vector field with associated function $X(p) = (1, 2)$

Definitions 18.2.3 (smooth). A function $f : \mathbb{R} \rightarrow \mathbb{R}$ is smooth if its partial derivatives of all orders exists and are continuous. A function $f : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$ is smooth if its component functions $f = (f_1, f_2, \dots, f_{n+1})$ are smooth. A vector field \mathbf{X} is smooth if the associated function $X(p)$ is smooth.

Definitions 18.2.4. Let $f : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$. Then the gradient of f at p is,

$$\nabla f(p) = \left(p, \frac{\partial f}{\partial x_1}(p), \frac{\partial f}{\partial x_2}(p), \dots, \frac{\partial f}{\partial x_{n+1}}(p) \right) \quad (18.5)$$

Remark. If f is a smooth function, then the gradient of f at p is a smooth vector field.

For example, $f : \mathbb{R}^2 \rightarrow \mathbb{R}$ defined by $f(x_1, x_2) = 2x_1x_2$ is a smooth function. We have, $\frac{\partial f}{\partial x_1} = 2x_2$ and $\frac{\partial f}{\partial x_2} = 2x_1$. And gradient of f at (x_1, x_2) is $(x_1, x_2, 2x_2, 2x_1)$. That is, $(2x_2, 2x_1)$ at (x_1, x_2) .

Calculations :

p	(x_1, x_2)	$(0, 0)$	$(1, 0)$	$(0, 1)$	$(-1, 0)$	$(0, -1)$
$X(p)$	$(2x_2, 2x_1)$	$(0, 0)$	$(0, 2)$	$(2, 0)$	$(0, -2)$	$(-2, 0)$

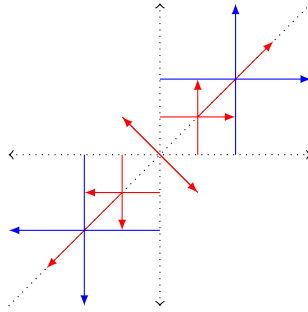


Figure 18.5: The gradient of $f(x_1, x_2) = 2x_1x_2$

Definitions 18.2.5. A parameterised curve is a function, $\alpha : I \rightarrow \mathbb{R}^{n+1}$ where I is some open interval in \mathbb{R} . The velocity vector of a parameterised curve $\alpha : I \rightarrow \mathbb{R}^{n+1}$ at a point $\alpha(t)$ is the tangent to the curve at that point.

$$\dot{\alpha}(t) = \left(\alpha(t), \frac{d\alpha}{dt}(t) \right) \quad (18.6)$$

For example, $\alpha : I \rightarrow \mathbb{R}^2$ defined by $\alpha(t) = (2t, t^2)$ is a parameterised curve. We have, $\frac{d\alpha}{dt} = \left(\frac{dx_1}{dt}(t), \frac{dx_2}{dt}(t) \right) = (2, 2t)$ where $\alpha(t) = (x_1(t), x_2(t))$. The velocity vector at $t = 3$ is $\dot{\alpha}(3) = (\alpha(3), \frac{d\alpha}{dt}) = (6, 9, 2, 6)$.

Definitions 18.2.6. Let \mathbf{X} be a vector field and let U be an open subset of \mathbb{R}^{n+1} . An integral curve α on U is a parameterised curve, $\alpha : I \rightarrow \mathbb{R}^{n+1}$ such that for each $\alpha(t) = p \in U$, the velocity vector $\dot{\alpha}(t)$ is the associated vector $\mathbf{X}(p)$ of the vector field \mathbf{X} at that point. Thus, for each $t \in I$, $\dot{\alpha}(t) = \mathbf{X}(\alpha(t))$.

$$\left(\alpha(t), \frac{d\alpha}{dt}(t) \right) = (\alpha(t), \mathbf{X}(\alpha(t))) \quad (18.7)$$

Let $\mathbf{X}(p) = (X_1(p), X_2(p), \dots, X_{n+1}(p))$ and $\alpha(t) = (x_1(t), x_2(t), \dots, x_{n+1}(t))$. Then, comparing components of the vector at $\alpha(t)$ we get the following system of equations,

$$\frac{dx_j}{dt}(t) = X_j(\alpha(t)), \quad j = 1, 2, \dots, (n+1) \quad (18.8)$$

For example, Consider $\alpha : (2, 3) \rightarrow \mathbb{R}^2$ defined by $\alpha(t) = (t, t^2)$. Then α is a parameterised curve in vector field, \mathbf{X} which has the associated function $X(x_1, x_2) = (1, 2x_1)$. Then, $\mathbf{X}(x_1, x_2) = (x_1, x_2, 1, 2x_1)$. And

$$\dot{\alpha}(t) = \left(\alpha(t), \frac{d\alpha}{dt}(t) \right) = \left(x_1(t), x_2(t), \frac{dx_1}{dt}(t), \frac{dx_2}{dt}(t) \right) = (t, t^2, 1, 2t)$$

Clearly, α is an integral curve of \mathbf{X} as $\dot{\alpha}(t) = X(\alpha(t))$ for every $t \in (2, 3)$.

Calculations:

p	(0, 0)	(1, 0)	(0, 1)	(1, 1)	(-1, 0)	(0, -1)	(-1, 1)	(1, -1)	(-1, -1)
$X(p)$	(1, 0)	(2, 2)	(1, 1)	(2, 3)	(0, -2)	(1, 1)	(0, -1)	(2, 1)	(0, -3)

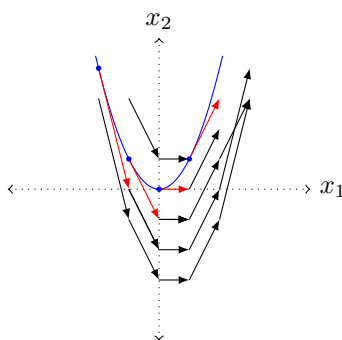


Figure 18.6: Integral Curve $\alpha(t) = (t, t^2)$ in \mathbf{X} with $X(x_1, x_2) = (1, 2x_1)$

Theorem 18.2.1. Let \mathbf{X} be a smooth vector field on an open set $U \subset \mathbb{R}^{n+1}$ and let $p \in U$. Then there exists an open interval I containing 0 and an integral curve $\alpha : I \rightarrow U$ such that

1. $\alpha(0) = p$
2. If $\beta : \tilde{I} \rightarrow U$ is any other integral curve with $\beta(0) = p$, then $\tilde{I} \subset I$ and $\beta(t) = \alpha(t)$, for all $t \in \tilde{I}$.

Proof. Let \mathbf{X} be a smooth vector field. Suppose α be an integral curve in \mathbf{X} . Then, $\dot{\alpha}(t) = \mathbf{X}(\alpha(t))$. Let $x_j(t)$ be the components of $\alpha(t)$ and $X_j(p)$ be the components of $X(p)$.

$$\begin{aligned} \dot{\alpha}(t) &= \left(\alpha(t), \frac{d\alpha}{dt}(t) \right) \\ &= \left(x_1(t), \dots, x_{n+1}(t), \frac{dx_1}{dt}(t), \dots, \frac{dx_{n+1}}{dt}(t) \right) \\ \mathbf{X}(\alpha(t)) &= (\alpha(t), X(\alpha(t))) \\ &= (x_1(t), \dots, x_{n+1}(t), X_1(\alpha(t)), \dots, X_{n+1}(\alpha(t))) \end{aligned}$$

Thus, we have a system of $n + 1$ first order differential equations in $n + 1$ unknowns

satisfying the initial condition $\alpha(0) = p$.

$$\begin{aligned}\frac{dx_1}{dt}(t) &= X_1(\alpha(t)) \\ \frac{dx_2}{dt}(t) &= X_2(\alpha(t)) \\ &\vdots \\ \frac{dx_{n+1}}{dt}(t) &= X_{n+1}(\alpha(t))\end{aligned}$$

□

By the theorem on solution of systems of first order ordinary differential equations, there exists an interval I containing 0 and a solution — a family of functions $\{x_1(t), x_2(t), \dots, x_{n+1}(t)\}$ satisfying the above system of equations satisfying the initial condition $\alpha(0) = p$.

Define $\alpha : I \rightarrow U$ using the component functions of α as x_j s in the above solution. Then, we have a integral curve of the vector field \mathbf{X} satisfying the initial condition $\alpha(0) = p$.

Let $\beta : \tilde{I} \rightarrow U$ be another integral curve with $\beta(0) = p$. Then by the uniqueness of the solution for the system of first order ordinary differential equations with an initial condition, $\beta(t) = \alpha(t)$ for every $t \in I \cup \tilde{I}$.

Let $\{\beta_1, \beta_2, \dots\}$ be the family of integral curves with $\beta_j : I_j \rightarrow U$ satisfying $\beta_j(0) = p$. Consider $I = \bigcup_{j \in \mathbb{N}} I_j$.

Define $\alpha : I \rightarrow U$ by $\alpha(t) = \beta_j(t)$ where $t \in I_j$ for some $j \in \mathbb{N}$. Then α is well-defined and is a maximal integral curve in \mathbf{X} such that $\alpha(0) = p$.

Definitions 18.2.7. A smooth vector field \mathbf{X} on $U \subset \mathbb{R}^{n+1}$ is **complete** if for every $p \in U$, the maximal integral curve through p has domain equal to \mathbb{R} .

Definitions 18.2.8. The **divergence** of a smooth vector field \mathbf{X} on $U \subset \mathbb{R}^{n+1}$ is the function $\text{div } \mathbf{X} : U \rightarrow \mathbb{R}$ defined by

$$\text{div } X(x_1, x_2, \dots, x_{n+1}) = \sum_{i=1}^{n+1} \frac{\partial X_i}{\partial x_i}$$

where X_i are the component function of the associated function X of the vector field \mathbf{X} .

For example, Consider \mathbf{X} with associated function $X : \mathbb{R}^2 \rightarrow \mathbb{R}^2$ defined by $X(x_1, x_2) = (2x_1, x_1x_2)$. Then $\text{div } \mathbf{X}(x_1, x_2) = \frac{\partial X_1}{\partial x_1} + \frac{\partial X_2}{\partial x_2} = 2 + x_1$.

18.3 The Tangent Space

Definitions 18.3.1. Let $f : U \rightarrow \mathbb{R}$ be a smooth function. Suppose $f^{-1}(c)$ is nonempty and $p \in f^{-1}(c)$. A vector \mathbf{v} is a **vector tangent to $f^{-1}(c)$ at p** if there exists a parametrized curve $\alpha : I \rightarrow U$ such that α is contained in U , $\alpha(0) = p$ and $\dot{\alpha}(0) = \mathbf{v}$.

Lemma 18.3.1. The gradient of f at $p \in f^{-1}(c)$ is orthogonal to all vectors tangent to $f^{-1}(c)$ at p .

Proof.

$$\begin{aligned}
 f \circ \alpha(t) &= f(x_1(t), x_2(t), \dots, x_{n+1}(t)) \\
 \frac{df \circ \alpha(t)}{dt} &= \frac{\partial f}{\partial x_1} \frac{d}{dt} x_1(t) + \frac{\partial f}{\partial x_2} \frac{d}{dt} x_2(t) + \dots + \frac{\partial f}{\partial x_{n+1}} \frac{d}{dt} x_{n+1}(t) \\
 &= \left(\frac{\partial f(\alpha(t_0))}{\partial x_1}, \frac{\partial f(\alpha(t_0))}{\partial x_2}, \dots, \frac{\partial f(\alpha(t_0))}{\partial x_{n+1}} \right) \cdot \left(\frac{dx_1(t)}{dt}, \frac{dx_2(t)}{dt}, \dots, \frac{dx_{n+1}(t)}{dt} \right) \\
 &= (\nabla f)(\alpha(t_0)) \cdot \dot{\alpha}(t_0)
 \end{aligned}$$

Clearly, $f \circ \alpha(t_0) = c$ and $\frac{d}{dt} f \circ \alpha(t_0) = \nabla f(\alpha(t_0)) \cdot \dot{\alpha}(t_0) = 0$. Since parametrized curve α is arbitrary, the gradient vector $\nabla f(\alpha(t_0))$ is orthogonal to every vector tangent to $f^{-1}(c)$ at $\alpha(t_0) = p$. \square

Theorem 18.3.2. *Let U be an open set, $f : U \rightarrow \mathbb{R}$ be a smooth function and $p \in U$ be a regular point of f . If $f(p) = c$, then the set all vectors tangent to $f^{-1}(c)$ at p is equal to $[\nabla f(p)]^\perp$.*

Proof. A vector \mathbf{v} is tangent to $f^{-1}(c)$ at p if there exists a parametrised curve α on $f^{-1}(c)$ such that $\alpha(0) = p$ and $\dot{\alpha}(0) = \mathbf{v}$. We know that $f(p) = c$. By chain rule, $f'(\alpha(0)) = \nabla f(\alpha(0)) \cdot \dot{\alpha}(0) = \nabla f(p) \cdot \mathbf{v} = 0$. Clearly, $\mathbf{v} \in [\nabla f(p)]^\perp$.

It is enough to prove that if $\mathbf{v} = (p, v) \in [\nabla f(p)]^\perp$ then there exists a parametrized curve α contained in $f^{-1}(c)$ which passes through p and has tangent vector v at p .

Let $\mathbf{v} = (p, v) \in [\nabla f(p)]^\perp$. Consider constant vector field \mathbf{X} on U defined by, $\mathbf{X}(q) = (q, v)$. Consider another vector field \mathbf{Y} defined by subtracting the component of \mathbf{X} along $\nabla f(q)$.

$$\mathbf{Y}(q) = \mathbf{X}(q) - \frac{\mathbf{X}(q) \cdot \nabla f(q)}{\|\nabla f(q)\|^2} \nabla f(q)$$

The domain of \mathbf{Y} is the open subset of U such that $\nabla f(q) \neq 0$. Since p is a regular point of f , p is in the domain of \mathbf{Y} . And $\mathbf{Y}(p) = \mathbf{X}(p)$ since $\mathbf{X}(p) \cdot \nabla f(p) = 0$. By construction, \mathbf{Y} is a smooth vector field orthogonal to $\nabla f(q)$ for every point q in its domain and $\mathbf{Y}(p) = \mathbf{v}$.

Since \mathbf{Y} is smooth vector field defined on an open set, there exists a maximal integral curve through p , say α . Thus, $\alpha : I \rightarrow \mathcal{D}(\mathbf{Y})$ such that $\dot{\alpha}(t) = \mathbf{Y}(\alpha(t))$ and $\alpha(0) = p$, $0 \in I$. Clearly, $\dot{\alpha}(0) = \mathbf{Y}(p) = \mathbf{v}$.

$$\begin{aligned}
 \frac{d}{dt} f(\alpha(t)) &= \nabla f(\alpha(t)) \cdot \dot{\alpha}(t) \text{ by chain rule} \\
 &= \nabla f(\alpha(t)) \cdot \mathbf{Y}(\alpha(t)) \text{ since } \dot{\alpha}(t) = \mathbf{Y}(\alpha(t)) \\
 &= \mathbf{0} \text{ since } \mathbf{Y} \perp \nabla f(\alpha(t))
 \end{aligned}$$

Clearly, $f(\alpha(t))$ is a constant and $f(\alpha(0)) = f(p) = c$ since $p \in f^{-1}(c)$. Therefore, $f(\alpha(t)) = c$, $\forall t \in I$. Clearly, α is contained in $f^{-1}(c)$. Therefore, there exists a parametrized curve α such that $\dot{\alpha}(0) = \mathbf{v}$. Since \mathbf{v} is arbitrary, every vector orthogonal to $\nabla f(p)$ is a tangent vector to $f^{-1}(c)$ at p . \square

Definitions 18.3.2. Let S be an n -surface and $p \in S$. The set of all vector tangent to S at p is called the **tangent space** S_p . If p is a regular point of f and $f(p) = c$. Then the **tangent space** S_p of the surface $S = f^{-1}(c)$ at p is spanned by the vector $\nabla f(p)$. That is, $S_p = [\nabla f(p)]^\perp$.

18.4 Surfaces

Definitions 18.4.1. A **surface** of dimension n , or n -surface in \mathbb{R}^{n+1} is a nonempty subset of the form $S = f^{-1}(c)$ where $f : U \rightarrow \mathbb{R}$ such that U is an open subset of \mathbb{R}^{n+1} , f is a smooth function and $\nabla f(p) \neq 0$, $\forall p \in S$.

The n -surfaces are usually called **plane curves**, **surfaces** or **hypersurfaces** when $n = 1$, $n = 2$ and $n > 2$ respectively.

Remark. Let S be an n -surface and $p \in S$. Then it has the tangent space S_p which is an n dimensional subspace of the vector space \mathbb{R}^{n+1} .

It is important to note that, the tangent space S_p depends on the surface S , the point p and is independent of the the function f .

Definitions 18.4.2. The n -**sphere** S^n is the level set f^{-1} where $f : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$ defined by $f(x_1, x_2, \dots, x_{n+1}) = x_1^2 + x_2^2 + \dots + x_{n+1}^2$.

The 1-sphere, S^1 is the unit circle.

Definitions 18.4.3. The n -**plane** is the level set $f^{-1}(c)$ where $f : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$ defined by $f(x_1, x_2, \dots, x_{n+1}) = a_1x_1 + a_2x_2 + \dots + a_{n+1}x_{n+1}$.

We may express $f(x) = a \cdot x$ where $a, x \in \mathbb{R}^{n+1}$. The n -planes are **lines**, **planes** and **hyperplanes** when $n = 1$, $n = 2$ and $n > 2$ respectively.

Definitions 18.4.4. Let S be an $(n-1)$ -surface. Then $S = f^{-1}(c)$ for some function $f : \mathbb{R}^n \rightarrow \mathbb{R}$. The **cylinder over the surface** S is the n -surface $g^{-1}(c)$ where $g(x_1, x_2, \dots, x_{n+1}) = f(x_1, x_2, \dots, x_n)$.

In general, let $S = f^{-1}(c)$ where $f : U \rightarrow \mathbb{R}$. Then the cylinder over S is an n -surface given by $g^{-1}(c)$ where $g : U \times \mathbb{R} \rightarrow \mathbb{R}$ defined by $g(u, t) = f(u)$ where $u \in U$, $t \in \mathbb{R}$.

The name comes from the cylinder over the unit circle.

Definitions 18.4.5. Let C be a curve in \mathbb{R}^2 which lies above x_1 axis. Then $C = f^{-1}(c)$ where $f : U \rightarrow \mathbb{R}$, $\nabla f(p) \neq 0$, $p \in C$ and U is contained in the upper half plane $x_2 > 0$. Then **surface of revolution obtained by rotating C about the x_1 axis** is given $S = g^{-1}(c)$ where $g(x_1, x_2, x_3) = f(x_1, (x_2^2 + x_3^2)^{\frac{1}{2}})$.

Theorem 18.4.1 (Lagrange's Multiplier). *Let S be an n -surface in \mathbb{R}^{n+1} , $S = f^{-1}(c)$ where $f : U \rightarrow \mathbb{R}$ is such that $\nabla f(q) \neq 0$, $\forall q \in S$. Suppose $g : U \rightarrow \mathbb{R}$ is a smooth function and $p \in S$ is an extreme point of g on S . Then there exists a real number λ such that $\nabla g(p) = \lambda \nabla f(p)$.*

Proof. Let S be an n -surface with $p \in S$ as an extreme point of smooth function g on S . We know that, $S_p = [\nabla f(p)]^\perp$. Clearly, S_p^\perp is 1-dimensional in \mathbb{R}^{n+1}

spanned by $\nabla f(p) \neq 0$.

Let $\mathbf{v} \in S_p$. Then there exists a parametrized curve α on S such that $\alpha(0) = p$ and $\dot{\alpha}(0) = \mathbf{v}$. Clearly, $g'(\alpha(0)) = g'(p) = 0$ since p is an extreme point of g on S .

$$g'(\alpha(0)) = \nabla g(\alpha(0)) \cdot \dot{\alpha}(0) = \nabla g(p) \cdot \mathbf{v} = 0$$

Thus, $\nabla g(p) \in S_p^\perp$. Therefore, $\nabla g(p) = \lambda \nabla f(p)$. \square

Remark. Let p be an extreme point of a smooth function g defined on an n -surface $f^{-1}(c)$. The **Lagrange's multiplier** λ is a real number such that $\nabla g(p) = \lambda \nabla f(p)$. If S is compact, then every smooth function on S attains both minimum and maximum on S .

The function $g(x_1, x_2) = ax_1^2 + 2bx_1x_2 + cx_2^2$ attains extrema on unit circle, $f^{-1}(1)$ where $f(x_1, x_2) = x_1^2 + x_2^2$. We know that

$$\begin{aligned}\nabla f(x_1, x_2) &= (x_1, x_2, 2x_1, 2x_2) \text{ and} \\ \nabla g(x_1, x_2) &= (x_1, x_2, 2ax_1 + 2bx_2, 2bx_1 + 2cx_2)\end{aligned}$$

By Lagrange's multiplier theorem, $\nabla g(x_1, x_2) = \lambda \nabla f(x_1, x_2)$. Thus,

$$\begin{aligned}2ax_1 + 2bx_2 &= 2\lambda x_1 \\ 2bx_1 + 2cx_2 &= 2\lambda x_2\end{aligned}$$

We may write, $\begin{pmatrix} a & b \\ b & c \end{pmatrix} \begin{pmatrix} x_1 \\ x_2 \end{pmatrix} = \lambda \begin{pmatrix} x_1 \\ x_2 \end{pmatrix}$. We know that, $\lambda, \begin{pmatrix} x_1 \\ x_2 \end{pmatrix}$ satisfying the condition are the eigenvalues and eigenvectors of the matrix $\begin{pmatrix} a & b \\ b & c \end{pmatrix}$ respectively. Clearly, the extreme points of the function g are the eigenvectors of the matrix.

Exercise

12. Show that the extreme values of the function $g(\mathbf{x}) = \sum_{i,j} a_{i,j}x_i x_j$ on the unit sphere $x_1^2 + x_2^2 + \dots + x_{n+1}^2 = 1$, where $(a_{i,j})$ is a symmetric matrix of real numbers, are eigen values of the matrix $(a_{i,j})$.

$$\begin{aligned}\nabla f(p) &= (p, 2p) \\ \nabla g(p) &= (p, 2 \sum_j a_{1,j}x_j, \dots, 2 \sum_j a_{n+1,j}x_j)\end{aligned}$$

By Lagrange's theorem, $\nabla g(p) = \lambda \nabla f(p)$. Thus,

$$\begin{pmatrix} a_{1,1} & a_{1,2} & \dots & a_{n+1,1} \\ a_{2,1} & a_{2,2} & \dots & a_{n+1,2} \\ \vdots & \vdots & \ddots & \vdots \\ a_{n+1,1} & a_{n+1,2} & \dots & a_{n+1,n+1} \end{pmatrix} \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_{n+1} \end{pmatrix} = \lambda \begin{pmatrix} x_1 \\ x_2 \\ \vdots \\ x_{n+1} \end{pmatrix}$$

Let $p = (p_1, p_2, \dots, p_{n+1})$. Then, $\|p\|^2 = 1$ since $p \in S^2$.

$$g(p) = [P][a_{i,j}][P]' = [P]\lambda[P]' = \lambda\|p\|^2 = \lambda$$

We know that all eigenvalues of a real symmetric matrix is real. Therefore, the minimum and maximum values of g are the minimum and maximum eigenvalues of the matrix.

15. Suppose $(x_1, x_2, x_3, x_4) \in \mathbb{R}^4$ is represented by $\begin{pmatrix} x_1 & x_2 \\ x_3 & x_4 \end{pmatrix}$. Then $SL(2, \mathbb{R})$ is a 3-surface in \mathbb{R}^4 .

Special Linear Group, $SL(2, \mathbb{R})$ is defined as the set of all 2×2 real matrices with determinant 1. That is, $x_1x_4 - x_2x_3 = 1$. Consider the function $f : \mathbb{R}^4 \rightarrow \mathbb{R}$ defined by $f(x_1, x_2, x_3, x_4) = x_1x_4 - x_2x_3$. Then $SL(2, \mathbb{R})$ is nothing but $f^{-1}(1)$.

We know that the domain of f is open, f is a smooth function and $\nabla f(x_1, x_2, x_3, x_4) = (x_1, x_2, x_3, x_4, -x_3, -x_2, x_1)$. Thus, $\nabla f(p) = 0$ if and only if $p = 0$. But $0 \notin SL(2, \mathbb{R})$. Therefore, $SL(2, \mathbb{R})$ is a 3-surface.

18.5 Vector Fields on Surfaces; Orientation

Definitions 18.5.1. Let S be an n -surface and X be a smooth function defined S such that $X(p) \in \mathbb{R}^{n+1}$, $\forall p \in S$. Then $\mathbf{X}(p) = (p, X(p))$ is a **vector field on the n -surface S** .

We know that $\nabla f(p) \neq 0$, $\forall p \in S$. Then there exists an open set U containing S such that $\nabla f(q) \neq 0$, $\forall q \in U$. The construction of U is as follows, Let $p \in S$. Then $\nabla f(p) \neq 0$. Since f is smooth, there exists a neighbourhood P of p in which $\nabla f(p) \neq 0$. Clearly $\bigcup_{p \in S} P$ is an open set containing S in which $\nabla f(p) \neq 0$.

Definitions 18.5.2. A **tangent vector field \mathbf{X} on an n -surface S** is a vector field on S which assigns to each point $p \in S$ a vector tangent to S at p . A **normal vector field \mathbf{X} on an n -surface S** is a vector field on S which assigns to each point $p \in S$ a vector orthogonal to S at p .

Definitions 18.5.3. A vector field \mathbf{X} on an n -surface S is **smooth** if \mathbf{X} it is a restriction of a smooth vector field defined on an open set.

Theorem 18.5.1 (maximal integral curve on a surface). *Let S be an n -surface in \mathbb{R}^{n+1} , let X be a smooth tangent vector field on S , and let $p \in S$. Then there exists an interval I containing 0 and a parametrized curve $\alpha : I \rightarrow S$ such that*

1. $\alpha(0) = p$.
2. $\dot{\alpha}(t) = \mathbf{X}(\alpha(t))$, $\forall t \in I$.
3. If $\beta : \tilde{I} \rightarrow S$ is any other parametrized curve satisfying above conditions, then $\tilde{I} \subset I$ and $\beta(t) = \alpha(t)$, $\forall t \in I \cap \tilde{I}$.

Proof. Let S be an n -surface. And \mathbf{X} be a smooth tangent vector field on S . By the definition of smooth vector field, \mathbf{X} should be a restriction of a smooth vector field $\tilde{\mathbf{X}}$ on an open set V containing S . Let function $f : U \rightarrow \mathbb{R}$, $c \in \mathbb{R}$ such that $f^{-1}(c) = S$ and $\nabla f(q) \neq 0$, $\forall q \in S$. Let $W = \{q \in U \cap V : \nabla f(q) \neq 0\}$. Clearly W is open.¹ And contains S . And $\tilde{\mathbf{X}}, f$ are defined on W . Consider the vector field \mathbf{Y} on W defined by

$$\mathbf{Y}(q) = \tilde{\mathbf{X}}(q) - \frac{\tilde{\mathbf{X}} \cdot \nabla f(q)}{\|\nabla f(q)\|^2} \nabla f(q)$$

Then \mathbf{Y} is a vector field on W such that it is tangent to level set of f since we have removed the components orthogonal to the level sets from each vector assigned.

Clearly, $\mathbf{Y}(p) = \tilde{\mathbf{X}}(p)$, $\forall p \in S$ since $\tilde{\mathbf{X}}(p) \cdot \nabla f(p) = 0$ for each $p \in S$. We already know that, $\tilde{\mathbf{X}}(p) = \mathbf{X}(p)$, $\forall p \in S$. Now \mathbf{Y} is a smooth vector defined on an open set W .

By maximal integral theorem, there exists a maximal integral on \mathbf{Y} through p , say $\alpha : I \rightarrow W$. Then $(f \circ \alpha)'(t) = \nabla f(\alpha(t)) \cdot \dot{\alpha}(t) = \nabla f(\alpha(t)) \cdot \mathbf{Y}(\alpha(t)) = 0$, since \mathbf{Y} is tangent to the level sets of f by construction. Thus, $f(\alpha(t))$ is constant. However, $f(\alpha(0)) = f(p) = c$. Thus, $f(\alpha(t)) = c, \forall t \in I$. Therefore, $\alpha(t) \in f^{-1}(c) = S$, $\forall t \in I$. Clearly, α is contained in S . Now, we have an integral curve on S .

Suppose $\beta : \tilde{I} : W$ be another integral curve on W through p . Then $\alpha(t) = \beta(t)$, $\forall t \in I \cap \tilde{I}$ by the uniqueness of the integral curve on \mathbf{Y} . Let $\{\alpha_j : i \in I\}$ be the set of all integral curves on \mathbf{Y} passing through p . Let I be union of the domain of α_j 's. Clearly, I is an interval containing 0. Then $\alpha : I \rightarrow W$ is a maximal integral on \mathbf{Y} passing through p where $\alpha(t) = \alpha_j(t)$, $t \in \mathcal{D}(\alpha_j)$. \square

Corollary 18.5.1.1. *Let S be an n -surface where $S = f^{-1}(c)$, $f : U \rightarrow \mathbb{R}$ and $\nabla f(q) \neq 0, \forall q \in S$. Let \mathbf{X} be a smooth vector field on U whose restriction to S is a tangent vector field on S . If $\alpha : I \rightarrow U$ is an integral curve on \mathbf{X} such that $\alpha(t_0) \in S$ for some $t_0 \in I$, then $\alpha(t) \in S$, $\forall t \in I$.*

Proof. \square

Theorem 18.5.2 (orientation). *Let S be a connected, n -surface. Then there exists on S exactly two unit normal vector fields \mathbf{N}_1 and \mathbf{N}_2 , and $\mathbf{N}_1(p) = -\mathbf{N}_2(p)$, $\forall p \in S$.*

Proof. \square

Definitions 18.5.4. Let S be an n -surface. A smooth, unit normal vector field \mathbf{N} on S is called an **orientation** on S . An n -surface S together with an orientation is called an **oriented n -surface**.

Remark. The Möbius strip is not an orientable 2-surface as it cannot have two orientations.

¹ $\{q \in \mathbb{R}^{n+1} : \nabla f(q) \neq 0\}$ is an open in \mathbb{R}^{n+1} since $\{q \in \mathbb{R}^{n+1} : \nabla f(q) = 0\}$ is closed.

Definitions 18.5.5. A unit vector $\mathbf{v} \in \mathbb{R}_p^{n+1}$ is called a **direction** at p . Clearly, $\mathbf{v} = (p, v)$ where $v \in S^n$ where S^n is the unit n -sphere.

Definitions 18.5.6. Let S be a 1-surface in \mathbb{R}^2 . The **positive tangent direction** at $p \in S$ is the direction obtained by rotating the orientation at p by $-\pi/2$ radians counterclockwise.

Definitions 18.5.7. Let S be a 2-surface in \mathbb{R}^3 . The **positive θ -rotation** at $p \in S$ is the linear map $R_\theta : S_p \rightarrow S_p$ defined by $R_\theta(\mathbf{v}) = (\cos \theta)\mathbf{v} + (\sin \theta)\mathbf{N}(p) \times \mathbf{v}$ where \mathbf{N} is the orientation. In other words, R_θ is the right-handed rotation about $\mathbf{N}(p)$ through angle θ .

Definitions 18.5.8. Let S be an n -surface in \mathbb{R}^{n+1} where $n > 2$. Let $p \in S$. An ordered orthonormal basis $\{e_1, e_2, \dots, e_n\}$ for the tangent space S_p is **right-handed** if the determinant

$$\det \begin{pmatrix} e_1 \\ e_2 \\ e_3 \\ \vdots \\ \mathbf{N}(p) \end{pmatrix}$$

is positive. And the basis **left-handed** if the determinant is negative.

Definitions 18.5.9. An ordered basis $\{v_1, v_2, \dots, v_n\}$ for the tangent space S_p is **consistent** if the determinant

$$\det \begin{pmatrix} v_1 \\ v_2 \\ \vdots \\ v_n, \mathbf{N}(p) \end{pmatrix}$$

is positive. And the basis **inconsistent** if the determinant is negative.

Exercise

9. The **cross product** $\mathbf{v} \times \mathbf{w} = v \times w$ where $\mathbf{v} = (p, v)$ and $\mathbf{w} = (p, w)$. The properties of cross product in tangent space S_p of an n -surface S are

- (a) $(\mathbf{v} + \mathbf{w}) \times \mathbf{x} = \mathbf{v} \times \mathbf{x} + \mathbf{w} \times \mathbf{x}$.
- (b) $\mathbf{v} \times (\mathbf{w} + \mathbf{x}) = \mathbf{v} \times \mathbf{w} + \mathbf{v} \times \mathbf{x}$.
- (c) $(c\mathbf{v}) \times \mathbf{w} = c(\mathbf{v} \times \mathbf{w})$.
- (d) $\mathbf{v} \times (c\mathbf{w}) = c(\mathbf{v} \times \mathbf{w})$.
- (e) $\mathbf{v} \times \mathbf{w} = -\mathbf{w} \times \mathbf{v}$.
- (f) $\mathbf{u} \cdot (\mathbf{v} \times \mathbf{w}) = \mathbf{v} \cdot (\mathbf{w} \times \mathbf{u}) = \mathbf{w} \cdot (\mathbf{u} \times \mathbf{v})$.
- (g) $\mathbf{v} \times \mathbf{w}$ is orthogonal to both \mathbf{v} and \mathbf{w} .

18.6 The Gauss Map

Suppose S is an n -surface. From the definition of an n -surface, there exists a smooth function $f : U \rightarrow \mathbb{R}$ where U is an open subset of \mathbb{R}^{n+1} such that $S = f^{-1}(c)$ for some real value $c \in \mathbb{R}$ and every point on S is a regular point of

f . That is $\nabla f(p) \neq \mathbf{0}$ for every point p on the surface S .

We have proved that every n -surface has exactly two orientations \mathbf{N}_1 and \mathbf{N}_2 . These orientations are $\frac{\nabla f}{\|\nabla f\|}$ and $\frac{-\nabla f}{\|\nabla f\|}$. Given an orientation \mathbf{N} (either \mathbf{N}_1 or \mathbf{N}_2), the surface together with that orientation is collectively referred as an oriented n -surface.

Since orientation \mathbf{N} is a smooth, unit normal vector field. The vector field \mathbf{N} has an associated function $N : U \rightarrow \mathbb{R}^{n+1}$ such that $\mathbf{N}(p) = (p, N(p))$. And we already have, $\mathbf{N}(p) = (p, N(p)) = (p, \frac{\pm \nabla f}{\|\nabla f\|})$. This associated function restricted to the n -surface S is the **Gauss Map**. That is, $N : S \rightarrow \mathbb{R}^{n+1}$.

From the definition of orientation, we know that this function is actually assigning direction to each point on that surface S . If you don't remember, the directions are vector in \mathbb{R}^{n+1} of unit length. That is $\|v\| = 1$. Thus, the range of Gauss Map is a subset of the set of all directions. And unit sphere S^n is \mathbb{R}^{n+1} is the set of all directions in \mathbb{R}^{n+1} .

Thus, we may write Gauss Map, $N : S \rightarrow S^n$

18.6.1 Spherical Image

We already saw that, the Gauss Map $N : S \rightarrow S^n$ is a function which maps directions/unit vectors to each point on that oriented surface S .

Do we need an oriented surface ? Yes. The Gauss Map is defined by this orientation. If we are provided with an oriented n Surface S , then we have a unit vector/orientation assigned to each point p on that surface. And Gauss Map assigns this unit vector to the point p on surface S .

We already saw that the range of the Gauss Map is a subset of the unit n Sphere S^n . In other words, the Gauss Map assigns each point on the oriented n -surface S into a subset of the unit n sphere S^n . Thus, **range of the Gauss Map is referred as the spherical image of the oriented n -surface S** .

$$N(S) = \{q \in S^n : q = N(p), p \in S\} \quad (18.9)$$

18.6.2 Compact, connected, oriented n Surface

Suppose we have a compact, connected, oriented n -surface S . The compact subsets in Euclidean spaces are closed and bounded subsets. And connected subsets in Euclidean Spaces are path connected.

Theorem 18.6.1 (Spherical Image of Compact, Connected, Oriented Surface). *The Gauss map of a compact, connected, oriented n -surface is surjective.*

Proof. Let $v \in S^n$ be a direction in \mathbb{R}^{n+1} . Let S be a compact, connected, oriented n -surface with orientation \mathbf{N} such that $S = f^{-1}(c)$ and every point $p \in S$ are regular points of the smooth function $f : U \rightarrow \mathbb{R}$ where U is an open subset of \mathbb{R}^{n+1} . Thus, we have the Gauss Map $N : S \rightarrow S^n$ defined by the

orientation on S .

Since v is arbitrary, it is enough to prove that $v \in N(S)$. Suppose there exists $v \in S^n$ such that $v \notin N(S)$, then the Gauss Map is not surjective. In other words, N is surjective if for every $v \in S^n$, $v \in N(S)$ OR for every $v \in S^n$, there exists $p \in S$ such that $v = N(p)$.

Let $g : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$ defined by $g(p) = p \cdot v$. Then g is a smooth function since first order partial derivatives are constant functions and all other partial derivatives of higher orders vanishes.

Since S is compact, g restricted to S is a continuous function defined on a compact interval. And thus it attains maximum and minimum values, say p and q . The maximum and minimum values of the dot product $p \cdot v$ are $\pm v$.

By Lagrange's multiplier theorem, $\nabla g(p) = \lambda \nabla f(p)$ and $\nabla g(q) = \lambda \nabla f(q)$. From the definition of the Gauss Map, we have $\nabla g(p) = \lambda \nabla f(p) = \lambda \|\nabla f(p)\| N(p) = \lambda \|\nabla f(p)\| (p, v)$. Thus, v and $N(p)$ are multiples of one another. Similarly, $\nabla g(q) = \lambda \|\nabla f(q)\| N(q)$. Therefore $N(p) = \pm v$ and $N(q) = \pm v$.

It remains to show that $N(p) \neq N(q)$. Suppose $N(p) = N(q)$. If there exists continuous function α such that $\alpha : [0, 1] \rightarrow \mathbb{R}^{n+1}$, $\alpha(0) = p$, $\alpha(1) = q$, $\dot{\alpha}(0) = (p, v)$ and $\dot{\alpha}(1) = (q, v)$. And α maps the interior, the open interval $(0, 1)$ outside the surface S . Then by intermediate value theorem, we arrive at a contradiction. And thus $N(p) \neq N(q)$.

Let $\alpha_1 : [0, x] \rightarrow \mathbb{R}^{n+1}$ defined by $\alpha_1(t) = p + tv$. Let $\alpha_2 : [y, 1] \rightarrow \mathbb{R}^{n+1}$ defined by $\alpha_2(t) = q + (t - 1)v$. Let $\alpha_3 : [x, y] \rightarrow S_1$ where S_1 is an n sphere properly containing S . Such an n sphere exists, since S is compact (bounded). And there exists such a function α_3 , the image of which is a compact subset of S_1 .

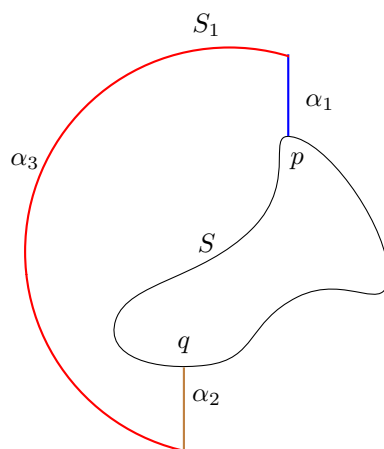
Now consider $\alpha : [0, 1] \rightarrow \mathbb{R}^{n+1}$ defined by

$$\alpha(t) = \begin{cases} \alpha_1(t) & t \in [0, x) \\ \alpha_3(t) & t \in [x, y] \\ \alpha_2(t) & t \in (y, 1] \end{cases} \quad (18.10)$$

Clearly, α is a smooth function with $\alpha(t) \notin S$, $t \in (0, 1)$ and

$$\begin{aligned} \alpha(0) &= \alpha_1(0) = p + 0v = p \\ \alpha(1) &= \alpha_2(1) = q + (1 - 1)v = q \\ \dot{\alpha}(0) &= \frac{d\alpha_1}{dt}(0) = \frac{d(p + tv)}{dt}(0) = v \\ \dot{\alpha}(1) &= \frac{d\alpha_2}{dt}(1) = \frac{d(q + (t - 1)v)}{dt} = v \end{aligned}$$

We have, $f(\alpha(0)) = f(0) = c$ since $p \in S = f^{-1}(c)$. Similarly, $f(\alpha(1)) = c$. And $(f \circ \alpha)'(0) = \nabla f \circ \alpha(0) \dot{\alpha}(0) = \nabla f(p) \cdot \dot{\alpha}(0) = \|\nabla f(p)\| N(p) \cdot v$. Similarly, $(f \circ \alpha)'(1) = \|\nabla f(q)\| N(q) \cdot v$. We have assumed that $N(p) = N(q)$. Then, $f \circ \alpha$

Figure 18.7: Construction of α

is either increasing at both 0 and 1 OR decreasing at both 0 and 1.

Without Loss of Generality, Suppose that $f \circ \alpha$ is increasing at either points. Then, there exists a sufficiently small $\epsilon > 0$ such that $f(\alpha(\epsilon)) > c$ and $f(\alpha(1 - \epsilon)) < c$. Since, $f \circ \alpha$ must have a value greater than c immediately after 0 and should have a value less than c just before reaching 1 as the function is increasing at either points (and in some small neighbourhood of those points).

By Intermediate Value theorem, there exists $t \in (0, 1)$ such that $f \circ \alpha(t) = c$ since the composition of smooth functions f and α is also smooth. But, it is clear from the construction that $\alpha(t)$ doesn't belong to the surface S . And therefore, $\alpha(t) \neq c \implies f(\alpha(t)) \neq c$ for any $t \in (0, 1)$. Thus by contradiction, $N(p) \neq N(q)$. And if N achieves v at p . Then it achieves $-v$ at q . And since $v \in S^n$ is arbitrary, $N(S) = S^n$ and the spherical image is the entire n sphere OR the Gauss map is surjective. \square

Given a compact, connected oriented, n -surface S , the Gauss Map on S is surjective. In other words, the spherical image of such a surface is the unit n sphere S^n itself.

Connectedness is not that critical (in my opinion). For a compact, orientated n -surface with multiple components, the above observation is valid for each connected component. And thus for any compact, oriented surface. Again, n -surfaces are always closed. Thus, the restriction practically reduces to boundedness of the n -surface.

18.7 Geodesics

We already know that our earth is not flat. Still, we feel like we move in straight lines. And our 'straight lines' are curved for an observer who is not on earth.

Geodesics are straightlines on an n -surface S .

vector field along α is function which assigns $X(t)$ at $\alpha(t)$ for each $t \in I$.

The definition of vector field doesn't allow you to assign multiple vectors at a point. But, vector field along α allows you to assign vectors to points on a parametrised curve depending on the value of parameter t .

function along α is function with the same domain I as α .

derivative of vector field X along α is a vector field along α given by $\dot{X}(t) = \left(\alpha(t), \frac{dX}{dt}(t) \right)$ where $X(t) = (\alpha(t), X(t))$.

velocity of α is a vector field along α defined by $\dot{\alpha}(t) = \left(\alpha(t), \frac{d\alpha}{dt}(t) \right)$.

Suppose $\alpha : I \rightarrow \mathbb{R}^2$ is defined by $\alpha(t) = (3t, t^2)$.

Then velocity of α is $\dot{\alpha}(t) = \left(\alpha(t), \frac{d\alpha}{dt}(t) \right) = (3t, t^2, 3, 2t)$.

speed of α is $\|\dot{\alpha}(t)\|$.

Speed of α is $\|\dot{\alpha}(t)\| = \sqrt{9 + 4t^2}$.

acceleration α is a vector field along α defined by $\ddot{\alpha}(t) = \left(\alpha(t), \frac{d^2\alpha}{dt^2}(t) \right)$.

Acceleration of α is $\ddot{\alpha}(t) = (3t, t^2, 0, 2)$

18.7.1 Properties of differentiation

Let X, Y be smooth vector fields along parametrised curve $\alpha : I \rightarrow \mathbb{R}^{n+1}$.

- $(X + Y) = \dot{X} + \dot{Y}$

$$(X + Y)(t) = (\alpha(t), X(t)) + (\alpha(t), Y(t)) = (\alpha(t), X(t) + Y(t))$$

$$\begin{aligned} (\dot{X + Y})(t) &= \left(\alpha(t), \frac{d}{dt}X(t) + Y(t) \right) \\ &= \left(\alpha(t), \frac{d}{dt}X(t) \right) + \left(\alpha(t), \frac{d}{dt}Y(t) \right) \\ &= \dot{X}(t) + \dot{Y}(t) \end{aligned}$$

- $(f\dot{X}) = f'X + f\dot{X}$

$$\begin{aligned} fX(t) &= f(t)(\alpha(t), X(t)) = (\alpha(t), f(t)X(t)) \\ (f\dot{X})(t) &= \left(\alpha(t), \frac{d}{dt}f(t)X(t) \right) \\ &= \left(\alpha(t), f'(t)X(t) + f(t)\frac{d}{dt}X(t) \right) \\ &= (\alpha(t), f'(t)X(t)) + \left(\alpha(t), f(t)\frac{d}{dt}X(t) \right) \\ &= f'(t)(\alpha(t), X(t)) + f(t)\left(\alpha(t), \frac{d}{dt}X(t) \right) \\ &= f'X(t) + f\dot{X}(t) \end{aligned}$$

$$\bullet (\mathbf{X} \cdot \mathbf{Y})' = \dot{\mathbf{X}} \cdot \mathbf{Y} + \mathbf{X} \cdot \dot{\mathbf{Y}}$$

$$(\mathbf{X} \cdot \mathbf{Y}) = (\alpha(t), X(t)) \cdot (\alpha(t), Y(t)) = \sum_{k=1}^{n+1} X_k(t) Y_k(t)$$

$$\begin{aligned} (\mathbf{X} \cdot \mathbf{Y})' &= \frac{d}{dt} \sum_{k=1}^{n+1} X_k(t) Y_k(t) \\ &= \sum_{k=1}^{n+1} \frac{d}{dt} X_k(t) Y_k(t) \\ &= \sum_{k=1}^{n+1} X'_k(t) Y_k(t) + X_k(t) Y'_k(t) \end{aligned}$$

$$\dot{\mathbf{X}}(t) \cdot \mathbf{Y}(t) = \left(\alpha(t), \frac{d}{dt} X(t) \right) \cdot (\alpha(t), Y(t)) = \sum_{k=1}^{n+1} X'_k(t) Y_k(t)$$

$$\mathbf{X}(t) \cdot \dot{\mathbf{Y}}(t) = (\alpha(t), X(t)) \cdot \left(\alpha(t), \frac{d}{dt} Y(t) \right) = \sum_{k=1}^{n+1} X_k(t) Y'_k(t)$$

Definitions 18.7.1 (geodesic). Let S be an n -surface. A Geodesic on S is a parametrised curve $\alpha : I \rightarrow S$ whose acceleration is orthogonal to S everywhere.

$$\ddot{\alpha}(t) \in S_{\alpha(t)}^\perp \quad (18.11)$$

18.7.2 An illustrative example

We know that S^1 given by $x_1^2 + x_2^2 = 1$ is a 1-surface in \mathbb{R}^2 . Consider the cylinder C over S^1 , $x_1^2 + x_2^2 = 1$. Clearly, C is a 2 surface in \mathbb{R}^3 . Also, $f : \mathbb{R}^3 \rightarrow \mathbb{R}$ defined by $f(x_1, x_2, x_3) = x_1^2 + x_2^2$ is a smooth function such that $C = f^{-1}(1)$ and $\nabla f(x_1, x_2, x_3) = (x_1, x_2, x_3, 2x_1, 2x_2, 0) \neq \mathbf{0}$.

Clearly, every vector orthogonal to the surface in a scalar multiple of ∇f at that point. Therefore, vectors orthogonal to the surface C at $\alpha(t)$ is of the form $(x_1, x_2, x_3, kx_1, kx_2, 0)$ where $k \in \mathbb{R}$.

If there exists a geodesic α in C , then $\ddot{\alpha}(t) \in S_{\alpha(t)}^\perp$. That is, $\ddot{\alpha}(t) = (x_1, x_2, x_3, kx_1, kx_2, 0)$. Thus, we need component functions $x_1(t), x_2(t), x_3(t)$ satisfying

$$\frac{d^2}{dt^2} x_1(t) = kx_1(t) \quad (18.12)$$

$$\frac{d^2}{dt^2} x_2(t) = kx_2(t) \quad (18.13)$$

$$\frac{d^2}{dt^2} x_3(t) = 0 \quad (18.14)$$

We have, $\frac{d^2}{dt^2} \cos t = -\cos t$ and $\frac{d^2}{dt^2} \sin t = -\sin t$. Thus, the parametrised curve $\alpha : I \rightarrow \mathbb{R}^3$ defined by $\alpha(t) = (\cos t, \sin t, t)$ is a geodesic in C since $\ddot{\alpha}(t) = (\cos t, \sin t, t, -\cos t, -\sin t, 0)$.

18.7.3 Maximal Geodesic

The conditions $\alpha(0) = p$, $\dot{\alpha}(0) = \mathbf{v}$ says that parametric curves are unique except for linear transformations on the parameter. That is, Suppose there exists another parametrised curve β in S through p with initial velocity \mathbf{v} with $\beta(t_0) = p$ and $\dot{\beta}(t_0) = \mathbf{v}$. Then, there exists a real number κ such that $\alpha(t) = \beta(\kappa t + t_0)$.

In other words, both α and β passes through the same points and the vectors assigned at each point is the same. And the difference between such two parametrised curves doesn't have any impact on the properties we are interested in.

The condition, if $\beta : \tilde{I} \rightarrow S$ is any other geodesic in S with $\beta(0) = p$ and $\dot{\beta}(0) = \mathbf{v}$, then $\tilde{I} \subset I$ and $\beta(t) = \alpha(t)$, $\forall t \in \tilde{I}$. is another way of saying that α is maximal and uniquely defined.

In essence, the following theorem says that if you are standing on an n -surface S at a point, say p . You can move on that surface in straightline from p , with any initial velocity, \mathbf{v} . Note that the velocity allows you to choose both direction and speed.

Theorem 18.7.1 (maximal geodesic). *Let S be an n -surface in \mathbb{R}^{n+1} . Let $p \in S$ and $\mathbf{v} \in S_p$. Then there exists a unique, maximal geodesic $\alpha : I \rightarrow S$ in S through p with initial velocity \mathbf{v} .*

Proof. Let S be an n -surface, then there exists a smooth function $f : U \rightarrow \mathbb{R}$ where U is an open subset of \mathbb{R}^{n+1} . And every points on that surface is regular with respect to f . That is, $\nabla f(p) \neq 0$, $\forall p \in S$. WLOG assume that every points in U is regular.

Define $N = \frac{\nabla f}{\|\nabla f\|}$. A parametrised curve α in S is a geodesic if it satisfies $\ddot{\alpha}(t) \in S_{\alpha(t)}^\perp$. Therefore,

$$\ddot{\alpha}(t) = g(t)\mathbf{N}(\alpha(t)) \quad (18.15)$$

Why don't we write $\ddot{\alpha}(t) = \kappa\mathbf{N}(\alpha(t))$? The vectors $\kappa\mathbf{N}(\alpha(t))$ are orthogonal to S . However, the converse is not true. For a geodesic it is not necessary that $\ddot{\alpha}(t) = \kappa\mathbf{N}(\alpha(t))$. The acceleration could be any vector in $S_{\alpha(t)}^\perp$. Since $S_{\alpha(t)}^\perp$ is spanned by $\mathbf{N}(\alpha(t))$, at each point on $\alpha(t)$ the scalars may be different. We overcome this with the help of a real valued function $g : I \rightarrow \mathbb{R}$ along α .

We have $\ddot{\alpha} = g(\mathbf{N} \circ \alpha)$. Thus $\ddot{\alpha} \cdot (\mathbf{N} \circ \alpha) = g\|\mathbf{N} \circ \alpha\|^2 = g$ since orientation \mathbf{N} assigns directions (vectors of unit length) to each point of that surface.

$$[\dot{\alpha} \cdot (\mathbf{N} \circ \alpha)]' = \ddot{\alpha} \cdot (\mathbf{N} \circ \alpha) + \dot{\alpha} \cdot (\mathbf{N} \dot{\circ} \alpha) \text{ since } (\mathbf{X} \cdot \mathbf{Y})' = \dot{\mathbf{X}} \cdot \mathbf{Y} + \mathbf{X} \cdot \dot{\mathbf{Y}}.$$

Thus, $\ddot{\alpha} \cdot (\mathbf{N} \circ \alpha) = [\dot{\alpha} \cdot (\mathbf{N} \circ \alpha)]' - \dot{\alpha} \cdot (\mathbf{N} \dot{\circ} \alpha) = -\dot{\alpha} \cdot (\mathbf{N} \dot{\circ} \alpha)$ since $\dot{\alpha} \cdot (\mathbf{N} \circ \alpha) = 0$ as $\dot{\alpha} \in S_{\alpha(t)}$, $\mathbf{N} \circ \alpha \in S_{\alpha(t)}^\perp$ and $\dot{\alpha} \perp (\mathbf{N} \circ \alpha)$. That is, velocity vectors always belongs to the tangent space and $(\mathbf{N} \circ \alpha)$ is an orientation which is orthogonal to all the tangent vectors.

Substituting the value of g in equation(18.15). We get

$$\ddot{\alpha} + [\dot{\alpha} \cdot (\mathbf{N} \dot{\circ} \alpha)](\mathbf{N} \circ \alpha) = \mathbf{0} \quad (18.16)$$

We have,

$$\frac{dN_1}{dt} = \frac{\partial N_1}{\partial x_1} \frac{dx_1}{dt} + \frac{\partial N_1}{\partial x_2} \frac{dx_2}{dt} + \cdots + \frac{\partial N_1}{\partial x_{n+1}} \frac{dx_{n+1}}{dt} = \sum_{k=1}^{n+1} \frac{\partial N_1}{\partial x_k} \frac{dx_k}{dt} \quad (18.17)$$

Thus,

$$(N \circ \alpha) = \left(\alpha(t), \sum_{k=1}^{n+1} \frac{\partial N_1}{\partial x_k} \frac{dx_k}{dt}, \sum_{k=1}^{n+1} \frac{\partial N_2}{\partial x_k} \frac{dx_k}{dt}, \dots, \sum_{k=1}^{n+1} \frac{\partial N_{n+1}}{\partial x_k} \frac{dx_k}{dt} \right) \quad (18.18)$$

Equating the components on either sides of the equation (18.16), we get a system of $n+1$ second order differential equations,

$$\frac{d^2}{dt^2} x_i(t) + \left[\sum_{j,k=1}^{n+1} \frac{\partial N_j}{\partial x_k} \frac{dx_k}{dt} \frac{dx_j}{dt} \right] N_i \circ \alpha = 0 \quad (18.19)$$

By the existence theorem for solution of such system of second order differential equations, there exists an open interval I containing 0 and $(n+1)$ functions $x_k : I \rightarrow \mathbb{R}$ satisfying the above system. Then $\beta : I \rightarrow S$ defined by $\beta(t) = (x_1(t), x_2(t), \dots, x_{n+1}(t))$

By the uniqueness theorem for solution of such system of second order differential equations, if there exists another open interval \tilde{I} containing 0 and $\beta_1 : I_1 \rightarrow U$. Then $\beta(t) = \beta(t)$ for every $t \in I \cap I_1$.

Let β_1, β_2, \dots be geodesics through p with initial velocity \mathbf{v} with parameter domain I_1, I_2, \dots . Let $I = \cup_k I_k$ and $\alpha : I \rightarrow U$ defined by $\alpha(t) = \beta_k(t)$, $t \in I_k$. Clearly, α is unique and maximal.

Now it is enough to prove that α is parametrised curve in S . We have $(f \circ \alpha)'(t) = \nabla f(\alpha(t)) \cdot \dot{\alpha}(t) = \|\nabla f(\alpha(t))\| N(\alpha(t)) \cdot \dot{\alpha}(t) = 0$ since $\dot{\alpha} \perp (N \circ \alpha)$. Therefore, $f \circ \alpha : I \rightarrow \mathbb{R}$ is constant. Also, $f(\alpha(0)) = f(p) = c$ since $p \in S$ and $S = f^{-1}(c)$. Thus, $f \circ \alpha = c \implies \alpha \subset f^{-1}(c)$. Thus, α is a parametrised curve in n -surface S . \square

18.7.4 Properties of Geodesics

1. $\dot{\alpha} \perp \ddot{\alpha}$ since $\dot{\alpha} \in S_{\alpha(t)}$ and $\ddot{\alpha} \in S_{\alpha(t)}^\perp$

In other words, the acceleration is orthogonal to the velocity vector.

2. Constant Speed, $\|\mathbf{v}\|$

We have, $(\dot{\alpha} \cdot \dot{\alpha})' = \ddot{\alpha} \cdot \dot{\alpha} + \dot{\alpha} \cdot \ddot{\alpha} = 2\dot{\alpha} \cdot \ddot{\alpha} = 0$

Clearly, $\frac{d}{dt} \|\dot{\alpha}\|^2 = (\dot{\alpha} \cdot \dot{\alpha})' = 0$. Therefore, α has constant speed.

Remark : Geodesics in cylinder over S^1 are horizontal circles, vertical lines, helix or a constant.

18.8 Parallel Transport

covariant derivative Covariant derivative of \mathbf{X} is the vector field \mathbf{X}' tangent to S along α given by $\mathbf{X}'(t) = \dot{\mathbf{X}}(t) - [\dot{\mathbf{X}}(t) \cdot \mathbf{N}(\alpha(t))]\mathbf{N}(\alpha(t))$. And it is independent of the choice of the orientation.

covariant acceleration Let $\alpha : I \rightarrow S$ be a parametrised curve in S . Then covariant acceleration of α is $(\dot{\alpha})' = \ddot{\alpha} - [\ddot{\alpha} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha)$ along α .

18.8.1 Properties of Covariant Derivative

Let \mathbf{X}, \mathbf{Y} be smooth vector fields tangent to S along α . Then, $\mathbf{X} \cdot (\mathbf{N} \circ \alpha) = \mathbf{0}$.

$$1. (\mathbf{X} + \mathbf{Y})' = \mathbf{X}' + \mathbf{Y}'$$

$$\begin{aligned} (\mathbf{X} + \mathbf{Y})' &= (\dot{\mathbf{X}} + \dot{\mathbf{Y}}) - [(\dot{\mathbf{X}} + \dot{\mathbf{Y}}) \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= (\dot{\mathbf{X}} + \dot{\mathbf{Y}}) - [(\dot{\mathbf{X}} \cdot (\mathbf{N} \circ \alpha)) + (\dot{\mathbf{Y}} \cdot (\mathbf{N} \circ \alpha))](\mathbf{N} \circ \alpha) \\ &= \left(\dot{\mathbf{X}} - [\dot{\mathbf{X}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \right) + \left(\dot{\mathbf{Y}} - [\dot{\mathbf{Y}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \right) \\ &= \mathbf{X}' + \mathbf{Y}' \end{aligned}$$

$$2. (f\mathbf{X})' = f'\mathbf{X} + f\mathbf{X}'$$

$$\begin{aligned} (f\mathbf{X})' &= (f\dot{\mathbf{X}})' - [(f\dot{\mathbf{X}}) \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= (f'\mathbf{X} + f\dot{\mathbf{X}})' - [(f'\mathbf{X} + f\dot{\mathbf{X}}) \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= f'(\mathbf{X} - [\mathbf{X} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha)) + f(\dot{\mathbf{X}} - [\dot{\mathbf{X}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha)) \\ &= f'\mathbf{X} + f\dot{\mathbf{X}} \text{ since } \mathbf{X} \cdot (\mathbf{N} \circ \alpha) = \mathbf{0} \end{aligned}$$

$$3. (\mathbf{X} \cdot \mathbf{Y})' = \mathbf{X}' \cdot \mathbf{Y} + \mathbf{X} \cdot \mathbf{Y}'$$

$$\begin{aligned} (\mathbf{X} \cdot \mathbf{Y})' &= \dot{\mathbf{X}} \cdot \mathbf{Y} + \mathbf{X} \cdot \dot{\mathbf{Y}} \\ &= \dot{\mathbf{X}} \cdot \mathbf{Y} - [\dot{\mathbf{X}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \cdot \mathbf{Y} + \mathbf{X} \cdot \dot{\mathbf{Y}} - [\dot{\mathbf{Y}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \cdot \mathbf{X} \end{aligned}$$

Substituting $\mathbf{X} \cdot (\mathbf{N} \circ \alpha) = \mathbf{0}$ and $\mathbf{Y} \cdot (\mathbf{N} \circ \alpha) = \mathbf{0}$, we get

$$\begin{aligned} &= \dot{\mathbf{X}} \cdot \mathbf{Y} - [\dot{\mathbf{X}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \cdot \mathbf{Y} + \mathbf{X} \cdot \dot{\mathbf{Y}} - [\dot{\mathbf{Y}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \cdot \mathbf{X} \\ &= (\dot{\mathbf{X}} - [\dot{\mathbf{X}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha)) \cdot \mathbf{Y} + \mathbf{X} \cdot (\dot{\mathbf{Y}} - [\dot{\mathbf{Y}} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha)) \\ &= \mathbf{X}' \cdot \mathbf{Y} + \mathbf{X} \cdot \mathbf{Y}' \end{aligned}$$

18.8.2 Parallelism

We have seen earlier that, lines that look straight on a surface (geodesics) is not necessarily straight from outside. The same way, two vectors that look *parallel on a surface* (Levi-Civita parallel) is not necessarily parallel.

Euclidean parallel (p, v) and (q, w) are Euclidean parallel if $v = w$.

Let $\alpha : I \rightarrow \mathbb{R}^{n+1}$ be a parametrised curve. A vector field \mathbf{X} is Euclidean parallel along α if the associated function X is constant say, v . Then,

$$\dot{\mathbf{X}} = \left(\alpha(t), \frac{d}{dt} X(t) \right) = \left(\alpha(t), \frac{dv}{dt} \right) = (\alpha(t), 0) = \mathbf{0}$$

Levi-Civita parallel A vector field \mathbf{X} tangent to the surface S along α is (Levi-Civita) parallel if \mathbf{X} is a constant vector field along α with respect to the surface S . That is, $\mathbf{X}' = \mathbf{0}$.

Properties of Levi-Civita parallel

Applying the properties of covariant derivative, we get

1. If \mathbf{X} is parallel along α , then \mathbf{X} has constant length. That is, $\|\mathbf{X}\|' = 0$.

$$\frac{d}{dt} \|\mathbf{X}\|^2 = \frac{d}{dt} \mathbf{X} \cdot \mathbf{X} = 2\mathbf{X}' \cdot \mathbf{X} = 2(\mathbf{0} \cdot \mathbf{X}) = 0$$

2. If \mathbf{X}, \mathbf{Y} are parallel along α , then $\mathbf{X} \cdot \mathbf{Y}$ is constant along α .

$$(\mathbf{X} \cdot \mathbf{Y})' = \mathbf{X}' \cdot \mathbf{Y} + \mathbf{X} \cdot \mathbf{Y}' = \mathbf{0} \cdot \mathbf{Y} + \mathbf{X} \cdot \mathbf{0} = 0$$

3. If \mathbf{X}, \mathbf{Y} are parallel along α , then angle between them is constant along α .

$$\theta = \cos^{-1} \frac{\mathbf{X} \cdot \mathbf{Y}}{\|\mathbf{X}\| \|\mathbf{Y}\|} = \cos^{-1} \kappa, \text{ since } \|\mathbf{X}\|, \|\mathbf{Y}\|, \mathbf{X} \cdot \mathbf{Y} \text{ are constant}$$

4. If \mathbf{X}, \mathbf{Y} are parallel along α , then $\mathbf{X} + \mathbf{Y}, c\mathbf{X}$ are parallel along α .

$$(\mathbf{X} + \mathbf{Y})' = \mathbf{X}' + \mathbf{Y}' = \mathbf{0} + \mathbf{0} = \mathbf{0}$$

$$(c\mathbf{X})' = c\mathbf{X}' = c\mathbf{0} = \mathbf{0}$$

5. The velocity vector field along a parametrised curve α in S is parallel if and only if α is a geodesic.

$$(\dot{\alpha})' = \ddot{\alpha} - [\ddot{\alpha} \cdot (\mathbf{N} \circ \alpha)] (\mathbf{N} \circ \alpha) = \mathbf{0} \iff \ddot{\alpha} \in S_{\alpha(t)}^{\perp}$$

Parametrised curve α is geodesic in S if and only if covariant acceleration $(\dot{\alpha})'$ is zero along α since $\ddot{\alpha} \in S_{\alpha(t)}^{\perp}$ and $[\ddot{\alpha} \cdot (\mathbf{N} \circ \alpha)] (\mathbf{N} \circ \alpha) = \ddot{\alpha}$.

We have, the notation f' and \mathbf{X}' . You should keep a note of the fundamental differences. f' refers to the derivative of a function along α with respect to the parameter t . And \mathbf{X}' refers to the covariant derivative of a vector field tangent to S along α . You should always check whether it is real value OR vector at a point to understand what they really mean.

For example, $\|\mathbf{X}\|'$ is a derivative of a real valued function (derivative of the length of the vectors assigned at different points of a parametrised curve). And it has nothing to do with the covariant derivative \mathbf{X}' .

Theorem 18.8.1. *Let S be an n -surface and $\alpha : I \rightarrow S$ be a parametrised curve in S . Let $t_0 \in I$ and $\mathbf{v} \in S_{\alpha(t_0)}$. Then there exists a unique vector field \mathbf{V} tangent to S along α which is parallel and has $\mathbf{V}(t_0) = \mathbf{v}$.*

Proof. Let S be an n -surface with orientation \mathbf{N} . Suppose \mathbf{V} is a vector field tangent to S along α and is parallel. That is, $\mathbf{V}(t) \in S_{\alpha(t)}$ and $\mathbf{V}' = \mathbf{0}$.

$$\begin{aligned}\mathbf{V}' &= \dot{\mathbf{V}} - \left[\dot{\mathbf{V}} \cdot (\mathbf{N} \circ \alpha) \right] (\mathbf{N} \circ \alpha) \\ &= \dot{\mathbf{V}} - \left[(\mathbf{V} \cdot (\mathbf{N} \circ \alpha))' - \mathbf{V} \cdot (\mathbf{N} \dot{\circ} \alpha) \right] (\mathbf{N} \circ \alpha) \\ &= \dot{\mathbf{V}} + \left[\mathbf{V} \cdot (\mathbf{N} \dot{\circ} \alpha) \right] (\mathbf{N} \circ \alpha) \text{ since } \mathbf{V} \perp \mathbf{N} \\ \dot{\mathbf{V}} + [\mathbf{V} \cdot (\mathbf{N} \dot{\circ} \alpha)] (\mathbf{N} \circ \alpha) &= \mathbf{0}\end{aligned}\tag{18.20}$$

Equating the components on either sides of the equation(18.20), we get the following system of $n + 1$ first order differential equations,

$$\frac{dV_i}{dt} + \sum_{j=1}^{n+1} [V_j (\mathbf{N}_j \dot{\circ} \alpha)'] (\mathbf{N}_i \circ \alpha) = 0, \quad \forall i \tag{18.21}$$

By the existence and uniqueness theorem for first order differential equations, there exists $V_1(t), V_2(t), \dots, V_{n+1}(t)$ satisfying the sytem of equations with initial condition $\mathbf{V}(t_0) = (\alpha(t_0), V_1(t_0), V_2(t_0), \dots, V_{n+1}(t_0)) = \mathbf{v}$.

It remains to prove that \mathbf{V} is tangent to S along α . By taking dot product with $\mathbf{N} \circ \alpha$ on either sides of equation(18.20), we get

$$(\mathbf{V} \cdot \mathbf{N} \circ \alpha)' = \dot{\mathbf{V}} \cdot (\mathbf{N} \circ \alpha) + \mathbf{V} \cdot (\mathbf{N} \dot{\circ} \alpha) = 0 \tag{18.22}$$

Thus, $\mathbf{V} \cdot (\mathbf{N} \circ \alpha) = \kappa$ is constant along α . However, $\mathbf{V}(t_0) \cdot (\mathbf{N} \circ \alpha)(t_0) = \mathbf{v} \cdot \mathbf{N}(\alpha(t_0)) = 0$ since $\mathbf{v} \in S_{\alpha(t_0)}$ is a tangent vector and $\mathbf{v} \perp \mathbf{N}$. Therefore, $\mathbf{V} \cdot (\mathbf{N} \circ \alpha) = 0$. And since \mathbf{V} satisfies equation(18.20), this vector field is tangent to S along α and is parallel. \square

Corollary 18.8.1.1. *Let S be a 2-surface in \mathbb{R}^3 and let $\alpha : I \rightarrow S$ be a parametrised curve in S with $\dot{\alpha} \neq \mathbf{0}$. Then the vector field \mathbf{X} tangent to S along α is parallel along α if and only if both $\|\mathbf{X}\|$ and the angle between \mathbf{X} and $\dot{\alpha}$ are constant along α .*

Proof. Sufficient Part : Let \mathbf{X} be a tangent vector field tangent to S along α . And α is a geodesic in S . Then \mathbf{X} is parallel along α . Also, we have $\dot{\alpha}$ is also parallel along α since $(\dot{\alpha})' = \mathbf{0}$. Thus $\|\mathbf{X}\|$ is constant and the angle between \mathbf{X} and $\dot{\alpha}$ is constant by properties Levi-Civita parallelism.

Necessary Part : Let \mathbf{X} be a vector field tangent to S along α and α is a geodesic in S . Then $\|\dot{\alpha}\|$ is constant. Suppose $\|\mathbf{X}\|$ and the angle θ between \mathbf{X} and $\dot{\alpha}$ are constant. Since S is a 2-surface, the tangent space $S_{\alpha(t)}$ is spanned by two tangent vectors \mathbf{v} and $\dot{\alpha}(t)$ such that $\mathbf{v} \perp \dot{\alpha}(t)$ and $\|\mathbf{v}\| = 1$.

Let \mathbf{V} be the unique vector field tangent to S along α which is parallel, $\mathbf{V} \cdot \dot{\alpha} = 0$ and $\|\mathbf{V}\| = 1$. Then, any vector field \mathbf{X} tangent to S along α can be written as linear combination of \mathbf{V} and $\dot{\alpha}$. That is, $\mathbf{X} = f\dot{\alpha} + g\mathbf{V}$

$$\cos \theta = \frac{\mathbf{X} \cdot \dot{\alpha}}{\|\mathbf{X}\| \|\dot{\alpha}\|} = \frac{(f\dot{\alpha} + g\mathbf{V}) \cdot \dot{\alpha}}{\|\mathbf{X}\| \|\dot{\alpha}\|} = \frac{f\|\dot{\alpha}\|}{\|\mathbf{X}\|} \implies f \text{ is constant}$$

$$\|\mathbf{X}\|^2 = (f\dot{\alpha} + g\mathbf{V}) \cdot (f\dot{\alpha} + g\mathbf{V}) = f^2\|\dot{\alpha}\| + g^2 \implies g \text{ is constant}$$

Since \mathbf{V} and $\dot{\alpha}$ are parallel along α , their linear combination \mathbf{X} is also parallel along α by linearity property (#4) of Levi-Civita parallelism. \square

18.8.3 Transporting tangent vectors using parallelism

Suppose you are standing on an n -surface with your hand stretched out as in a pledge. Then you can move on that surface along any smooth road (without sharp turns, bumps or potholes) from one point to another keeping your hand in the same position. This is what a parallel transport does to tangent vectors. And we can calculate the direction you will be pointing, at each point of your journey.

Definitions 18.8.1 (Parallel transport). Let S be an n -surface. Let $p, q \in S$, and let $\alpha : [a, b] \rightarrow S$ be a (smooth) parametrised curve in S from p to q . **Parallel transport** is the function $P_\alpha : S_p \rightarrow S_q$ defined by $P_\alpha(\mathbf{v}) = \mathbf{V}(b)$ where $\mathbf{v} \in S_p$ and \mathbf{V} is the unique vector field tangent to S along α which is parallel and $\mathbf{V}(a) = \mathbf{v}$.

The following theorem says that, *Parallel transports along piecewise smooth parametrised curves are vector space isomorphisms preserving dot products.*

Theorem 18.8.2. Let S be an n -surface in \mathbb{R}^{n+1} . Let $p, q \in S$ and let α be a piecewise smooth parametrised curve from p to q . Then parallel transport $P_\alpha : S_p \rightarrow S_q$ along α is a vector space isomorphism which preserves dot products.

Proof. Let S be an n -surface and $p, q \in S$. Let $\alpha : [a, b] \rightarrow S$ be a piecewise smooth parametrised curve from p to q . Let $P_\alpha : S_p \rightarrow S_q$ be a parallel transport and $\mathbf{v}, \mathbf{w} \in S_p$. Clearly, $\mathbf{v} + \mathbf{w}, c\mathbf{v} \in S_p$

To prove that P_α is a vector space isomorphism preserving dot products, we need to prove the following

1. P_α is linear, $P_\alpha(\mathbf{v} + \mathbf{w}) = P_\alpha(\mathbf{v}) + P_\alpha(\mathbf{w})$ and $P_\alpha(c\mathbf{v}) = cP_\alpha(\mathbf{v})$

$$P_\alpha(\mathbf{v} + \mathbf{w}) = (\mathbf{V} + \mathbf{W})(b) = \mathbf{V}(b) + \mathbf{W}(b) = P_\alpha(\mathbf{v}) + P_\alpha(\mathbf{w})$$

$$P_\alpha(c\mathbf{v}) = (c\mathbf{V})(b) = c\mathbf{V}(b) = cP_\alpha(\mathbf{v})$$

2. P_α is bijective

$$\|P_\alpha(\mathbf{v})\| = \|\mathbf{V}(b)\| = 0 \implies \|\mathbf{v}\| = 0 \implies \ker(P_\alpha) = \{\mathbf{0}\}$$

Thus P_α is a linear map from n -dimensional vector space S_p into another n -dimensional vector space S_q . Thus, P_α is onto.

3. P_α preserves dot products, $P_\alpha(\mathbf{v}) \cdot P_\alpha(\mathbf{w}) = \mathbf{v} \cdot \mathbf{w}$, $\forall \mathbf{v}, \mathbf{w} \in S_p$
 We have, \mathbf{V} and \mathbf{W} are parallel along α . Then $\mathbf{V} \cdot \mathbf{W}$ is constant.
 That is, $\mathbf{V}(t) \cdot \mathbf{W}(t) = \kappa$ for every $t \in [a, b]$.

$$P_\alpha(\mathbf{v}) \cdot P_\alpha(\mathbf{w}) = \mathbf{V}(b) \cdot \mathbf{W}(b) = \kappa = \mathbf{V}(a) \cdot \mathbf{W}(a) = \mathbf{v} \cdot \mathbf{w}$$

□

18.9 The Weingarten Map

The Weingarten map L_p gives information about the shape of a surface S at a point $p \in S$.

$\nabla_{\mathbf{v}} f$ is the derivative of a function $f : U \rightarrow \mathbb{R}$ with respect to a vector tangent to S , say $\mathbf{v} = (p, v)$ where U is an open subset of \mathbb{R}^{n+1} , $p \in U$ and $\alpha : I \rightarrow S$ is any parametrised curve in S with $\dot{\alpha}(t_0) = \mathbf{v}$ and $\alpha(t_0) = p$. And $\nabla_{\mathbf{v}} f$ is given by,

$$\nabla_{\mathbf{v}} f = (f \circ \alpha)'(t_0) = \nabla f(\alpha(t_0)) \cdot \dot{\alpha}(t_0) = \nabla f(p) \cdot \mathbf{v}$$

For example : if $f(x_1, x_2) = x_1^2 - x_2^2$ and $\mathbf{v} = (1, 1, \cos \theta, \sin \theta)$ Then, $p = (1, 1)$, $v = (\cos \theta, \sin \theta)$ and $\nabla f = (x_1, x_2, 2x_1, -2x_2)$. Therefore, $\nabla_{\mathbf{v}} f = \nabla f(p) \cdot \mathbf{v} = (1, 1, 2, -2) \cdot (1, 1, \cos \theta, \sin \theta) = 2 \cos \theta - 2 \sin \theta$.

Note : $\nabla_{\mathbf{v}} f$ is independent of the choice of α .

$\nabla_{\mathbf{v}} \mathbf{X}$ is the derivative of a smooth vector field \mathbf{X} on open subset U of \mathbb{R}^{n+1} with respect to $\mathbf{v} \in S_{\alpha(t)}$. And $\nabla_{\mathbf{v}} \mathbf{X}$ is given by

$$\nabla_{\mathbf{v}} \mathbf{X} = (\alpha(t), \nabla_{\mathbf{v}} X_1, \nabla_{\mathbf{v}} X_2, \dots, \nabla_{\mathbf{v}} X_{n+1})$$

For example, if $\mathbf{X}(x_1, x_2) = (x_1, x_2, x_1 x_2, x_2^2)$ and $\mathbf{v} = (1, 0, 0, 1)$. Then $p = (1, 0)$ and the component functions of the associated function X are $X_1(x_1, x_2) = x_1 x_2$ and $X_2(x_1, x_2) = x_2^2$.

We have, $\nabla X_1(x_1, x_2) = (x_1, x_2, x_2, x_1)$, $\nabla X_2(x_1, x_2) = (x_1, x_2, 0, 2x_2)$. Thus, $\nabla_{\mathbf{v}} X_1 = \nabla X_1(1, 0) \cdot \mathbf{v} = (1, 0, 0, 1) \cdot (1, 0, 0, 1) = 1$ and $\nabla_{\mathbf{v}} X_2 = \nabla X_2(1, 0) \cdot \mathbf{v} = (1, 0, 0, 0) \cdot (1, 0, 0, 1) = 0$. Therefore, $\nabla_{\mathbf{v}} \mathbf{X} = (1, 0, \nabla_{\mathbf{v}} X_1, \nabla_{\mathbf{v}} X_2) = (1, 0, 1, 0)$

$D_{\mathbf{v}} \mathbf{X}$ is the covariant derivative of the smooth vector field \mathbf{X} with respect to $\mathbf{v} \in S_{\alpha(t)}$, where \mathbf{X} is tangent to S along α . And $D_{\mathbf{v}} \mathbf{X}$ is given by,

$$D_{\mathbf{v}} \mathbf{X} = \nabla_{\mathbf{v}} \mathbf{X} - [\nabla_{\mathbf{v}} \mathbf{X} \cdot (\mathbf{N} \circ \alpha)] (\mathbf{N} \circ \alpha)$$

It is the component of the derivative $\nabla_{\mathbf{v}} \mathbf{X}$ in the tangent space S_p .

18.9.1 Properties of differentiation, $\nabla_{\mathbf{v}} f$

We have, $\nabla_{\mathbf{v}} f : \mathbb{R}^{n+1} \rightarrow \mathbb{R}$ is a linear map.

$$1. \nabla_{\mathbf{v}+\mathbf{w}}f = \nabla_{\mathbf{v}}f + \nabla_{\mathbf{w}}f$$

$$\nabla_{\mathbf{v}+\mathbf{w}}f = \nabla f(p) \cdot (\mathbf{v} + \mathbf{w}) = \nabla f(p) \cdot \mathbf{v} + \nabla f(p) \cdot \mathbf{w}$$

$$2. \nabla_{c\mathbf{v}}f = c\nabla_{\mathbf{v}}f$$

$$\nabla_{c\mathbf{v}}f = \nabla f(p) \cdot c\mathbf{v} = c(\nabla f(p) \cdot \mathbf{v}) = c\nabla_{\mathbf{v}}f$$

Clearly, we have $\nabla_{\mathbf{v}}(f + g) = \nabla_{\mathbf{v}}f + \nabla_{\mathbf{v}}g$ and $\nabla_{\mathbf{v}}cf = c\nabla_{\mathbf{v}}f$.

18.9.2 Properties of differentiation, $\nabla_{\mathbf{v}}\mathbf{X}$

$$1. \nabla_{\mathbf{v}}(\mathbf{X} + \mathbf{Y}) = \nabla_{\mathbf{v}}\mathbf{X} + \nabla_{\mathbf{v}}\mathbf{Y}$$

$$\begin{aligned} \nabla_{\mathbf{v}}(\mathbf{X} + \mathbf{Y}) &= (\alpha(t), \nabla_{\mathbf{v}}(X_1 + Y_1), \nabla_{\mathbf{v}}(X_2 + Y_2), \dots, \nabla_{\mathbf{v}}(X_{n+1} + Y_{n+1})) \\ &= (\alpha(t), \nabla_{\mathbf{v}}X_1 + \nabla_{\mathbf{v}}Y_1, \nabla_{\mathbf{v}}X_2 + \nabla_{\mathbf{v}}Y_2, \dots, \nabla_{\mathbf{v}}X_{n+1} + \nabla_{\mathbf{v}}Y_{n+1}) \\ &= (\alpha(t), \nabla_{\mathbf{v}}X_1, \nabla_{\mathbf{v}}X_2, \dots, \nabla_{\mathbf{v}}X_{n+1}) + (\alpha(t), \nabla_{\mathbf{v}}Y_1, \nabla_{\mathbf{v}}Y_2, \dots, \nabla_{\mathbf{v}}Y_{n+1}) \\ &= \nabla_{\mathbf{v}}\mathbf{X} + \nabla_{\mathbf{v}}\mathbf{Y} \end{aligned}$$

$$2. \nabla_{\mathbf{v}}(f\mathbf{X}) = (\nabla_{\mathbf{v}}f)\mathbf{X}(p) + f(p)\nabla_{\mathbf{v}}\mathbf{X}$$

$$\begin{aligned} \nabla_{\mathbf{v}}(f\mathbf{X}) &= (\alpha(t), \nabla_{\mathbf{v}}fX_1, \nabla_{\mathbf{v}}fX_2, \dots, \nabla_{\mathbf{v}}fX_{n+1}) \\ &= (\alpha(t), (\nabla_{\mathbf{v}}f)X_1 + f\nabla_{\mathbf{v}}X_1, (\nabla_{\mathbf{v}}f)X_2 + f\nabla_{\mathbf{v}}X_2, \dots, (\nabla_{\mathbf{v}}f)X_{n+1} + f\nabla_{\mathbf{v}}X_{n+1}) \\ &= (\alpha(t), (\nabla_{\mathbf{v}}f)X_1, (\nabla_{\mathbf{v}}f)X_2, \dots, (\nabla_{\mathbf{v}}f)X_{n+1}) \\ &\quad + (\alpha(t), f\nabla_{\mathbf{v}}X_1, f\nabla_{\mathbf{v}}X_2, \dots, f\nabla_{\mathbf{v}}X_{n+1}) \\ &= \nabla_{\mathbf{v}}f(\alpha(t), X_1, X_2, \dots, X_{n+1}) + f(\alpha(t), \nabla_{\mathbf{v}}X_1, \nabla_{\mathbf{v}}X_2, \dots, \nabla_{\mathbf{v}}X_{n+1}) \\ &= (\nabla_{\mathbf{v}}f)\mathbf{X} + f\nabla_{\mathbf{v}}\mathbf{X} \end{aligned}$$

$$3. \nabla_{\mathbf{v}}(\mathbf{X} \cdot \mathbf{Y}) = \nabla_{\mathbf{v}}\mathbf{X} \cdot \mathbf{Y}(p) + \mathbf{X}(p) \cdot \nabla_{\mathbf{v}}\mathbf{Y}$$

$$\begin{aligned} \nabla_{\mathbf{v}}(\mathbf{X} \cdot \mathbf{Y}) &= \nabla_{\mathbf{v}}(X_1Y_1 + X_2Y_2 + \dots + X_{n+1}Y_{n+1}) \\ &= (\nabla_{\mathbf{v}}X_1)Y_1 + X_1\nabla_{\mathbf{v}}Y_1 + (\nabla_{\mathbf{v}}X_2)Y_2 + X_2\nabla_{\mathbf{v}}Y_2 \\ &\quad + \dots + (\nabla_{\mathbf{v}}X_{n+1})Y_{n+1} + X_{n+1}\nabla_{\mathbf{v}}Y_{n+1} \\ &= (\nabla_{\mathbf{v}}X_1)Y_1 + (\nabla_{\mathbf{v}}X_2)Y_2 + \dots + (\nabla_{\mathbf{v}}X_{n+1})Y_{n+1} \\ &\quad + X_1\nabla_{\mathbf{v}}Y_1 + X_2\nabla_{\mathbf{v}}Y_2 + \dots + X_{n+1}\nabla_{\mathbf{v}}Y_{n+1} \\ &= (\alpha(t), \nabla_{\mathbf{v}}X_1, \nabla_{\mathbf{v}}X_2, \dots, \nabla_{\mathbf{v}}X_{n+1}) \cdot (\alpha(t), Y_1, Y_2, \dots, Y_{n+1}) \\ &\quad + (\alpha(t), X_1, X_2, \dots, X_{n+1}) \cdot (\alpha(t), \nabla_{\mathbf{v}}Y_1, \nabla_{\mathbf{v}}Y_2, \dots, \nabla_{\mathbf{v}}Y_{n+1}) \\ &= \nabla_{\mathbf{v}}\mathbf{X} \cdot \mathbf{Y} + \mathbf{X} \cdot \nabla_{\mathbf{v}}\mathbf{Y} \end{aligned}$$

18.9.3 Properties of covariant differentiation, $D_{\mathbf{v}}\mathbf{X}$

$$1. D_{\mathbf{v}}(\mathbf{X} + \mathbf{Y}) = D_{\mathbf{v}}\mathbf{X} + D_{\mathbf{v}}\mathbf{Y}$$

$$\begin{aligned} D_{\mathbf{v}}(\mathbf{X} + \mathbf{Y}) &= \nabla_{\mathbf{v}}(\mathbf{X} + \mathbf{Y}) - [\nabla_{\mathbf{v}}(\mathbf{X} + \mathbf{Y}) \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= \nabla_{\mathbf{v}}\mathbf{X} + \nabla_{\mathbf{v}}\mathbf{Y} - [(\nabla_{\mathbf{v}}\mathbf{X} + \nabla_{\mathbf{v}}\mathbf{Y}) \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= \nabla_{\mathbf{v}}\mathbf{X} - [\nabla_{\mathbf{v}}\mathbf{X} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) + \nabla_{\mathbf{v}}\mathbf{Y} - [\nabla_{\mathbf{v}}\mathbf{Y} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= D_{\mathbf{v}}\mathbf{X} + D_{\mathbf{v}}\mathbf{Y} \end{aligned}$$

$$2. D_v(f\mathbf{X}) = (\nabla_v f)\mathbf{X}(p) + f(p)D_v\mathbf{X}$$

$$\begin{aligned} D_v(f\mathbf{X}) &= \nabla_v(f\mathbf{X}) - [\nabla_v(f\mathbf{X}) \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= (\nabla_v f)\mathbf{X} + f\nabla_v\mathbf{X} - [((\nabla_v f)\mathbf{X} + f\nabla_v\mathbf{X}) \cdot \mathbf{N} \circ \alpha](\mathbf{N} \circ \alpha) \\ &= (\nabla_v f)\mathbf{X} - [(\nabla_v f)\mathbf{X} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) + f\nabla_v\mathbf{X} - [f\nabla_v\mathbf{X} \cdot (\mathbf{N} \circ \alpha)](\mathbf{N} \circ \alpha) \\ &= \nabla_v f\mathbf{X} + fD_v\mathbf{X} \end{aligned}$$

$$3. \nabla_v(\mathbf{X} \cdot \mathbf{Y}) = D_v\mathbf{X} \cdot \mathbf{Y}(p) + \mathbf{X}(p) \cdot D_v\mathbf{Y}$$

$$\begin{aligned} \nabla_v(\mathbf{X} \cdot \mathbf{Y}) &= \nabla_v(X_1Y_1) + \nabla_v(X_2Y_2) + \cdots + \nabla_v(X_{n+1}Y_{n+1}) \\ &= (\nabla_v X_1)Y_1 + X_1\nabla_v Y_1 + (\nabla_v X_2)Y_2 + X_2\nabla_v Y_2 \\ &\quad + \cdots + (\nabla_v X_{n+1})Y_{n+1} + X_{n+1}\nabla_v Y_{n+1} \\ &= (\nabla_v X_1)Y_1 + (\nabla_v X_2)Y_2 + \cdots + (\nabla_v X_{n+1})Y_{n+1} \\ &\quad + X_1\nabla_v Y_1 + X_2\nabla_v Y_2 + \cdots + X_{n+1}\nabla_v Y_{n+1} \\ &= (\nabla_v \mathbf{X}) \cdot \mathbf{Y} + \mathbf{X} \cdot \nabla_v \mathbf{Y} \\ &= D_v\mathbf{X} \cdot \mathbf{Y} + \mathbf{X} \cdot D_v\mathbf{Y} \text{ since } \mathbf{X}, \mathbf{Y} \perp \mathbf{N} \end{aligned}$$

18.9.4 Weingarten Map, L_p

In the case of covariant derivatives along α , we came across a linear map – the parallel transport P_α . Here in the case of covariant derivative with respect to a vector, we have another linear map – the Weingarten map L_p given by

$$L_p(\mathbf{v}) = -\nabla_v \mathbf{N} \quad (18.23)$$

Weingarten map tells us how much the tangent space turns/tilts as we move through p along α on an n -surface S . It gives the information about the shape of S . And is the **shape operator** of S at p .

18.9.5 Properties of Weingarten Map, L_p

1. The normal component of acceleration is same for all parametrised curves in S passing through p with velocity \mathbf{v} , $\ddot{\alpha}(t_0) \cdot \mathbf{N}(p) = L_p(\mathbf{v}) \cdot \mathbf{v}$

$$\begin{aligned} 0 &= \dot{\alpha} \cdot (\mathbf{N} \circ \alpha) \text{ since } \dot{\alpha} \perp \mathbf{N} \\ 0 &= [\dot{\alpha} \cdot (\mathbf{N} \circ \alpha)]' \\ &= \ddot{\alpha} \cdot (\mathbf{N} \circ \alpha) + \dot{\alpha} \cdot (\mathbf{N} \dot{\circ} \alpha) \end{aligned}$$

Rearranging the terms and evaluating at $t = t_0$, we get

$$\begin{aligned} \ddot{\alpha}(t_0) \cdot (\mathbf{N} \circ \alpha(t_0)) &= -\dot{\alpha}(t_0) \cdot (\mathbf{N} \dot{\circ} \alpha)(t_0) \\ \ddot{\alpha}(t_0) \cdot \mathbf{N}(p) &= \mathbf{v} \cdot L_p(\mathbf{v}) \text{ since } (\mathbf{N} \dot{\circ} \alpha)(t_0) = \nabla_v \mathbf{N} = -L_p(\mathbf{v}) \end{aligned}$$

2. Weingarten map is self adjoint, $L_p(\mathbf{v}) \cdot \mathbf{w} = \mathbf{v} \cdot L_p(\mathbf{w})$

Clearly, we have $\mathbf{v}, \mathbf{w} \in S_{\alpha(t)}$ and $\nabla f \in S_{\alpha(t)}^\perp$. Also note that

$$\begin{aligned}\nabla f(p) &= \left(p, \frac{\partial f}{\partial x_1}(p), \frac{\partial f}{\partial x_2}(p), \dots, \frac{\partial f}{\partial x_{n+1}}(p) \right) \\ \nabla_{\mathbf{v}}(\nabla f(p)) &= \left(p, \nabla_{\mathbf{v}} \frac{\partial f}{\partial x_1}(p), \nabla_{\mathbf{v}} \frac{\partial f}{\partial x_2}(p), \dots, \nabla_{\mathbf{v}} \frac{\partial f}{\partial x_{n+1}}(p) \right) \\ &= \left(p, \nabla \left(\frac{\partial f}{\partial x_1} \right) (p) \cdot \mathbf{v}, \nabla \left(\frac{\partial f}{\partial x_2} \right) (p) \cdot \mathbf{v}, \dots, \nabla \left(\frac{\partial f}{\partial x_{n+1}} \right) (p) \cdot \mathbf{v} \right) \\ &= \left(p, \sum_{j=1}^{n+1} \frac{\partial^2 f}{\partial x_j \partial x_1}(p) v_j, \sum_{j=1}^{n+1} \frac{\partial^2 f}{\partial x_j \partial x_2}(p) v_j, \dots, \sum_{j=1}^{n+1} \frac{\partial^2 f}{\partial x_j \partial x_{n+1}}(p) v_j \right)\end{aligned}$$

$$\begin{aligned}L_p(\mathbf{v}) \cdot \mathbf{w} &= -\nabla_{\mathbf{v}} \mathbf{N} \cdot \mathbf{w} \\ &= -\nabla_{\mathbf{v}} \left(\frac{\nabla f}{\|\nabla f\|} \right) \cdot \mathbf{w} \\ &= \left[\left(-\nabla_{\mathbf{v}} \frac{1}{\|\nabla f\|} \right) \nabla f - \frac{1}{\|\nabla f\|} \nabla_{\mathbf{v}}(\nabla f) \right] \cdot \mathbf{w} \\ &= \left[-\frac{1}{\|\nabla f\|} \nabla_{\mathbf{v}}(\nabla f) \right] \cdot \mathbf{w} \text{ since } \nabla f \perp \mathbf{w} \\ &= -\frac{1}{\|\nabla f\|} \left[\sum_{k=1}^{n+1} \left(\sum_{j=1}^{n+1} \frac{\partial^2 f}{\partial x_j \partial x_k}(p) v_j \right) w_k \right] \\ &= -\frac{1}{\|\nabla f\|} \sum_{j,k=1}^{n+1} \frac{\partial^2 f}{\partial x_j \partial x_k}(p) v_j w_k\end{aligned}$$

Similarly,

$$L_p(\mathbf{w}) \cdot \mathbf{v} = -\frac{1}{\|\nabla f\|} \sum_{j,k=1}^{n+1} \frac{\partial^2 f}{\partial x_j \partial x_k}(p) w_j v_k$$

Since $\frac{\partial^2 f}{\partial x_j \partial x_k}(p) = \frac{\partial^2 f}{\partial x_k \partial x_j}(p)$, both the sums are equal

$$L_p(\mathbf{v}) \cdot \mathbf{w} = \mathbf{v} \cdot L_p(\mathbf{w})$$

18.10 The Curvature of Plane Curves

curvature Let C be plane curve in \mathbb{R}^2 . The curvature of C is the function $\kappa : \mathbb{R}^2 \rightarrow \mathbb{R}$ defined by $\kappa(p) = L_p(\mathbf{v}) \cdot \mathbf{v} / \|\mathbf{v}\|^2$

For a parametrised curve α in C with $\dot{\alpha} \neq \mathbf{0}$,

$$\kappa(\alpha(t)) = \frac{\ddot{\alpha}(t) \cdot (\mathbf{N} \circ \alpha(t))}{\|\dot{\alpha}(t)\|^2}$$

For unit speed parametrised curve passing through p , curvature at p is the normal component of acceleration at that point.

local parametrisation of plane curve Let C be a plane curve in \mathbb{R}^2 . Then C is an oriented 1-surface. Parametrisation of a segment of C containing p is a function $\alpha : I \rightarrow C$ such that

1. α is regular. That is, $\dot{\alpha}(t) \neq \mathbf{0}$, $\forall t \in I$
2. α is consistently oriented with C . That is, the orientation at $\alpha(t)$ should be the same as that of C at that point.
3. $p \in \alpha(I)$

circle of curvature Let C be a plane curve in \mathbb{R}^2 with orientation \mathbf{N} . The circle of curvature of C at a point p is the unique oriented circle O with orientation \mathbf{N}_1 such that

1. O is tangent to C at p , $C_p = O_p$.
2. O is oriented consistently with C , $\mathbf{N}(p) = \mathbf{N}_1(p)$ and
3. Its normal turns at the same rate at p as the normal of the plane curve C , $\nabla_{\mathbf{v}}\mathbf{N} = \nabla_{\mathbf{v}}\mathbf{N}_1$

This circle, O is the circle which hugs the curve C closest among all circles containing p .

α is regular. That is, $\dot{\alpha}(t) \neq \mathbf{0}$, $\forall t \in I$

radius of curvature The radius of the circle of curvature of C at point p . And radius of curvature, $r = 1/|\kappa(p)|$.

center of curvature The center of the circle of curvature of C at point p .

Meaning of Sign of Curvature If curvature at p is positive, $\kappa(p) > 0$ then the curve at p is turning towards $\mathbf{N}(p)$. If curvature at p is negative, $\kappa(p) < 0$ then the curve at p is turning away from $\mathbf{N}(p)$.

Construction of a local parametrisation We have, $C = f^{-1}(c)$ where $\nabla f(q) \neq \mathbf{0}$. Consider $\mathbf{X}(q) = \left(q, \frac{\partial f}{\partial x_2}(q), -\frac{\partial f}{\partial x_1}(q)\right)$. The maximal integral curve α through p in \mathbf{X} is a local parametrisation of a segment of C containing p .

For example, (refer : Exercise 10.3c) $f(x_1, x_2) = x_2 - ax_1^2$. Then $C = f^{-1}(c) = \{(x_1, x_2) \in \mathbb{R}^2 : x_2 - ax_1^2 = c, a \neq 0\}$. The global parametrisation of C with orientation $\nabla f/\|\nabla f\|$ can be obtained by constructing the maximal integral curve for each segment of a piecewise smooth plane curve. Since the given curve is smooth, it is sufficient to construct one integral curve through any point on that curve.

We have, $\frac{dx_1}{dt}(t) = -2ax_1(t)$ and $\frac{dx_2}{dt} = 1$. Thus, $x_1(t) = \cos 2at + c_1$ and $x_2(t) = t + c_2$. We have $(0, c) \in C$ since $c - a0^2 = c$. Therefore, $x_1(0) = 0 = c_1$ and $x_2(0) = c = c_2$. Thus, we have $\alpha : I \rightarrow \mathbb{R}^2$ given by $\alpha(t) = (\cos 2at, t + c)$ as a global parametrisation of C .

18.10.1 Unit speed local parametrisation of C is unique.

Theorem 18.10.1. *Local parametrisations of plane curves are unique upto reparametrisation.*

Proof. Let α be the unit speed local parametrisation which is the maximal integral curve through p . Let $\beta : \tilde{I} \rightarrow C$ be any parametrisation of a segment of C containing p . Then it is enough to prove that there exists a smooth function $h : \tilde{I} \rightarrow \mathbb{R}$ such that $h'(t) > 0$ and $\beta(t) = \alpha(h(t))$ for every $t \in \tilde{I}$.

Define $h : \tilde{I} \rightarrow \mathbb{R}$ defined by

$$h(t) = \int_{t_0}^t \|\dot{\beta}(u)\| du$$

Clearly, h is monotone, $h(t_0) = 0$ and $h'(t) = \|\dot{\beta}(t)\|$. Since h is monotone and strictly increasing, $h(h^{-1}(t)) = t$. Thus, $(h \circ h^{-1})'(t) = h'(h^{-1}(t))(h^{-1})'(t)$. Therefore, $(h^{-1})'(t) = 1/h'(h^{-1}(t))$.

Now, $\beta \circ h^{-1}$ is a reparametrised curve with velocity, $(\beta \circ h^{-1})(t)$ given by,

$$\begin{aligned} (\beta \circ h^{-1})(t) &= \dot{\beta}(h^{-1}(t)) (h^{-1})'(t) \text{ by chain rule} \\ &= \dot{\beta}(h^{-1}(t)) / h'(h^{-1}(t)) \\ &= \dot{\beta}(h^{-1}(t)) / \|\dot{\beta}(h^{-1}(t))\| \text{ since } h'(t) = \|\dot{\beta}(t)\| \end{aligned}$$

We know that, $\mathbf{X}(\beta(h^{-1}(t)))$ spans the tangent space $S_{\beta(h^{-1}(t))}$.

$$= \mathbf{X}(\beta(h^{-1}(t))) \text{ since } \beta \circ h^{-1}(t) / \|\beta \circ h^{-1}(t)\| = \mathbf{X}(\beta(h^{-1}(t)))$$

Thus, $\beta \circ h^{-1}$ is an integral curve through p in \mathbf{X} . By uniqueness of integral curves, $\beta \circ h^{-1}(t) = \alpha(t)$ for all $t \in \tilde{I}$. \square

18.11 Arc Length and Line Integrals

length of arc The length of parametrised arc α is the integral of its speed.

Let $\alpha : [a, b] \rightarrow \mathbb{R}^{n+1}$. Length of α is

$$l(\alpha) = \int_a^b \|\dot{\alpha}(t)\| dt \quad (18.24)$$

For Example (refer : Exercise 11.4) : Given $\alpha : [0, 2\pi] \rightarrow \mathbb{R}^4$, $\alpha(t) = (\cos t, \sin t, \cos t, \sin t)$. Then, $\dot{\alpha}(t) = (\alpha(t), -\sin t, \cos t, -\sin t, \cos t)$. We have, $\|\dot{\alpha}(t)\| = \sqrt{\sin^2 t + \cos^2 t + \sin^2 t + \cos^2 t} = \sqrt{2}$. Therefore, $l(\alpha) = \int_0^{2\pi} \sqrt{2} dt = 2\sqrt{2}\pi$.

Note : Length of arc is the total distance travelled.

For example : $\alpha_1 : [0, 2\pi] \rightarrow \mathbb{R}^2$ defined by $\alpha_1(t) = (\cos t, \sin t)$ has its length 2π , the perimeter of unit circle. But, for $\alpha_2 : [0, 4\pi] \rightarrow \mathbb{R}^2$ defined by $\alpha_2(t) = (\cos t, \sin t)$ has its length 4π . And $\alpha : \mathbb{R} \rightarrow \mathbb{R}^2$ defined by $\alpha(t) = (\cos t, \sin t)$ has its length $+\infty$.

18.11.1 Properties of Arc Length

1. Arc length is preserved under reparametrisation.

Let $\alpha : [a, b] \rightarrow \mathbb{R}^{n+1}$ be a parametrised curve. And let $\beta : [c, d] \rightarrow \mathbb{R}^{n+1}$ be a reparametrisation of α defined by $\beta(t) = \alpha(h(t))$ where $h : [c, d] \rightarrow [a, b]$. Then length of β is given by,

$$\begin{aligned} l(\beta) &= \int_c^d \|\dot{\beta}(t)\| dt \\ &= \int_c^d \|\dot{\alpha}(h(t))\| h'(t) dt \\ &= \int_a^b \|\alpha(u)\| du = l(\alpha) \end{aligned}$$

2. Unit speed arcs are parametrised by arc length.

Let α be a unit speed parametrised curve. That is, $\|\dot{\alpha}\| = 1$. Therefore $l(\alpha) = \int_a^b dt = b - a$. Clearly, length of the arc is the length of the parameter interval.

Theorem 18.11.1 (Existence of Global Parametrisation). *Let C be an oriented plane curve. Then there exists a global parametrisation of C if and only if C is connected.*

Proof. Sufficient Part : Suppose plane curve C has a global parametrisation, $\alpha : I \rightarrow \mathbb{R}^{n+1}$. Let $p, q \in C$, then $p = \alpha(t_1)$ and $q = \alpha(t_2)$ for some $t_1, t_2 \in I$. WLOG $t_1 < t_2$ and clearly, there exists a path from p to q obtained restricting α to $[t_1, t_2]$. Therefore, the plane curve C is path-connected and thus connected.

Necessary Part : Step 1 Construction of α

Let \mathbf{X} be the unit tangent vector field on C (obtained by rotating ∇f by an angle of $-\pi/2$ and orientation $\mathbf{N} = \nabla f / \|\nabla f\|$). Suppose the plane curve $C = f^{-1}(c)$ is connected. Let $p \in C$ and α be the local parametrisation of C at p , which is nothing but the maximal integral curve of \mathbf{X} through p . Let $p_1 \in C$. Parametrisation α is global if $p_1 \in \text{Image } \alpha$.

Step 2 : Construction of β

The plane curve C is connected. Thus there exists a continuous path β from p to p_1 where $\beta : [a, b] \rightarrow C$ with $\beta(a) = p$ and $\beta(b) = p_1$. We have, $\beta(a) \in \text{Image } \alpha$. But, $\beta(b) \notin \text{Image } \alpha$. Let $t_0 = \sup\{t \in [a, b] : \beta(t) \in \text{Image } \alpha\}$, the least upper bound of the points which are in $\text{Image } \alpha$. That is, if $t > t_0$, then $\beta(t) \notin \text{Image } \alpha$. And if $t < t_0$, then there exists an $\epsilon > 0$ such that $\beta(t + \epsilon) \in \text{Image } \alpha$. Otherwise, t_0 is not the supremum.

Step 3 : Construction of γ

Let γ be the maximal integral curve of \mathbf{X} through $\beta(t_0) = p_0$. Suppose there exists an open rectangle B about $\beta(t_0) = p_0$ such that $C \cap B \subset \text{Image } \gamma$. Then by the continuity of β , there exists $\delta > 0$ such that $\beta(t) \in \text{Image } \gamma$ for all $t \in (t_0 - \delta, t_0 + \delta)$. Thus, there exists $\epsilon > 0$ such that $\epsilon < \delta$ and $\beta(t_0 - \epsilon) \in \text{Image } \gamma$. Since t_0 is the least upper bound, $\beta(t_0 - \epsilon) \in \text{Image } \alpha$ as well. Thus, α and γ are maximal integral curves through a common point (in the

neighbourhood of p_0). And thus, $\text{Image } \alpha = \text{Image } \gamma$. Clearly, $p_1 \in \text{Image } \alpha$. Therefore, the parametrisation α is global. Therefore it is sufficient to prove that there exists an open box B such that $C \cap B \subset \text{Image } \gamma$.

Step 4 : Construction of A

Let $u = (\frac{\partial f}{\partial x_2}(p_0), -\frac{\partial f}{\partial x_1}(p_0))$ and $v = (\frac{\partial f}{\partial x_1}(p_0), \frac{\partial f}{\partial x_2}(p_0))$. Clearly, $\mathbf{u} = (p_0, u) \in C_{p_0}$ and $\mathbf{v} = (p_0, v) \perp C_{p_0}$. Consider the open rectangle A given by,

$$A = \{p_0 + ru + sv : |r| < \epsilon_1, |s| < \epsilon_2\}$$

where ϵ_1, ϵ_2 are so chosen that A is contained in the domain of f and $\nabla f(q) \cdot (q, v) > 0$ for every $q \in A$. Since $\nabla f(p_0) \cdot \mathbf{v} \neq 0$, by continuity of $\nabla f(q) \cdot (q, v)$ there exists a neighbourhood of p_0 in which this function is positive.

Step 5 : Construction of $\{g_r\}$

Consider the family of functions $\{g_r\}$ defined by $g_r(s) = f(p_0 + ru + sv)$. Since $\nabla f(q) \cdot (q, v) > 0$ for every $q \in A$, for $|r| < \epsilon_1$, $g_r(s)$ is strictly increasing in $(-\epsilon_2, \epsilon_2)$ as $g'_r(s) > 0$. That is, there exists at most one $s \in (-\epsilon_2, \epsilon_2)$ such that $f(p_0 + ru + sv) = c$.

Step 6 : Construction of B

Since the vectors \mathbf{u}, \mathbf{v} spans C_{p_0} , we have $\gamma(t) = p_0 + h_1(t)u + h_2(t)v$ where h_1, h_2 are real-valued functions. Clearly,

$$\begin{aligned} (\gamma(t) - p_0) \cdot \mathbf{u} &= h_1(t)\mathbf{u} \cdot \mathbf{u} + h_2(t)\mathbf{u} \cdot \mathbf{v} \\ &= h_1(t)\|\mathbf{u}\|^2 \text{ since } \mathbf{u} \perp \mathbf{v} \end{aligned}$$

$$\begin{aligned} (\gamma(t) - p_0) \cdot \mathbf{v} &= h_1(t)\mathbf{u} \cdot \mathbf{v} + h_2(t)\mathbf{v} \cdot \mathbf{v} \\ &= h_2(t)\|\mathbf{v}\|^2 \end{aligned}$$

Thus,

$$h_1(t) = \frac{(\gamma(t) - p_0) \cdot \mathbf{u}}{\|\mathbf{u}\|^2} \quad (18.25)$$

$$h_2(t) = \frac{(\gamma(t) - p_0) \cdot \mathbf{v}}{\|\mathbf{v}\|^2} \quad (18.26)$$

Now, $h'_1(0) = \dot{\gamma}(0) \cdot (p_0, u/\|u\|^2) = \mathbf{X}(p_0) \cdot \mathbf{X}(p_0)/\|\mathbf{X}(p_0)\|^2 = 1$. And $h_1(0) = 0$ and $h_2(0) = 0$. Thus, there exists (t_1, t_2) containing 0 such that $\gamma(t) \in A$ and $h'_1(t) > 0$ for every $t \in (t_1, t_2)$. Set $r_1 = h_1(t_1)$ and $r_2 = h_1(t_2)$. Since h_1 is continuous and strictly increasing, for every $r \in (r_1, r_2)$ there exists $t \in (t_1, t_2)$ such that $r = h_1(t)$ and $s = h_2(t)$.

$$\text{Define } B = \{p_0 + ru + sv : r \in (r_1, r_2), |s| < \epsilon_2\} \quad (18.27)$$

Then $p_0 + ru + sv \in B \cap C$ if and only if there exists $t \in (t_1, t_2)$ such that $r = h_1(t)$, and $s = h_2(t)$. That is, $p_0 + ru + sv \in \text{Image } \gamma$. Therefore, $C \cap B \subset \text{Image } \gamma$ as required in Step 3. \square

Theorem 18.11.2. *Let C be a connected oriented plane curve and let $\beta : I \rightarrow C$ be a unit speed global parametrisation of C . Then β is either one to one or periodic. Moreover, β is periodic if and only if C is compact.*

Proof. Suppose $\beta(t_1) = \beta(t_2)$ for some $t_1, t_2 \in I$ ($t_1 \neq t_2$). Let \mathbf{X} be the unit tangent vector field on C . And let α be the maximal integral curve of \mathbf{X} through $\beta(t_1)$. That is, $\alpha(0) = \beta(t_1) = \beta(t_2)$. Since β is a unit speed global parametrisation of C , β is also a maximal integral curve of \mathbf{X} . Thus, β is a reparametrisation of α . Thus, $\beta(t) = \alpha(t - t_1)$ and $\beta(t) = \alpha(t - t_2)$. Let $\tau = t_2 - t_1$. Then $\beta(t) = \alpha(t - t_1) = \alpha(t - t_2) = \alpha(t - t_2 + t_2 - t_1) = \alpha(t + \tau - t_1) = \beta(t + \tau)$. Therefore, β is periodic.

Suppose β is periodic, then its continuous image, $C = \beta[t, t + \tau]$ is compact. Suppose β is not periodic, then β is one to one. Then, C is not compact as β^{-1} doesn't attain its extrema. Thus, it is enough to prove that β^{-1} is continuous.

Given $t_0 \in I$ and $\epsilon > 0$, define $\gamma(t) = \beta(t + t_0)$ such that $|t| < \epsilon$ and $t + t_0 \in I$. Chose an open rectangle B about $p_0 = \beta(t_0)$ such that $C \cap B \subset \text{Image } \gamma$. Then, $|\beta^{-1}(p) - t_0| = |\gamma^{-1}(p)| < \epsilon$ for every $p \in C \cap B$.

Therefore it is enough to prove that such an open rectangle B exists. The construction of which is given in previous proof. \square

Fundamental Domain Let β be a periodic parametrised plane curve with period τ . Then any subset $[t_0, t_0 + \tau]$ of its domain is a fundamental domain of β .

18.11.2 1-Form

Definitions 18.11.1 (1-form). The 1-form is a function $\omega : U \times \mathbb{R}^{n+1} \rightarrow \mathbb{R}$ where $U \subset \mathbb{R}^{n+1}$ such that for every $p \in U$, $\mathbb{R}_p^{n+1} \subset U \times \mathbb{R}^{n+1}$.

Note : 1-form ω is smooth if ω is a smooth function. A few examples of 1-form are given below: $\omega_{\mathbf{X}}$, df and dx_i .

1-form dual Let \mathbf{X} be a vector field on U . Then 1-form dual of \mathbf{X} is $\omega_{\mathbf{X}}$ given by $\omega_{\mathbf{X}}(p, v) = \mathbf{X}(p) \cdot (p, v)$

differential of f Let $f : U \rightarrow \mathbb{R}$ be a smooth function. Then differential of f is df given by $df(\mathbf{v}) = \nabla f(p) \cdot \mathbf{v}$ where $\mathbf{v} = (p, v)$.

Cartesian Coordinate Function Let $U \subset \mathbb{R}^{n+1}$. Then i th cartesian coordinate function x_i is given by $x_i : U \rightarrow \mathbb{R}$, $x_i(a_1, a_2, \dots, a_{n+1}) = a_i$.

dx_i The differential of the cartesian coordinate function x_i is given by dx_i where $dx_i : U \times \mathbb{R}^{n+1} \rightarrow \mathbb{R}$

$$dx_i = \nabla x_i(p) \cdot \mathbf{v} = (p, 0, 0, \dots, 1, \dots, 0) \cdot (p, v_1, v_2, \dots, v_{n+1}) = v_i$$

18.11.3 1-form

Sum of two 1-forms $\omega_1 + \omega_2$

$$\begin{aligned} (\omega_1 + \omega_2)(\mathbf{v}) &= \nabla(\omega_1 + \omega_2) \cdot \mathbf{v} \\ &= \nabla\omega_1 \cdot \mathbf{v} + \nabla\omega_2 \cdot \mathbf{v} \\ &= \omega_1(\mathbf{v}) + \omega_2(\mathbf{v}) \end{aligned}$$

Product of function and 1-form $f\omega$

$$\begin{aligned}
f\omega(\mathbf{v}) &= \nabla(f\omega) \cdot \mathbf{v} \\
&= f\nabla\omega \cdot \mathbf{v} \\
&= f(p)\omega(\mathbf{v})
\end{aligned}$$

Note : Let $\omega(\mathbf{X}) : U \rightarrow \mathbb{R}$, $(\omega(\mathbf{X}))(p) = \omega(\mathbf{X}(p))$.
If ω and \mathbf{X} are smooth, then function $\omega(\mathbf{X})$ is also smooth.

Proposition 18.11.1 (1-form representation). *For every 1-form ω on U , there exists unique functions $f_i : U \rightarrow \mathbb{R}$ such that*

$$\omega = \sum_{i=1}^{n+1} f_i dx_i \quad (18.28)$$

And ω is smooth if and only if each f_i is smooth.

Proof. Let \mathbf{X}_j be a smooth vector field on U where $\mathbf{X}_j(p) = (p, 0, 0, \dots, 1, \dots, 0)$. Then, $dx_i(\mathbf{X}_j) = \delta_{ij}$. Suppose $\omega = \sum f_i dx_i$. Then

$$\omega(\mathbf{X}_j) = \sum_{i=1}^{n+1} (f_i dx_i) \mathbf{X}_j = \sum_{i=1}^{n+1} f_i(p) dx_i(\mathbf{X}_j) = \sum_{i=1}^{n+1} f_i(p) \delta_{ij} = f_j(p)$$

Thus f_j is unique for each j if they exist. And for each 1-form ω there exist functions f_i such that $\omega = \sum f_i dx_i$ exists where $f_i = \omega(\mathbf{X}_i)$. And thus, ω is smooth if and only if each f_i is smooth. \square

Corollary 18.11.2.1. *Let $f : U \rightarrow \mathbb{R}$. Then*

$$df = \sum_{i=1}^{n+1} \frac{\partial f}{\partial x_i} dx_i \quad (18.29)$$

Proof.

$$df(\mathbf{X}_j) = \nabla f \cdot \mathbf{X}_j = \sum_{i=1}^{n+1} \frac{\partial f}{\partial x_i} \delta_{ij} = \frac{\partial f}{\partial x_j} = f_j \implies df = \sum_{i=1}^{n+1} \frac{\partial f}{\partial x_i} dx_i$$

\square

18.11.4 Line Integral

Definitions 18.11.2 (Line Integral). Let ω be a 1-form and $\alpha : [a, b] \rightarrow U$ be a parametrised curve on U . Then the line integral of ω over α is given by,

$$\int_{\alpha} \omega = \int_a^b \omega(\dot{\alpha}(t)) \quad (18.30)$$

Note : The line integral of 1-form ω over a parametrised curve is invariant of reparametrisation.

Let $\beta : [c, d] \rightarrow U$ be reparametrisation of $\alpha : [a, b] \rightarrow U$ where $\beta(t) = \alpha(h(t))$ and function $h : [c, d] \rightarrow [a, b]$. Then ,

$$\begin{aligned} \int_{\beta} \omega &= \int_c^d \omega(\dot{\beta}(t)) \\ &= \int_c^d \omega(\dot{\alpha}(h(t))h'(t)) \\ &= \int_c^d \omega(\dot{\alpha}(h(t)))h'(t) \\ &= \int_a^b \omega(\dot{\alpha}(u)) = \int_{\alpha} \omega \end{aligned}$$

Theorem 18.11.3. Let η be a 1-form on $\mathbb{R}^2 - \{0\}$ defined by

$$\eta = \frac{-x_2}{x_1^2 + x_2^2} dx_1 + \frac{x_1}{x_1^2 + x_2^2} dx_2$$

Then for any closed, piecewise smooth, parametrised smooth curve $\alpha : [a, b] \rightarrow \mathbb{R} - \{0\}$, the line integral of η over α

$$\int_{\alpha} \eta = 2\pi k$$

Proof. Consider $\varphi : [a, b] \rightarrow \mathbb{R}$ defined by $\varphi(t) = \varphi(a) + \int_{\alpha_t} \eta$ where α_t is the restriction of α to $[a, t]$ and $\varphi(a)$ is so chosen that $\alpha(a)/\|\alpha(a)\| = (\cos \varphi(a), \sin \varphi(a))$. We claim that

$$\frac{\alpha(t)}{\|\alpha(t)\|} = (\cos \varphi(t), \sin \varphi(t)), \quad \forall t \in [a, b] \quad (18.31)$$

Let $t_0 = \sup\{t \in [a, b] : \alpha(t)/\|\alpha(t)\| = (\cos \varphi(t), \sin \varphi(t))\}$. By continuity the claim should be true for t_0 as well. Therefore it is enough to prove that $t_0 = b$.

Define $v = -\alpha(t_0)/\|\alpha(t_0)\|$ and $V = \mathbb{R}^2 - \{rv : r \geq 0\}$. And $\theta_V : V \rightarrow \mathbb{R}$ is defined in such a way that $v = (\cos \theta_v, \sin \theta_v)$ and $\theta_v \in [0, 2\pi)$. In other words, θ_V a real-valued function which give the angle of the ray from origin through the points.

Then $(\cos \theta_V(\alpha(t_0)), \sin \theta_V(\alpha(t_0))) = \alpha(t_0)/\|\alpha(t_0)\| = (\cos \varphi(t_0), \sin \varphi(t_0))$. Clearly, $\varphi(t_0) - \theta_V(\alpha(t_0)) = 2\pi m$ for some integer m .

Consider a δ neighbourhood of t_0 such that $\alpha(t) \in V$ for every point t in it. Then, α is smooth in that neighbourhood.

$$\frac{d}{dt} (\varphi(t) - \theta_V(\alpha(t))) = \eta(\dot{\alpha}(t)) = d\theta_V(\dot{\alpha}(t)) = 0$$

Thus, $\varphi(t) - \theta_V(\alpha(t)) = 2\pi m$ for every $t \in (t_0 - \delta, t_0 + \delta)$. Therefore, the claim is true. Suppose $t_0 < b$, then the claim is true for $t_0 + \epsilon \in (t_0 - \delta, t_0 + \delta)$ which is a contradiction to the choice of t_0 as least upper bound of all such values.

Thus, $(\cos \varphi(a), \sin \varphi(a)) = (\cos \varphi(b), \sin \varphi(b))$. Therefore, $\varphi(b) - \varphi(a) = 2\pi k$ for some integer k . And $\int_a^b \eta = \int_a^b \eta(\dot{\alpha}(t)) dt = \varphi(b) - \varphi(a) = 2\pi k$. \square

Winding Number The winding number of α is $k(\alpha) = \frac{1}{2\pi} \int_a^b \eta$

Note : Winding number is the number of times α winds around origin.

18.12 Curvature of Surfaces

Normal Curvature The number $\kappa(\mathbf{v}) = L_p(\mathbf{v}) \cdot \mathbf{v}$ is the normal curvature of the n -surface S at point p in the direction \mathbf{v} .

If $\kappa(\mathbf{v}) > 0$, then S bends toward orientation at p . And if $\kappa(\mathbf{v}) < 0$, then S bends away from orientation at p .

We have, $\kappa : S_p \rightarrow \mathbb{R}$. When $n = 1$, then $\kappa(\mathbf{v}) = \kappa(p)$. Clearly S_p is compact and κ attains its extrema, which are the eigen values of the Weingarten Map.

Normal Section The normal section of an n -surface S with orientation N is $\mathcal{N}(\mathbf{v}) \subset \mathbb{R}^{n+1}$ given by

$$\mathcal{N}(\mathbf{v}) = \{q \in \mathbb{R}^{n+1} : q = p + xv + yN(p), (x, y) \in \mathbb{R}^2\} \quad (18.32)$$

You take an apple and place a knife at any point on that apple in any direction tangent to its surface and cut it. The apple cuts into two parts. The 2-dimensional space of that cut, is a **Normal Section** of that apple. For any $n \geq 2$, the normal section of an n -surface is always 2 dimensional.

$S \cap \mathcal{N}(\mathbf{v})$ is the intersection of both the surface and its normal section.

In above example, $S \cap \mathcal{N}(\mathbf{v})$ is plane curve traced by the skin around cross-section of that apple.

Theorem 18.12.1 (Components of $S \cap \mathcal{N}(\mathbf{v})$). *Let S be an oriented n -surface and \mathbf{v} is a tangent direction at p , $\mathbf{v} \in S_p$. Then there exists an open V containing p such that $S \cap \mathcal{N}(\mathbf{v}) \cap V$ is a plane curve. Moreover, the curvature at p of this curve is the normal curvature $\kappa(\mathbf{v})$ of S at p in the direction \mathbf{v} .*

Proof. Let $f : U \rightarrow \mathbb{R}^n$ such that $S = f^{-1}(c)$ and $\nabla f(q) \neq \mathbf{0}$, $\forall q \in S$. Let $\widetilde{\nabla} f(q) = \begin{pmatrix} q, \widetilde{\nabla} f(q) \end{pmatrix}$. Let $p \in S$ and $\mathbf{v} \in S_p$. Let $i : \mathbb{R}^2 \rightarrow \mathbb{R}^{n+1}$ defined by $i(x, y) = p + xv + yN(p)$. Clearly, the image of the function i is the normal section, $\mathcal{N}(\mathbf{v}) = i(\mathbb{R}^2)$.

Let V be the space spanned by v and $N(p)$. That is,

$$V = \left\{ q \in U : \widetilde{\nabla} f(q) \cdot v \neq 0 \text{ OR } \widetilde{\nabla} f(q) \cdot N(p) \neq 0 \right\} \quad (18.33)$$

Then $\nabla(f \circ i)(x, y) = \begin{pmatrix} x, y, \widetilde{\nabla} f(i(x, y)) \cdot v, \widetilde{\nabla} f(i(x, y)) \cdot N(p) \end{pmatrix} \neq \mathbf{0}$ for every $(x, y) \in i^{-1}(V)$. Therefore, $C = i^{-1}(S \cap \mathcal{N}(\mathbf{v})) = (f \circ i)^{-1}(c) \cap i^{-1}(V)$ is a plane curve. In other words, $S \cap \mathcal{N}(\mathbf{v}) \cap V$ is a plane curve. \square

Lemma 18.12.2 (Extrema of Curvature are Eigenvalue of Weingarten Map). *Let V be a finite dimensional vector space with dot product and let $L : V \rightarrow V$ be a self-adjoint linear transformation on V . Let $S = \{v \in V : v \cdot v = 1\}$ and define $f : S \rightarrow \mathbb{R}$ by $f(v) = L(v) \cdot v$. Suppose f is stationary at $v_0 \in S$. Then $L(v_0) = f(v_0)v_0$.*

Proof. Since f is stationary at v_0 , $f \circ \alpha'(0) = 0$ for every parametrised curves in S with $\alpha(0) = v_0$. Let $v \perp v_0$. Then $\alpha(t) = (\cos t)v_0 + (\sin t)v$. Then,

$$\begin{aligned} 0 &= (f \circ \alpha)'(0) \\ &= \frac{d}{dt} (L(\alpha(t)) \cdot \alpha(t)) (0) \\ &= \frac{d}{dt} (\cos^2 t L(v_0) \cdot v_0 + 2 \sin t \cos t L(v_0) \cdot v + \sin^2 t L(v) \cdot v) \\ &= (-\sin 2t L(v_0) \cdot v_0 + 2 \cos 2t L(v_0) \cdot v + \sin 2t L(v) \cdot v)|_{t=0} \\ &= 2L(v_0) \cdot v \end{aligned}$$

Clearly, $L(v_0) \perp v$ for every $v \in v_0^\perp$. In other words, $L(v_0) \perp v_0^\perp$. Thus, v_0 can span the space $L(v_0)$. That is, $L(v_0) = \lambda v_0$ for some $\lambda \in \mathbb{R}$. Thus v_0 is an eigen vector of $L(v_0)$. And the corresponding eigen vector λ can be obtained, $\lambda = \lambda v_0 \cdot v_0 = L(v_0) \cdot v_0 = f(v_0)$. \square

Theorem 18.12.3. *Let V be a finite dimensional vector space with dot product and let $L : V \rightarrow V$ be a self-adjoint linear transformation on V . Then there exists an orthonormal basis for V consisting of eigen vectors of L .*

Proof. The proof is by mathematical induction. For $n = 1$, the theorem is true by lemma. Suppose that the theorem is true for $n = k$. It is enough to prove that theorem is then true for $n = k + 1$.

By lemma, there exists a unit vector $v_1 \in V$ which is an eigen vector of L . This is done by selecting v_1 such that $L(v_1) \cdot v_1 \geq L(v) \cdot v$ for every unit vector $v \in V$. Define $W = v_1^\perp$. Then $L(w) \cdot v_1 = w \cdot L(v_1) = \lambda w \cdot v_1 = 0$ for every $w \in W$. (Remember : L self-adjoint, $L(w) \cdot v = w \cdot L(v)$)

Thus the restriction of L , $L|_W : W \rightarrow W$ is surjective. And clearly self-adjoint. The dimension of W is k , thus by induction hypothesis there exists an orthonormal basis $\{v_2, v_3, \dots, v_{k+1}\}$ for W with eigen vectors of $L|_W$. Clearly, these are eigen vectors of L and $\{v_1, v_2, \dots, v_{k+1}\}$ is an orthonormal basis of V . \square

Principal Curvature Directions are the vectors in the orthonormal basis $\{v_1, v_2, \dots, v_n\}$ for S_p obtained as eigen vectors of the Weingarten Map. (Hint : \mathbb{R}^n is a vector space with dot product and L_p is self-adjoint, linear transformation)

Principal Curvature are the eigen values of the Weingarten Map. That is, $k_i(p)$ is the eigen value of $L_p : S_p \rightarrow S_p$ such that $L_p(w) \cdot v_i = \kappa_i(p)w \cdot v_i$.

Theorem 18.12.4. *Let S be an oriented n -surface in \mathbb{R}^{n+1} . Let $p \in S$ and $\{\kappa_1(p), \kappa_2(p), \dots, \kappa_n(p)\}$ be the principal curvatures of S at p with corresponding orthogonal principal curvature directions $\{v_1, v_2, \dots, v_n\}$. Then the normal*

curvature $\kappa(\mathbf{v})$ in the direction $\mathbf{v} \in S_p$ is given by

$$\kappa(v) = \sum_{i=1}^n \kappa_i(p) (\mathbf{v} \cdot \mathbf{v}_i)^2 = \sum_{i=1}^n \kappa_i(p) \cos^2 \theta_i \quad (18.34)$$

where θ_i is the angle between \mathbf{v} and \mathbf{v}_i .

Proof. We have $v \in V$ and $\{\mathbf{v}_1, \mathbf{v}_2, \dots, \mathbf{v}_n\}$ is an orthonormal basis for V . Then,

$$\mathbf{v} = \sum_{i=1}^n (\mathbf{v} \cdot \mathbf{v}_i) \mathbf{v}_i = \sum_{i=1}^n \cos \theta_i \mathbf{v}_i$$

We have,

$$\begin{aligned} \kappa(\mathbf{v}) &= L_p(\mathbf{v}) \cdot \mathbf{v} \\ &= \sum_{i=1}^n \cos \theta_i L_p(\mathbf{v}_i) \cdot \mathbf{v} \\ &= \sum_{i=1}^n \kappa_i(\mathbf{v}_i) \mathbf{v}_i \cdot \mathbf{v} \\ &= \sum_{i=1}^n \kappa_i \cos^2 \theta_i \end{aligned}$$

□

Direction Cosines Let $\{\mathbf{v}_1, \mathbf{v}_2, \dots, \mathbf{v}_n\}$ be an orthonormal basis for S_p . Then θ_i are the angles \mathbf{v} makes with each vector in that basis. And direction cosines are the components of \mathbf{v} in those directions. That is, $\cos \theta_i = \mathbf{v} \cdot \mathbf{v}_i$.

Quadratic Form Let V be a finite dimensional vector space with dot product and $L : V \rightarrow V$ be a self-adjoint, linear transformation on V . Then quadratic form associated with L is a $\mathcal{L} : V \rightarrow \mathbb{R}$ defined by $\mathcal{L}(v) = L(v) \cdot v$.

First Fundamental Form, \mathcal{I}_p of surface S is the quadratic form associated with the identity transformation on S_p .

$$\mathcal{I}_p(v) = id(v) \cdot v = v \cdot v = \|v\|^2$$

Second Fundamental Form \mathcal{S}_p of surface S is the quadratic form associated with the Weingarten Map.

$$\mathcal{S}_p(v) = L_p(v) \cdot v$$

A quadratic forms is

positive definite if $\mathcal{L}(v) > 0, \forall v \neq 0$.

negative definite if $\mathcal{L}(v) < 0, \forall v \neq 0$.

definite if either positive definite or negative definite.

indefinite if not definite.

positive semi-definite if $\mathcal{L}(v) \geq 0$, $\forall v \neq 0$.

negative semi-definite if $\mathcal{L}(v) \leq 0$, $\forall v \neq 0$.

semi-definite if either positive semi-definite or negative semi-definite.

For example, first fundamental form of any surface is positive definite.

Theorem 18.12.5. *On each compact, oriented n -surface S in \mathbb{R}^{n+1} there exists a point p such that the second fundamental form at p is definite.*

Proof. Let S^n be an n -sphere containing S and S^n touches S at p . Then \mathcal{S}_p is definite.

Define $g : \mathbb{R}^n \rightarrow \mathbb{R}$ by $g(x_1, x_2, \dots, x_n) = x_1^2 + x_2^2 + \dots + x_n^2$. Since S is compact, g attains maximum at $p \in S$. By Lagrange multiplier theorem, $\nabla g(p) = \lambda \nabla f(p) = \mu N(p)$ where $\lambda, \mu \in \mathbb{R}$ and $\mu = \pm \lambda \|\nabla f(p)\|$. WLOG Suppose that $\mu < 0$. Then $\mu = -|\mu| = -\|\mu N(p)\| = -\|\nabla g(p)\| = -2\|p\|$. Thus, $N(p) = \frac{1}{\mu} \nabla g(p) = \frac{-1}{\|p\|}(p, p)$.

Let $\mathbf{v} \in S_p$ such that $\|\mathbf{v}\| = 1$. Let $\alpha : I \rightarrow S$ defined by $\dot{\alpha}(t_0) = \mathbf{v}$. Then $g \circ \alpha(t_0) \geq g \circ \alpha(t)$, $\forall t \in I$.

$$\begin{aligned} 0 &\geq \frac{d^2}{dt^2} (g \circ \alpha)(t_0) \\ &= \frac{d}{dt} (\nabla g(\alpha(t)) \cdot \dot{\alpha}(t))(t_0) \\ &= \frac{d}{dt} \left(2\alpha(t) \cdot \frac{d\alpha}{dt}(t) \right) (t_0) \\ &= 2\dot{\alpha}(t_0) \cdot \dot{\alpha}(t_0) + 2\alpha(t_0) \cdot \ddot{\alpha}(t_0) \\ &= 2(\|\dot{\alpha}(t_0)\|^2 + (\alpha(t_0), \alpha(t_0)) \cdot \ddot{\alpha}(t_0)) \\ &= 2(\|\mathbf{v}\|^2 + (p, p) \cdot \ddot{\alpha}(t_0)) \\ &= 2(1 - \|p\| \mathbf{N}(p) \cdot \ddot{\alpha}(t_0)) \\ &= 2(1 - \|p\| \kappa(\mathbf{v})) \end{aligned}$$

Thus, $0 \geq 2 - 2\|p\| \kappa(\mathbf{v}) \implies 2\|p\| \kappa(\mathbf{v}) \geq 2 \implies \kappa(\mathbf{v}) \geq \frac{1}{\|p\|}$, $\forall \mathbf{v} \in S_p$. If S is oriented so that $\mu > 0$, then $\kappa(\mathbf{v}) \leq -\frac{1}{\|p\|}$. \square

Gauss Kronecker Curvature, $K(p)$ of surface S in \mathbb{R}^{n+1} at a point p is $K(p) = \det L_p$. It is equal to the product of the principal curvatures at p .

$$K(p) = k_1(p)k_2(p) \cdots k_n(p) \quad (18.35)$$

Mean Curvature $H(p)$ of surface S in \mathbb{R}^{n+1} at a point p is the average value of principal curvatures at p .

$$H(p) = \frac{1}{n}(\kappa_1(p) + \kappa_2(p) + \dots + \kappa_n(p)) \quad (18.36)$$

Theorem 18.12.6. *Let S be an n -surface in \mathbb{R}^{n+1} . Let $p \in S$. Let \mathbf{Z} be any non-zero normal vector field on S such that $\mathbf{N} = \mathbf{Z}/\|\mathbf{Z}\|$ and let $\{\mathbf{v}_1, \mathbf{v}_2, \dots, \mathbf{v}_n\}$ be any basis for S_p . Then*

$$K(p) = \frac{(-1)^n \det \begin{pmatrix} \nabla_{v_1} \mathbf{Z} \\ \nabla_{v_2} \mathbf{Z} \\ \vdots \\ \nabla_{v_n} \mathbf{Z} \\ \mathbf{Z}(p) \end{pmatrix}}{\|\mathbf{Z}(p)\|^n \det \begin{pmatrix} \mathbf{v}_1 \\ \mathbf{v}_2 \\ \vdots \\ \mathbf{v}_n \\ \mathbf{Z}(p) \end{pmatrix}} \quad (18.37)$$

Proof. $\mathbf{Z} = \|\mathbf{Z}\|\mathbf{N}$.

$$\begin{aligned} \det \begin{pmatrix} \nabla_{v_1} \mathbf{Z} \\ \nabla_{v_2} \mathbf{Z} \\ \vdots \\ \nabla_{v_n} \mathbf{Z} \\ \mathbf{Z}(p) \end{pmatrix} &= \det \begin{pmatrix} (\nabla_{v_1} \|\mathbf{Z}\|) \mathbf{N}(p) + \|\mathbf{Z}(p)\| \nabla_{v_1} \mathbf{N} \\ (\nabla_{v_2} \|\mathbf{Z}\|) \mathbf{N}(p) + \|\mathbf{Z}(p)\| \nabla_{v_2} \mathbf{N} \\ \vdots \\ (\nabla_{v_n} \|\mathbf{Z}\|) \mathbf{N}(p) + \|\mathbf{Z}(p)\| \nabla_{v_n} \mathbf{N} \\ \|\mathbf{Z}(p)\| \mathbf{N}(p) \end{pmatrix} \\ &= \|\mathbf{Z}(p)\|^n \det \begin{pmatrix} \nabla_{v_1} \mathbf{N} \\ \nabla_{v_2} \mathbf{N} \\ \vdots \\ \nabla_{v_n} \mathbf{N} \\ \mathbf{N}(p) \end{pmatrix} \\ &= (-1)^n \|\mathbf{Z}(p)\|^n \det \begin{pmatrix} L_p(\mathbf{v}_1) \\ L_p(\mathbf{v}_2) \\ \vdots \\ L_p(\mathbf{v}_n) \\ \mathbf{Z}(p) \end{pmatrix} \\ &= (-1)^n \|\mathbf{Z}(p)\|^n \det \begin{pmatrix} \cdots & 0 \\ A^t & \vdots \\ 0 & \cdots & 0 & 1 \end{pmatrix} \begin{pmatrix} \mathbf{v}_1 \\ \mathbf{v}_2 \\ \vdots \\ \mathbf{v}_n \\ \mathbf{Z}(p) \end{pmatrix} \\ &= (-1)^n \|\mathbf{Z}(p)\|^n \det A \det \begin{pmatrix} \mathbf{v}_1 \\ \mathbf{v}_2 \\ \vdots \\ \mathbf{v}_n \\ \mathbf{Z}(p) \end{pmatrix} \end{aligned}$$

$$= (-1)^n \|Z(p)\|^n K(p) \det \begin{pmatrix} \mathbf{v}_1 \\ \mathbf{v}_2 \\ \vdots \\ \mathbf{v}_n \\ Z(p) \end{pmatrix}$$

where A is the matrix for L_p with respect to the orthonormal basis $\{\mathbf{v}_1, \mathbf{v}_2, \dots, \mathbf{v}_n\}$ for S_p and A^t is its transpose. \square

local property is a property of S which is valid in the neighbourhood of a particular point.

global property is a property of S which is valid everywhere on S .

local theorem is a theorem on local behaviour of S .

global theorem is a theorem on global behaviour of S .

Theorem 18.12.7 (Global Characterisation of Surfaces with Definite Second Fundamental Form). *Let S be a compact, connected, oriented n -surface in \mathbb{R}^{n+1} . Then $\forall p \in S, K(p) \neq 0 \iff \forall p \in S, \mathcal{S}_p$ is definite.*

Proof. If \mathcal{S}_p is definite then normal curvature $\kappa(\mathbf{v}) = \mathcal{S}_p(\mathbf{v})$ is nonzero in every direction $\mathbf{v} \in S_p$. Thus, all principal curvatures are nonzero and their product Gauss Kronecker Curvatures $K(p)$ is also nonzero.

Let $\kappa_1 \leq \kappa_2 \leq \dots \leq \kappa_n$. Let \mathcal{S}_{p_0} be definite. Since every compact, connected, oriented n -surface S has a point p_0 at which its second fundamental form \mathcal{S}_{p_0} is definite. Suppose \mathcal{S}_{p_0} is positive definite. Then minimal principal curvature κ_1 is positive at p_0 . Clearly, κ_i is nowhere zero and continuous. Thus, every principal curvature κ_i are positive everywhere on S . Therefore, their product $K(p)$ is positive everywhere on S . That is, \mathcal{S}_p is positive definite.

Suppose \mathcal{S}_{p_0} is negative definite. Then the maximal principal curvature $\kappa_n(p)$ is negative at p_0 . And κ_n is nonzero everywhere and continuous. Thus, κ_n is negative definite. Thus every principal curvature is negative definite. Thus Gauss Kronecker Curvature $K(p)$ is either everywhere negative or positive depending on the parity on n . Therefore, $K(p)$ is definite. \square

18.14 Parameterized Surfaces

tangent bundle is the set $T(U) = U \times \mathbb{R}^{n+1}$ where U is an open subset of \mathbb{R}^{n+1} such that $\forall p \in U, \mathbb{R}_p^{n+1} \subset U \times \mathbb{R}^{n+1}$.

Definitions 18.14.1 (Differential of parametrisations). Let $\varphi : U \rightarrow \mathbb{R}^{n+1}$ be parametrisation of a surface. Then its differential $d\varphi : T(U) \rightarrow T(\mathbb{R}^{n+1})$ is given by

$$d\varphi(\mathbf{v}) = (\varphi \circ \alpha)(t_0) \quad (18.38)$$

Note : $d\varphi$ is independent of the choice of α .

Definitions 18.14.2 (Differential of surface parametrisation). Let S be an n -surface in \mathbb{R}^{n+1} . Then $d\varphi_p$ is the restriction of $d\varphi$ to S_p .

There are two different meanings for $d\varphi$ depending on its domain.

1. $d\varphi : I \rightarrow \mathbb{R}$
2. $d\varphi : I \times \mathbb{R} \rightarrow \mathbb{R}^2$

$$d\varphi(t, u) = (\varphi(t), d\varphi(t, u)) \quad (18.39)$$

Author could have used dark font for this second version, $d\varphi$. But he chose not to.

18.14.1 Parametrized Surfaces

parametrised n -surface is a smooth map $\varphi : U \rightarrow \mathbb{R}^n$ where U is a connected and open subset of \mathbb{R}^n .

regular surface A parametrised n -surface is regular if $d\varphi$ is non-singular.

parametrised 2-sphere $\varphi : U \rightarrow \mathbb{R}^3$ defined by $\varphi(\theta, \phi) = (r \cos \theta \sin \phi, r \sin \theta \sin \phi, r \cos \phi)$

parametrised n -plane Let $L : \mathbb{R}^n \rightarrow \mathbb{R}^{n+k}$ be non-singular linear map. Then $\varphi : U \rightarrow \mathbb{R}^{n+k}$ defined by $\varphi(p) = L(p) + w$ is a parametrised n -plane through w in \mathbb{R}^{n+k} .

cylinder over n -surface φ Let $\varphi : U \rightarrow \mathbb{R}^{n+k}$ be a parametrised n -surface in \mathbb{R}^{n+k} where U is an open subset of \mathbb{R}^n . The cylinder over φ is an $(n+1)$ -surface. $\tilde{\varphi} : U \times \mathbb{R} \rightarrow \mathbb{R}^{n+k+1}$ defined by

$$\tilde{\varphi}(u_1, u_2, \dots, u_{n+1}) = (\varphi(u_1, u_2, \dots, u_n), u_{n+1}) \quad (18.40)$$

Surface of revolution Let $\alpha : I \rightarrow \mathbb{R}^2$ be a parametrised curve in \mathbb{R}^2 whose image lies above x_1 axis. The surface obtained by revolving $\alpha(t) = (x_1(t), x_2(t))$ around x_1 axis is an 2-surface in \mathbb{R}^3 .

$$\varphi : I \times \mathbb{R} \rightarrow \mathbb{R}^3 \text{ defined by } \varphi(t, \theta) = (x_1(t), x_2(t) \cos \theta, x_2(t) \sin \theta) \quad (18.41)$$

Torus in \mathbb{R}^3 Consider the circle $\alpha(\phi) = (a + b \cos \phi, b \sin \phi)$ with center at $(a, 0)$ with radius b ($b > a$). Then the surface of revolution of α about x_2 axis,

$$\varphi(\phi, \theta) = (a + b \cos \phi \cos \theta, a + b \cos \phi \sin \theta, b \sin \phi) \quad (18.42)$$

Torus in \mathbb{R}^4 is a parametrised 2-surface $\varphi : \mathbb{R}^2 \rightarrow \mathbb{R}^4$ defined by

$$\varphi(\theta, \phi) = \{(\cos \theta, \sin \theta, \cos \phi, \sin \phi) : \theta, \phi \in \mathbb{R}\} \quad (18.43)$$

Möbius Band is a parametrised 2-surface $\varphi : I \times \mathbb{R} \rightarrow \mathbb{R}^3$ defined by

$$\varphi(t, \theta) = \left(\left(1 + t \cos \frac{\theta}{2} \right) \cos \theta, \left(1 + t \cos \frac{\theta}{2} \right) \sin \theta, t \sin \frac{\theta}{2} \right) \quad (18.44)$$

18.14.2 Vector Field along φ

Definitions 18.14.3. Let $\varphi : U \rightarrow \mathbb{R}^{n+k}$ be a smooth function. A vector field along φ is a map \mathbf{X} which assigns a vector $\mathbf{X}(p) \in \mathbb{R}_{\varphi(p)}^{n+k}$ to each point $p \in U$.

Definitions 18.14.4. A vector field along φ is a tangent vector field along φ if there exists a vector field \mathbf{Y} on U such that $\mathbf{X}(p) = d\varphi_p(\mathbf{Y}(p))$.

Definitions 18.14.5. Let $\varphi : U \rightarrow \mathbb{R}^{n+k}$ be a smooth map where U is an open subset of \mathbb{R}^n . The coordinate vector fields along φ are the tangent fields \mathbf{E}_i defined by

$$\mathbf{E}_i(p) = d\varphi_p(p, 0, \dots, 1, \dots, 0) \quad (18.45)$$

Theorem 18.14.1 (Basis for Tangent Space Image $d\varphi_p$). *Let $\varphi : U \rightarrow \mathbb{R}^{n+k}$ be smooth. Then the set of all coordinate vector fields $\{\mathbf{E}_i(p) : i = 1, 2, \dots, n\}$ form a basis for the tangent space image $d\varphi_p$.*

Definitions 18.14.6. Let $\varphi : U \rightarrow \mathbb{R}^{n+k}$ be smooth. And \mathbf{X} be a smooth vector field along φ . Let $p \in U$ and $\mathbf{v} \in \mathbf{R}_p^n$. The derivative $\nabla_{\mathbf{v}}\mathbf{X} \in \mathbb{R}_{\varphi(p)}^{n+k}$ defined by

$$\nabla_{\mathbf{v}}\mathbf{X} = \left(\varphi(p), \left(\frac{d}{dt} X \circ \alpha \right) (t_0) \right) \quad (18.46)$$

Definitions 18.14.7 (orientation vector field). Let $\varphi : U \rightarrow \mathbb{R}^{n+1}$ be smooth parametrised n -surface in \mathbb{R}^{n+1} . Let $p \in U$. An orientation vector field along φ is the unique vector field \mathbf{N} along φ such that $\mathbf{N}(p)$ are unit vectors and $\mathbf{N}(p) \perp d\varphi_p(\mathbf{v})$. And \mathbf{N} is consistently oriented if the following determinant is positive.

$$\det \begin{pmatrix} \mathbf{E}_1(p) \\ \mathbf{E}_2(p) \\ \vdots \\ \mathbf{E}_n(p) \\ \mathbf{N}(p) \end{pmatrix} > 0 \quad (18.47)$$

18.14.3 Weingarten Map

Definitions 18.14.8 (Weingarten map). The Weingarten map is the linear function $L_p : \text{Image } d\varphi_p \rightarrow \text{Image } d\varphi_p$ defined by $L_p(d\varphi(\mathbf{v})) = -\nabla_{\mathbf{v}}\mathbf{N}$.

L_p is self-adjoint $L_p(\mathbf{v}) \cdot \mathbf{w} = \mathbf{v} \cdot L_p(\mathbf{w})$

principal curvatures are the eigen values of L_p

principal curvature directions are the unit eigen vectors of L_p

Gauss-Kronecker Curvature is the determinant.

Mean Curvature is $1/n$ times its trace.

Subject 19

ME800402 Algorithmic Graph Theory

19.1 An Introduction to Graphs

19.1.1 What is a Graph ?

Definitions 19.1.1 (graph). A graph G is a finite nonempty set $V(G)$ and a family $E(G)$ of 2-element subsets of $V(G)$.

Remark. Elements of $V(G)$ are called **vertices** and elements of $E(G)$ are called **edges** of the graph G . Suppose $\{u, v\}$ be an edge of a graph G . We may represent this edge by uv .

Two vertices u, v are **adjacent** if there exists an edge $e = uv$ joining them.

Definitions 19.1.2 (multigraph). A graph with parallel edges is called a multigraph.

Definitions 19.1.3 (psuedograph). A graph with parallel egdes and loops is called a psuedograph.

Definitions 19.1.4. The number of vertices in graph G is the order of the graph G . The number of edge in graph G is the size of the graph G .

Remark. A graph G of order p and size q is called a (p, q) graph.

19.1.2 The Degree of a Vertex

Definitions 19.1.5 (neighbourhood). Let v be a vertex of graph G . The neighbourhood of v , $N(v)$ is the set of all vertices adjacent to it.

$$N(v) = \{u \in V(G) : uv \in E(G)\}$$

Remark. The **degree** of a vertex v is the cardinality of the neighbourhood of v in graph G .

$$\deg v = |N(v)|$$

An edge of zero degree is called an **isolated vertex**. An edge of odd(even) degree is called an **odd(even) edge**.

Remark. Let $e = uv$ be an edge of a graph G . Then e is **incident** on both u and v .

Theorem 19.1.1 (First Theorem). *Let G be a graph of order p and size q with vertex set $V(G) = \{v_1, v_2, \dots, v_p\}$. Then,*

$$\sum_{i=1}^p \deg v_i = 2q$$

Proof. Since degrees of vertices are counted, every edge is counted twice (from both ends). Thus the sum of degrees is twice the size of the graph. \square

Corollary 19.1.1.1. *Every graph has even number of odd vertices.*

Proof. By first theorem of graph theory, the sum of degree is an even integer. Suppose there are odd number of odd vertices. Then the sum of degrees is odd which is a contradiction. Therefore, there are even number of odd vertices in any graph G . \square

Definitions 19.1.6 (regular). A graph G is r -regular if every vertex of G has degree r .

Definitions 19.1.7 (complement). Let G be a graph. The complement \bar{G} is the graph with vertex set $V(\bar{G}) = V(G)$ and uv is an edge of \bar{G} if uv is not an edge of G .

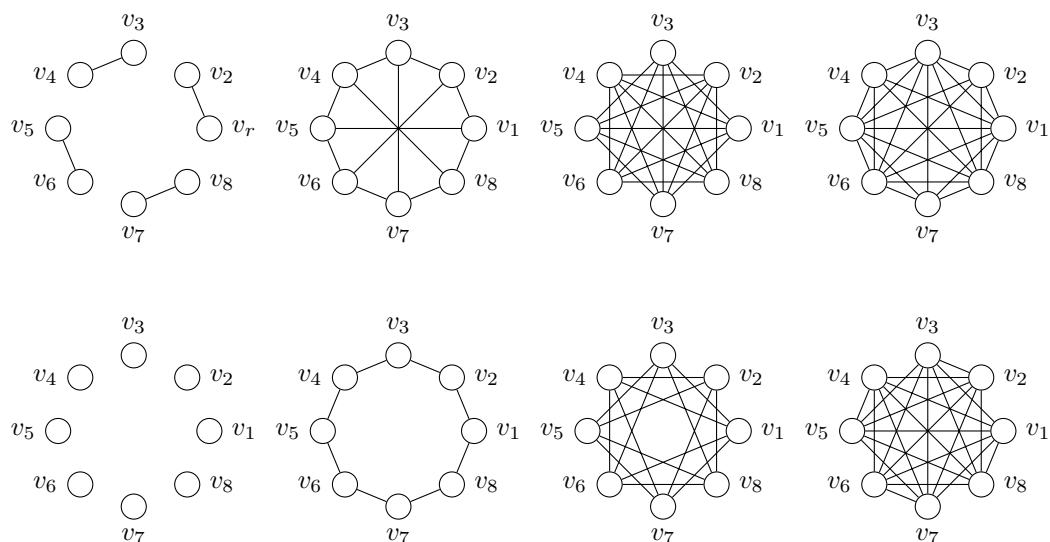
Problem Set 1.2

1. Prove that every graph of order $n \geq 2$ has at least two vertices of the same degree.

Proof. Let G be a graph of order p . Suppose every vertex is of different degree. Then there exists a vertex of degree 0 and a vertex of degree $p - 1$. This is a contradiction as the vertex of degree 0 is not adjacent to any other vertex and the vertex of degree $p - 1$ is adjacent to every other vertex. \square

7. (a) Construct an r -regular graph of order 8 for each r , $0 \leq r < 8$.
(See Figure 19.1)
- (b) Determine the complement of each graph constructed in (a).
(complements are already there in the list)
- (c) Prove that if G is a regular graph. Then \bar{G} is regular.

Proof. Let G be an r -regular graph of order p . Let $v \in V(G)$. Then $\deg v = r$. Thus, there are r vertices adjacent to v in graph G . In other words, there are $p - r - 1$ vertices non-adjacent to v in G . Therefore, $\deg v$ in the complement graph \bar{G} is $p - r - 1$. Since $v \in V(G)$ is arbitrary, \bar{G} is $p - r - 1$ -regular graph. \square

Figure 19.1: r -regular graph of order 8

19.1.3 Isomorphic Graphs

Definitions 19.1.8 (isomorphism). Two graphs G_1, G_2 are isomorphic if there exists a bijection

$$\phi : V(G_1) \rightarrow V(G_2) \text{ such that } uv \in E(G_1) \iff \phi(u)\phi(v) \in E(G_2)$$

Remark. Two graphs G_1, G_2 are equal if $V(G_1) = V(G_2)$ and $E(G_1) = E(G_2)$. Clearly, equal graphs are isomorphic. However, isomorphic graphs are not equal.

Problem Set 1.3

- Find two non-isomorphic 3-regular graphs of order 6 and size 9.
There are exactly two non-isomorphic 3-regular graphs of order 6. By first theorem of graph theory, the size of such a graph is always 9.(see Figure 19.2).

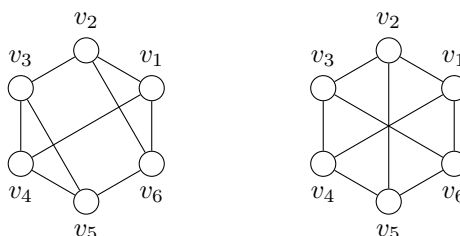


Figure 19.2: 3-regular graph of order 6

- (a) Draw all eleven nonisomorphic graphs of order 4
 $P_4, C_4, \bar{C}_4, K_4, \bar{K}_4, K_4 - e, K_2 + 2K_1, K_{1,3}, \bar{K}_{1,3}, P_2 + K_1, C_3 + e$

- (b) Let S be a set of 23 graphs of order 4, then S has at least three graphs that are pairwise isomorphic.

Proof. There are at most eleven non-isomorphic graphs of order r . By pigeonhole principle, in a family of 23 graphs of order 4 at least three graphs are pairwise isomorphic. \square

3. Give an example of a graph G of order 5 such that $G \cong \bar{G}$.

There are only two such graphs upto isomorphism. (see Figure 19.3)

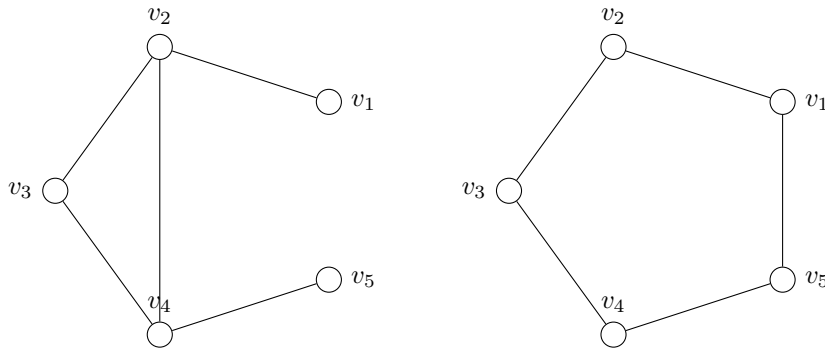


Figure 19.3: Self-complementary graphs of order 5

4. There are three regular graphs of order 5, and eight regular graphs of order 6. Draw all of them
 Three regular graphs of order 5 are C_5 , K_5 and \bar{K}_5 .
 Eight regular graphs of order 6 are $3K_2$, C_6 , $2K_3$, K_6 and complements.
5. Prove that two graph G_1 and G_2 are isomorphic if and only if their complements are isomorphic.

Proof.

$$\begin{aligned}
 G_1 \cong G_2 &\iff \phi : G_1 \rightarrow G_2, [uv \in E(G_1) \iff \phi(u)\phi(v) \in E(G_2)] \\
 &\iff \phi : G_1 \rightarrow G_2, [uv \notin E(G_1) \iff \phi(u)\phi(v) \notin E(G_2)] \\
 &\iff \bar{G}_1 \cong \bar{G}_2
 \end{aligned}$$

\square

6. Draw all nonisomorphic 4-regular graphs of order 7.
 There are only two distinct 2-regular graphs of order 7 upto isomorphism. The graphs are C_7 and $C_3 + C_4$. Every 4-regular graphs of order 7 is complement of some 2-regular graph of order 7. Thus there are only two 4-regular graphs of order 7. The graphs are \bar{C}_7 and $\overline{C_3 + C_4}$.

There are two trivial regular graphs K_p and \bar{K}_p . In order to count regular graphs of order p , it is always easier to count r -regular graphs for $0 \leq r \leq \lfloor (p-1)/2 \rfloor$.

For example, $p = 8$. We need to construct all nonisomorphic r -regular graphs of order 8 for $r = 0, 1, 2, 3$. There are one graph each for $r = 0, 1$. The 2-regular graphs are always disjoint union of cycles. Thus, the only possibilities are $C_3 + C_5$, $2C_4$ and C_8 . Finding 3-regular graphs of order 8 is the hard part.

You may find various classifications of all small graphs (upto order 13) at graphclasses.org. And there is a combinatorial result on number of regular graphs of order p available at oeis.org.

As per OEIS there are six nonisomorphic 3 regular graph of order 8. Thus, there are twenty two nonisomorphic regular graphs of order 8.

9. Show that the relation ‘is isomorphic to’ is an equivalence relation on the set of all graphs.

Proof. Trivially $G \cong G$ as G is equal to itself and equal graphs are always isomorphic.

Suppose $G_1 \cong G_2$. Then there exists a bijection $\phi : V(G_1) \rightarrow V(G_2)$ such that $uv \in E(G_1) \iff \phi(u)\phi(v) \in E(G_2)$. Clearly, $\phi^{-1} : V(G_2) \rightarrow V(G_1)$ is an isomorphism. Thus, $G_2 \cong G_1$.

Suppose $G_1 \cong G_2$ and $G_2 \cong G_3$. Then there exist bijection $\phi : V(G_1) \rightarrow V(G_2)$ such that $uv \in E(G_1) \iff \phi(u)\phi(v) \in E(G_2)$ and bijection $\psi : V(G_2) \rightarrow V(G_3)$ such that

$$\phi(u)\phi(v) \in E(G_2) \iff \psi(\phi(u))\psi(\phi(v)) \in E(G_3)$$

Clearly, $\psi \circ \phi : V(G_1) \rightarrow V(G_3)$ is a bijection such that

$$uv \in E(G_1) \iff \psi \circ \phi(u)\psi \circ \phi(v) \in E(G_3)$$

Therefore, $G_1 \cong G_3$. □

19.1.4 Subgraphs

Definitions 19.1.9. A graph H is a subgraph of G if $V(H) \subset V(G)$ and $E(H) \subset E(G)$.

Definitions 19.1.10. Let G be a graph and S be a nonempty subset of $V(G)$. The vertex set S induced subgraph H of graph G is the maximal subgraph of G with vertex set S .

Remark. Let G be a graph and let S be a nonempty subset of $V(G)$. If graph H is a vertex induced subgraph of G , then $V(H) = S$ and $E(H) \subset E(G)$.

Definitions 19.1.11. Let G be a graph and X be a nonempty subset of $E(G)$. A graph H is an edge induced subgraph of G if H is the minimal subgraph of G with edge set X .

Remark. Let G be a graph and let X be a nonempty subset of $E(G)$. If graph H is a vertex induced subgraph of G , then $V(H)$ is the set of all vertices incident with some edge in X and $E(H) \subset X$.

Definitions 19.1.12. Let G be a graph and S be a nonempty, *proper* subset of G . The graph $G - S$ is the maximal subgraph of G with vertex set $V(G) - S$.

Remark. When a vertex v is deleted from G , then all the edges incident with v (if any) are also deleted from G .

Definitions 19.1.13. Let G be a graph. A graph H is a spanning subgraph of G if $V(H) = V(G)$.

Definitions 19.1.14. Let G be a graph and X be a nonempty, *proper* subset of $E(G)$. The graph $G - X$ is the maximal subgraph of G with edge set $E(G) - X$.

Remark. Vertices of graph G are not deleted when its edges are deleted.

Definitions 19.1.15. Let G be a graph with vertices u and v . Suppose u and v are non-adjacent in G . Graph $G + \{uv\}$ is the graph with vertex set $V(G)$ and edge set $E(G) + \{uv\}$.

Problem Set 1.4

5. If $H = \langle E(G) \rangle$, does it follow that $H = \langle V(G) \rangle$?

Solution : No. Suppose G has an isolated vertex v . Then, subgraph induced by edge set $E(G)$ does not contain the vertex v .

6. Let G be a $\langle p, q \rangle$ graph. How many different edge-induced subgraphs does G have?

Solution : The size of the graph $q = |E(G)|$, the cardinality of the edge set of G . The number of nonempty subsets of $E(G)$ is $2^q - 1$. There exists a unique edge-induced subgraph for each subset of $E(G)$. Therefore, G has $2^q - 1$ edge-induced subgraphs.

19.1.5 Degree Sequences

Definitions 19.1.16. Let G be a graph of order p . The degree sequence $\mathcal{D}(G)$ is the finite sequence d_1, d_2, \dots, d_p of length p such that $d_1 \geq d_2 \geq \dots \geq d_p$ and $d_j = \deg(v_j)$ where $\{v_j : 1 \leq j \leq p\} = V(G)$.

Definitions 19.1.17. The smallest term d_p in the degree sequence of G is the minimum degree of G , $\delta(G)$. The largest term d_1 in the degree sequence of G is the maximum degree of G , $\Delta(G)$.

Theorem 19.1.2 (Havel-Hakimi). *A sequence d_1, d_2, \dots, d_p of nonnegative integers with $d_1 \geq d_2 \geq \dots \geq d_p$ where $p \geq 2$ and $d_1 \geq 1$ is graphical if and only if the sequence $d_2 - 1, d_3 - 1, \dots, d_{d_1+1} - 1, d_{d_1+2}, \dots, d_p$ is graphical.*

Proof. Suppose the sequence $s_1 : d_2 - 1, d_3 - 1, \dots, d_{d_1+1} - 1, d_{d_1+2}, \dots, d_p$ is graphical with vertex set $\{v_2, v_3, \dots, v_p\}$. Introduce a new vertex v_1 which is adjacent to $v_2, v_3, \dots, v_{d_1+1}$. The new graph has degree sequence $s : d_1, d_2, \dots, d_p$. Therefore, the sequence s is graphical.

Let G be a graph with degree sequence s . We claim that v_1 is adjacent to next d_1 vertices, say $v_2, v_3, \dots, v_{d_1+1}$. Suppose this is not true, then there exists $j < k$ such that $\deg(v_j) > \deg(v_k)$, but v_1 is adjacent to v_k and v_1 is non-adjacent to v_j . Since $\deg(v_j) > \deg(v_k)$, there exists a vertex v_n which is adjacent to v_j but nonadjacent to v_k . The degree sequence of the graph remains unchanged if we delete edges v_1v_k and v_jv_n and add edges v_1v_j and v_kv_n . Therefore, the claim is true. Now, the graph $G - v_1$ has degree sequence s_1 since $v_1 \in V(G)$ was adjacent to next d_1 vertices in G . \square

Definitions 19.1.18 (algorithm). An algorithm is a finite set of well-defined instructions to solve a particular family of problems.

Algorithm 19.1.1 (Graphical Degree Sequence). *The algorithm to check whether a finite sequence is graphical.*

Step 1 If all the integers in the sequence are 0, then the sequence is graphical. If the sequence contains a negative integer, then the sequence is not graphical. Otherwise, continue to Step 2

Step 2 Reorder the number in the current sequence in nonincreasing order.

Step 3 Delete the first number, say n from the sequence and subtract 1 from next n numbers in the sequence. Return to Step 1.

In Step 2, we don't have to check whether the sequence is nonincreasing order as it takes almost the same amount of time (effort) to reorder the sequence. We safely assume that if the number n is larger than the length of the sequence then 0 are added to the end of the sequence so that 1 can be subtracted.

Problem Set 1.5

1. Determine whether the degree sequence s is graphical and, if so, find a graph having s as its degree sequence

- (a) $s : 5, 5, 4, 4, 3, 3, 1, 0, 0$
 $s_1 : 4, 3, 3, 2, 2, 1, 0, 0, 0$, $s_2 : 2, 2, 1, 1, 1, 0, 0, 0$, $s_3 : 1, 1, 1, 0, 0, 0, 0$,
 $s_4 : 1, 0, 0, 0, 0, 0$, $s_5 : -1, 0, 0, 0, 0$. Thus, s is not graphical.
 Also the number of odd vertices in s is odd, which is not possible.
- (b) $s : 7, 6, 2, 2, 2, 1, 0, 0$
 $s_1 : 5, 1, 1, 1, 1, 0, -1, -1$. Thus, s is not graphical.
- (c) $s : 5, 3, 3, 3, 2, 2, 1, 1$
 $s_1 : 2, 2, 2, 1, 1, 1, 1, 1$, $s_2 : 1, 1, 1, 1, 1, 1, 1$, $s_3 : 1, 1, 1, 1, 0, 0$,
 $s_4 : 1, 1, 0, 0, 0, 0$, $s_5 : 0, 0, 0, 0$

In figure 19.4 delete the vertices in the order v_1, v_2, \dots, v_5 to get degree sequences s_1, s_2, \dots, s_5 .

Draw all those five graphs G_1, G_2, \dots, G_5 . From Havel-Hakimi theorem, we have a method for constructing the graph G . We start from G_5 with degree sequence s_5 and vertex set $\{v_6, v_7, v_8\}$. Graph G_5 has order 3 and size 0. Now we may add vertices v_5, v_4, \dots, v_1 and edges as required by the degree sequences s_4, s_3, \dots, s_1, s to get graphs G_4, G_3, \dots, G_1, G respectively.

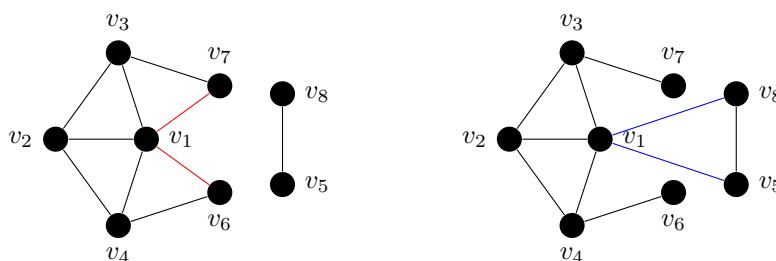


Figure 19.4: Two graphs with Degree Sequence 5, 3, 3, 3, 2, 2, 1, 1

(d) $s : 3, 3, 2, 2, 2, 2, 1$

$s_1 : 2, 2, 2, 1, 1, 1, s_2 : 1, 1, 1, 1, 1, s_3 : 1, 1, 1, 0$

$s_4 : 1, 0, 0, s_5 : -1, 0$. Thus, s is not graphical.

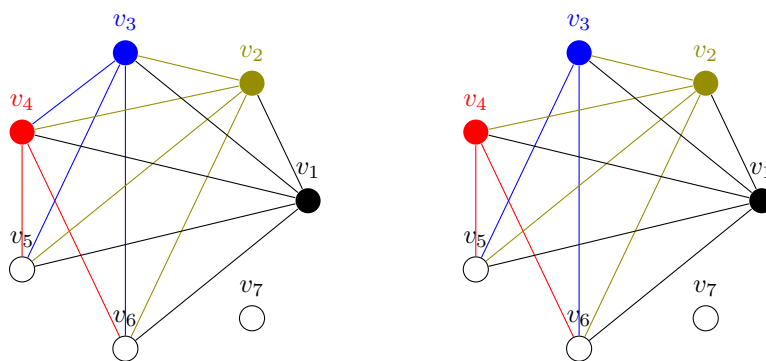
Also the number of odd vertices in s is odd, which is not possible.

2. Find a graph with degree sequence $\mathcal{D}(G) = \{0, 4, 5\}$.

The number of odd vertices should be even. And the degree of vertex should be less than the order of the graph. Since there is a vertex with degree 5 and a vertex of degree 0, we need at least a graph of order 7 with exactly one vertex of degree 0. The possibilities for graph of order seven are 5, 5, 5, 5, 4, 4, 0 and 5, 5, 4, 4, 4, 4, 0. If both these sequences are nongraphical, we will have to consider graph of higher order.

Suppose $s : 5, 5, 5, 5, 4, 4, 0$. Then $s_1 : 4, 4, 4, 3, 3, 0, s_2 : 3, 3, 2, 2, 0, s_3 : 2, 1, 1, 0, s_4 : 0, 0, 0$. Clearly, $s : 5, 5, 5, 5, 4, 4, 0$ is graphical.

Suppose $s : 5, 5, 4, 4, 4, 4, 0$. Then $s_1 : 4, 3, 3, 3, 3, 0, s_2 : 2, 2, 2, 2, 0, s_3 : 2, 1, 1, 0, s_4 : 0, 0, 0$. Clearly, $s : 5, 5, 4, 4, 4, 4, 0$ is also graphical.

Figure 19.5: Graphs of order 7 with degree sequence $\mathcal{D}(G) = \{0, 4, 5\}$

3.

4.

5.

- 6.
- 7.
- 8.
- 9.

19.1.6 Connected Graphs

Definitions 19.1.19 (walk). A walk of length n is an alternating sequence $W : v_0, e_1, v_1, \dots, e_n, v_n$ where $e_j = v_{j-1}v_j$.

Remark. Walk of length zero is a trivial walk. In the case of walk, we count the number of edges involved.

Definitions 19.1.20 (trail). A trail is a walk in which no edges are repeated.

Definitions 19.1.21 (path). A path is a walk in which no vertices are repeated.

Theorem 19.1.3. Every $u - v$ walk contains a $u - v$ path.

Proof. Suppose $u \neq v$. If $u = v$ then the walk contains the trivial path $u - u$.

Let $u - v : u_0, u_1, \dots, u_n$ where $u_0 = u$ and $u_n = v$ be a $u - v$ walk in graph G . Suppose $u - v$ has vertex $u_i = u_j$ for some $i \neq j$. If there are no such intermediate vertices then the $u - v$ walk is a path. Consider the walk $W_1 : u_0, u_1, \dots, u_{i-1}, u_j, u_{j+1}, \dots, u_n$ is a walk contained in previous walk with. The walk W_1 is shorter in length compared to W . If W_1 has repeated vertices, we repeat the process until there are no vertices repeated. \square

The walk has only finite length. And in each step the number of walk reduces. It is quite important from algorithmic point of view as it ensures that the algorithm will terminate in a finite number steps.

Definitions 19.1.22 (cycle). A cycle is a closed walk. That is, $v_0 = v_n$.

Definitions 19.1.23 (circuit). A circuit is a nontrivial closed trail.

Definitions 19.1.24. Let G be a graph with vertices u, v . The vertex u is connected to the vertex v if there exists a $u - v$ path in G .

Definitions 19.1.25. A graph G is connected if every pair of vertices u, v are connected in G .

Definitions 19.1.26 (component). A component of G is a maximal connected subgraph of G .

Remark. $k(G)$ is the number of components of graph G .

Problem Set 1.6

- 1.
- 2.
- 3.
- 4.
- 5.
- 6.
- 7.
- 8.
- 9.
- 10.
- 11.

19.1.7 Cut-Vertices and Bridges

Definitions 19.1.27. A vertex v in a graph G is a cut-vertex if $k(G-v) > k(G)$.

Definitions 19.1.28. An edge e in a graph G is a bridge if $k(G-e) > k(G)$.

Remark. If v is a cut-vertex of a connected graph G , then $k(G-v) \geq 2$. However, if e is a bridge of a connected graph G , then $k(G-e) = 2$.

Theorem 19.1.4. *An edge e of a connected graph G is a bridge if and only if e does not lie on a cycle of G .*

Proof. Suppose $e = uv$ is a bridge of G . Suppose e lies on a cycle u, v, w, \dots, x, u . Clearly, G has a $u-v$ path in it, say u, x, \dots, w, v .

Let $u_1, v_1 \in V(G)$. Since G is connected, there is a $u_1 - v_1$ path in G . Suppose the path does not contain e , then the path remains in $G - e$. If the path contains edge e , say $u_1, \dots, u, v, \dots, v_1$, then $u_1, \dots, u, x, \dots, w, v, \dots, v_1$ is a $u_1 - v_1$ walk in $G - e$. We know that, every $u_1 - v_1$ walk contains a $u_1 - v_1$ path. Thus, $G - e$ is connected. Therefore, e does not lie on any cycle.

Let G be a connected graph. Suppose $e = uv$ lies on no cycle of G . Then $G - e$ contains no $u - v$ path. Suppose $G - e$ has a $u - v$ path P , then $P + uv$ is a cycle in G which is a contradiction. Therefore, e is a bridge of G . \square

Definitions 19.1.29. A graph G is nonseparable if G has no cut-vertex.

Definitions 19.1.30. A block B of a graph G is a maximal nonseparable subgraph of G .

Definitions 19.1.31. An end-block of a graph G is a block that contains exactly one cut-vertex.

Theorem 19.1.5. *Let G be a connected graph with at least one cut-vertex, then G has at least two end-blocks.*

Problem Set 1.7

- 1.
- 2.
- 3.
- 4.
- 5.
- 6.
- 7.
- 8.

19.1.8 Special Graphs

Definitions 19.1.32. A complete graph is a graph in which every pair of vertices are adjacent.

Remark. A complete graph of order p , K_p is $(p - 1)$ -regular and has size $\binom{p}{2}$.

Definitions 19.1.33. A graph of order n which is itself a path is a path of order n , say P_n .

Definitions 19.1.34. A graph of order n which is itself a cycle is an n -cycle, C_n .

Remark. An odd(even) path is a path of odd(even) length. An odd(even) cycle is an n -cycle where n is odd(even).

Definitions 19.1.35. A graph G is bipartite if $V(G)$ can be partitioned into two nonempty subsets V_1 and V_2 such that every edge of G joins a vertex of V_1 with a vertex of V_2 .

Theorem 19.1.6. A nontrivial graph G is bipartite if and only if G has no odd cycles.

Proof. Suppose G is bipartite. Suppose G has an n -cycle $C : v_1, v_2, \dots, v_n, v_1$. Suppose $v_1 \in V_1$. Then $v_2 \in V_2$ since G is bipartite. Similarly, $v_3 \in V_1, v_4 \in V_2, \dots$. Since there is an edge $v_n v_1$ in C , $v_n \in V_2$. Clearly, n is even. Thus, every n -cycle in bipartite graph G is an even cycle.

Suppose G has no odd cycles. Assume without loss of generality that G is a connected graph. Let v be a vertex of G . Define $V_1 \subset V(G)$ as the set of all vertices v such that the minimal $v_1 - v$ is of odd length. Define $V_2 = V(G) - V_1$. Suppose there exists an edge between vertices in V_1 , say u, v . Then $(v_1 - u \text{ path} + uv + v - v_1 \text{ path})$ is a cycle of odd length. Similarly, if there exists an edge between vertices in V_2 , say u, v . Then there exists an odd cycle. Therefore, every edge in G joins a vertex of V_1 and a vertex of V_2 . Clearly, graph G is bipartite. \square

Definitions 19.1.36. Let G be a bipartite graph with vertex set partition V_1 and V_2 . Graph G is complete bipartite, if every vertex in V_1 is adjacent to every vertex in V_2 .

Remark. Let G be a bipartite graph with $|V_1| = n$ and $|V_2| = m$. Then $G = K_{m,n}$. The graph $K_{1,n}$ is a star.

Definitions 19.1.37. Let G be a graph with vertex set partition V_1, V_2, \dots, V_n such that no edge joins vertices in the same partite set.

Definitions 19.1.38. A complete n -partite graph is an n -partite graph in which every vertex in V_j is adjacent to every vertex in V_k where $j \neq k$.

Definitions 19.1.39 (union). $G_1 \cup G_2$ is the graph with

$$V(G_1 \cup G_2) = V(G_1) \cup V(G_2) \text{ and } E(G_1 \cup G_2) = E(G_1) \cup E(G_2)$$

Remark. If $G_1 \cong G_2 \cong G$, then $G_1 \cup G_2 = 2G$.

Definitions 19.1.40 (join). $G_1 + G_2$ is the graph with

$$V(G_1 + G_2) = V(G_1) \cup V(G_2)$$

$$uv \in E(G_1 + G_2) \text{ if } \begin{cases} uv \in E(G_1) \\ u \in V(G_1), v \in V(G_2) \\ uv \in E(G_2) \end{cases}$$

Definitions 19.1.41 (product). $G_1 \times G_2$ is the graph with

$$V(G_1 \times G_2) = V(G_1) \times V(G_2)$$

A vertex of $G_1 \times G_2$ is of the form (u, v) where $u \in V(G_1)$ and $v \in V(G_2)$.

$$(u_1, u_2)(v_1, v_2) \in E(G_1 \times G_2) \text{ if } \begin{cases} u_1 = v_1, u_2 v_2 \in E(G_2) \\ u_2 = v_2, u_1 v_1 \in E(G_1) \end{cases}$$

Remark. By symmetry of the definitions, $G_1 \cup G_2 = G_2 \cup G_1$, $G_1 + G_2 = G_2 + G_1$ and $G_1 \times G_2 = G_2 \times G_1$.

Problem Set 1.8

1.

19.1.9 Digraphs

$id(v)$ indegree of a vertex v is the number of arcs entering the vertex v .

$od(v)$ outdegree of a vertex v is the number of arcs leaving the vertex v .

$$deg(v) = id(v) + od(v).$$

semiwalk is an alternating sequence of vertices and arcs, say $W : u_0 e_0 u_1 e_1 \dots u_n e_n$ where each edge e_k is either a (u_{n-1}, u_n) arc(forward arc) or a (u_n, u_{n-1}) arc(reverse arc).

unilaterally connected A digraph is unilaterally connected if there is either a directed $u - v$ path or a directed $v - u$ path for each pair of vertices u and v .

strongly connected A digraph is strongly connected if for each pair of vertices u and v , there exists both $u - v$ directed path and $v - u$ directed path.

(weakly) connected A digraph is connected or weakly connected if for each pair of vertices u and v there is a $u - v$ semiwalk(or a $u - v$ semipath).

symmetric A digraph is symmetric if each arc (u, v) there exists an opposite arc (v, u) in it.

tournament A digraph in which each pair of vertices is joined by exactly one arc. The underlying graph of a tournament is a complete graph.

G^* The symmetric digraph obtained by replacing each edge uv of a graph G with symmetric arcs (u, v) and (v, u) .

parallel arcs Two arcs are parallel if both arcs are incident with the same vertices and have the same direction.

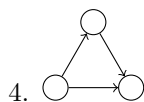
r-regular A digraph D is r -regular if $id(v) = od(v) = r$ for each vertex $v \in V(D)$.

Theorem 19.1.7.

$$\sum od(v) = \sum id(v) = q$$

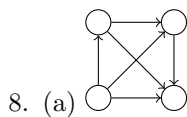
Proof. Every arc should enter some vertex and should leave some vertex. Counting the arcs is same as counting the number vertices it enters(or leaves). \square

Problem Set 1.9



4. is a unilaterally connected digraph with no (directed) cycles.

5. The digraph in problem 1.9.2 is not strongly connected. But, every vertex belongs to a directed cycle.



8. (a) (b) No, (c) No, \neq

9. Any tournament of order six have 15 arcs.

(a) Total outdegree is 15. Impossible to have 6 vertices of same outdegree.

(b) Total indegree is 15. Impossible to have 6 vertices of same indegree.

(c) Total degree in a tournament of order 6 is 30. Thus the degree of each vertex should be 5. Any tournament of order 6 will have the same degree for all its vertices. Draw K_6 and give some direction to all its edges to get a tournament of order 6.

10. Let D be a tournament with two vertices u, v of outdegree 0. Then there is no arc in between u and v . This is a contradiction as D is a tournament.

19.2 An Introduction to Algorithms

19.2.1 Algorithmic Complexity

Definitions 19.2.1. $f = O(g)$ if there exists $c \in \mathbb{R}^+$ and exists $N \in \mathbb{N}$ such that $\forall n > N, f(n) \leq cg(n)$.

Example, $f : 3n + 5$ and $g : n^2$. Then $f = O(g)$, consider $c = 1$ and $N = 4$. If $f = O(g)$, then order of f is less than or equal to the order of g .

Remark. f, g have same order if $f = O(g)$ and $g = O(f)$.

An algorithm of complexity f is **more efficient** than an algorithm of complexity g if $f = O(g)$ and $g \neq O(f)$.

The hierarchy of increasing order of complexity is,

$$O(1), O(\log n), O(n), O(n \log n), O(n^2), O(n^3), O(2^n), O(n!)$$

An algorithm is efficient if its complexity is a polynomial in n . If there exists an efficient algorithm to solve a problem, then that problem is **tractable**. Tiling problem is intractable.

Problem Set 2.1

3. $c = 1, N = 10$.
4. $c = 1, N = 5$.
5. To prove $f = O(g)$ take $c = 2, N = 5$ and to prove $g = O(f)$ take $c = 1, N = 1$. This is very important from complexity point of view. The result says that you can safely ignore terms with less complexity.
6. $\forall n \in \mathbb{N}, 2^n > n \implies \log 2^n > \log n \implies n \log 2 > \log n$. Take $c = \log 2$ and $N = 1$, then $\log n = O(n)$.
7. Use result from Problem 2.1.5 (a) $f = O(n^5)$ (b) $f = O(n^3 \log n)$ (c) $f = O(2^n)$
8. (a) $f = O(n^3) + O(n^{2.5}) = O(n^3)$ and $g = O(n^2 \log n) + O(n) = O(n^2 \log n)$. By hierarchy, $g = O(f)$ and $f \neq O(g)$.
(b) $f = O(2^n) + O(n^{10}) = O(2^n)$ and $g = O(n!) + O(3^n) = O(n!)$. By hierarchy, $f = O(g)$ and $g \neq O(f)$.
9. The complexity of algorithm 2.1 is $O(n)$. The problem to find first word is tractable.
10. Complexity is $O(n)$.
11. Complexity is $O(n)$.

Algorithm 19.2.1 (Largest Number). Let $N(1), N(2), \dots, N(n)$ be the list of numbers from which LARGEST number and its position are to be found.

1. LARGEST $\leftarrow N(1)$.
2. $p \leftarrow 1$.

3. For $k = 2$ to n .
 - (a) IF $N(k) < LARGEST$ then
 - i. $LARGEST \leftarrow N(k)$
 - ii. $p \leftarrow k$
4. Output $LARGEST, p$.

Algorithm 19.2.2 (First two Words). .

1. IF $W(1) < W(2)$ then
 - (a) $FIRST \leftarrow W(1)$
 - (b) $SECOND \leftarrow W(2)$
2. Else :
 - (a) $FIRST \leftarrow W(2)$
 - (b) $SECOND \leftarrow W(1)$
3. For $k = 3$ to n
 - (a) IF $W(1) > W(k)$ then
 - i. $FIRST \leftarrow W(k)$.
 - ii. $SECOND \leftarrow FIRST$.
 - (b) Else IF $W(2) > W(k)$ then
 - i. $SECOND \leftarrow W(k)$.
4. Output $FIRST, SECOND$.

19.2.2 Search Algorithms

Problem Set 2.2

1. Sequential Search
 - (a) $W(1) = MEYER$? No
 - (b) $W(2) = MEYER$? No
 - (c) ...
 - (e) $W(5) = MEYER$? Yes
 - i. $p \leftarrow 5$.
 - ii. Output p and Stop.

Binary Search

- (a) $f \leftarrow 1, l \leftarrow 11, k \leftarrow 6$,
 $MITCHELL = MEYER$? No
- (b) $MITCHELL > MEYER$? Yes $l \leftarrow 5, k \leftarrow 3$,
 $CHAPMAN = MEYER$? No
- (c) $CHAPMAN < MEYER$? Yes $f \leftarrow 4, k \leftarrow 4$,
 $DONALDSON = MEYER$? No

- (d) $DONALDSON < MEYER$? Yes $f \leftarrow 5$, $k \leftarrow 5$,
 $MEYER = MEYER$? Yes Output 5 and Stop

2. Binary Search

- (a) $f \leftarrow 1$, $l \leftarrow 12$, $k \leftarrow 6$,
 $HAWAII = MICHIGAN$? No.
- (b) $HAWAII < MICHIGAN$? Yes $f \leftarrow 7$, $k \leftarrow 9$,
 $LOUISIANA = MICHIGAN$? No.
- (c) $LOUISIANA < MICHIGAN$? Yes $f \leftarrow 10$, $k \leftarrow 11$,
 $NEVADA = MICHIGAN$? No
- (d) $NEVADA > MICHIGAN$? Yes $l \leftarrow 10$, $k \leftarrow 10$,
 $MICHIGAN = MICHIGAN$? Yes Output 10 and Stop.

19.2.3 Sorting Algorithms

Problem Set 2.3

1. Sequential Sort

- (a) $FIRST \leftarrow CARTER$, $p \leftarrow 4$
 $W(1) \leftarrow CARTER$, $W(4) \leftarrow DENNIS$.
- (b) $FIRST \leftarrow DENNIS$, $p \leftarrow 4$
 $W(2) \leftarrow DENNIS$, $W(4) \leftarrow THOMAS$.
- (c) $FIRST \leftarrow JACKSON$, $p \leftarrow 5$
 $W(3) \leftarrow JACKSON$, $W(5) \leftarrow MATTHEWS$.
- (d) $FIRST \leftarrow MATTHEWS$, $p \leftarrow 5$
 $W(4) \leftarrow MATTHEWS$, $W(5) \leftarrow THOMAS$.

Bubble Sort

- (a) $JACKSON < CARTER$ No.
 $CARTER < MATTHEWS$ Yes.
 $W(3) \leftarrow CARTER$, $W(4) \leftarrow MATTHEWS$.
 $\rightarrow DENNIS, THOMAS, CARTER, MATTHEWS, JACKSON$
 $CARTER < THOMAS$ Yes.
 $W(2) \leftarrow CARTER$, $W(3) \leftarrow THOMAS$.
 $\rightarrow DENNIS, CARTER, THOMAS, MATTHEWS, JACKSON$
 $CARTER < DENNIS$ Yes.
 $W(1) \leftarrow CARTER$, $W(2) \leftarrow DENNIS$.
 $\rightarrow CARTER, DENNIS, THOMAS, MATTHEWS, JACKSON$
- (b) $JACKSON < MATTHEWS$ Yes.
 $W(4) \leftarrow JACKSON$, $W(5) \leftarrow MATTHEWS$.
 $\rightarrow CARTER, DENNIS, THOMAS, JACKSON, MATTHEWS$
 $JACKSON < THOMAS$ Yes.
 $W(3) \leftarrow JACKSON$, $W(4) \leftarrow THOMAS$.
 $\rightarrow CARTER, DENNIS, JACKSON, THOMAS, MATTHEWS$
 $JACKSON < DENNIS$ No.
 $\rightarrow CARTER, DENNIS, JACKSON, THOMAS, MATTHEWS$

- (c) $MATTHEWS < THOMAS$ Yes.
 $W(4) \leftarrow MATTHEWS, W(5) \leftarrow THOMAS$.
 $\rightarrow CARTER, DENNIS, JACKSON, MATTHEWS, THOMAS$
 $MATTHEWS < JACKSON$ No.
 $\rightarrow CARTER, DENNIS, JACKSON, MATTHEWS, THOMAS$
- (d) $THOMAS < MATTHEWS$ No.
 $\rightarrow CARTER, DENNIS, JACKSON, MATTHEWS, THOMAS$

19.2.5 Greedy Algorithms

Problem Set 2.5

- 1.

19.2.6 Representing Graphs in a Computer

Problem Set 2.6

- 1.

19.3 Trees

19.3.1 Properties of Trees

Theorem 19.3.1. *Every tree of order p has size $p - 1$.*

19.3.2 Rooted Trees

19.3.3 Depth-First Search

Algorithm 19.3.1 (DFS Algorithm). .

1. $dfi(u) \leftarrow 0$.
2. $i \leftarrow 1$.
3. $S \leftarrow \phi$.
4. If $dfi(r) \neq 0$ for every vertex r , then output S and stop. Otherwise, let r be the first vertex such that $dfi(r) = 0$, $w \leftarrow r$ and continue.
5. $dfi(w) \leftarrow i$.
6. $i \leftarrow i + 1$.
7. 7.1 If $dfi(v) = 0$ for some vertex v in the adjacency list of w , then continue. Otherwise and goto Step 7.4
- 7.2 $PARENT(v) \leftarrow w$ and $w \leftarrow v$.
- 7.3 $S \leftarrow S \cup \{vw\}$ and goto Step 7.1
- 7.4 If $w \neq r$, then $w \leftarrow PARENT(w)$ and goto Step 7.1 Otherwise goto Step 4.

19.3.4 Depth-First Search : A Tool for Finding Blocks

19.3.5 Breadth-First Search

Algorithm 19.3.2 (Component finding Algorithm). . Let G be a graph. Let $bfi(v)$ denote the BFS index of each vertex, S be the set of vertices in the component of G containing the current vertex v and Q be the queue of all vertices of G whose children are yet to be visited.

1. (All vertices are given breadth first search index of 0. And i is initialised and it will be assigned to the i th vertex visited during the search.)
 - (a) $bfi(v) \leftarrow 0$ for each vertex $v \in V(G)$.
 - (b) $i \leftarrow 1$.
2. (We start with a vertex v from an unvisited component of G .)
 If $bfi(v) \neq 0$ for all $v \in V(G)$, then stop since all the components of G have been determined. Otherwise, let v be the first vertex such that $bfi(v) = 0$. Assign $bfi(v) \leftarrow i$, $i \leftarrow i + 1$, $S \leftarrow \{v\}$ and add v to Q .
3. For each vertex y adjacent to v such that $bfi(y) = 0$, assign $bfi(y) \leftarrow i$, $i \leftarrow i + 1$, $S \leftarrow S \cup \{y\}$, add y to Q .
4. If $Q \neq \phi$, then delete a vertex x from Q . Assign $v \leftarrow x$ and go to Step 3. Otherwise output S and go to Step 5.

19.3.6 The Minimum Spanning Tree Problem

19.4 Paths and Distance in Graphs

19.4.1 Distance in Graphs

19.4.2 Distance in Weighted Graphs

19.4.3 The Center and Median of a Graph

19.4.4 Activity Digraphs and Critical Paths

19.5 Networks

19.5.1 An Introduction to Networks

Definitions 19.5.1. A **network** N is a digraph D with two special vertices source s and sink t together with a capacity function $c : E(D) \rightarrow \mathbb{Z}$ such that for every arc $a = (u, v)$ of the digraph, $c(u, v)$ is non-negative.

Remark. Mathematical Modeling using Network,

1. There is no restriction on indegree/outdegree of source/sink vertices of the digraph D of a network N .
2. Applications of Network : Transportation problem.

$c(u, v)$ is the capacity of the arc (u, v) of D

$N^+(x) = \{y \in V(D) : (x, y) \in E(D)\}$ is the out-neighbourhood of x .

$N^-(x) = \{y \in V(D) : (y, x) \in E(D)\}$ is the in-neighbourhood of x .

Definitions 19.5.2. A flow \mathbf{f} in a network \mathbf{N} is function $f : E(D) \rightarrow \mathbb{Z}$ such that 1. each edge satisfies capacity constraint and 2. each vertex except source and sink satisfies conservation equation.

capacity constraint

$$0 \leq f(a) \leq c(a) \text{ for every arc } a \in V(D) \quad (19.1)$$

conservation equation

$$\sum_{y \in N^+(x)} f(x, y) = \sum_{y \in N^-(x)} f(y, x), \quad \forall \text{ vertex } x \in V(D) - \{s, t\} \quad (19.2)$$

net flow out of x

$$\sum_{y \in N^+(x)} f(x, y) - \sum_{y \in N^-(x)} f(y, x)$$

net flow into x

$$\sum_{y \in N^-(x)} f(y, x) - \sum_{y \in N^+(x)} f(x, y)$$

Definitions 19.5.3. The flow \mathbf{f} in a network \mathbf{N} is the net flow out of source s .

Remark. 1. net flow out of/into $x \in V(D) - \{s, t\}$ is zero.

2. Without loss of generality¹, underlying digraph is always assymetric.

$(X, Y) = \{(x, y) \in E(D) : x \in X, y \in Y\}$.

Let X, Y be non-empty subsets of $V(D)$ such that X, Y are disjoint. Then (X, Y) is the set of all arcs from X to Y .

flow from X to Y is the sum of flow on each arc in (X, Y)

$$f(X, Y) = \sum_{(x, y) \in (X, Y)} f(x, y) \quad (19.3)$$

capacity of the partition (X, Y) is the total capacity of arcs in (X, Y)

$$c(X, Y) = \sum_{(x, y) \in (X, Y)} c(x, y) \quad (19.4)$$

cut Let $P \subset V(D)$ such that $s \in P$ and $t \notin P$ and $\bar{P} = V(D) - P$, then (P, \bar{P}) is a cut.

¹If underlying digraph of a network is symmetric, then by replacing an arc (u, v) with a new vertex w and two arcs $(u, w), (w, v)$ gives an assymetric digraph.[Gray Chartrand,]pp.131

flow from P to \bar{P} is the sum of flow on each arc in (P, \bar{P}) .

$$f(P, \bar{P}) = \sum_{(x,y) \in (P, \bar{P})} f(x, y) \quad (19.5)$$

flow from \bar{P} to P is the sum of flow on each arc in (\bar{P}, P)

$$f(\bar{P}, P) = \sum_{(x,y) \in (\bar{P}, P)} f(x, y) \quad (19.6)$$

capacity of the cut (P, \bar{P}) is the total capacity of the arcs in (P, \bar{P})

$$c(P, \bar{P}) = \sum_{(x,y) \in (P, \bar{P})} c(x, y) \quad (19.7)$$

Theorem 19.5.1. *For any cut (P, \bar{P}) , the flow in N is $f(N) = f(P, \bar{P}) - f(\bar{P}, P)$.*

Synopsis. The net flow out of source s is the flow $f(N)$ in the network N . Let (P, \bar{P}) be a cut of N , then $s \in P$ and $t \notin P$. Suppose $P = \{s\}$, then the theorem is true. Suppose P is not singleton, then for each vertex $x \in P$, $x \neq s$, the net flow out of x is zero by flow conservation equation. And flow between vertices in P cancels out each other. Thus adding net flow out of each vertex in P , will be same as the net flow out of source which is the flow in the network, $f(N)$.

Proof.

$$\text{Flow, } f = \sum_{y \in N^+(s)} f(s, y) - \sum_{y \in N^-(s)} f(y, s) \quad (19.8)$$

By conservation equation, we have $\forall x \in P$, $x \neq s$,

$$\sum_{y \in N^+(x)} f(x, y) - \sum_{y \in N^-(x)} f(y, x) = 0 \quad (19.9)$$

By above equations,

$$\begin{aligned} \text{Flow, } f &= \sum_{x \in P} \sum_{y \in N^+(x)} f(x, y) - \sum_{x \in P} \sum_{y \in N^-(x)} f(y, x) \\ &= \sum_{(x,y) \in (P, \bar{P})} f(x, y) - \sum_{(y,x) \in (\bar{P}, P)} f(y, x) \end{aligned} \quad (19.10)$$

□

Corollary 19.5.1.1. *Flow cannot exceed the capacity of any cut (P, \bar{P}) . Further, $f(N) \leq \min c(P, \bar{P})$.*

Synopsis. Let (P, \bar{P}) be a cut in network N , then by theorem the flow $f(N) = \text{flow from } P \text{ to } \bar{P} - \text{flow from } \bar{P} \text{ to } P$. Since the flow from \bar{P} to P is non-negative, $f(N) \leq \text{flow from } P \text{ to } \bar{P}$. Clearly, $f(x, y) \leq c(x, y)$ by the capacity constraint. Thus $f(N) \leq f(P, \bar{P}) \leq c(P, \bar{P}) \leq \min c(P, \bar{P})$.

Proof.

$$\begin{aligned}
 f(N) &= \sum_{(x,y) \in (P, \bar{P})} f(x,y) - \sum_{(y,x) \in (\bar{P}, P)} f(y,x) \\
 &\leq \sum_{(x,y) \in (P, \bar{P})} f(x,y) = f(P, \bar{P}) \\
 &\leq \sum_{(x,y) \in (P, \bar{P})} c(x,y) = c(P, \bar{P}), \quad \because \forall x, y \in V(D), f(x,y) \leq c(x,y) \\
 &\leq \min c(P, \bar{P})
 \end{aligned}$$

□

Corollary 19.5.1.2. *In a network N flow is the net flow into the sink of N .*

Synopsis. Let $\bar{P} = \{t\}$, then by theorem $f(N)$ is the net flow into the sink.

Proof. Suppose $P = V(D) - \{t\}$. Then by theorem, we have

$$\begin{aligned}
 f(N) &= \sum_{(x,y) \in (P, \bar{P})} f(x,y) - \sum_{(y,x) \in (\bar{P}, P)} f(y,x) \\
 &= \sum_{x \in N^-(t)} f(x,t) - \sum_{x \in N^+(t)} f(t,x)
 \end{aligned}$$

□

Remark. Exercise 5.1

4. Let N be a network with underlying digraph D which has a vertex $v \in V(D) - \{s, t\}$ with zero indegree. Clearly the flow into v is zero. Thus flow out of v is also zero by flow conservation equation. Let N' be the network obtained from N by deleting the vertex v . Then $f(N) = f(N')$.

19.5.2 The Max-Flow Min-Cut Theorem

maximum flow A flow f in network N is maximum flow in N , if $f(N) \geq f'(N)$ for each flow f' in N .

minimum cut A cut (P, \bar{P}) in network N is minimum cut of N , if $c(P, \bar{P}) \leq c(X, \bar{X})$ for each cut (X, \bar{X}) in N .

f -unsaturated Let f be a flow in network N with underlying digraph D , and $Q = u_0, a_1, u_1, a_2, \dots, u_{n-1}, a_n, u_n$ be a semipath in D such that every forward arc $a_i = (u_{i-1}, u_i)$ has flow not upto its capacity, $f(a_i) < c(a_i)$ and every reverse arc $a_i = (u_i, u_{i-1})$ has some positive flow in it, $f(a_i) > 0$

f -augmenting semipath Let f be a flow in a network N with underlying digraph D . Suppose semipath $Q = s, a_1, u_1, a_2, \dots, u_{n-1}, a_n, t$ (from source to sink) is f -unsaturated, then Q is an f -augmenting semipath.

Theorem 19.5.2. *Let f be a flow in a network N with underlying digraph D . The flow f is maximum in N iff there is no f -augmenting semipath in D .*

Synopsis. Suppose Q is an f -augmenting semipath in D , then there exists a flow f^* in N such that $f(N) + \Delta = f^*(N)$. Therefore, f is not a maximum flow in N . Suppose there is no f -augmenting semipath in D , then there exists a cut (P, \bar{P}) such that $f(a) = c(a) \forall a \in (P, \bar{P})$ and $f(a) = 0 \forall a \in (\bar{P}, P)$. Suppose f^* is a maximum flow in N , then $f(N) \leq f^*(N) \leq c(P, \bar{P}) = f(N)$.

Proof. Let f be a flow in a network N with underlying digraph D and $Q = s, a_1, u_1, a_2, u_2, \dots, u_{n-1}, a_n, t$ be an f -augmenting semipath in D .

$$\text{define } \Delta_i = \begin{cases} c(a_i) - f(a_i) & \text{for every forward arc } a_i \in Q, \\ f(a_i) & \text{for every reverse arc } a_i \in Q, \end{cases}$$

Define $\Delta = \min\{\Delta_i\}$. Also define $f^* : E(D) \rightarrow \mathbb{Z}$ such that

$$f^*(a_i) = \begin{cases} f(a_i) + \Delta, & \text{for every forward arc } a_i \in Q, \\ f(a_i) - \Delta, & \text{for every reverse arc } a_i \in Q, \\ f(a_i), & \text{for every arc of } D \text{ which are not in } Q. \end{cases}$$

Since Q is an f -augmenting semipath in D , $\Delta > 0$ and $f(N) + \Delta = f^*(N)$.

Clearly $f(N) < f^*(N)$, and it is enough to show that f^* is a flow in N . f^* is a flow if it satisfies 1. capacity constraint and 2. conservation equation. For any arc $a_i \notin Q$, $f^*(a_i) = f(a_i) \leq c(a_i)$. Suppose $a_i \in Q$. If $a_i = (u_{i-1}, u_i)$, a_i is a forward arc and we have $f^*(a_i) = f(a_i) + \Delta \leq f(a_i) + \Delta_i = f(a_i) + c(a_i) - f(a_i) = c(a_i)$. If $a_i = (u_i, u_{i-1})$, then a_i is a reverse arc and we have $f^*(a_i) = \Delta \leq \min\{\Delta_i\} = \Delta_i = c(a_i)$. Thus f^* satisfies capacity constraint on every arc of D .

Let $x \in V(D) - \{s, t\}$. Suppose $x \notin Q$,

$$\begin{aligned} \text{Net flow out of } x &= \sum_{y \in N^+(x)} f^*(x, y) - \sum_{y \in N^-(x)} f^*(y, x) \\ &= \sum_{y \in N^+(x)} f(x, y) - \sum_{y \in N^-(x)} f(y, x) \\ &= 0 \end{aligned}$$

Suppose $x = u_i \in Q$, then Q has two arcs having vertex x say, a_{i-1} , and a_i . There are four possibilities for these two arcs,

1. Both a_{i-1} , a_i are forward arcs.
2. Arc a_{i-1} is forward, but arc a_i is reverse.
3. Arc a_{i-1} is reverse, but arc a_i is forward.
4. Both a_{i-1} , a_i are reverse arcs.

Case 1 $a_{i-1} = (u_{i-1}, u_i)$ and $a_i = (u_i, u_{i+1})$.

$$\begin{aligned}
 \text{Net flow out of } x &= \sum_{y \in N^+(x)} f^*(x, y) - \sum_{y \in N^-(x)} f^*(y, x) \\
 &= \sum_{\substack{y \in N^+(x) \\ y \neq u_{i+1}}} f^*(x, y) + f^*(u_i, u_{i+1}) - \left(\sum_{\substack{y \in N^-(x) \\ y \neq u_{i-1}}} f^*(y, x) + f^*(u_{i-1}, u_i) \right) \\
 &= \sum_{\substack{y \in N^+(x) \\ y \neq u_{i+1}}} f(x, y) + f(u_i, u_{i+1}) + \Delta - \left(\sum_{\substack{y \in N^-(x) \\ y \neq u_{i-1}}} f(y, x) + f(u_{i-1}, u_i) \right) - \Delta \\
 &= \sum_{y \in N^+(x)} f(x, y) - \sum_{y \in N^-(x)} f(y, x) \\
 &= 0
 \end{aligned}$$

Case 2 $a_{i-1} = (u_{i-1}, u_i)$ and $a_i = (u_{i+1}, u_i)$.

$$\begin{aligned}
 \text{Net flow out of } x &= \sum_{y \in N^+(x)} f^*(x, y) - \sum_{y \in N^-(x)} f^*(y, x) \\
 &= \sum_{\substack{y \in N^+(x) \\ y \neq u_{i+1}, u_{i-1}}} f^*(x, y) + f^*(u_i, u_{i+1}) + f^*(u_i, u_{i-1}) - \sum_{y \in N^-(x)} f^*(y, x) \\
 &= \sum_{\substack{y \in N^+(x) \\ y \neq u_{i+1}, u_{i-1}}} f(x, y) + f(u_i, u_{i+1}) + \Delta + f(u_i, u_{i-1}) - \Delta - \sum_{y \in N^-(x)} f(y, x) \\
 &= \sum_{y \in N^+(x)} f(x, y) - \sum_{y \in N^-(x)} f(y, x) \\
 &= 0
 \end{aligned}$$

Case 3 $a_{i-1} = (u_i, u_{i-1})$ and $a_i = (u_i, u_{i+1})$.

$$\begin{aligned}
 \text{Net flow out of } x &= \sum_{y \in N^+(x)} f^*(x, y) - \sum_{y \in N^-(x)} f^*(y, x) \\
 &= \sum_{y \in N^+(x)} f^*(x, y) - \left(\sum_{\substack{y \in N^-(x) \\ y \neq u_{i-1}, u_{i+1}}} f^*(y, x) + f^*(u_{i-1}, u_i) + f^*(u_{i+1}, u_i) \right) \\
 &= \sum_{y \in N^+(x)} f^*(x, y) - \left(\sum_{\substack{y \in N^-(x) \\ y \neq u_{i-1}, u_{i+1}}} f(y, x) + f(u_{i-1}, u_i) + \Delta + f(u_{i+1}, u_i) - \Delta \right) \\
 &= \sum_{y \in N^+(x)} f(x, y) - \sum_{y \in N^-(x)} f(y, x) \\
 &= 0
 \end{aligned}$$

Case 4 $a_{i-1} = (u_i, u_{i-1})$ and $a_i = (u_{i+1}, u_i)$.

$$\begin{aligned}
 \text{Net flow out of } x &= \sum_{y \in N^+(x)} f^*(x, y) - \sum_{y \in N^-(x)} f^*(y, x) \\
 &= \sum_{\substack{y \in N^+(x) \\ y \neq u_{i-1}}} f^*(x, y) + f^*(u_i, u_{i-1}) - \left(\sum_{\substack{y \in N^-(x) \\ y \neq u_{i+1}}} f^*(y, x) + f^*(u_{i+1}, u_i) \right) \\
 &= \sum_{\substack{y \in N^+(x) \\ y \neq u_{i-1}}} f(x, y) + f(u_i, u_{i-1}) - \Delta - \left(\sum_{\substack{y \in N^-(x) \\ y \neq u_{i+1}}} f(y, x) + f(u_{i+1}, u_i) \right) + \Delta \\
 &= \sum_{y \in N^+(x)} f(x, y) - \sum_{y \in N^-(x)} f(y, x) \\
 &= 0
 \end{aligned}$$

Therefore, f^* is a flow on N . We have $f(N) < f^*(N)$. Thus f is not maximum flow in N due to the existence of an f -augmenting semipath in D .

Conversely, assume that there is no f -augmenting semipath in D . Now, we construct a cut (P, \bar{P}) of N . Let P be the set of all vertices $x \in V(D)$ such that there is an f -unsaturated $s - x$ semipath in D . Trivially, $s \in P$. And $t \notin P$ since there are no f -augmenting semipath in D .² Clearly, (P, \bar{P}) is a cut of the network N .

We claim that $c(P, \bar{P}) = f(N)$. Suppose there is a forward arc $(x, y) \in (P, \bar{P})$, then flow in it is saturated. If $f(x, y) < c(x, y)$, then there is an f -unsaturated $s - y$ semipath in D . ie, $s - x$ semipath + arc (x, y) . Thus every forward arc $(x, y) \in (P, \bar{P})$ is saturated. Suppose there is a reverse arc $(y, x) \in (\bar{P}, P)$, then there is no flow in it (saturated reversed arc). If $f(y, x) > 0$, then there is an f -unsaturated $s - y$ semipath in D . ie, $s - x$ semipath + arc (y, x) . Thus every reverse arc $(y, x) \in (\bar{P}, P)$ is saturated. And we have,

$$\begin{aligned}
 \sum_{(x,y) \in (P, \bar{P})} f(x, y) &= \sum_{(x,y) \in (P, \bar{P})} c(x, y) \\
 \sum_{(y,x) \in (\bar{P}, P)} f(y, x) &= 0 \\
 f(N) &= \sum_{(x,y) \in (P, \bar{P})} f(x, y) - \sum_{(y,x) \in (\bar{P}, P)} f(y, x) \\
 &= \sum_{(x,y) \in (P, \bar{P})} c(x, y) \\
 &= c(P, \bar{P})
 \end{aligned}$$

Suppose f^* is maximum flow in network N and (X, \bar{X}) is minimum cut of N . Then $f(N) \leq f^*(N)$. Thus we have, $f(N) \leq f^*(N) \leq c(X, \bar{X}) \leq c(P, \bar{P}) =$

²An f -augmenting semipath is an f -unsaturated $s - t$ semipath in D .

$f(N)$. Therefore, $f(N) = f^*(N)$. ie, the flow f is maximum in network N if there are no f -augmenting semipaths in D . \square

Theorem 19.5.3 (maximum-flow, min-cut). *In every network, the value of maximum flow equals capacity of minimum cut.*

Proof. Suppose flow f in network N is maximum, then by previous theorem there is no f -augmenting semipath in D . And $f(N) \leq c(X, \bar{X})$ for any cut (X, \bar{X}) in N . We can construct a cut (P, \bar{P}) in N such that $f(N) = c(P, \bar{P})$. Let P be the set of all vertices x in D such that there is an f -unsaturated $s-x$ semipath in D . Clearly $s \in P$ and $t \notin P$. Also $f(P, \bar{P}) = c(P, \bar{P})$ and $f(\bar{P}, P) = 0$. Then the cut (P, \bar{P}) is minimum cut of N . Suppose there is a cut (X, \bar{X}) such that $c(X, \bar{X}) < c(P, \bar{P})$. Then $f(N) = f(P, \bar{P}) - f(\bar{P}, P) = c(P, \bar{P}) < c(X, \bar{X})$ which is a contradiction. Therefore, the value of maximum flow equals capacity of minimum cut. \square

Remark. Exercise 5.2

1. Suppose (X, \bar{X}) is a cut of N such that $f(a) = c(a)$, $\forall a \in (X, \bar{X})$ and $f(a) = 0$, $\forall a \in (\bar{X}, X)$. By the definition of cut, $s \in X$ and $t \in \bar{X}$. Thus there is no f -augmenting semipath in D . Suppose there is an f -augmenting semipath Q in D , then there is either (a) a forward arc $(x, y) \in (X, \bar{X})$ such that $f(x, y) < c(x, y)$ or (b) a reverse arc $(y, x) \in (\bar{X}, X)$ such that $f(y, x) > 0$ which is a contradiction. Therefore, the flow $f(N)$ is maximum and the given cut (X, \bar{X}) is minimum as shown in the proof of the maximum-flow min-cut theorem.
3. The algorithm suggested in the hint of this exercise won't work if two subnetworks have a common arc such that the direction of flow in which is not consistent. Suppose, the generalized network is not supposed to have any common arcs. Then construct subnetworks for each pair (s, t) with all those arcs which are on some $s-t$ semipath. Define subnetwork capacity function $c'(a) = c(a)$ for every arc in N' .

Let N be a generalized network with set of sources S and set of sinks T . A flow in N is maximum if there is not f -augmenting $s-t$ semipath for each pair $(s, t) \in S \times T$.

19.5.3 A max-flow min-cut algorithm

Theorem 19.5.4. *Let N be a network with underlying digraph D , source s , sink t , capacity function c and flow f . Let D' be the digraph with same vertex set as D and arc set defined by $E(D') = \{(x, y) : (x, y) \in E(D), c(x, y) > f(x, y) \text{ or } (y, x) \in E(D), f(y, x) > 0\}$. ie, D' has only the unsaturated arcs of D . Then D' has an $s-t$ directed path iff D has an f -augmenting semipath. Moreover, shortest $s-t$ path in D' has the same length as shortest f -augmenting semipath in D .*

Synopsis. Each directed $s-t$ path in D' has respective f -augmenting semipath in D and vice versa. Clearly, they have the same length.

Proof. Let N be a network with underlying digraph D , capacity c and flow f . Let D' be the digraph with vertex set $V(D') = V(D)$ and arc set $E(D') = \{(x, y) : \text{either } (x, y) \text{ or } (y, x) \text{ is unsaturated in } N\}$.

Suppose D' has a directed $s - t$ path $Q' : s, u_1, u_2, \dots, u_{n-1}, t$. Then by the construction of D' , for each $u_i \in Q$, there exists an f unsaturated arc a_i in D . ie, either forward arc $a_i = (u_{k-1}, u_k)$ such that $f(u_{k-1}, u_k) < c(u_{k-1}, u_k)$ or reverse arc $a_i = (u_k, u_{k-1})$ such that $f(u_k, u_{k-1}) > 0$. Therefore, we have an $s - t$ semipath $Q : s, a_1, u_1, a_2, \dots, u_{n-1}, a_n, t$ in D such that Q is an f -augmenting semipath since every arc in Q is f -unsaturated. Clearly, Q, Q' are of the same length.

Conversely, suppose that the digraph D has an f -augmenting semipath $Q : s, a_1, u_1, a_2, \dots, u_{n-1}, a_n, t$. Then each arc $a_i \in Q$ are f -unsaturated and by the construction of D' , there exists a directed $s - t$ path $Q' = s, u_1, u_2, \dots, u_{n-1}, t$ in D' . And Q, Q' are of the same length.

There is a one-one correspondence between the directed $s - t$ paths in D' and f -augmenting semipaths in D . Clearly, they have the same length. Thus shortest directed $s - t$ path in D and shortest f -augmenting semipath in D' are of the same length. \square

saturation arc of N with respect to the flow f is an arc a_j in an f -augmenting semipath Q with $\Delta_j = \Delta$.

augmentation path is an f -augmenting semipath Q in D .

Algorithm 19.5.1 (max-flow min-cut). *An algorithm to find maximum flow and minimum cut of a network N with underlying digraph D , source s , sink t , capacity function c and initial flow f .*

1. Construct digraph D' with vertex set $V(D') = V(D)$ and arc set $E(D') = \{(x, y) : (x, y) \in E(D) \& f(x, y) < c(x, y) \text{ or } (y, x) \in E(D) \& f(y, x) > 0\}$
2. Find (shortest) $s - t$ directed path in D' using Moore's breadth first search(BFS) algorithm. If D' doesn't have an $s - t$ path, then proceed to step 5. Otherwise, let $Q' : s, u_1, u_2, \dots, u_{n-1}, t$ be a (shortest) $s - t$ path in D' .
3. Let $Q : s, a_1, u_1, a_2, \dots, u_{n-1}, a_n, t$ be the respective semipath in D such that $f(a_j) < c(a_j)$ for forward arcs and $f(a_i) > 0$ for reverse arcs. Let $\Delta_j = c(a_j) - f(a_j)$ for forward arcs and $\Delta_j = f(a_j)$ for reverse arcs. And let $\Delta = \min\{\Delta_j\}$. And augment flow f by Δ ie, $f(a_j) \leftarrow f(a_j) + \Delta$ for forward arcs and $f(a_j) \leftarrow f(a_j) - \Delta$ for reverse arcs.
4. Goto step 1 (Proceed with new flow f and find whether there are any directed $s - t$ paths in D' . If any, augment the flow along the new augmentation path Q by saturating the flow along the saturation arc.)
5. There is no $s - t$ directed path in D' . Thus there is no f -augmenting semipath in D . Therefore the flow f in N is maximum. Let P be the set of all vertices in D' with non-zero breadth first index(bfi) from Moore's BFS algorithm applied in step 2. (P, \bar{P}) is minimum cut of N .

Remark. Validity of the algorithm is proved in the previous theorem.

19.5.5 Connectivity and Edge-Connectivity

edge cutset is the set U subset of $E(G)$ such that $G - U$ is disconnected.

vertex cutset is the set S subset of $V(G)$ such that $G - S$ is disconnected.

edge connectivity $\lambda(G)$ is the minimum cardinality of all edge cutsets of G .

connectivity $\kappa(G)$ is the minimum cardinality of all vertex cutsets of G .

Theorem 19.5.5. For every graph G , $\kappa(G) \leq \lambda(G) \leq \delta(G)$

Proof. Suppose graph G is disconnected then $\kappa(G) = \lambda(G) = \delta(G) = 0$. Let G be a connected graph. Then G has at least one vertex v with degree $\delta(G)$. Therefore $\lambda(G) \leq \delta(G)$ since edges incident with v form an edge cutset of G and $\lambda(G)$ is the cardinality of all edge cutsets.

Let G be a graph with edge connectivity $\lambda(G) = c$. Let U be an edge cutset with cardinality c and let edge $uv \in U$. Construct a set of vertices $S \subset V(G)$ such that (S is of minimal cardinality and) for each edge in U other than uv , S has a vertex incident with it. Cardinality of S is at most $c - 1$, since we can select one vertex each for each edge in U other than uv . If $G - S$ is a disconnected graph, then $\kappa(G) < \lambda(G)$. Suppose $G - S$ is a connected graph, then delete a non-pendent vertex u or v from $G - S$, say v . Since $G - S$ is a connected graph with a singleton edge cutset, $\{uv\}$. We have a vertex cutset $S \cup \{v\}$ of G . Therefore, $\kappa(G) \leq c = \lambda(G)$. \square

Theorem 19.5.6. If G is a graph of diameter 2, then $\lambda(G) = \delta(G)$

n -edge connected G is n -edge connected if $\lambda(G) \geq n$.

n connected G is n -connected if $\kappa(G) \geq n$.

Theorem 19.5.7. Let G be a graph of order p and n be an integer such that $1 \leq n \leq p - 1$. If $\delta(G) \geq \frac{p+n-2}{2}$, then G is n -connected.

connection number $c(G)$ is the smallest integer such that $2 \leq c(G) \leq p$ and every subgraph of order n in G is connected.

l -connectivity $\kappa_l(G)$ is minimum number of vertices whose removal will produce a disconnected graph with at least l components or a graph with fewer than l vertices.

(n, l) -connected A graph G is (n, l) -connected if $\kappa_l(G) \geq n$.

Remark. Exercises 5.5

1. $\lambda(K_{m,n}) = \kappa(K_{m,n}) = m$

8. $c(K_p) = 2$, $c(K_{m,n}) = n + 1$, $c(C_p) = p - 1$

Every two vertices of complete graph of order p are adjacent. For complete bi-partite graph $K_{m,n}$ such that $1 \leq m \leq n$, there exists a totally disconnected subgraph of order n . Therefore $c(K_{m,n}) \geq n + 1$. And with $n + 1$ vertices, both partitions have at least two vertices each and therefore the graph is connected and $c(K_{m,n}) \leq n + 1$. For cycle C_p , any

subgraph is disconnected if two non-adjacent vertices are deleted. Therefore $c(C_p) \geq p - 1$. And C_p remains connected even after deletion of any vertex, therefore $c(C_p) \leq p - 1$.

9.

$$\delta(G) \geq \frac{p + (l - 1)(n - 2)}{l} \implies \kappa_l(G) \geq n$$

19.5.6 Menger's Theorem

Theorem 19.5.8. For a non-trivial graph G , $\lambda(u, v) = M'(u, v)$ for every pair (u, v) of vertices of G .

Corollary 19.5.8.1. Graph G is n -edge connected iff every two vertices of G are connected by at least n edge disjoint paths.

Theorem 19.5.9. For every pair of non-adjacent vertices u, v in graph G , $\kappa(u, v) = M(u, v)$.

Corollary 19.5.9.1. Graph G is n -connected iff every pair of vertices of G are connected by at least n internally disjoint paths.

Algorithm 19.5.2 (connectivity $\kappa(G)$). .

1. If degree of every vertex is $p - 1$, then output $\kappa = p - 1$ and stop. Otherwise, continue.
2. If G is disconnected, output $\kappa = 0$ and stop. Otherwise, continue.
3. $\kappa \leftarrow p$
4. $i \leftarrow 0$
5. If $i \leq \kappa$, then $i \leftarrow i + 1$ and continue. Otherwise, output κ and stop.
6. $j \leftarrow i + 1$
7. (1) If $j = p + 1$, then return to step 5. Otherwise continue.
 (2) If $v_i v_j \notin E(G)$, construct network N with digraph D as follows :
 for each vertex $v \in V(G)$, there are two vertices $v', v'' \in V(D)$ and an arc $(v', v'') \in E(D)$. And for each edge $uv \in E(G)$, there are two arcs $(u'', v), (v'', u) \in E(D)$. The capacity function is given by, $c(v', v'') = 1$ for every $v \in V(G)$ and $c(a) = \infty$ for every other arc in D . Set source $s = v_i''$ and sink $t = v_j'$ and find maximum flow in N using max-flow min-cut algorithm. Otherwise proceed to step 7d
 (3) If $f(N) < \kappa$, then $\kappa \leftarrow f(N)$. Otherwise, continue.
 (4) $j \leftarrow j + 1$ and return to step 7a

19.6 Matchings and Factorizations

19.6.1 An Introduction to Matching

Marriage Problem Given a collection of men and women, where each woman knows some of the men. Can every woman marry a man she knows ?

Assignment Problem Given several job openings and applicants for one or more of these positions. Find an assignment so that maximum positions are filled ?

Optimal Assignment Problem Given several job openings and applicants for one or more of these positions. The benefits of employing these applicants on those positions are also given. Find an assignment of maximum benefit to the company ?

matching in G is a 1-regular³ subgraph of G .

maximum matching in G is a matching of G with maximum cardinality.

perfect matching in G is a matching of cardinality $p/2$. ie, $p/2$ edges.

maximum weight matching in a weighted graph G is a matching with maximum weight.

Definitions 19.6.1. Let M be a matching in a graph G ,

matched edge is an edge in subgraph M of G .

unmatched edge is an edge of G that doesn't belong to M .

matched vertex with respect to M is a vertex incident with an edge of M .

single vertex is a vertex that is not incident with any edge of M .

alternating path in G is a path with edges alternately matched and unmatched.

augmenting path in G is a non-trivial alternating path with single vertices as end vertices.

Theorem 19.6.1. Let M_1, M_2 be two matchings in G such that there is a spanning subgraph H of G with edges that are either in M_1 or M_2 , but not both. Then the components of H are either 1. isolated vertex 2. even cycle with edge alternately from M_1 and M_2 3. a non-trivial path with edges alternately from M_1 and M_2 such that each end vertex is single with respect to either M_1 or M_2 , but not both.

Synopsis. $\Delta(H) \leq 2$ by Pigeonhole principle. Any component of H is either a path or a cycle. A cycle with edge alternately from M_1 and M_2 is even. If an end vertex of a non-trivial path is matched with respect to M_1 (WLOG), then it is there in $M_1 - M_2$ ie, it is not there in M_2 . If there is another edge in M_2 incident with it, then it has to be in H and it will cease to be an end vertex of path component. Therefore, it is unmatched with respect to M_2 .

³A graph G is k -regular, if every vertex of G has degree k .

Theorem 19.6.2. *A matching M in a graph G is maximum iff there is no augmenting path with respect to M in G .*

Synopsis. If M is maximum matching and P an M -augmenting path. Since both end-vertices are single, length of P is odd. Let M', M'' be edges of P which are in M and not in M respectively. Then $M - M' + M''$ is a matching of cardinality one greater than that of M which is a contradiction since M is maximum.

Conversely, suppose M be a matching such that there no M -augmenting paths in G . Let M' be a maximum matching in G . Then a nontrivial path component of the graph induced by $M \Delta M'$ is of even length otherwise both end-vertices are matched with respect to one of the matching M or M' which is a contradiction. Again every cycle components are even. Therefore $|M| = |M'|$, since $M \Delta M'$ doesn't have a nontrivial component of another kind.

Definitions 19.6.2. Let U_1, U_2 be two nonempty, disjoint, subsets of the vertex set of a graph G . Then U_1 is **matched to** U_2 if there exists a matching M in G such that every edge in M incident with a vertex in U_1 and a vertex in U_2 . And every vertex of U_1 (or U_2) is incident with some edge in M . Suppose M^* be a matching such that $M \subset M^*$, then U_1 is **matched under M^* to** U_2 .

Definitions 19.6.3. Let U be a nonempty set of vertices of a graph G . U is **nondeficient**,⁴ if $|N(S)| \geq |S|$ for every nonempty subset S of U .

Theorem 19.6.3. *Let G be a bipartite graph with partite sets V_1, V_2 . The set V_1 can be matched to a subset of V_2 iff V_1 is nondeficient.*

Corollary 19.6.3.1. *Every r -regular bipartite multigraph has a perfect matching.*

Theorem 19.6.4. *A collection S_1, S_2, \dots, S_n of finite non-empty sets has a system of distinct representatives iff for each k , $0 \leq k \leq n$, the union of any k of these sets contains at least k elements.*

Remark (Hall's Marriage Theorem). Suppose there are n women. Then every women can marry a man she knows iff each subset of k women ($1 \leq k \leq n$) collectively knows atleast k men.

Remark. Let W be a set of n women. Then there are $2^n - 1$ nonempty subset for W . Thus, Hall's Marriage Theorem suggests that we ensure $|N(S)| \geq |S|$ for every nonempty subset S of W . This method has complexity $O(2^n)$.

19.6.2 Maximum Matching in Bipartite Graphs

Definitions 19.6.4. Let M be a matching in a graph G and P is an augmenting path with respect to M . Let M' be set of edge in P and M . And M'' be the set of edges in P and not in M . Then $M_1 = (M - M') \cup M''$ is the **matching obtained by augmenting M along path P** .

Remark. $|M_1| = |M| + 1$

⁴ $N(S)$ is the neighbourhood set of all vertices adjacent to some vertex in S

Theorem 19.6.5. *Let M be a matching of a graph G that is not maximum, and let v be a single vertex with respect to M . Let M_1 denote the matching obtained by augmenting M along some augmenting path. If G contains an augmenting path with respect to M_1 that has v as an end-vertex, then G contains an augmenting path with respect to M that has v as an end-vertex.*

Corollary 19.6.5.1. *Let M be a matching of a graph G . Suppose that $M = M_1, M_2, \dots, M_k$ is a finite sequence of matchings of G such that M_i ($2 \leq i \leq k$) is obtained by augmenting M_{i-1} along some augmenting path. Suppose v is a single vertex with respect to M for which there exists no augmenting path starting at v . Then G does not contain an augmenting path with respect to M_i ($2 \leq i \leq k$) that has v as an end-vertex.*

Definitions 19.6.5. An **alternating tree** with respect to a matching M is a tree such that every path from its root is an alternating path with respect to M .

Algorithm 19.6.1 (Maximum Matching Algorithm for Bipartite Graphs). .

1. $i \leftarrow 1$ and $M \leftarrow M_1$
2. If $i < p$, then continue; otherwise stop.
3. If v_i is matched, then $i \leftarrow i + 1$ and return to Step 2; otherwise, $v \leftarrow v_i$ and Q is initialized to contain v only.
4. (1) For $j = 1, 2, \dots, p$ and $j \neq i$, let $TREE(v_j) \leftarrow F$. Also, $TREE(v_j) \leftarrow T$.
 (2) If $Q = \phi$, then $i \leftarrow i + 1$ and return to Step 2; otherwise, delete a vertex x from Q and continue.
 (3) (1) Suppose that $N(x) = \{y_1, y_2, \dots, y_k\}$. Let $j \leftarrow 1$.
 (2) If $j \leq k$, then $y \leftarrow y_j$; otherwise return to Step 4.2
 (3) If $TREE(y) = T$, then $j \leftarrow j + 1$ and return to Step 4.3.2. Otherwise, continue.
 (4) If y is incident with a matching edge yz , then $TREE(y) \leftarrow T$, $TREE(z) \leftarrow T$, $PARENT(y) \leftarrow x$, $PARENT(z) \leftarrow y$ and add z to Q , $j \leftarrow j + 1$ and return to Step 4.3.2. Otherwise, y is a single vertex and continue.
 (5) Use $PARENT$ to determine the alternating $v - x$ path P' in the alternating tree. Let P be the augmenting path obtained from P' by adding the path x, y . Proceed to Step 5
5. Augment M along P to obtain a new matching M' . Let $M \leftarrow M'$, $i \leftarrow i + 1$, and return to Step 2.

Definitions 19.6.6. Let G be a weighted complete bipartite graph with partite sets V_1 and V_2 . A **feasible vertex labeling** is a real function $l : V(G) \rightarrow \mathbb{R}$ on vertex set of G such that $l(v) + l(u) \geq w(vu)$ where $v \in V_1$ and $u \in V_2$.

Definitions 19.6.7. Consider the function $l : V(G) \rightarrow \mathbb{R}$ such that $\forall v \in V_1$, $l(v) = \max\{w(vu) : u \in V_2\}$ and $\forall u \in V_2$, $l(u) = 0$. Then l is a feasible vertex labeling on $V(G)$. And,

E_l is the set of all edge of the weighted complete bipartite graph G such that $l(v) + l(u) = w(vu)$.

H_l is the spanning subgraph of G induced by the edge set E_l .

Theorem 19.6.6. *Let l be a feasible vertex labeling of a weighted complete bipartite graph G . If H_l contains a perfect matching M' , then M' is a maximum weight matching of G .*

Algorithm 19.6.2 (Kuhn-Munkres). .

1. (1) For each $v \in V_1$, let $l(v) \leftarrow \max\{w(uv) : v \in V_2\}$.
 (2) For each $u \in V_2$, let $l(u) \leftarrow 0$.
 (3) Let H_l be the spanning subgraph of G with edge set E_l .
 (4) Let G_l be the underlying graph of H_l .
2. Apply Algorithm 19.6.1 to determine a maximum matching M in G_l .
3. (1) If every vertex v of V_1 is matching with respect to M , output M and stop. Otherwise, continue.
 (2) Let x be the first single vertex of V_1 .
 (3) Construct an alternating tree with respect to M that is rooted at x . If an augmenting path P is discovered, then augmenting M along P and return to Step 3.1. Otherwise, let T be the alternating tree with respect to M and rooted at x that cannot be expanded further in G_l .
4. Compute $m_l \leftarrow \min\{l(v) + l(u) - w(vu) : v \in V_1 \cap V(T), u \in V_2 - V(T)\}$.
 Let

$$l'(v) = \begin{cases} l(v) - m_l & \text{for } v \in V_1 \cap V(T) \\ l(v) + m_l & \text{for } v \in V_2 \cap V(T) \\ l(v) & \text{otherwise} \end{cases}$$

5. Let $l \leftarrow l'$, construct G_l and return to Step 3.3.

19.6.4 Factorizations

Definitions 19.6.8. A **factor** of a graph G is a spanning⁵ subgraph of G .

Definitions 19.6.9. Let G_1, G_2, \dots, G_n be edge-disjoint factors of G such that $E(G) = \cup_{i=1}^n E(G_i)$. Then G is **factorable** and $G = G_1 \oplus G_2 \oplus \dots \oplus G_n$.

Definitions 19.6.10. An r -regular factor of G is an **r-factor** of G .

Definitions 19.6.11. If G has a factorisation to r -factors, then G is **r-factorable**.

Remark. $K_{3,3}$ is 1-factorable. K_5 is 2-factorable.

Definitions 19.6.12. An **odd component of G** is a component of G with odd number of vertices. And an **even component of G** is a component of G of with even number of vertices.

⁵Spanning subgraph of a graph G has every vertex of G

Theorem 19.6.7 (Tutte). *A nontrivial graph G has a 1-factor iff for every proper subset S of $V(G)$, the number of odd components of $G - S$ does not exceed $|S|$.*

Remark. There exist cubic graphs that don't have a 1-factor.

Theorem 19.6.8 (Petersen). *Every bridgeless cubic graph contains a 1-factor.*

Remark. Every bridgeless cubic graph has a 1-factor. Let G be a bridgeless cubic graph. Consider every pair of factors G_1, G_2 such that $G = G_1 \oplus G_2$ where G_1 is a 1-factor and G_2 is a 2-factor. G is not 1-factorable only if every such G_2 doesn't have a 1-factor.

Theorem 19.6.9. *Petersen graph is not 1-factorable.*

Theorem 19.6.10. *Every r -regular bipartite multigraph ($r \geq 1$) is 1-factorable.*

Remark (Application of 1-factorisation). For even number p , a 1-factorisation of K_p corresponds to the schedule of a round of the round robin tournament among p teams. If p is odd, consider K_{p+1} where v_{p+1} is an imaginary team called bye team. A game with bye team is a bye.

Definitions 19.6.13. A **hamiltonian cycle** is a spanning cycle. And, **Hamiltonian graph** is a graph containing a hamiltonian cycle.

Theorem 19.6.11. *Complete graph K_{2n+1} can be factored into n hamiltonian cycles.*

Remark. For $n = 3$, K_7 can be factored into three hamiltonian cycles.

Theorem 19.6.12. *Let $0 \leq r < p$. Then there exists an r -regular graph of order p iff pr is even.*

Definitions 19.6.14. Let $\{E_1, E_2, \dots, E_n\}$ be partition of $E(G)$. And let H_i be subgraph of G induced by the edge set E_i . A **decomposition** of a graph G is a collection of these subgraphs H_1, H_2, \dots, H_n . And $G = H_1 \oplus H_2 \oplus \dots \oplus H_n$.

Definitions 19.6.15. Let $G = H_1 \oplus H_2 \oplus \dots \oplus H_n$ be a decomposition of G such that $H \cong H_i$. Then G is H -decomposable.

Remark. $K_{3,3}$ is $3K_2$ -decomposable. K_5 is C_5 -decomposable. K_{2n} is nK_2 -decomposable. K_{2n+1} is C_{2n+1} -decomposable.

Every graph is K_2 -decomposable. Every complete bipartite graph $K_{m,n}$ is $K_{1,m}$ -decomposable and $K_{1,n}$ -decomposable.

19.6.5 Block Designs

Definitions 19.6.16. A block design on a set V is a collection of k -element subsets of V such that each element of V appears exactly in r subsets.

variety The elements of V are called varieties.

block k -element subsets of V are called blocks.

balanced design If each variety appears in exactly r blocks and each pair of varieties appears in exactly λ blocks.

incomplete design If blocks are proper subsets of V . ie, $k < v$.

Definitions 19.6.17. A balanced incomplete block design of v varieties in b blocks of cardinality k such that each variety appears in exactly r blocks and each pair of varieties appears in exactly λ blocks is a (b, v, r, k, λ) -design.

Theorem 19.6.13. $bk = vr$

Theorem 19.6.14. $\lambda(v - 1) = r(k - 1)$

Corollary 19.6.14.1. $\lambda < r$

Theorem 19.6.15 (Fisher's Inequality). $b \geq v$

symmetric design If $b = v$

Theorem 19.6.16. In a symmetric (b, v, r, k, λ) -design with even v , $r - \lambda$ is a perfect square.

steiner triple system (b, v, r, k, λ) -design with $k = 3$, $\lambda = 1$.

Remark. (b, v, r, k, λ) -designs are incomplete. However, a complete block design (ie, $v = 3$) is also included as a Steiner triple system.

Theorem 19.6.17. Steiner triple system with v varieties exists iff $v = 6n + 1$ or $v = 6n + 3$ or $v = 3$.

Definitions 19.6.18 (Kirkman's Schoolgirls Problem). A class of 15 girls. Parade 15 girls in five rows (3 girls in a row). Is it possible to plan 7 days parade so that two girls are together in a row exactly once ?

kirkman triple system Steiner triple system with $v = 6n + 3$.

Remark. It is proved that kirkman triple system exists with $v = 6n + 3$ for every $n \geq 0$.

Theorem 19.6.18. The code consisting of the rows of the incidence matrix of a (b, v, r, k, λ) -design ($b = v$, $r=k$) is t -error correcting, where $t = k - \lambda - 1$.

Subject 20

ME800403 Combinatorics

20.1 Permutations and Combinations

20.1.1 Two Basic Counting Principles

“Who decides when to add/multiply numbers ?”

Definitions 20.1.1 (addition principle). Suppose event E can be partitioned into k pairwise disjoint events. And event E_j has n_j ways to occur where $j = 1, 2, \dots, k$. Then the number of ways for the event E to occur is $n_1 + n_2 + \dots + n_k$.

For example : Let E be the event of throwing a dice. Event E can be partitioned into two subevents E_1 and E_2 where E_1 is the event of getting an even number and E_2 is the event of getting an odd number. E_1 has 3 ways to occur and E_2 has 3 ways to occur. And E has 6 ways to occur.

Remark (Set Theoretical Equivalent). Suppose set A is disjoint union of the sets A_1, A_2, \dots, A_k where $A_i \cap A_j = \phi$, ($i \neq j$). Then,

$$|A| = \left| \bigcup_{j=1}^k A_j \right| = \sum_{j=1}^k |A_j|$$

Definitions 20.1.2 (multiplication principle). Suppose event E can be decomposed into k subevents. And event E_j has n_j ways to occur where $j = 1, 2, \dots, k$. Then number of ways for event E to occur is $n_1 \times n_2 \times \dots \times n_k$.

For example : Let E be the event throwing 3 coins. Event E can be decomposed into three events E_1 , E_2 and E_3 where E_j is the event of throwing j th coin. Each subevent has 2 ways to occur. And E has 8 ways to occur.

Remark (Set Theoretical Equivalent). Suppose $A = A_1 \times A_2 \times \dots \times A_k$. Then

$$|A| = \left| \prod_{j=1}^k A_j \right| = \prod_{j=1}^k |A_j|$$

Problems

1. The number of ordered pair (x, y) such that $x^2 + y^2 \leq 5$.
(hint : Addition Principle, disjoint Events E_j such that $x^2 + y^2 \leq j$,
($j = 0, 1, 2, \dots, 5$)) [Chen Chuan Chong, , Example 1.1.2]
2. The number of k -ary sequences of length n .
(hint : Multiplication Principle, subevents E_j of choosing j th term in the
sequence from $\{0, 1, \dots, (k-1)\}$) [Chen Chuan Chong, , Example 1.1.4]
3. The number of positive divisors of 600.
(hint : Multiplication Principle, $600 = 2^3 \times 3^1 \times 5^2$, subevents E_j of dif-
ferent powers of prime j dividing 600, ($j = 2, 3, 5$)) [Chen Chuan Chong, ,
Example 1.1.5]
4. The cardinality of $S = \{(a, b, c) : 0 \leq a, b, c \leq 100, a < b, a < c\}$.
(hint : Multiplication Principle, subevents E_a for each choice of a .) [Chen Chuan Chong, ,
Example 1.1.6]
5. The number of pairs $\{a, b\}$ such that $|a - b| = 5$.
(hint : Addition Principle, disjoint events E_j where $\min\{a, b\} = j$.)
[Chen Chuan Chong, , Exercise 1.1(i)]
6. The number of pairs $\{a, b\}$ such that $|a - b| \leq 5$.
(hint : Multiplication Principle, subevents E_k for $|a - b| = k$.) [Chen Chuan Chong, ,
Exercise 1.1(ii)]

20.1.2 Permutations

Definitions 20.1.3 (Permutation). The r -permutation from a set of n objects is the number of ways of arranging r objects out of n objects in a row.

Theorem 20.1.1. $P_r^n = n \times (n-1) \times \cdots \times (n+1-r) = \frac{n!}{(n-r)!}$

Proof. The event of r -permutation can be decomposed into r subevents. The subevents E_j 's are the event of placing an object from a set of $n+1-j$ objects in the j th position.

For example, E_1 has n ways to occur as we are placing an object from n objects in the first position and E_2 has $n-1$ ways to occur as we are placing an object from the remaining $n-1$ objects in the second position. Continuing like this, E_r has $(n+1-r)$ ways to occur as we are placing an object from $n+1-r$ objects in the r th position.

Clearly, each subevent E_j has $n+1-j$ ways to occur. By multiplication principle, E has $n \times (n-1) \times (n+1-r)$ ways to occur. Clearly, $P_r^n = n \times (n-1) \times \cdots \times (n+1-r)$. Multiplying both numerator and denominator with $(n-r)!$, we get $P_r^n = \frac{n!}{(n-r)!}$. \square

Remark. $0! = 1$

Problems

1. Number of five letter words where first and last letter are distinct vowels and remaining three letters are distinct consonants.
(hint: Multiplication Principle, Subevents E_v has P_2^5 ways to occur and E_c has P_3^{21} ways to occur.)

20.1.3 Circular Permutations

Definitions 20.1.4 (circular permutation). The number of r -circular permutations from a set of n objects, Q_r^n is the number of ways of arranging r objects out of n objects around a circle.

Theorem 20.1.2. $Q_r^n = P_r^n / r$

Proof. Let x_1, x_2, \dots, x_r be any r -circular permutation. Consider the following r -permutations obtained from the r -circular permutation by rotation.

- x_1, x_2, \dots, x_r
- $x_2, x_3, \dots, x_r, x_1$
- \vdots
- $x_r, x_1, x_2, \dots, x_{r-1}$

Each r -circular permutations represents a subset of r number of r -permutations. Clearly, the subsets corresponding to two different r -circular permutations are disjoint. Therefore, the number of r -circular permutations is P_r^n / r . \square

Remark. $Q_n^n = P_n^n / n = (n-1)!$.

Problems

1. Number of ways to seat 5 boys and 3 girls around a table is $Q_8^8 = 7!$.

Definitions 20.1.5 (Principle of Complementation). If event E has n ways to occur and its subevent E_1 has r ways to occur. Let E_2 be the complement event of E_1 . Then E_2 has $n - r$ ways to occur.

20.1.4 Combinations

Definitions 20.1.6. An r -combination of a set of n objects is a subset of r elements.

Theorem 20.1.3. The number of r -combinations of n objects $C_r^n = P_r^n / r!$.

Proof. Let E be the event of r -permutations of n objects. We may decompose this event into two subevents, event E_1 : the r -combinations of n objects and event E_2 : the r -permutations of those r objects. Thus, $P_r^n = C_r^n \times P_r^r$. Therefore, $C_r^n = P_r^n / P_r^r = P_r^n / r!$. \square

Remark. $C_r^n = \binom{n}{r} = \frac{n!}{r!(n-r)!}$.

Properties

1. $\binom{n}{r} = \binom{n}{n-r}$, by the symmetry of the factorial expression.
2. $\binom{n}{r} = \binom{n-1}{r-1} + \binom{n-1}{r}$.

Proof. Let event E be the r -combinations of n objects. Event E can be partitioned into two disjoint events, event E_1 : a particular element e should be there in the r -combination and event E_2 : element e should not be there in the r -combination. Clearly E_1 has $\binom{n-1}{r-1}$ ways to occur and E_2 has $\binom{n-1}{r}$ ways to occur. By addition principle, E has $\binom{n-1}{r-1} + \binom{n-1}{r}$ ways to occur. \square

Problem

1. Number of binary sequence of length seven with four 0's is $\binom{7}{4}$.
2. Number of ways to form a committee of 5 from a group of 11 is $\binom{11}{5}$.

Definitions 20.1.7. The **Stirling numbers of the first kind** $s(r, n)$ is the number of ways to distribute r distinct objects around n identical circles in such a way that each circle has at least one object.

Properties

1. $s(r, 0) = 0$ if $r \geq 1$, trivial as there is no table.
2. $s(r, r) = 1$ if $r \geq 0$, unique distribution as the tables are identical.
3. $s(r, 1) = (r-1)!$ is r -circular permutation.
4. $s(r, r-1) = \binom{r}{2}$ as some table has a 2-combination of r objects.
5. $s(r, n) = s(r-1, n-1) + s(r-1, n)s(r-1, n)$.

Proof. The event E of distributing r distinct objects around n identical circles can be partitioned into two disjoint events, event E_1 : a particular element e is the only element in its circle and event E_2 : element e is not alone in its circle. And event E_2 can be further decomposed into two subevents, event E_{21} : distributing $r-1$ objects around n identical circles and event E_{22} : placing element e on the immediate right of an element.

Clearly event E_1 has $s(r-1, n-1)$ ways to occur, event E_{21} has $s(r-1, n)$ ways to occur and event E_{22} has $r-1$ ways to occur. Thus, E has $s(r-1, n-1) + (r-1)s(r-1, n)$ ways to occur. \square

20.1.5 The Injection and Bijection Principles**Problems**

1. The number shortest routes an $m \times n$ street map is $\binom{m+n}{n}$.
2. The number of m -separated r -combinations of N_n is

20.1.6 Arrangements and Selections with Repetitions

20.1.7 Distribution Problems

20.2 Binomial Coefficients and Multinomial Coefficients*

20.3 The Pigeonhole Principle and Ramsey Numbers

20.3.1 Introduction

20.3.2 The Pigeonhole Principle

20.3.3 More Examples

20.3.4 Ramsey Type Problems and Ramsey Numbers

20.3.5 Bounds for Ramsey Numbers

20.4 The Principle of Inclusion and Exclusion

20.4.1 Introduction

20.4.2 The Principle

20.4.3 A Generalization

20.4.4 Integer Solutions and Shortest Routes

20.4.5 Surjective Mappings and Stirling Numbers of the Second Kind

20.4.6 Derangements and A Generalization

20.4.7 The Sieve of Eratosthenes and Euler ϕ -function*

20.4.8 The ‘Problème des Ménages’*

20.5 Generating Functions

20.5.1 Ordinary Generating Functions

20.5.2 Some Modelling Problems

20.5.3 Partitions of Integers

20.5.4 Exponential Generating Functions

20.6 Recurrence Relations

20.6.1 Introduction

20.6.2 Two Examples

20.6.3 Linear Homogeneous Recurrence Relations

20.6.4 General Linear Recurrence Relations

20.6.5 Two Applications

20.6.6 A System of Linear Recurrence Relations

20.6.7 The Method of Generating Functions

Subject 21

Probability Theory

Subject 22

Operational Research

Subject 23

Coding Theory

Subject 24

Commutative Algebra

Subject 25

Ordinary Differential Equations

Subject 26

Classical Mechanics

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