

MATH 20700 (Honors Analysis in \mathbb{R}^n I) Notes

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December 2, 2021

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Part I

Linear Algebra

Chapter 1

Basic Notions

1.1 Notes

- 9/27:
- **Vector space:** Basically, a set for which you have an addition and multiplication.
 - \mathbb{F}^d is used for \mathbb{R}^d or \mathbb{C}^d in Treil (2017).
 - \mathbb{P}_n is the vector space of polynomials up to degree n .
 - $C([0, 1])$ is the set of continuous functions defined on $[0, 1]$, an infinite-dimensional vector space.
 - **Generating set:** A subset of a vector space, all linear combinations of which generate the vector space. *Also known as spanning set.*
 - Any element of VS is a linear comb. of elements of the generating set.
 - **Linearly independent** (list): A list of vectors $\mathbf{v}_1, \dots, \mathbf{v}_k \in V$ such that $\sum_{i=1}^k \alpha_i \mathbf{v}_i = 0$ implies $\alpha_i = 0$ for all i .
 - **Base:** A generating set consisting of linearly independent vectors.
 - Any element of a VS can be written as a *unique* linear combination of the vectors in a base.
 - If $\mathbf{x} = \sum_{i=1}^k \alpha_i \mathbf{v}_i = \sum_{i=1}^k \beta_i \mathbf{v}_i$, then $\alpha_i = \beta_i$ for all i .
 - **Linear transformation:** A function $T : X \rightarrow Y$, where X, Y are VSs, such that

$$T(\alpha \mathbf{x} + \beta \mathbf{y}) = \alpha T\mathbf{x} + \beta T\mathbf{y}$$

for all $\mathbf{x} \in X, \mathbf{y} \in Y$.

- Examples of linear transformations:
 - Consider \mathbb{P}_n . Let $Tp_n = p'_n$. This T is linear.
 - Rotation in \mathbb{R}^d .
 - Think graphically about two vectors \mathbf{x}, \mathbf{y} .
 - Rotating and summing them is the same as summing and rotating. Same for scaling.
 - Thus, rotation is actually linear!
 - Reflection as well.
- Consider $T : \mathbb{R} \rightarrow \mathbb{R}$.
 - Any linear map on the line is a line.
 - We must have $Tx = \alpha x$: $Tx = T(1x) = xT(1) = x\alpha$.

- Consider $T : \mathbb{R}^n \rightarrow \mathbb{R}^m$ linear.
 - Any linear map between \mathbb{R}^n and \mathbb{R}^m is linear.
 - Thus, $T(\mathbf{x}) = A\mathbf{x}$ for all $\mathbf{x} \in \mathbb{R}^n$, where A is an $m \times n$ matrix.
- To find A , do the same calculation as for $Tx = \alpha x$ but more carefully:
 - Let $\{\mathbf{e}_1, \dots, \mathbf{e}_n\}$ be a basis.
 - So $\mathbf{x} = \sum_{i=1}^n \alpha_i \mathbf{e}_i$.
 - Thus, $T\mathbf{x} = \sum_{i=1}^n \alpha_i T(\mathbf{e}_i)$.
 - Each $T(\mathbf{e}_i)$ is part of the matrix that we multiply by the column vector representing \mathbf{x} .
- Multiplication of matrices is equivalent to composition of linear maps.
- Consider $T_1 : \mathbb{R}^n \rightarrow \mathbb{R}^m$ and $T_2 : \mathbb{R}^m \rightarrow \mathbb{R}^r$.
 - $T_2 \circ T_1$ is equivalent to BA , if A represents T_1 and B represents T_2 . In other words, $(T_2 \circ T_1)(\mathbf{x}) = BA\mathbf{x}$ for all \mathbf{x} .
- Recall that if $A = (\alpha_{ij})$ and $B = (\beta_{ij})$, then $(BA)_{ij} = (\sum \beta_{ik} \alpha_{kj})$.
- Properties of multiplication:

$$(AB)C = A(BC)$$

$$A(B + C) = AB + AC$$

$$(A + B)C = AC + BC$$

- However, it is not true in general that $AB = BA$.

- **Trace** (of an $n \times n$ matrix A): The sum of the diagonal entries of A . Denoted by $\text{tr}(A)$. Given by

$$\text{tr}(A) = \sum \alpha_{ii}$$

- It is true that $\text{tr}(AB) = \text{tr}(BA)$.
 - Indeed, on the diagonals, multiplication is commutative; it's the other terms that mess you up in general.
- Invertibility of matrices.
 - In general, matrices are not invertible: Not every system of equations is solveable; $Ax = b$ does not always have a solution $x = A^{-1}b$.
- C is the inverse from the left: $CA = I$. B is the inverse from the right: $AB = I$. A matrix can have a left and a right inverse and still not be invertible. A matrix is invertible iff $C = B$.
- Any time we write “inverse,” we do so under the assumption that it exists.
- $(AB)^{-1} = B^{-1}A^{-1}$ — easy proof by multiplication.
- If $A = (a_{ij})$, $A^T = (a_{ji})$.
 - $(A^{-1})^T = (A^T)^{-1}$.
 - $(AB)^T = B^T A^T$.
- Let X, Y VS.
 - $X \cong Y^{[1]}$ if there exists a linear $T : X \rightarrow Y$ that is one-to-one and onto.
 - Check: $A(\text{basis of } X) = \text{basis of } Y$. Prove by definition and expression of elements as linear combinations.
- **Subspace**: A subset of a vector space which happens to be a vector space, itself.

¹“ X is isomorphic to Y .”

1.2 Chapter 1: Basic Notions

From Treil (2017).

- 10/24:
- **Coordinates** (of $\mathbf{v} \in V$ wrt. a basis $\mathbf{v}_1, \dots, \mathbf{v}_n$ of V): The unique scalars $\alpha_1, \dots, \alpha_n \in \mathbb{F}$ such that $\mathbf{v} = \alpha_1 \mathbf{v}_1 + \dots + \alpha_n \mathbf{v}_n$.
 - **Spanning system**: A list of vectors that spans V . *Also known as **generating system**, **complete system**.*
 - **Trivial** (linear combination): A linear combination $\alpha_1 \mathbf{v}_1 + \dots + \alpha_n \mathbf{v}_n$ of vectors such that $\alpha_k = 0$ for each $k = 1, \dots, n$.
 - **Transformation**: A function $T : X \rightarrow Y$. *Also known as **transform**, **mapping**, **map**, **operator**, and **function**.*
 - The matrix of a linear transformation T is often denoted by $[T]$.
 - To compute the reflection of vectors over an arbitrary line through the origin in \mathbb{R}^2 , represent the overall transformation as a composition of rotating the line to be the x -axis, reflecting over the x -axis, and rotating back.
 - Theorem 1.5.1: If A is an $m \times n$ matrix and B is an $n \times m$ matrix, then

$$\text{tr}(AB) = \text{tr}(BA)$$

- Theorem 1.6.1: If a linear transformation is invertible, then its left and right inverses are unique and coincide.
- The column $(1, 1)^T$ is left-invertible, with one possible left inverse being $(1/2, 1/2)$.
 - Note that it is not right invertible since its left inverses are not unique (see Theorem 1.6.1).
- An invertible matrix must be square.
- **Isomorphic** (vector spaces): Two vectors spaces V, W such that there exists an isomorphism $A : V \rightarrow W$. *Denoted by $V \cong W$.*
- Theorem 1.6.8: $A : X \rightarrow Y$ is invertible if and only if for any right side $\mathbf{b} \in Y$, the equation

$$A\mathbf{x} = \mathbf{b}$$

has a unique solution $\mathbf{x} \in X$.

- Corollary 1.6.9: An $m \times n$ matrix is invertible if and only if its columns form a basis in \mathbb{F}^m .
- **Linear span** (of $\mathbf{v}_1, \dots, \mathbf{v}_n \in V$): The set of all linear combinations of $\mathbf{v}_1, \dots, \mathbf{v}_n$. *Denoted by $\mathcal{L}\{\mathbf{v}_1, \dots, \mathbf{v}_n\}$, $\text{span}(\mathbf{v}_1, \dots, \mathbf{v}_n)$.*

Chapter 2

Systems of Linear Equations

2.1 Notes

9/29: • Row elimination:

– Let

$$A = \begin{pmatrix} 1 & 2 & 3 & 1 \\ 3 & 1 & 2 & 7 \\ 2 & 1 & 2 & 1 \end{pmatrix}$$

– Then the **echelon form** matrix

$$A_e = \begin{pmatrix} 1 & 2 & 3 & 1 \\ 0 & 1 & 2 & -1 \\ 0 & 0 & 2 & -4 \end{pmatrix}$$

– Lastly, the **reduced echelon form** matrix

$$A_{re} = \begin{pmatrix} 1 & 0 & 0 & 7 \\ 0 & 1 & 0 & 3 \\ 0 & 0 & 1 & -2 \end{pmatrix}$$

- **echelon form:**

- All zero rows are below nonzero rows.
- For any nonzero row, its leading element is strictly to the left of the nonzero entry of the next row.

- **Reduced echelon form:**

- All pivots are 1.
- Used to solve systems of the form $Ax = b$.

- **Inconsistent** (system of equations): A system with no solution.

- If the last row is of the form $(0, \dots, 0, b)$ where $b \neq 0$, then there is no solution.

- Unique solution if A_e has a pivot in every column.

- There exists a solution for every b if there is a pivot in every row?

- Let $A : \mathbb{R}^n \rightarrow \mathbb{R}^m$ be a matrix. Then $\ker A = \{x \in \mathbb{R}^n : Ax = 0\}$ (subspace of \mathbb{R}^n) and $\text{range } A = \{Ax : x \in \mathbb{R}^n\}$ (subspace of \mathbb{R}^m).

- Also consider $\ker(A^T)$ and $\text{range}(A^T)$, the basis of the kernel and range, and dimension.

- Finite-dimensional vector spaces:

- A basis is a generating set (so every element of V can be written uniquely as a linear combination of the basis) the length of which is equal to the dimension of V .
- All bases of finite-dimensional vector spaces have the same number of elements.

- Let v_1, v_2, v_3 and w_1, w_2 be two generating sets of V .
- Then

$$v_1 = \lambda_{11}w_1 + \lambda_{12}w_2$$

$$v_2 = \lambda_{21}w_1 + \lambda_{22}w_2$$

$$v_3 = \lambda_{31}w_1 + \lambda_{32}w_2$$

- Suppose the only solution to $\alpha_1 v_1 + \alpha_2 v_2 + \alpha_3 v_3 = 0$ is $\alpha_1 = \alpha_2 = \alpha_3 = 0$.
- But this is not true, as we can find another one in terms of the λ s.
- If you have a list of linearly independent vectors, you can complete it into a basis.
 - If there exists a vector that can't be written as a linear combination of the list, add it to the list.
- If you find any particular solution to a system $Ax = b$, and you add to it any element of $\ker A$, you will obtain another solution.
 - $Ax_1 = b$ and $Ax_h = 0$ implies that $A(x_1 + x_h) = b$.
 - $Ax_1 = b$ and $Ax_2 = b$ imply that $A(x_1 - x_2) = 0$, i.e., that $x_1 - x_2 \in \ker A$.
- If $A : \mathbb{R}^n \rightarrow \mathbb{R}^m$ and $\dim \text{range } A = m$, then $Ax = b$ is solvable for all $b \in \mathbb{R}^m$.
- Let $\text{rank } A = \dim \text{range } A$.
- Rank theorem:
 - $\text{rank } A = \text{rank } A^T$.
 - Let $A : \mathbb{R}^n \rightarrow \mathbb{R}^m$. We know that $\dim \ker A + \dim \text{range } A = n$.
 - $\dim \ker A^T + \text{rank } A^T = m$.
 - This theorem survives linear algebra and enters functional analysis under the name **Fredholm's alternative**.

- **Fredholm's alternative:** $Ax = b$ has a solution for all $b \in \mathbb{R}^n$ iff $\dim \ker A^T = 0$.

- $\dim \ker A^T = 0$ implies $\text{rank } A^T = m$ implies $\text{rank } A = m$ implies $\dim \text{range } A = m$, as desired.

- **Pivot column** (of A): A column of A where A_e has pivots.
- The **pivot columns** of A give a basis for $\text{range } A$.
- The pivot rows of A_e give a basis for $\text{range } A^T$.
- A basis for the kernel is enough to solve $Ax = 0$.
- If you take these three things as givens, you can prove the rank theorem.

2.2 Chapter 2: Systems of Linear Equations

From Treil (2017).

- 10/24:
- A system is inconsistent iff the echelon form of the augmented matrix has a row of the form $(0 \ \cdots \ 0 \ b)$.
 - A solution to $Ax = b$ is unique iff there are no free variables, i.e., iff there is a pivot in every column.
 - $Ax = b$ is consistent iff the echelon form of the coefficient matrix has a pivot in every row.

- $A\mathbf{x} = \mathbf{b}$ has a unique solution for any \mathbf{b} iff the echelon form of the coefficient matrix A has a pivot in every row and column.
- Proposition 2.3.1: Let $\mathbf{v}_1, \dots, \mathbf{v}_m \in \mathbb{F}^n$, and let $A = [\mathbf{v}_1 \ \cdots \ \mathbf{v}_m]$ be an $n \times m$ matrix with columns $\mathbf{v}_1, \dots, \mathbf{v}_m$. Then
 1. The system $\mathbf{v}_1, \dots, \mathbf{v}_m$ is linearly independent iff the echelon form of A has a pivot in every column.
 2. The system $\mathbf{v}_1, \dots, \mathbf{v}_m$ is complete iff the echelon form of A has a pivot in every row.
 3. The system $\mathbf{v}_1, \dots, \mathbf{v}_m$ is a basis of \mathbb{F}^n iff the echelon form of A has a pivot in every column and in every row.
- Proposition 2.3.6: A matrix A is invertible if and only if its echelon form has a pivot in every column and every row.
- Corollary 2.3.7: An invertible matrix must be square ($n \times n$).
- Proposition 2.3.8: If a square ($n \times n$) matrix is left invertible or if it is right invertible, then it is invertible. In other words, to check the invertibility of a square matrix A , it is sufficient to check only one of the conditions $AA^{-1} = I$, $A^{-1}A = I$.
- Any invertible matrix is row-equivalent to (can be row-reduced to) the identity matrix.
- **Homogeneous** (system of linear equations): A system of the form $A\mathbf{x} = \mathbf{0}$.
- Theorem 2.6.1: Let a vector \mathbf{x}_1 satisfy the equation $A\mathbf{x} = \mathbf{b}$. and let H be the set of all solutions of the associated homogeneous system $A\mathbf{x} = \mathbf{0}$. Then the set

$$\{\mathbf{x}_1 + \mathbf{x}_h : \mathbf{x}_h \in H\}$$

is the set of all solutions to the equation $A\mathbf{x} = \mathbf{b}$.

- The pivot columns are a basis of range A . The pivot rows are a basis of range A^T . The solutions to the equation $A\mathbf{x} = \mathbf{0}$ are a basis of $\ker A$.
- Theorem 2.7.3: Let A be an $m \times n$ matrix. Then the equation $A\mathbf{x} = \mathbf{b}$ has a solution for every $\mathbf{b} \in \mathbb{R}^m$ iff the dual equation $A^T\mathbf{x} = \mathbf{0}$ has a unique (only the trivial) solution.
 - Note that this is a corollary to the rank theorem.
- Change of coordinates formula:
 - Let $T : V \rightarrow W$ be a linear transformation, and let $\mathcal{V} = \{\mathbf{v}_1, \dots, \mathbf{v}_n\}$ and $\mathcal{W} = \{\mathbf{w}_1, \dots, \mathbf{w}_n\}$ be bases of V and W , respectively.
 - The $m \times n$ matrix of T with respect to these bases is $[T]_{\mathcal{W}\mathcal{V}}$, and relates the coordinates of $[T\mathbf{v}]_{\mathcal{W}}$ and $[\mathbf{v}]_{\mathcal{V}}$ via

$$[T\mathbf{v}]_{\mathcal{W}} = [T]_{\mathcal{W}\mathcal{V}}[\mathbf{v}]_{\mathcal{V}}$$

- Change of coordinates matrix: If \mathcal{A}, \mathcal{B} are two bases of V , then we can convert the coordinates of a vector in \mathcal{B} to its in \mathcal{A} with the identity matrix (with respect to the appropriate bases). In particular,

$$[\mathbf{v}]_{\mathcal{B}} = [I]_{\mathcal{B}\mathcal{A}}[\mathbf{v}]_{\mathcal{A}}$$

- Note that the k^{th} column of $[I]_{\mathcal{B}\mathcal{A}}$ is the coordinate representation in \mathcal{B} of \mathbf{a}_k , i.e., $[\mathbf{a}_k]_{\mathcal{B}}$.
- The change of coordinates matrix from a basis \mathcal{B} to the standard basis \mathcal{S} is easy to compute; by the above, it's just

$$[I]_{\mathcal{S}\mathcal{B}} = [\mathbf{b}_1 \ \cdots \ \mathbf{b}_n]$$

- It follows that $[I]_{\mathcal{B}\mathcal{S}} = ([I]_{\mathcal{S}\mathcal{B}})^{-1}$.

- This allows us to compute $[I]_{\mathcal{B}\mathcal{A}}$ as $[I]_{\mathcal{B}\mathcal{S}}[I]_{\mathcal{S}\mathcal{A}}$
- If $T : V \rightarrow W$, $\mathcal{A}, \tilde{\mathcal{A}}$ are bases of V , and $\mathcal{B}, \tilde{\mathcal{B}}$ are bases of W , and we have $[T]_{\mathcal{B}\mathcal{A}}$, then

$$[T]_{\tilde{\mathcal{B}}\tilde{\mathcal{A}}} = [I]_{\tilde{\mathcal{B}}\mathcal{B}}[T]_{\mathcal{B}\mathcal{A}}[I]_{\mathcal{A}\tilde{\mathcal{A}}}$$

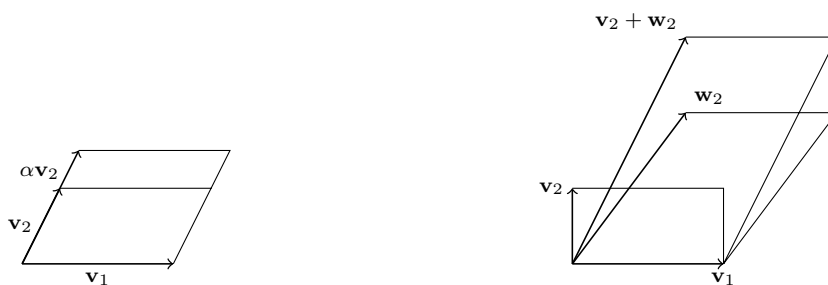
- Change of basis ends up at similarity; two operators are similar if we can change the basis of one into another.

Chapter 3

Determinants

3.1 Notes

- 9/29:
- The determinant, geometrically, is the volume of the object (in \mathbb{R}^3) you get when you take linear combinations of the vectors.
 - In 2D:
 - Let v_1, v_2 be two vectors. Put tail to tail and forming a parallelogram, the determinant of the matrix (v_1, v_2) is the area of said parallelogram.
 - Linearity 1: $D(av_1, v_2, \dots, v_n) = aD(v_1, \dots, v_n)$ is the same as saying that if you stretch one vector by a , you scale up the area by that much, too.
 - Linearity 2: $D(v_1, \dots, v_{k+} + v_{k-}, \dots, v_n) = D(-) + D(+)$.
 - Antisymmetry: $D(v_1, \dots, v_k, \dots, v_j, \dots, v_n) = -D(v_1, \dots, v_j, \dots, v_k, \dots, v_n)$. Interchanging columns flips the sign of the determinant.
 - Basis: $D(e_1, \dots, e_n) = 1$.
 - Determinant: Denoted by $D(v_1, \dots, v_n)$, where (v_1, \dots, v_n) is an $n \times n$ matrix.
- 10/1:
- Consider an $n \times n$ matrix A consisting of n columns containing vectors $\mathbf{v}_1, \dots, \mathbf{v}_n \in \mathbb{R}^n$.
 - $D(A)$ is the volume of the solid $V = \sum_{i=1}^n \alpha_i v_i$.
 - $D(\mathbf{e}_1, \dots, \mathbf{e}_n) = 1$.



(a) $D(\mathbf{v}_1, \alpha \mathbf{v}_2) = \alpha D(\mathbf{v}_1, \mathbf{v}_2)$.

(b) $D(\mathbf{v}_1, \mathbf{v}_2 + \mathbf{w}_2) = D(\mathbf{v}_1, \mathbf{v}_2) + D(\mathbf{v}_1, \mathbf{w}_2)$.

Figure 3.1: Visualizing properties of determinants.

- Basic properties of the determinant.
 - If A has a zero column, then $\det A = 0$: Scalar property.

- If A has two equal columns, then $\det A = 0$: Multiply one by minus and add.
- If A has a column which is a multiple of another, then $\det A = 0$: Pull out the multiple and then you have the previous one.
- If columns are linearly dependent, then $\det A = 0$: Decompose it into sums, split, add back up with previous properties.
- The determinant is preserved under column reduction.
- $\det A^T = \det A$: Put everything in rref.
- If A is not invertible, then $\det A = 0$ (not invertible implies linearly dependent columns, implies $\det A = 0$).
- $\det(AB) = \det A \det B$.
- Determinant of...
 - A diagonal matrix: The product of the diagonal entries (pull out the terms, and then note that the remaining identity matrix has determinant 1).
 - An upper triangular matrix: The product of the diagonal entries (column reduction to make it into a diagonal matrix, and then the property above).

3.2 Chapter 3: Determinants

From Treil (2017).

- 10/24: • Let $A_{j,k}$ denote the $(n-1) \times (n-1)$ matrix obtained from A by crossing out row j and column k and pushing it together.

- **Cofactors** (of A): The numbers $C_{j,k}$, one per entry, defined by

$$C_{j,k} = (-1)^{j+k} \det A_{j,k}$$

- **Cofactor matrix** (of A): The matrix

$$C = \{C_{j,k}\}_{j,k=1}^n$$

- Theorem 3.5.2: Let A be an invertible matrix and let C be its cofactor matrix. Then

$$A^{-1} = \frac{1}{\det A} C^T$$

- **Cramer's rule**: If A is invertible and $A\mathbf{x} = \mathbf{b}$, then

$$x_k = \frac{\det B_k}{\det A}$$

where B_k is obtained from A by replacing column k of A by the vector \mathbf{b} .

- **Minor** (of order k of A): The determinant of a $k \times k$ submatrix of A .
- Theorem 3.6.1: The rank of a nonzero matrix A is equal to the largest integer k such that there exists a nonzero minor of order k .

Chapter 4

Introduction to Spectral Theory

4.1 Notes

- 10/1:
- **Difference equation:** Like a differential equation, but instead of writing a differentials, you write differences.
 - Suppose we want to solve $x_{n+1} = Ax_n$ with x_0 given.
 - You will find that $x_n = A^n x_0$.
 - This gets hard to compute, so we want to find a way to simplify the computation.
 - Thus, we want to diagonalize the matrix, and this concept is inherently linked to eigenvalues and eigenvectors.
 - If you can decompose the x_0 into a linear combination of eigenvectors, then you can simplify the computation a lot:
$$x_n = \sum \alpha_i A^n v_i = \sum \alpha_i \lambda_i^n v_i$$
 - An $n \times n$ matrix will have n eigenvalues. You want n linearly independent eigenvectors, creating an eigenbasis.
 - To find eigenvalues and eigenvectors, we need to solve $Ax = \lambda x$, i.e., $(A - \lambda I)x = 0$. Thus, $\ker(A - \lambda I) \neq \{0\}$, so $\det(A - \lambda I) = 0$.
 - The eigenvalues of A are independent of the choice of basis of the domain of A or the range.
- 10/4:
- We need to know everything in Treil (2017).
 - We don't need to know the applications sections, but you should be interested.
 - **Spectral theory:** Decomposing a linear operator.
 - Let $A : V \rightarrow V$ be a linear operator. $\lambda \in \mathbb{C}$ is an eigenvalue if there exists $x \in V$ nonzero such that $Ax = \lambda x$.
 - Let A be an $n \times n$ matrix over \mathbb{C} or \mathbb{R} .
 - The eigenvalues are the roots of the polynomial $\det(A - \lambda I) = 0$ in λ .
 - Things we want to do:
 - Given A , find the eigenvalues and eigenvectors (solve $(A - \lambda I)x = 0$).

- In order to simplify A , make it a diagonal matrix:

$$A = S \begin{pmatrix} \lambda_1 & & 0 \\ & \ddots & \\ 0 & & \lambda_n \end{pmatrix} S^{-1}$$

- Eigenvalues are independent of the choice of basis.

- From the book, we have that

$$[A]_{\mathcal{A}\mathcal{A}} = [S]_{\mathcal{A}\mathcal{B}}[B]_{\mathcal{B}\mathcal{B}}[S]_{\mathcal{A}\mathcal{B}}^{-1}$$

- It follows that

$$A - \lambda I = [S]_{\mathcal{A}\mathcal{B}}(B - \lambda I)[S]_{\mathcal{A}\mathcal{B}}^{-1}$$

so

$$\det(A - \lambda I) = \det([S]_{\mathcal{A}\mathcal{B}}(B - \lambda I)[S]_{\mathcal{A}\mathcal{B}}^{-1}) = \det([S]_{\mathcal{A}\mathcal{B}}[S]_{\mathcal{A}\mathcal{B}}^{-1}(B - \lambda I)) = \det(B - \lambda I)$$

- If $p(z) = (z - \lambda)^k q(z)$, then k is the **algebraic multiplicity** of λ . The **geometric multiplicity** of λ is $\dim \ker(A - \lambda I)$.

- These terms are not always the same, but they are related.

- Diagonalization:

- Given A that corresponds to $T : V \rightarrow V$, can we find a basis of V in which the operator is a diagonal matrix?
- $A = SDS^{-1}$ iff there exists a basis of V consisting of the eigenvectors of A .
- Proves $A^N = SD^N S^{-1}$ via $A^2 = SDS^{-1}SDS^{-1} = SDIDS^{-1} = SD^2 S^{-1}$.

- Let A be an $n \times n$ matrix over \mathbb{F} . If $\lambda_1, \dots, \lambda_r$ are distinct eigenvalues, then their eigenvectors are linearly independent.

- Prove with induction contradiction argument. Assume true for \mathbf{v}_{r-1} . Then

$$0 = (A - \lambda_r I)[\mathbf{v}_1 + \dots + \mathbf{v}_r] = (\lambda_1 - \lambda_r)\mathbf{v}_1 + \dots + (\lambda_{r-1} - \lambda_r)\mathbf{v}_{r-1}$$

- Implies $\lambda_r = \lambda_i$ for all $i \in [r-1]$, a contradiction.
- If A has n distinct eigenvalues, then A is diagonalizable.

- If $A : V \rightarrow V$ has n complex eigenvalues, then A is diagonalizable iff the algebraic multiplicity equals the geometric multiplicity for each eigenvalue.

- Goes through a sample diagonalization with $\begin{pmatrix} 1 & 2 \\ 8 & 1 \end{pmatrix}$.

- We have

$$A - \lambda I = \begin{pmatrix} 1 - \lambda & 2 \\ 8 & 1 - \lambda \end{pmatrix}$$

so

$$0 = \det(A - \lambda I) = (1 - \lambda)^2 - 16$$

- It follows that $\lambda = 5, -3$.
- This yields

$$\begin{pmatrix} 1 & 2 \\ 8 & 1 \end{pmatrix} = \begin{pmatrix} 1 & 1 \\ 2 & -2 \end{pmatrix} \begin{pmatrix} 5 & 0 \\ 0 & -3 \end{pmatrix} \begin{pmatrix} 1 & 1 \\ 2 & -2 \end{pmatrix}^{-1}$$

by inspection.

- As another example, consider $\begin{pmatrix} 1 & 2 \\ -2 & 1 \end{pmatrix}$.

- Here, we have $\lambda = 1 \pm 2i$.

4.2 Chapter 4: Introduction to Spectral Theory

From Treil (2017).

10/24:

- **Spectrum** (of A): The set of all eigenvalues of A . Denoted by $\sigma(A)$.
- Proposition 4.1.1: The geometric multiplicity of an eigenvalue cannot exceed its algebraic multiplicity.
- Theorem 4.2.1: A matrix A (with values in \mathbb{F}) admits a representation $A = SDS^{-1}$ where D is a diagonal matrix and S is invertible if and only if there exists a basis of \mathbb{F}^n of eigenvectors of A . Moreover, in this case diagonal entries of D are the eigenvalues of A and columns of S are the corresponding eigenvectors.
- Any operator on a complex vector space has n eigenvalues (counting multiplicities).
 - Think n necessary roots of the characteristic polynomial, or the necessary upper triangular representation.
- Theorem 4.2.8: Let an operator $A : V \rightarrow V$ have exactly $n = \dim V$ eigenvalues (counting multiplicities). Then A is diagonalizable if and only if for each eigenvalue λ , the dimension of the eigenspace $\ker(A - \lambda I)$ (i.e., the geometric multiplicity of λ) coincides with the algebraic multiplicity of λ .
- Theorem 4.2.9: A real $n \times n$ matrix A admits a real factorization (i.e., a real representation $A = SDS^{-1}$ where S and D are real matrices, D is diagonal, and S is invertible) if and only if it admits a complex factorization and all eigenvalues of A are real.
- Example of a nondiagonalizable matrix:

$$A = \begin{pmatrix} 1 & 1 \\ 0 & 1 \end{pmatrix}$$

- $p(\lambda) = (1 - \lambda)^2$, so $\lambda = 1$ with algebraic multiplicity 2.
- However, $\dim \ker(A - I) = 1$ since $A - I$ has only one pivot, hence $2 - 1 = 1$ free variable.
- Thus, apply Theorem 4.2.8.

Chapter 5

Inner Product Spaces

5.1 Notes

10/6: • We define

$$\ell^2(\mathbb{R}) = \left\{ \{a_n\}_{n \geq 1} \subset \mathbb{R} : \sum_1^\infty |a_n|^2 < \infty \right\}$$

• **Inner product:** A map $V \times V \rightarrow \mathbb{F}$ that takes $(\mathbf{x}, \mathbf{y}) \mapsto \mathbf{x} \cdot \mathbf{y}$. Denoted by $\cdot, (\cdot, \cdot), \langle \cdot, \cdot \rangle$.

• Properties of the inner product:

- $(\mathbf{x}, \mathbf{y}) = \overline{(\mathbf{y}, \mathbf{x})}$ (symmetry).
- $(\alpha \mathbf{x} + \beta \mathbf{y}, \mathbf{z}) = \alpha(\mathbf{x}, \mathbf{z}) + \beta(\mathbf{y}, \mathbf{z})$ (linearity).
- $(\mathbf{x}, \mathbf{x}) \geq 0$.
- $(\mathbf{x}, \mathbf{x}) = 0$ iff $\mathbf{x} = 0$.

• If $\mathbf{x}, \mathbf{y} \in \mathbb{R}^n$, then

$$(\mathbf{x}, \mathbf{y}) = \sum_{i=1}^n x_i y_i$$

• If $\mathbf{x}, \mathbf{y} \in \mathbb{C}^n$, then

$$(\mathbf{x}, \mathbf{y}) = \sum_{i=1}^n x_i \bar{y}_i$$

• If $f, g \in \mathbb{P}_n(t)$, then

$$(f, g) = \int_{-1}^1 f \bar{g} dt$$

- The conjugate of a polynomial is the polynomial with the conjugate of the coefficients of the original polynomial. Symbolically, if $f = \sum_{i=0}^n \alpha_i t^i$ is a polynomial, then $\bar{f} = \sum_{i=0}^n \bar{\alpha}_i t^i$.

• It is a fact that

$$\left| \sum_{n=1}^\infty a_n \bar{b}_n \right| \leq \| (a_n)_{n \geq 1} \| \| (b_n)_{n \geq 1} \|$$

• Suppose we want to define the inner product between two matrices.

- A common one is

$$(A, B) = \text{tr}(B^* A)$$

where $B^* = \bar{B}^T = \overline{B^T}$ is the conjugate transpose.

- We define the norm as a function $V \rightarrow [0, \infty)$ given by

$$\|\mathbf{x}\| = \sqrt{(\mathbf{x}, \mathbf{x})}$$

- Properties of the norm.

- $\|\alpha\mathbf{x}\| = |\alpha|\|\mathbf{x}\|$.
- $\|\mathbf{x} + \mathbf{y}\| \leq \|\mathbf{x}\| + \|\mathbf{y}\|$.
- $\|\mathbf{x}\| = 0$ iff $\mathbf{x} = 0$.

- In \mathbb{R}^n ,

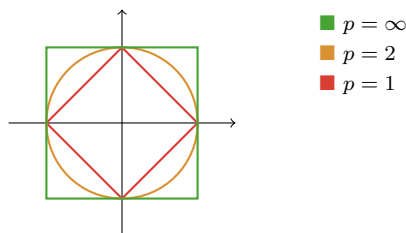


Figure 5.1: The unit ball of norms corresponding to $p = 1, 2, \infty$.

- The standard norm is

$$\|\mathbf{x}\| = \sqrt{\sum |x_i|^2}$$

- We can also define

$$\|\mathbf{x}\|_p = \sqrt[p]{\sum |x_i|^p}$$

- We can even define

$$\|\mathbf{x}\|_\infty = \max |x_i|$$

- And we can prove that all of these are valid norms.
- Only the norm corresponding to ℓ^2 is given by an inner product, but all the other quantities are still norms as defined by the properties (see Treil (2017)).
- Figure 5.1 shows the unit ball of each norm, i.e., the set of all points which have norm 1.

- The parallelogram rule:

$$\|\mathbf{x} + \mathbf{y}\|^2 + \|\mathbf{x} - \mathbf{y}\|^2 = 2(\|\mathbf{x}\|^2 + \|\mathbf{y}\|^2)$$

- Orthogonality: Given \mathbf{v}, \mathbf{w} , if $\mathbf{v} \perp \mathbf{w}$, then $(\mathbf{v}, \mathbf{w}) = 0$.

- In particular, if $\mathbf{v} \perp \mathbf{w}$, then

$$\|\mathbf{v} + \mathbf{w}\|^2 = \|\mathbf{v}\|^2 + \|\mathbf{w}\|^2$$

- Let E be a subspace of V . If $\mathbf{v} \perp E$, then $\mathbf{v} \perp \mathbf{e}$ for all $\mathbf{e} \in E$, i.e., $\mathbf{v} \perp$ a set of vectors spanning E .
- Any set of orthogonal vectors is linearly independent. Thus, if V is n dimensional, then $\mathbf{v}_1, \dots, \mathbf{v}_n$ orthogonal is a basis.
- Let E be a subspace of V . Take $\mathbf{v} \in V$. We want to define the projection $P_E \mathbf{v}$ of \mathbf{v} onto E .
 - We have that $P_E \mathbf{v} \in E$ and $\mathbf{v} - P_E \mathbf{v} \perp E$.
 - Additionally, we have that

$$\|\mathbf{v} - P_E \mathbf{v}\| \leq \|\mathbf{v} - \mathbf{e}\|$$

for all $\mathbf{e} \in E$.

- Lastly, we have that $P_E \mathbf{v}$ is unique.
- If we receive a basis of a vector space, how do we create out of that a basis that is orthogonal? The process of doing this is called **Gram-Schmidt orthogonalization**.
 - We keep \mathbf{v}_1 , subtract $P_{\mathbf{v}_1} \mathbf{v}_2$ from \mathbf{v}_2 , subtract $P_{\{\mathbf{v}_1, \mathbf{v}_2\}} \mathbf{v}_3$ from \mathbf{v}_3 , and on and on.
- If we are given a set of orthogonal vectors, we can normalize them by dividing each by its norm. This creates an orthonormal list. The standard basis is orthonormal.
- Let

$$E^\perp = \{v \in V : v \perp E\}$$

- It follows that $V = E \oplus E^\perp$.
- How close can we come to solving $A\mathbf{x} = \mathbf{b}$ if we cannot solve it exactly (i.e., if the columns are not linearly independent)?
 - Let A be an $m \times n$ matrix, and let $\mathbf{b} \in \mathbb{R}^m$.
 - Then the best solution is given by minimizing $\|A\mathbf{x} - \mathbf{b}\|$. We minimize this with projections. A special case of this is least squares regression! More details in Treil (2017).

10/8:

- Soug is gonna send us a hefty amount of reading for the weekend.
- Least square approximation:
 - If we want to minimize $\|A\mathbf{x} - \mathbf{b}\|$, the best we can do is project \mathbf{b} onto the range of A .
 - Let $\mathbf{v}_1, \dots, \mathbf{v}_k$ be an orthogonal basis of range A .
 - Then

$$\text{Proj}_{\text{range } A} \mathbf{b} = \sum_{k=1}^k \frac{(\mathbf{b}, \mathbf{v}_k)}{\|\mathbf{v}_k\|^2} \mathbf{v}_k$$

- Matrix equation form:

$$\text{Projection}_{\text{range } A} = A(A^*A)^{-1}A^*$$

if A^*A is invertible, where $A^* = \bar{A}^T$.

■ Soug never uses this though.

- The minimum is found when $\mathbf{b} - A\mathbf{x} \perp \text{range } A$. Implies that $\mathbf{b} - A\mathbf{x} \perp \mathbf{a}_k$ for all k . Implies $(\mathbf{b} - A\mathbf{x}, \mathbf{a}_k) = \bar{\mathbf{a}}_k^T (\mathbf{b} - A\mathbf{x}) = 0$.
- Note that we're letting $\bar{\mathbf{a}}_k^T$ be the row vector

$$\bar{\mathbf{a}}_k^T = (\bar{a}_{1,k} \quad \cdots \quad \bar{a}_{n,k})$$

- We also have $\bar{A}^T (\mathbf{b} - A\mathbf{x}) = 0$, from which it follows that $A^*A\mathbf{x} = A^*\mathbf{b}$, so $\mathbf{x} = (A^*A)^{-1}A^*\mathbf{b}$. Thus, $\text{Proj}_{\text{range } A} = Ax$, so $\text{Proj}_{\text{range } A} = A(A^*A)^{-1}A^*$.
- Adjoint of a linear map $T : V \rightarrow W$ is the A^* discussed above.
 - First, we'll do this for matrices. And then we'll do it for any finite-dimensional vector space.
 - Let A be an $m \times n$ matrix. We claim then that

$$(A\mathbf{x}, \mathbf{y}) = (\mathbf{x}, A^*\mathbf{y})$$

for all $\mathbf{x} \in \mathbb{C}^n, \mathbf{y} \in \mathbb{C}^m$. Proof:

$$\begin{aligned} (A\mathbf{x}, \mathbf{y}) &= \bar{\mathbf{y}}^T A\mathbf{x} \\ &= \mathbf{y}^* A\mathbf{x} \\ &= (A^*\mathbf{y})^* \mathbf{x} \\ &= (\mathbf{x}, A^*\mathbf{y}) \end{aligned}$$

- Properties of the adjoint:

$$(AB)^T = B^T A^T$$

$$(AB)^* = B^* A^*$$

$$(A^*)^* = A$$

- A^* is the unique matrix B such that $(A\mathbf{x}, \mathbf{y}) = (\mathbf{x}, B\mathbf{y})$.
- Let $\mathbf{v}_1, \dots, \mathbf{v}_n$ be a basis of V , and let $\mathbf{w}_1, \dots, \mathbf{w}_m$ be a basis of W .
- Definition of A^* : If $(A\mathbf{x}, \mathbf{y}) = (y, A^*\mathbf{x})$ for all $\mathbf{x} \in V$ and $\mathbf{y} \in W$.
- But it's not enough to define something; we have to check that it exists.
- If $[A]_{AB}$, then $[A^*]_{AB}$.
- More properties (give criteria for solving systems of equations):

$$\ker A^* = (\text{range } A)^\perp$$

$$\ker A = (\text{range } A^*)^\perp$$

$$\text{range } A = (\ker A^*)^\perp$$

$$\text{range } A^* = (\ker A)^\perp$$

■ Soug proves these.

• Isometries and unitary operators.

- $U : X \rightarrow Y$ is an isometry if $\|\mathbf{x}\| = \|U\mathbf{x}\|$ for all $\mathbf{x} \in X$. It is an isometry because it preserves the distance between points.
- It immediately follows that $\|\mathbf{x}_1 - \mathbf{x}_2\| = \|U\mathbf{x}_1 - U\mathbf{x}_2\| = \|U(\mathbf{x}_1 - \mathbf{x}_2)\|$.
- This definition is equivalent to an inner product one: $(\mathbf{x}, \mathbf{y}) = (U\mathbf{x}, U\mathbf{y})$. This follows from the definition of the norm.
- We have

$$(\mathbf{a}, \mathbf{b}) = \frac{1}{4} \sum_{\alpha=\pm 1, \pm i} \alpha \|\mathbf{a} + \alpha \mathbf{b}\|^2$$

■ $(a+b)^2 - (a-b)^2 = 4ab$ for any $a, b \in \mathbb{R}$, so $ab = \frac{1}{4}[(a+b)^2 - (a-b)^2]$. Thus, in a real inner product space,

$$(\mathbf{a}, \mathbf{b}) = \frac{1}{4} \left(\|\mathbf{a} + \mathbf{b}\|^2 - \|\mathbf{a} - \mathbf{b}\|^2 \right)$$

■ It follows that isometries preserve inner products.

- U is an isometry if and only if $U^*U = I$. Proof:

$$(\mathbf{x}, \mathbf{x}) = (U\mathbf{x}, U\mathbf{y}) = (U^*U\mathbf{x}, \mathbf{x})$$

$$(\mathbf{x}, \mathbf{y}) = (U\mathbf{x}, U\mathbf{y}) = (U^*U\mathbf{x}, \mathbf{y})$$

$$(\mathbf{x}, \mathbf{y}) = (U\mathbf{x}, U\mathbf{y}) = (\mathbf{x}, \mathbf{y})$$

for all \mathbf{y} .

- An isometry is unitary if it is invertible.

■ Thus, $U : X \rightarrow Y$ an isometry is unitary iff $\dim X = \dim Y$.

- Note that it follows that $U^* = U^{-1}$ for U an isometry.
- U unitary implies $|\det U| = 1$, so λ an eigenvalue of U implies that $|\lambda| = 1$.
- A is diagonalizable iff it has an orthogonal basis of eigenvectors.

5.2 Chapter 5: Inner Product Spaces

From Treil (2017).

- 10/24: • **Standard inner product** (on \mathbb{C}^n): The inner product (\mathbf{z}, \mathbf{w}) defined by

$$(\mathbf{z}, \mathbf{w}) = \mathbf{w}^* \mathbf{z}$$

- Corollary 5.1.5: Let \mathbf{x}, \mathbf{y} be vectors in an inner product space V . The equality $\mathbf{x} = \mathbf{y}$ holds if and only if

$$(\mathbf{x}, \mathbf{z}) = (\mathbf{y}, \mathbf{z})$$

for all $\mathbf{z} \in V$.

- Corollary 5.1.6: Suppose two operator $A, B : X \rightarrow Y$ satisfy

$$(A\mathbf{x}, \mathbf{y}) = (B\mathbf{x}, \mathbf{y})$$

for all $\mathbf{x} \in X$ and $\mathbf{y} \in Y$. Then $A = B$.

- **Normed space:** A vector space V equipped with a norm that satisfies properties of homogeneity, the triangle inequality, non-negativity, and non-degeneracy.
- Any inner product space is naturally a normed space.
- If $1 \leq p < \infty$, we can define a corresponding norm on \mathbb{R}^n or \mathbb{C}^n by

$$\|\mathbf{x}\|_p = \left(\sum_{k=1}^n |x_k|^p \right)^{1/p}$$

- We can also define the norm for $p = \infty$ by

$$\|\mathbf{x}\|_\infty = \max\{|x_k| : k = 1, \dots, n\}$$

- Note that the norm of this form for $p = 2$ is the usual norm.
- These norms are heavily associated with Figure 5.1.

- **Minkowski inequality:** One of the triangle inequalities for norms with $p \neq 2$.
- Theorem 5.1.11: A norm in a normed space is obtained from some inner product if and only if it satisfies the Parallelogram Identity

$$\|\mathbf{u} + \mathbf{v}\|^2 + \|\mathbf{u} - \mathbf{v}\|^2 = 2(\|\mathbf{u}\|^2 + \|\mathbf{v}\|^2)$$

for all $\mathbf{u}, \mathbf{v} \in V$.

- It follows that norms with $p \neq 2$ do not have associated inner products, since such norms fail to satisfy the parallelogram identity.

- Lemma 5.2.5 (Generalized Pythagorean Identity): Let $\mathbf{v}_1, \dots, \mathbf{v}_n$ be an orthogonal system. Then

$$\left\| \sum_{k=1}^n \alpha_k \mathbf{v}_k \right\|^2 = \sum_{k=1}^n |\alpha_k|^2 \|\mathbf{v}_k\|^2$$

- Proposition 5.3.3: Let $\mathbf{v}_1, \dots, \mathbf{v}_r$ be an orthogonal basis in E . Then the orthogonal projection $P_E \mathbf{v}$ of a vector \mathbf{v} is given by the formula

$$P_E \mathbf{v} = \sum_{k=1}^r \frac{(\mathbf{v}, \mathbf{v}_k)}{\|\mathbf{v}_k\|^2} \mathbf{v}_k$$

– It follows that

$$\begin{aligned} P_E \mathbf{v} &= \sum_{k=1}^r \frac{\mathbf{v}_k^* \mathbf{v}}{\|\mathbf{v}_k\|^2} \mathbf{v}_k \\ &= \sum_{k=1}^r \frac{1}{\|\mathbf{v}_k\|^2} \mathbf{v}_k \mathbf{v}_k^* \mathbf{v} \\ &= \left(\sum_{k=1}^r \frac{1}{\|\mathbf{v}_k\|^2} \mathbf{v}_k \mathbf{v}_k^* \right) \mathbf{v} \end{aligned}$$

– Thus, we have that

$$P_E = \sum_{k=1}^r \frac{1}{\|\mathbf{v}_k\|^2} \mathbf{v}_k \mathbf{v}_k^*$$

- **Gram-Schmidt orthogonalization:** Let $\mathbf{x}_1, \dots, \mathbf{x}_n$ be a linearly independent system of vectors to orthogonalize. Then $\mathbf{v}_1 = \mathbf{x}_1$, $\mathbf{v}_2 = \mathbf{x}_2 - P_{\text{span}\{\mathbf{v}_1\}} \mathbf{x}_2$, $\mathbf{v}_3 = \mathbf{x}_3 - P_{\text{span}\{\mathbf{v}_1, \mathbf{v}_2\}} \mathbf{x}_3$, and on and on.
- To find the least squares solution to $A\mathbf{x} = \mathbf{b}$, solve $A\mathbf{x} = P_{\text{range } A} \mathbf{b}$.
 - We can do this by finding an orthogonal basis of range A and then applying the projection formula.
 - Alternatively, we can use the following formula to speed things up if A^*A is invertible:

$$P_{\text{range } A} \mathbf{b} = A(A^*A)^{-1}A^*\mathbf{b}$$

- Theorem 5.4.1: For an $m \times n$ matrix A ,

$$\ker A = \ker(A^*A)$$

- Thus, A^*A is invertible iff A is invertible iff A is full rank. This gives us a condition on when we can use the projection formula.
- Theorem 5.6.1: An operator $U : X \rightarrow Y$ is an isometry if and only if it preserves the inner product, i.e., if and only if

$$(\mathbf{x}, \mathbf{y}) = (U\mathbf{x}, U\mathbf{y})$$

for all $\mathbf{x}, \mathbf{y} \in X$.

- Lemma 5.6.2: An operator $U : X \rightarrow Y$ is an isometry if and only if $U^*U = I$.
- **Unitary** (operator): An invertible isometry.
- Proposition 5.6.3: An isometry $U : X \rightarrow Y$ is a unitary operator iff $\dim X = \dim Y$.
- **Orthogonal** (matrix): A unitary matrix with real entries.
- Unitary operator properties:
 1. $U^{-1} = U^*$.
 2. U unitary implies $U^* = U^{-1}$ unitary.
 3. If $\mathbf{v}_1, \dots, \mathbf{v}_n$ is orthonormal, $U\mathbf{v}_1, \dots, U\mathbf{v}_n$ is orthonormal.
 4. U, V unitary implies UV unitary.
- A matrix U is an isometry iff its columns form an orthonormal system.
- Proposition 5.6.4: Let U be a unitary matrix. Then
 1. $|\det U| = 1$. In particular, if U is orthogonal, then $\det U = \pm 1$.
 2. $|\lambda| = 1$ for every eigenvalue λ of U .
- Proposition 5.6.5: A matrix A is unitarily equivalent to a diagonal one iff it has an orthogonal (or-thonormal) basis of eigenvectors.

Chapter 6

Structure of Operators on Inner Product Spaces

6.1 Notes

10/11:

- Spectral decomposition of self-adjoint linear maps.
 - Can we write a map in term of the eigenvalues only?
 - Let $A : X \rightarrow X$ be linear and self-adjoint. Where $\dim X < \infty$.
 - Let A have eigenvalues $\lambda_1, \dots, \lambda_n$ and eigenvectors $\mathbf{v}_1, \dots, \mathbf{v}_n$. Then there is an orthonormal basis of X consisting of eigenvectors of A . An operator is self-adjoint if $A = A^*$.
 - If A is self-adjoint, then A can be written as diagonal with the eigenvalues on the diagonal with respect to some orthonormal basis of eigenvectors.
 - Let $\mathbb{F} = \mathbb{C}$.
- If there exists an orthonormal basis u_1, \dots, u_n of X such that A is triangular, then $A = UTU^*$ where U is unitary and T is upper triangular.
 - Proved with induction on $\dim X$.
 - $\dim X = 1$ is clear.
 - Assume for $\dim X = n - 1$, WTS for $\dim X = n$.
 - The subspace has a basis $\mathbf{v}_1, \dots, \mathbf{v}_{n-1}$ such that A has a diagonal form.
 - Let $u \in X$ be linearly independent of $\mathbf{v}_1, \dots, \mathbf{v}_{n-1}$.
 - Let λ be the remaining eigenvalue and u the corresponding eigenvector. Let $E = \text{span}(u)$. Then make the matrix λ in the upper left corner, and block diagonal with “ A_{n-1} ” in the bottom right corner, zeroes everywhere else.
- **Self-adjoint** (matrix A): A linear map $A : X \rightarrow X$ where $\dim X < \infty$ such that $A = A^*$.
 - Similarly, $(Ax, y) = (x, Ay)$.
 - A self-adjoint implies all eigenvalues are real, eigenvectors corresponding to different eigenvalues are orthogonal.
 - Soug proves this.
- **Strictly positive** (operator A): A self-adjoint operator $A : X \rightarrow X$ such that $(Ax, x) > 0$ for all $x \neq 0$. Also known as **positive definite**.
 - Implies that all eigenvalues are strictly positive.

- **Nonnegative** (operator A): A self-adjoint operator $A : X \rightarrow X$ such that $(Ax, x) \geq 0$ for all $x \neq 0$. Also known as **definite**.

- All eigenvalues are nonnegative.

- Suppose $A \geq 0$ is self-adjoint. Then there exists a unique self-adjoint $B \geq 0$ such that $B^2 = A$.

- A self-adjoint is diagonal (wrt. some basis).
- A positive means that all eigenvalues (diagonal entries) are positive.
- Thus, take

$$B = \begin{pmatrix} \sqrt{\lambda_1} & 0 & 0 \\ 0 & \ddots & 0 \\ 0 & 0 & \sqrt{\lambda_n} \end{pmatrix}$$

- Suppose $B^2 = A$, $C^2 = A$. Then we have an orthonormal basis corresponding to B and an orthonormal basis corresponding to C . It follows that $B^2 = C^2 = A$. Write B^2x and C^2x in terms of their bases; will necessitate that the bases are the same.

10/13:

- If we get yes/no questions, we don't have to justify.
- Cauchy-Schwarz inequality:

$$|(\mathbf{x}, \mathbf{y})| \leq \|\mathbf{x}\| \|\mathbf{y}\|$$

- Real spaces, V vs. (\cdot, \cdot) inner product.
- Proof:

$$\begin{aligned} 0 &\leq \|\mathbf{x} + t\mathbf{y}\|^2 \\ &= t^2 \|\mathbf{y}\|^2 + 2t(\mathbf{x}, \mathbf{y}) + \|\mathbf{x}\|^2 \end{aligned}$$

Thus, the discriminant must be less than zero (because the whole polynomial is positive, so the discriminant [the opposite of the x^0 term of the factored form of the polynomial] must be less than zero so the polynomial doesn't get dragged down to negative values):

$$(\mathbf{x}, \mathbf{y})^2 - \|\mathbf{x}\|^2 \|\mathbf{y}\|^2 \leq 0$$

Taking square roots of both sides proves the desired inequality.

- Recall that if $A^* = A$, then all eigenvalues are real and all eigenvectors of distinct eigenvalues are orthogonal to each other.
- **Normal** (matrix): A matrix N such that $N^*N = NN^*$.
 - Examples: Diagonal, self-adjoint, and unitary operators are all normal.
- Any normal operator in a complex vector space has an orthonormal set of eigenvectors, e.g., $N = UDU^*$.
 - Proof: N is upper triangular wrt. some basis (because all matrices are). WTS any normal upper triangular matrix is diagonal. Done by induction on the dimension of N from $n = 2$.
 - Assume the claim for every $(n - 1) \times (n - 1)$ normal upper triangular matrix.
 - Let

$$N = \begin{pmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ 0 & & & \\ 0 & & & \\ \vdots & & N_1 & \\ 0 & & & \end{pmatrix}$$

(we know every normal matrix can be written in this upper triangular form)

- Then just compute NN^* and N^*N . Knowing they have to be equal, we have that $a_{12} = \dots = a_{1n} = 0$.
- We can also prove from the above (block diagonal multiplication) that N_1 is normal. Thus, it's diagonal, too. Therefore, the whole thing is diagonal.
- N is normal if and only if $\|N\mathbf{x}\| = \|N^*\mathbf{x}\|$.
 - Proof: $(N\mathbf{x}, N\mathbf{y}) = (N^*N\mathbf{x}, \mathbf{y}) = (NN^*\mathbf{x}, \mathbf{y}) = (N^*\mathbf{x}, N^*\mathbf{y})$. This is equivalent to the desired condition.
- If A is nonnegative and $(A\mathbf{e}_k, \mathbf{e}_k) = a_{kk}$, then

$$\sum_{i,j=1}^n a_{ij} \mathbf{x}_i \mathbf{x}_j$$

- **Positive definite** (matrix): An $n \times n$ self-adjoint matrix such that $(A\mathbf{x}, \mathbf{x}) > 0$ for all $\mathbf{x} \in X$.
- Let $A : X \rightarrow Y$, $\dim X = \dim Y$. Then AA^* is positive semidefinite. And there exists a unique square root $R = \sqrt{A^*A}$.
 - Proof: $(A^*A\mathbf{x}, \mathbf{x}) = (A\mathbf{x}, A\mathbf{x}) = \|A\mathbf{x}\|^2 \geq 0$.
- **Modulus** (of A): The matrix $|A| = \sqrt{A^*A}$.
- Check $\| |A|\mathbf{x} \| = \|A\mathbf{x}\|$.

$$\| |A|\mathbf{x} \|^2 = (|A|\mathbf{x}, |A|\mathbf{x}) = (|A|^*|A|\mathbf{x}, \mathbf{x}) = (A^*A\mathbf{x}, \mathbf{x}) = (A\mathbf{x}, A\mathbf{x}) = \|A\mathbf{x}\|^2$$

- Let $A : X \rightarrow X$ be a linear operator. Then $A = U|A|$ where U is unitary.
- Look at singular matrices.

10/15:

- Recall that if $A : X \rightarrow Y$, we have that A^*A is semidefinite, positive, and self adjoint.
 - Thus, there exists a unique matrix $R = \sqrt{A^*A} \geq 0$, which we define to be $|A| = \sqrt{A^*A}$.
- Polar form of a matrix:

$$A = U|A|$$

- This may not be unique!
- Proof: Suppose $A\mathbf{x} = U(|A|\mathbf{x})$. $A\mathbf{x} \in \text{range } A$, and $|A|\mathbf{x} \in \text{range } (|A|)$. $\mathbf{x} \in \text{range } (|A|)$ implies that there exists $\mathbf{v} \in X$ such that $\mathbf{x} = |A|\mathbf{v}$.
- Define $U\mathbf{x} = A\mathbf{x}$. U is a well-defined linear map.
- $\|U\mathbf{x}\| = \|A\mathbf{x}\| = \||A|\mathbf{v}\| = \|\mathbf{x}\|$.
- U is an isometry.
- $\text{range } |A| \rightarrow X$.
- Use $\ker A = \ker |A| = (\text{range } A)^\perp$ to extend U_0 to U : $U = U_0 + U_1$.
- **Singular values** (of a matrix): The eigenvalues of $|A|$.
 - So if $\lambda_1, \dots, \lambda_n$ are the eigenvalues of A^*A , the singular values of A are $\sqrt{\lambda_1}, \dots, \sqrt{\lambda_n}$.
- Let $A : X \rightarrow Y$ be a linear map.
 - Let $\sigma_1, \dots, \sigma_n$ be the singular values of A . Then $\sigma_1, \dots, \sigma_n > 0$.
 - Additionally, if $\mathbf{v}_1, \dots, \mathbf{v}_n$ is an orthonormal basis of eigenvectors of A^*A , then the list of n vectors $\mathbf{w}_1, \dots, \mathbf{w}_n$ defined by $\mathbf{w}_k = 1/\sigma_k A\mathbf{v}_k$ for each $k = 1, \dots, n$ is orthonormal.

■ Proof:

$$(\mathbf{w}_k, \mathbf{w}_j) = \frac{1}{\sigma_k \sigma_j} (A\mathbf{v}_k, A\mathbf{v}_j) = \frac{1}{\sigma_k \sigma_j} = \frac{1}{\sigma_k \sigma_j} (A^* A \mathbf{v}_k, \mathbf{v}_j) = \frac{\sigma_k^2}{\sigma_k \sigma_j} (\mathbf{v}_k, \mathbf{v}_j) = 0$$

and

$$\|\mathbf{w}_k\| = \frac{1}{\sigma_k} \|A\mathbf{v}_k\| = \frac{1}{\sigma_k} \||A|\mathbf{v}_k\| = 1$$

– Schmidt decomposition of A :

$$A\mathbf{x} = \sum_{k=1}^r \sigma_k (\mathbf{x}, \mathbf{v}_k) \mathbf{w}_k$$

■ This is because $\mathbf{x} = \sum (\mathbf{x}, \mathbf{v}_k) \mathbf{v}_k$, so by the above,

$$A\mathbf{x} = \sum_{k=1}^n (\mathbf{x}, \mathbf{v}_k) A\mathbf{v}_k = \sum_{k=1}^r \sigma_k (\mathbf{x}, \mathbf{v}_k) \mathbf{w}_k$$

• **Operator norm:** $\|A\| = \max\{\|A\mathbf{x}\| : \|\mathbf{x}\| \leq 1\}$.

• Properties of the operator norm:

- $\|A\mathbf{x}\| \leq \|A\| \|\mathbf{x}\|$.
- $\|\alpha A\| = |\alpha| \|A\|$.
- $\|A + B\| \leq \|A\| + \|B\|$.
- $\|A\| \geq 0$.
- $\|A\| = 0$ iff $A = 0$.

• **Frobenius norm:** The norm $\|A\|_2^2 = \text{tr}(A^* A)$.

• The operator norm is always less than or equal to the Frobenius norm.

• If $A : \mathbb{F}^n \rightarrow \mathbb{F}^n$, then $A = W\Sigma V^*$ where σ is a diagonal matrix of nonzero singular values.

• The operator norm of A is the largest of the singular values.

• An orthogonal matrix can be decomposed to a block-diagonal matrix of rotations.

10/18:

• Soug tests what he teaches and doesn't give super tricky questions.

• Structure of orthogonal matrices.

• **Orthogonal (matrix):** A unitary matrix U with all elements real and $|\det U| = 1$.

• Theorem: Let U be an orthogonal operator on \mathbb{R}^n such that $\det U = 1$. Then there exists an orthonormal basis $\mathbf{v}_1, \dots, \mathbf{v}_n$ such that with respect to this basis,

$$U = \begin{pmatrix} R_{\phi_1} & & & \mathbf{0} \\ & \ddots & & \\ & & R_{\phi_k} & \\ \mathbf{0} & & & I_{n-2k} \end{pmatrix}$$

where each R_{ϕ_i} is a 2×2 rotation matrix.

- If you are in \mathbb{R}^7 for example, you would be able to express U as a composition of at most 3 rotation maps and the identity map.
- Each rotation map acts on two orthonormal vectors.
- Proof: $P(\lambda)$ is the n -degree characteristic polynomial $\det(U - \lambda I) = 0$. The eigenvalues are the roots of it.

- $p(\lambda) = 0$ if and only if $p(\bar{\lambda}) = 0$.
 - $\lambda \in \mathbb{C}$ is an eigenvalue with eigenvector $\mathbf{u} \neq 0$ iff $U\mathbf{u} = \lambda\mathbf{u}$ and $U\bar{\mathbf{u}} = \bar{\lambda}\bar{\mathbf{u}}$.
- Recall that U unitary implies $|\lambda| = 1$.
 - Proof^[1]: $\|U\mathbf{x}\| = \|\mathbf{x}\|$ and $U\mathbf{x} = \lambda\mathbf{x}$. Thus,

$$\|U\mathbf{x}\| = \|\lambda\mathbf{x}\| = |\lambda|\|\mathbf{x}\| = \|\mathbf{x}\|$$

and since $\mathbf{x} \neq 0$, we can divide by $\|\mathbf{x}\|$, so $|\lambda| = 1$.

- Let $\mathbf{u} = \text{Re } \mathbf{u} + i \text{Im } \mathbf{u}$.
- It follows that we may define

$$\mathbf{x} = \text{Re } \mathbf{u} = \frac{\mathbf{u} + \bar{\mathbf{u}}}{2} \qquad \mathbf{y} = \text{Im } \mathbf{u} = \frac{\mathbf{u} - \bar{\mathbf{u}}}{2}$$

- Thus, $\mathbf{u} = \mathbf{x} + i\mathbf{y}$ and $\bar{\mathbf{u}} = \mathbf{x} - i\mathbf{y}$.
- Since $U\mathbf{x} = \frac{U\mathbf{u} + U\bar{\mathbf{u}}}{2} = \frac{\lambda\mathbf{u} + \bar{\lambda}\bar{\mathbf{u}}}{2}$, $U\mathbf{y} = \text{Im}(\lambda\mathbf{u}) = \text{Re}(\lambda\mathbf{u})$.
- Since $|\lambda| = 1$, $\lambda = e^{i\alpha}$ and $\bar{\lambda} = e^{-i\alpha}$.
- It follows that $U\mathbf{x} = (\cos \alpha)\mathbf{x} - (\sin \alpha)\mathbf{y}$ and $U\mathbf{y} = (\cos \alpha)\mathbf{y} + (\sin \alpha)\mathbf{x}$.
- Thus, since $U\mathbf{x} = \text{Re } \lambda\mathbf{u}$, we have that

$$\begin{aligned} \lambda\mathbf{u} &= (\cos \alpha + i \sin \alpha)(\mathbf{x} + i\mathbf{y}) \\ &= (\cos \alpha)\mathbf{x} - (\sin \alpha)\mathbf{y} + i[(\cos \alpha)\mathbf{y} + (\sin \alpha)\mathbf{x}] \end{aligned}$$

- If E_λ is a 2 dimensional space spanned by \mathbf{x} and \mathbf{y} and invariant by U . Thus, any block of the desired matrix leaves its desired sub-block invariant.
- We also know that the eigenvectors of a unitary matrix corresponding to different eigenvalues are orthogonal.
- Thus, $\|\mathbf{x}\| = \|\mathbf{y}\| = \sqrt{2}/2\|\mathbf{u}\|$, $\mathbf{x} \perp \mathbf{y}$.
- Let \mathbf{x}, \mathbf{y} complete the theorem to form a basis of \mathbb{R}^n .
- It will follow that

$$U = \begin{pmatrix} R_\alpha & \mathbf{0} \\ \mathbf{0} & U_1 \end{pmatrix}$$

where U_1 is orthogonal, and we may repeat the process.

6.2 Chapter 6: Structure of Operators on Inner Product Spaces

From Treil (2017).

10/24:

- Theorem 6.1.1: Let $A : X \rightarrow X$ be an operator acting in a complex inner product space. Then there exists an orthonormal basis $\mathbf{u}_1, \dots, \mathbf{u}_n$ of X such that the matrix of A in this basis is upper triangular. In other words, any $n \times n$ matrix A can be represented as $A = UTU^*$, where U is unitary and T is upper-triangular.
- Theorem 6.1.2: Let $A : X \rightarrow X$ be an operator acting on a real inner product space. Suppose that all eigenvalues of A are real. Then there exists an orthonormal basis $\mathbf{u}_1, \dots, \mathbf{u}_n$ in X such that the matrix of A in this basis is upper triangular. In other words, any real $n \times n$ matrix A with all real eigenvalues can be represented as $T = UTU^* = UTU^T$, where U is orthogonal and T is a real upper-triangular matrix.

¹This would be a good exam question.

- Theorem 6.2.1: Let $A = A^*$ be a self-adjoint operator in an inner product space X (the space can be complex or real). Then all eigenvalues of A are real and there exists an orthonormal basis of eigenvectors of A in X .

Equivalently (see Theorem 6.2.2), A can be represented as $A = UDU^*$ where U is a unitary matrix and D is a diagonal matrix with real entries. Moreover, if A is real, U can be chosen to be real, i.e., orthogonal.

- Proposition 6.2.3: Let $A = A^*$ be a self-adjoint operator and let $\lambda, \mathbf{u}, \mu, \mathbf{v}$ be such that $A\mathbf{u} = \lambda\mathbf{u}$ and $A\mathbf{v} = \mu\mathbf{v}$. Then if $\lambda \neq \mu$, $\mathbf{u} \perp \mathbf{v}$.
- Since complex multiplication is commutative,

$$D^*D = DD^*$$

for every diagonal matrix D .

– It follows that $A^*A = AA^*$ if the matrix of A in some orthonormal basis is diagonal.

- Theorem 6.2.4: Any normal operator N in a complex vector space has an orthonormal basis of eigenvectors.

Equivalently, any matrix N satisfying $N^*N = NN^*$ can be represented as $N = UDU^*$ where U is unitary and D is diagonal.

- Proposition 6.2.5: An operator $N : X \rightarrow X$ is normal iff

$$\|N\mathbf{x}\| = \|N^*\mathbf{x}\|$$

for all $\mathbf{x} \in X$.

- **Hermitian square** (of A): The matrix A^*A .
- **Modulus** (of A): The unique positive semidefinite square root $\sqrt{A^*A}$.
- Proposition 6.3.3: For a linear operator $A : X \rightarrow Y$,

$$\| |A| \mathbf{x} \| = \| A \mathbf{x} \|$$

- Corollary 6.3.4: $\ker A = \ker |A|$.
- Theorem 6.3.5: Let $A : X \rightarrow X$ be an operator (square matrix). Then A can be represented as

$$A = U|A|$$

where U is a unitary operator.

- **Singular value** (of A): An eigenvalue of $|A|$.
 - A positive square root of an operator of A^*A .
- Proposition 6.3.6: Let $\sigma_1, \dots, \sigma_n$ be the singular values of A , ordered such that $\sigma_1, \dots, \sigma_r$ are the nonzero singular values, and let $\mathbf{v}_1, \dots, \mathbf{v}_n$ be an orthonormal basis of eigenvectors of A^*A . Then the system

$$\mathbf{w}_k = \frac{1}{\sigma_k} A \mathbf{v}_k$$

for $k = 1, \dots, r$ is orthonormal.

- **Schmidt decomposition** (of A): The decompositions

$$A = \sum_{k=1}^r \sigma_k \mathbf{w}_k \mathbf{v}_k^*$$

and

$$A\mathbf{x} = \sum_{k=1}^r \sigma_k (\mathbf{x}, \mathbf{v}_k) \mathbf{w}_k$$

- Note that these can be verified by plugging $\mathbf{x} = \mathbf{v}_j$ for each $j = 1, \dots, n$ into the latter equation.

10/25:

- Lemma 6.3.7: A can be represented as the Schmidt decomposition

$$A = \sum_{k=1}^r \sigma_k \mathbf{w}_k \mathbf{v}_k^*$$

where $\sigma_k > 0$ for any orthonormal systems $\mathbf{v}_1, \dots, \mathbf{v}_r$ and $\mathbf{w}_1, \dots, \mathbf{w}_r$.

- Corollary 6.3.8: Let $A = \sum_{k=1}^r \sigma_k \mathbf{w}_k \mathbf{v}_k^*$ be a Schmidt decomposition of A . Then

$$A^* = \sum_{k=1}^r \sigma_k \mathbf{v}_k \mathbf{w}_k^*$$

is a Schmidt decomposition of A^* .

- **Reduced singular value decomposition** (of A): The decomposition

$$A = \tilde{W} \tilde{\Sigma} \tilde{V}^*$$

where $A : \mathbb{F}^n \rightarrow \mathbb{F}^m$ has the Schmidt decomposition $A = \sum_{k=1}^r \sigma_k \mathbf{w}_k \mathbf{v}_k^*$, $\tilde{\Sigma} = \text{diag}\{\sigma_1, \dots, \sigma_r\}$, and \tilde{V}, \tilde{W} are matrices with columns $\mathbf{v}_1, \dots, \mathbf{v}_r$ and $\mathbf{w}_1, \dots, \mathbf{w}_r$, respectively. *Also known as **compact singular value decomposition**.*

- Note that \tilde{V} is an $n \times r$ matrix, $\tilde{\Sigma}$ is an $r \times r$ matrix, and \tilde{W} is an $m \times r$ matrix.
- Since $\mathbf{v}_1, \dots, \mathbf{v}_r$ and $\mathbf{w}_1, \dots, \mathbf{w}_r$ are orthonormal, \tilde{V}, \tilde{W} are isometries.
- Note that $r = \text{rank } A$ (see Problem 6.3.1).
 - It follows that if A is invertible, then $m = n = r$, so \tilde{V}, \tilde{W} are unitary and $\tilde{\Sigma}$ is an invertible diagonal matrix.
- However, A need not be invertible for us to get a representation similar to $A = \tilde{W} \tilde{\Sigma} \tilde{V}^*$.
 - Complete $\mathbf{v}_1, \dots, \mathbf{v}_r$ and $\mathbf{w}_1, \dots, \mathbf{w}_r$ to bases of \mathbb{F}^n and \mathbb{F}^m , respectively.
 - Then we get the following.
- **Singular value decomposition** (of A): The decomposition

$$A = W \Sigma V^*$$

where $V \in M_{n \times n}^{\mathbb{F}}$ and $W \in M_{m \times m}^{\mathbb{F}}$ are unitary matrices with columns $\mathbf{v}_1, \dots, \mathbf{v}_n$ and $\mathbf{w}_1, \dots, \mathbf{w}_m$, respectively, and $\Sigma \in M_{m \times n}^{\mathbb{R}^+}$ is a “diagonal” matrix such that

$$\Sigma_{j,k} = \begin{cases} \sigma_k & j = k \leq r \\ 0 & \text{otherwise} \end{cases}$$

- Notice that if $A = W\Sigma V^*$, then

$$A^*A = (W\Sigma V^*)^*(W\Sigma V^*) = V\Sigma^*W^*W\Sigma V^* = V\Sigma^2V^*$$

proving that the singular values of A , squared, are the eigenvalues of A^*A .

- If A is invertible, the reduced SVD is the matrix form of the Schmidt decomposition is the SVD.
- If $A = W\Sigma V^*$ is $n \times n$, then

$$A = \underbrace{(WV^*)}_U \underbrace{(V\Sigma V^*)}_{|A|}$$

is a polar decomposition of A .

- Consider the unit ball $B = \{\mathbf{x} \in \mathbb{R}^n : \|\mathbf{x}\| \leq 1\}$.
 - We want to describe $A(B)$, i.e., the image of the unit ball under A .
 - Let $\mathbf{x} = (x_1, \dots, x_n)^T$ and let $\mathbf{y} = (y_1, \dots, y_n)^T$. If $A = \text{diag}\{\sigma_1, \dots, \sigma_n\}$, we have $\mathbf{y} \in A(B)$ iff $\mathbf{y} = A\mathbf{x}$ where $\mathbf{x} \in B$ iff

$$\sum_{k=1}^n \frac{y_k^2}{\sigma_k^2} = \sum_{k=1}^n x_k^2 = \|\mathbf{x}\|^2 \leq 1$$

- Thus, $A(B)$ is an ellipsoid with half-axes $\sigma_1, \dots, \sigma_n$.
- In the more general case, if $A = W\Sigma V^*$, then since V^* is unitary, $V^*(B) = B$. $\Sigma V^*(B) = \Sigma(B)$ is thus by the above an ellipsoid in range Σ with half-axes $\sigma_1, \dots, \sigma_r$. Thus, since isometries don't change geometry, $W(\Sigma(B))$ is also an ellipsoid with the same half-axes, but in range A .
- Conclusion: The image $A(B)$ of the closed unit ball B is an ellipsoid in range A with half-axes $\sigma_1, \dots, \sigma_r$, where r is the number of nonzero singular values, i.e., the rank of A .
- Finding the maximum of $\|A\mathbf{x}\|$ for $\mathbf{x} \in B$.
 - For a diagonal matrix Σ with nonnegative entries, the maximum is clearly the maximal diagonal entry: In this case if s_1 is the maximal diagonal entry, then since

$$\Sigma\mathbf{x} = \sum_{k=1}^r s_k x_k \mathbf{e}_k$$

we have that

$$\|A\mathbf{x}\|^2 = \sum_{k=1}^r s_k^2 |x_k|^2 \leq s_1^2 \sum_{k=1}^r |x_k|^2 = s_1^2 \cdot \|\mathbf{x}\|^2$$

- We get the following by a similar logic to before.
- Conclusion: The maximum of $\|A\mathbf{x}\|$ on the unit ball B is the maximal singular value of A .
- **Operator norm** (of A): The following quantity. Denoted by $\|A\|$. Given by

$$\|A\| = \max\{\|A\mathbf{x}\| : \mathbf{x} \in X, \|\mathbf{x}\| \leq 1\}$$

- $\|A\|$ clearly satisfies the four properties of a norm.
- Additionally,

$$\|A\mathbf{x}\| \leq \|A\| \cdot \|\mathbf{x}\|$$

- Alternate definition: The operator norm $\|A\|$ is the smallest number $C \geq 0$ such that $\|A\mathbf{x}\| \leq C\|\mathbf{x}\|$.

- **Frobenius norm:** The following norm. *Also known as Hilbert-Schmidt norm.* Denoted by $\|A\|_2$. Given by

$$\|A\|_2^2 = \text{tr}(A^*A)$$

- If we let s_1, \dots, s_n be the singular values of A and let s_1 be the largest value, then we have

$$\|A\|^2 = s_1^2 \leq \sum_{k=1}^n s_k^2 = \text{tr}(A^*A) = \|A\|_2^2$$

- Conclusion: The operator norm of a matrix cannot be more than its Frobenius norm.
- Suppose we want to solve $A\mathbf{x} = \mathbf{b}$ where A is invertible, but there is some (experimental) error $\Delta\mathbf{b}$ in \mathbf{b} . Then we are really solving for an approximate solution $\mathbf{x} + \Delta\mathbf{x}$ to the equation

$$A(\mathbf{x} + \Delta\mathbf{x}) = \mathbf{b} + \Delta\mathbf{b}$$

- It follows since A is invertible that $\mathbf{x} = A^{-1}\mathbf{b}$ and $\Delta\mathbf{x} = A^{-1}\Delta\mathbf{b}$.
- To estimate the relative error $\|\Delta\mathbf{x}\|/\|\mathbf{x}\|$ in the solution in comparison with the relative error $\|\Delta\mathbf{b}\|/\|\mathbf{b}\|$ in the data, use

$$\frac{\|\Delta\mathbf{x}\|}{\|\mathbf{x}\|} = \frac{\|A^{-1}\Delta\mathbf{b}\|}{\|\mathbf{b}\|} \frac{\|A\mathbf{x}\|}{\|\mathbf{x}\|} \leq \frac{\|A^{-1}\| \cdot \|\Delta\mathbf{b}\|}{\|\mathbf{b}\|} \frac{\|A\| \cdot \|\mathbf{x}\|}{\|\mathbf{x}\|} = \|A^{-1}\| \cdot \|A\| \cdot \frac{\|\Delta\mathbf{b}\|}{\|\mathbf{b}\|}$$

- **Condition number** (of A): The following quantity. *Given by*

$$\|A\| \cdot \|A^{-1}\|$$

- If s_1 is the largest singular value of A and s_n is the smallest, then

$$\|A\| \cdot \|A^{-1}\| = s_1 \cdot \frac{1}{s_n} = \frac{s_1}{s_n}$$

- **Well-conditioned** (matrix): A matrix the condition number of which is not “too big.”
- **Ill-conditioned** (matrix): A matrix that is not well-conditioned.
- Theorem 6.5.1: Let U be an orthogonal operator on \mathbb{R}^n and let $\det U = 1$. Then there exists an orthonormal basis $\mathbf{v}_1, \dots, \mathbf{v}_n$ such that the matrix of U in this basis has the block diagonal form

$$\begin{pmatrix} R_{\varphi_1} & & & 0 \\ & \ddots & & \\ & & R_{\varphi_k} & \\ 0 & & & I_{n-2k} \end{pmatrix}$$

where each R_{φ_j} is a two-dimensional rotation

$$R_{\varphi_j} = \begin{pmatrix} \cos \varphi_j & -\sin \varphi_j \\ \sin \varphi_j & \cos \varphi_j \end{pmatrix}$$

and I_{n-2k} represents the $(n-2k) \times (n-2k)$ identity matrix.

- Alternate interpretation: Any rotation in \mathbb{R}^n can be represented as a composition of at most $n/2$ commuting planar rotations.

- Theorem 6.5.2: Let U be an orthogonal operator on \mathbb{R}^n and let $\det U = -1$. Then there exists an orthonormal basis $\mathbf{v}_1, \dots, \mathbf{v}_n$ such that the matrix of U in this basis has block diagonal form

$$\begin{pmatrix} R_{\varphi_1} & & & & 0 \\ & \ddots & & & \\ & & R_{\varphi_k} & & \\ & & & I_r & \\ 0 & & & & -1 \end{pmatrix}$$

where $r = n - 2k - 1$ and each R_{φ_j} is a two-dimensional rotation

$$R_{\varphi_j} = \begin{pmatrix} \cos \varphi_j & -\sin \varphi_j \\ \sin \varphi_j & \cos \varphi_j \end{pmatrix}$$

- Corollary: An orthogonal 2×2 matrix U with determinant -1 is always a reflection.
- Theorem 6.5.3: Any rotation U (i.e., any orthogonal transformation U with $\det U = 1$) can be represented as a product of at most $n(n-1)/2$ elementary rotations.
- Consider the following orthonormal bases of \mathbb{R}^2 .



Figure 6.1: Orientation in \mathbb{R}^2 .

- Notice that a rotation will get you from the standard basis (a) to basis (b), but not from the standard basis (a) to basis (c).
- This is the motivation for defining orientation.
- More formally, we know that there is a unique linear transformation U such that $U\mathbf{e}_k = \mathbf{v}_k$ for each $k = 1, 2$. In particular, the matrix of U with respect to the standard basis is orthogonal with columns $\mathbf{v}_1, \mathbf{v}_2$.
- By Theorems 6.5.1 and 6.5.2, if $\det U = 1$, then U is a rotation, and if $\det U = -1$, then U is not a rotation.
- **Similarly oriented** (bases \mathcal{A}, \mathcal{B}): Two bases \mathcal{A}, \mathcal{B} of a real vector space such that the change of coordinates matrix $[I]_{\mathcal{B}\mathcal{A}}$ has a positive determinant.
- **Differently oriented** (bases \mathcal{A}, \mathcal{B}): Two bases \mathcal{A}, \mathcal{B} of a real vector space that are not similarly oriented (i.e., $[I]_{\mathcal{B}\mathcal{A}}$ has a negative determinant).
- We usually let the standard basis of \mathbb{R}^n have a **positive orientation**.
 - In an abstract vector space, we need only fix a basis and declare its orientation to be positive.
- **Continuously transformable** (bases \mathcal{A}, \mathcal{B}): Two bases \mathcal{A}, \mathcal{B} such that $\mathcal{A} = \{\mathbf{a}_1, \dots, \mathbf{a}_n\}$ can be continuously transformed to a basis $\mathcal{B} = \{\mathbf{b}_1, \dots, \mathbf{b}_n\}$. In particular, there exists a **continuous family of bases** $\mathcal{V}(t) = \{\mathbf{v}_1(t), \dots, \mathbf{v}_n(t)\}$, $t \in [a, b]$, such that

$$\mathbf{v}_k(a) = \mathbf{a}_k \qquad \mathbf{v}_k(b) = \mathbf{b}_k$$

for each $k = 1, \dots, n$.

- **Continuous family of bases:** A family of bases $\mathcal{V}(t) = \{\mathbf{v}_1(t), \dots, \mathbf{v}_n(t)\}$, $t \in [a, b]$, such that the vector-functions $\mathbf{v}_k(t)$ are continuous (their coordinates in some bases are continuous functions) and the system $\mathbf{v}_1(t), \dots, \mathbf{v}_n(t)$ is a basis for all $t \in [a, b]$.
- Theorem 6.6.1: Two bases $\mathcal{A} = \{\mathbf{a}_1, \dots, \mathbf{a}_n\}$ and $\mathcal{B} = \{\mathbf{b}_1, \dots, \mathbf{b}_n\}$ have the same orientation if and only if one of the bases can be continuously transformed to the other.

Chapter 7

Bilinear and Quadratic Forms

7.1 Notes

10/18: • **Bilinear form:** A function $L : \mathbb{R}^n \times \mathbb{R}^n \rightarrow \mathbb{R}$ such that

$$L(\alpha \mathbf{x}_1 + \beta \mathbf{x}_2, \mathbf{y}) = \alpha L(\mathbf{x}_1, \mathbf{y}) + \beta L(\mathbf{x}_2, \mathbf{y}) \quad L(\mathbf{x}, \alpha \mathbf{y}_1 + \beta \mathbf{y}_2) = \alpha L(\mathbf{x}, \mathbf{y}_1) + \beta L(\mathbf{x}, \mathbf{y}_2)$$

$$- L(\mathbf{x}, \mathbf{y}) = (A\mathbf{x}, \mathbf{y}).$$

• **Quadratic form:** A bilinear form $L(\mathbf{x}, \mathbf{x})$.

$$- (\mathbf{x}, \mathbf{x}) \text{ is a polynomial of degree 2 in } \mathbf{x}_1, \dots, \mathbf{x}_n:$$

$$L(\lambda \mathbf{x}, \lambda \mathbf{x}) = (\lambda \mathbf{x}, \lambda \mathbf{x}) = \lambda^2 (\mathbf{x}, \mathbf{x})$$

• We have that

$$(A\mathbf{x}, \mathbf{x}) = (A\lambda \mathbf{x}, \lambda \mathbf{x}) = \lambda^2 (A\mathbf{x}, \mathbf{x}) = \sum_{j,i=1}^n \alpha_{j,i} \mathbf{x}_i \mathbf{x}_j$$

• The general form of a quadratic form:

$$- \text{Can any quadratic form on } \mathbb{R}^n \text{ be written as } (A\mathbf{x}, \mathbf{x})?$$

10/20: • Bilinear forms are linear in each argument when keeping the other fixed.

• Quadratic forms $Q(\mathbf{x}) = L(\mathbf{x}, \mathbf{x})$ are quadratic polynomials in the coordinates of x .

$$- \text{In particular, } Q(\lambda \mathbf{x}) = |\lambda|^2 Q(\mathbf{x}).$$

• If Q quadratic is real, then $Q(\mathbf{x}) = (A\mathbf{x}, \mathbf{x})$ where A is some square matrix.

$$- \text{If } \mathbf{e}_1, \dots, \mathbf{e}_n \text{ is an orthonormal basis of } \mathbb{R}^n, \text{ then there exists a unique } A = A^* \text{ such that } (A)_{ij} = L(\mathbf{e}_i, \mathbf{e}_j).$$

$$- \text{Keeping } \mathbf{x} = \sum_{i=1}^n \mathbf{x}_i \mathbf{e}_i \text{ fixed, we have}$$

$$\begin{aligned} Q(\mathbf{x}) &= L(\mathbf{x}, \mathbf{x}) \\ &= L\left(\sum_{i=1}^n \mathbf{x}_i \mathbf{e}_i, \sum_{j=1}^n \mathbf{x}_j \mathbf{e}_j\right) \\ &= \sum_{i=1}^n \mathbf{x}_i L\left(\mathbf{e}_i, \sum_{j=1}^n \mathbf{x}_j \mathbf{e}_j\right) \\ &= \sum_{i,j=1}^n \mathbf{x}_i \mathbf{x}_j \underbrace{L(\mathbf{e}_i, \mathbf{e}_j)}_{A_{ij}} \end{aligned}$$

- We have that

$$\begin{aligned}(A\mathbf{x}, \mathbf{x}) &= (UDU^{-1}\mathbf{x}, \mathbf{x}) \\ &= (DU^{-1}\mathbf{x}, U^{-1}\mathbf{x}) \\ &= \sum_{i=1}^n \lambda_i \underbrace{(U^{-1}\mathbf{x})_i}_{\mathbf{y}_i} \underbrace{(U^{-1}\mathbf{x})_i}_{\mathbf{y}_i}\end{aligned}$$

- Can we characterize the set $\{\mathbf{x} : (A\mathbf{x}, \mathbf{x}) = 1\}$?
 - Note that this set is equivalent to $\{\mathbf{y} : (D\mathbf{y}, \mathbf{y}) = 1\}$ by the above. This set is a rotation of the previous one. Ellipse?
- Positive quadratic form:
 - Q is positive definite if $Q(\mathbf{x}) > 0$ for all $\mathbf{x} \neq \mathbf{0}$ and Q is positive semidefinite if $Q(\mathbf{x}) \geq 0$ for all $\mathbf{x} \neq \mathbf{0}$.
 - Take a self-adjoint matrix $A = A^*$. It is positive definite if $Q(\mathbf{x}) = (A\mathbf{x}, \mathbf{x})$ is positive definite.
- Theorem: If $A = A^*$, then
 1. A is positive definite if and only if all eigenvalues of A are positive.
 2. A is positive semidefinite if and only if all eigenvalues of A are nonnegative.
 3. A is negative semidefinite if and only if all eigenvalues of A are nonpositive.
 4. A is negative definite if and only if all eigenvalues of A are negative.
 5. A is indefinite if and only if the eigenvalues of A have positive and negative values.
- Theorem: $A = A^*$ is positive definite iff $\det A_k > 0$ for all $k = 1, \dots, n$ where A_k is the upper left $k \times k$ submatrix.
- Minimax representation of eigenvalues of a self-adjoint A .
 - Let E be a subspace of X where $\dim X < \infty$. We define $\text{codim}(E) = \dim E^\perp$.
 - Thus, $\dim E + \text{codim } E = \dim X$.
 - Theorem: Let $A = A^*$, $\lambda_1 \geq \dots \geq \lambda_n$ eigenvalues of A . Then

$$\lambda_k = \max_{\substack{E \text{ subspace} \\ \dim E = k}} \min_{\substack{\mathbf{x} \in E \\ \|\mathbf{x}\|=1}} (A\mathbf{x}, \mathbf{x}) = \min_{\substack{F \text{ subspace} \\ \text{codim } F = k-1}} \max_{\substack{\mathbf{x} \in F \\ \|\mathbf{x}\|=1}} (A\mathbf{x}, \mathbf{x})$$

- Proof: A diagonal equals $(\lambda_1, \dots, \lambda_n)$.
- An orthonormal basis of X such that $\dim E = k$, $\text{codim } F = k-1$, $\dim F = n-k+1$.
- There exists an $\mathbf{x}_0 \neq \mathbf{0}$ such that $\mathbf{x}_0 \in E \cap F$.
- Note that if $B = B^*$, then the max and min of $(B\mathbf{x}, \mathbf{x})$ over the unit sphere is the maximal and minimal eigenvalue of B .
- Thus,

$$\min_{\substack{\mathbf{x} \in E \\ \|\mathbf{x}\|=1}} (A\mathbf{x}, \mathbf{x}) \leq (A\mathbf{x}_0, \mathbf{x}_0) \leq \max_{\substack{\mathbf{x} \in F \\ \|\mathbf{x}\|=1}} (A\mathbf{x}, \mathbf{x})$$

- This is true for any E, F subspaces. $\dim E = k$, $\text{codim } F = k-1$, $E_0 = \text{span}(\mathbf{e}_1, \dots, \mathbf{e}_k)$ and $F_0 = \text{span}(\mathbf{e}_k, \dots, \mathbf{e}_n)$.
- Thus,

$$\min_{\substack{E_0 \\ \|\mathbf{x}\|=1}} (A\mathbf{x}, \mathbf{x}) = \lambda_k = \max_{\substack{F_0 \\ \|\mathbf{x}\|=1}} (A\mathbf{x}, \mathbf{x})$$

■ Additionally,

$$\lambda_{k_1} \leq \max_{\dim E=k} \min_{\mathbf{x}} (A\mathbf{x}, \mathbf{x}) \leq \min_{\text{codim } F=k-1} \max_{\mathbf{x}} (A\mathbf{x}, \mathbf{x}) \leq \lambda_k$$

- Corollary: Let $A = A^* = (a_{jk})_{1 \leq j, k \leq n}$ with eigenvalues $\lambda_1, \dots, \lambda_n$ listed in decreasing order. Let $\tilde{A} = (a_{j,k})_{1 \leq j, k \leq n-1}$ with eigenvalues μ_1, \dots, μ_{n-1} listed in decreasing order. Then $\lambda_1 \geq \mu_1 \geq \lambda_2 \geq \mu_2 \geq \dots \geq \mu_{n-1} \geq \lambda_n$.
 - Consider $(A\mathbf{x}, \mathbf{x})$ on $\{\mathbf{e}_1, \dots, \mathbf{e}_n\}$, but then restrict yourself to $\mathbf{x} \in \mathbb{R}^{n-1}$ on $\{\mathbf{e}_1, \dots, \mathbf{e}_{n-1}\}$.

7.2 Chapter 7: Bilinear and Quadratic Forms

From Treil (2017).

- 10/25: • **Bilinear form** (on \mathbb{R}^n): A function $L(\mathbf{x}, \mathbf{y})$ of two arguments $\mathbf{x}, \mathbf{y} \in \mathbb{R}^n$ that is linear in each argument.
- Linearity in each argument:

$$L(\alpha\mathbf{x}_1 + \beta\mathbf{x}_2, \mathbf{y}) = \alpha L(\mathbf{x}_1, \mathbf{y}) + \beta L(\mathbf{x}_2, \mathbf{y}) \quad L(\mathbf{x}, \alpha\mathbf{y}_1 + \beta\mathbf{y}_2) = \alpha L(\mathbf{x}, \mathbf{y}_1) + \beta L(\mathbf{x}, \mathbf{y}_2)$$

- If $\mathbf{x} = (x_1, \dots, x_n)^T$ and $\mathbf{y} = (y_1, \dots, y_n)^T$, then

$$\begin{aligned} L(\mathbf{x}, \mathbf{y}) &= \sum_{j,k=1}^n a_{j,k} x_k y_j \\ &= (A\mathbf{x}, \mathbf{y}) \\ &= \mathbf{y}^T A\mathbf{x} \end{aligned}$$

where

$$A = \begin{pmatrix} a_{1,1} & \cdots & a_{1,n} \\ \vdots & & \vdots \\ a_{n,1} & \cdots & a_{n,n} \end{pmatrix}$$

- A is uniquely determined by L .
- **Quadratic form** (on \mathbb{R}^n): The diagonal of a bilinear form L , i.e., a bilinear form $Q[\mathbf{x}] = L(\mathbf{x}, \mathbf{x}) = (A\mathbf{x}, \mathbf{x})$.
 - Alternatively: A homogeneous polynomial of degree 2, i.e., a polynomial in x_1, \dots, x_n with only ax_k^2 and cx_jx_k terms.
- There are infinitely many ways to write a quadratic form as $(A\mathbf{x}, \mathbf{x})$.
 - However, there is a unique representation $(A\mathbf{x}, \mathbf{x})$ where A is a (real) symmetric matrix.
- **Quadratic form** (on \mathbb{C}^n): A function of the form $Q[\mathbf{x}] = (A\mathbf{x}, \mathbf{x})$ where A is self-adjoint.
- Lemma 7.1.1: Let $(A\mathbf{x}, \mathbf{x})$ be real for all $\mathbf{x} \in \mathbb{C}^n$. Then $A = A^*$.
- To classify quadratic forms, consider the set of points $\mathbf{x} \in \mathbb{R}^n$ defined by $Q[\mathbf{x}] = 1$ for some quadratic form Q .
 - If the matrix of Q is diagonal, i.e., $Q[\mathbf{x}] = a_1x_1^2 + \dots + a_nx_n^2$, then the set of points can easily be visualized.
- The standard method of diagonalizing a quadratic form is change of variables.
- Orthogonal diagonalization.

- Let $Q[\mathbf{x}] = (A\mathbf{x}, \mathbf{x})$ in \mathbb{F}^n .
- Suppose $\mathbf{y} = S^{-1}\mathbf{x}$ where S is an invertible $n \times n$ matrix. Then

$$Q[\mathbf{x}] = Q[S\mathbf{y}] = (AS\mathbf{y}, S\mathbf{y}) = (S^*AS\mathbf{y}, \mathbf{y})$$

so in the new variables \mathbf{y} , the quadratic form has matrix S^*AS .

- Thus, we can let $A = UDU^*$, choose $D = U^*AU$ as our new (diagonal) matrix, and let this matrix act on the variables $\mathbf{y} = U^*\mathbf{x}$.

- Non-orthogonal diagonalization:

- Completing the square:
 - Eliminate all $x_i x_j$ terms by completing the square. Then substitute in a y_k for each squared term.
- Row/column operations:
 - Augment $(A|I)$. Row reduce A to D . Then $I \rightarrow S^*$.

10/28:

- **Sylvester's Law of Inertia:** For a Hermitian matrix A (i.e., for a quadratic form $Q[\mathbf{x}] = (A\mathbf{x}, \mathbf{x})$) and any of its diagonalizations $D = S^*AS$, the number of positive, negative, and zero diagonal entries of D depends only on A , but not on a particular choice of diagonalization.
- **Positive** (subspace $E \subset \mathbb{F}^n$ corresponding to A): A subspace E such that $(A\mathbf{x}, \mathbf{x}) > 0$ for all nonzero $\mathbf{x} \in E$. Also known as **A-positive**.
- **Negative** (subspace $E \subset \mathbb{F}^n$ corresponding to A): A subspace E such that $(A\mathbf{x}, \mathbf{x}) < 0$ for all nonzero $\mathbf{x} \in E$. Also known as **A-negative**.
- **Neutral** (subspace $E \subset \mathbb{F}^n$ corresponding to A): A subspace E such that $(A\mathbf{x}, \mathbf{x}) = 0$ for all nonzero $\mathbf{x} \in E$. Also known as **A-neutral**.
- Theorem 7.3.1: Let A be an $n \times n$ Hermitian matrix, and let $D = S^*AS$ be its diagonalization by an invertible matrix S . Then the number of positive (resp. negative) diagonal entries of D coincides with the maximal dimension of an A -positive (resp. A -negative) subspace.
- Lemma 7.3.2: Let $D = \text{diag}\{\lambda_1, \dots, \lambda_n\}$. Then the number of positive (resp. negative) diagonal entries of D coincides with the maximal dimension of a D -positive (resp. D -negative) subspace.
- **Positive definite** (quadratic form Q): A quadratic form Q such that $Q[\mathbf{x}] > 0$ for all $\mathbf{x} \neq \mathbf{0}$.
- **Positive semidefinite** (quadratic form Q): A quadratic form Q such that $Q[\mathbf{x}] \geq 0$ for all \mathbf{x} .
- **Negative definite** (quadratic form Q): A quadratic form Q such that $Q[\mathbf{x}] < 0$ for all $\mathbf{x} \neq \mathbf{0}$.
- **Negative semidefinite** (quadratic form Q): A quadratic form Q such that $Q[\mathbf{x}] \leq 0$ for all \mathbf{x} .
- **Indefinite** (quadratic form Q): A quadratic form Q for which there exist $\mathbf{x}_1, \mathbf{x}_2$ such that $Q[\mathbf{x}_1] > 0$ and $Q[\mathbf{x}_2] < 0$.
- **Positive definite** (Hermitian matrix A): A matrix A for which the corresponding quadratic form $Q[\mathbf{x}] = (A\mathbf{x}, \mathbf{x})$ is positive definite.
 - Positive semidefinite, negative definite, negative semidefinite, and indefinite Hermitian matrices are defined similarly.
- Theorem 7.4.1: Let $A = A^*$. Then
 1. A is positive definite iff all eigenvalues of A are positive.
 2. A is positive semidefinite iff all eigenvalues of A are non-negative.

3. A is negative definite iff all eigenvalues of A are negative.
 4. A is negative semidefinite iff all eigenvalues of A are non-positive.
 5. A is indefinite iff it has both positive and negative eigenvalues.
- **Upper left submatrix** (of A): A $k \times k$ matrix A_k composed of all entries of A from row (column) 1 through k in the same arrangement.
 - Theorem 7.4.2 (Sylvester's Criterion of Positivity): A matrix $A = A^*$ is positive definite if and only if $\det A_k > 0$ for all $k = 1, \dots, n$.
 - To check if a matrix A is negative definite, check that the matrix $-A$ is positive definite.
 - Theorem 7.4.3 (Minimax characterization of eigenvalues): Let $A = A^*$ be an $n \times n$ matrix and let $\lambda_1 \geq \dots \geq \lambda_n$ be its eigenvalues taken in decreasing order. Then

$$\lambda_k = \max_{E: \dim E = k} \min_{\mathbf{x} \in E: \|\mathbf{x}\|=1} (A\mathbf{x}, \mathbf{x}) = \min_{F: \text{codim } F = k-1} \max_{\mathbf{x} \in F: \|\mathbf{x}\|=1} (A\mathbf{x}, \mathbf{x})$$

- Corollary 7.4.4 (Intertwining of eigenvalues): Let $A = A^* = \{a_{j,k}\}_{j,k=1}^n$ be a self-adjoint matrix and let $\tilde{A} = \{a_{j,k}\}_{j,k=1}^{n-1}$ be its submatrix of size $(n-1) \times (n-1)$. Let $\lambda_1, \dots, \lambda_n$ and μ_1, \dots, μ_{n-1} be the eigenvalues of A and \tilde{A} respectively, taken in decreasing order. Then

$$\lambda_1 \geq \mu_1 \geq \lambda_2 \geq \mu_2 \geq \dots \geq \lambda_{n-1} \geq \mu_{n-1} \geq \lambda_n$$

Chapter 8

Dual Spaces and Tensors

8.1 Notes

10/22: • **Functional:** A linear bounded map $L : H \rightarrow F$, where H is finite dimensional (equivalent to \mathbb{R}^n).

• **Dual space:** The set of bounded linear functionals on H . Denoted by H' , H^* .

• If $l \leq p < \infty$, then

$$l^p = \left\{ (a_n)_{n \in \mathbb{N}} : \sum_{n=1}^{\infty} |a_n|^p < \infty \right\}$$

• Back to finite dimensions, $H' \approx \mathbb{R}^n$.

• Let $\mathbf{a}_1, \dots, \mathbf{a}_n$ be a basis of H . Then $L\mathbf{x} = (L\mathbf{a}_1, \dots, L\mathbf{a}_n) \approx \mathbb{R}^n$.

• Let $L((a_n)_{n \in \mathbb{N}}) = \sum_{n=1}^{\infty} a_n b_n$. Then $L((a_n)_{n \in \mathbb{N}})$ will be bounded if and only if $(b_n)_{n \in \mathbb{N}} \in l^q$ where $1 < p < q$ where $\frac{1}{q} + \frac{1}{p} = 1$.

• **Young's inequality:** The statement

$$ab \leq \frac{a^p}{p} + \frac{b^q}{q}$$

• We have $|\sum a_n b_n| \leq \|a_n\|_p \|b_n\|_q$.

• Conclusion:

$$\sum \frac{|a_n| |b_n|}{\|a_n\|_p \|b_n\|_q} = 1$$

• We can define H'' , too. This contains linear functionals on H' .

• We know that $L(x) = \langle x, L \rangle = x(L)$. $x \in H''$.

• Riesz representation theorem: Let H have an inner product. $L \in H'$ if and only if there exists a unique $y \in H$ such that $L(x) = (x, y)$.

– Gives us a way to identify all bounded linear functionals on H .

– In finite dimensions, $L(x)$, where $x = \sum_1^n \alpha_i a_i$ gives us $L(x) = \sum_1^n \alpha_i L(a_i)$.

8.2 Chapter 8: Dual Spaces and Tensors

10/28:

- Linear functionals are denoted by L .
 - L is given by a $1 \times n$ matrix denoted by $[L]$.
- The collection of all $[L]$ (the dual space) is isomorphic to \mathbb{R}^n via $[L] \mapsto [L]^T$.
 - However, the objects are different: Let $[I]_{\mathcal{B}\mathcal{A}}$ be the change of coordinates matrix in \mathbb{R}^n . We thus have that

$$[\mathbf{v}]_{\mathcal{B}} = [I]_{\mathcal{B}\mathcal{A}}[\mathbf{v}]_{\mathcal{A}}$$

but we also have that

$$[L]_{\mathcal{B}} = [L]_{\mathcal{A}}[I]_{\mathcal{A}\mathcal{B}}$$

so that

$$[L]_{\mathcal{B}}^T = ([L]_{\mathcal{A}}[I]_{\mathcal{A}\mathcal{B}})^T = [I]_{\mathcal{A}\mathcal{B}}^T [L]_{\mathcal{A}}^T$$

- Essentially, “if S is the change of coordinate matrix in $X \dots$ then the change of coordinate matrix in the dual space X' is $(S^{-1})^T$ ” (Treil, 2017, p. 219).
- Lemma 8.1.3: Let $\mathbf{v} \in V$. If $L(\mathbf{v}) = 0$ for all $L \in V'$, then $\mathbf{v} = \mathbf{0}$. As a corollary, if $L(\mathbf{v}_1) = L(\mathbf{v}_2)$ for all $L \in V'$, then $\mathbf{v}_1 = \mathbf{v}_2$.
- The second dual V'' is canonically (i.e., in a natural way) isomorphic to V .
- **Dual basis** (to $\mathbf{b}_1, \dots, \mathbf{b}_n \in V$): The system of vectors $\mathbf{b}'_1, \dots, \mathbf{b}'_n \in V'$ uniquely defined by the following equation. *Also known as biorthogonal basis.*

$$\mathbf{b}'_k(\mathbf{b}_j) = \delta_{kj}$$

- The k^{th} coordinate of a vector \mathbf{v} in a basis $\mathbf{b}_1, \dots, \mathbf{b}_n$ is $\mathbf{b}'_k(\mathbf{v})$.
 - This is a baby version of the **abstract non-orthogonal Fourier decomposition** of \mathbf{v} .
- Theorem 8.2.1 (Riesz representation theorem): Let H be an inner product space. Given a linear functional L on H , there exists a unique vector $\mathbf{y} \in H$ such that

$$L(\mathbf{v}) = (\mathbf{v}, \mathbf{y})$$

for all $\mathbf{v} \in H$.

- If V is a real inner product space, we can define an isomorphism from V to V' by $\mathbf{y} \mapsto L_{\mathbf{y}} = (\mathbf{v}, \mathbf{y})$.
 - If V is complex, this function is not linear since if α is complex,

$$L_{\alpha\mathbf{y}}(\mathbf{v}) = (\mathbf{v}, \alpha\mathbf{y}) = \bar{\alpha}(\mathbf{v}, \mathbf{y}) = \bar{\alpha}L_{\mathbf{y}}(\mathbf{v})$$

- It follows by such a mapping that $\mathbf{b}'_k = \mathbf{b}_k$ for each k .
- **Conjugate linear** (transformation): A transformation T such that

$$T(\alpha\mathbf{x} + \beta\mathbf{y}) = \bar{\alpha}T\mathbf{x} + \bar{\beta}T\mathbf{y}$$

- It is customary to write outputs of linear functionals $L(\mathbf{v})$ in the form $\langle \mathbf{v}, L \rangle$.
 - This expression is linear in both arguments, unlike the inner product.
- Defines the dual transformation as the unique transformation such that

$$\langle A\mathbf{x}, \mathbf{y}' \rangle = \langle \mathbf{x}, A'\mathbf{y}' \rangle$$

for all $\mathbf{x} \in X, \mathbf{y}' \in Y'$.

- It's matrix in the standard bases equals A^T .
- Annihilators are denoted by E^\perp here.
- Proposition 8.3.6: The annihilator of the annihilator of E equals E .
- Let $A : X \rightarrow Y$ be an operator acting from one vector space to another. Then
 1. $\ker A' = (\text{range } A)^\perp$.
 2. $\ker A = (\text{range } A')^\perp$.
 3. $\text{range } A = (\ker A')^\perp$.
 4. $\text{range } A' = (\ker A)^\perp$.

Chapter 9

Advanced Spectral Theory

9.1 Notes

- 10/22:
- Let $p(z) = \sum_{i=0}^n a_i z^i$ be a polynomial. Let A be an $n \times n$ matrix. We let $p(A) = \sum_{i=0}^n a_i A^i$.
 - Theorem: If A is an $n \times n$ and $p(\lambda) = \det(A - \lambda I)$, then $p(A) = 0$.
 - We know that $p(\lambda) = a(z - \lambda_1) \cdots (z - \lambda_n)$ where $\lambda_1, \dots, \lambda_n$ are the eigenvalues.
 - Thus $p(A) = a(A - \lambda_1 I) \cdots (A - \lambda_n I)$.
 - If you are in \mathbb{R}^n and have this property, you can factorize your matrix.
 - Thus, $p(A)\mathbf{x} = \mathbf{0}$ since \mathbf{x} can be decomposed into a linear combination of eigenvectors of A , which will be taken to 0 one by one by the terms of $p(A)$.
 - $\sigma(B) = \{\text{eigenvalues of } B\}$ is known as the **spectrum** of B .
 - If p is an arbitrary polynomial and A is $n \times n$, then μ is an eigenvalue of $p(A)$ if and only if $\mu = p(\lambda)$ where λ is an eigenvalue of A . In essence, $\sigma(p(A)) = p(\sigma(A))$.
 - Chapter 9 will not be on the exam. We don't have to know the generalization to infinite dimensional spaces.
- 10/25:
- If A is an $n \times n$ square matrix and $p(\lambda) = \det(A - \lambda I)$, then $p(A) = 0$.
 - Proof: WLOG, let A be an upper triangular matrix with diagonal entries equal to the eigenvalues.
 - Think of $p(z) = (-1)^n (z - \lambda_1) \cdots (z - \lambda_n)$.
 - Thus, $p(A) = (-1)^n (A - \lambda_1 I) \cdots (A - \lambda_n I)$.
 - WTS: $p(A)\mathbf{x} = 0$ for all $\mathbf{x} \in V$.
 - Let $E_k = \text{span}(e_1, \dots, e_k)$ be the span of the first k eigenvectors of A , where e_1, \dots, e_n is a standard basis in \mathbb{C}^n .
 - A triangular implies $AE_k \subset E_k$. Thus, $(A - \lambda I)E_k \subset E_k$, so E_k is invariant under $A - \lambda I$ for all λ .
 - If we apply $A - \lambda_k I$ to a vector in E_k , we are left with a vector in E_{k-1} .
 - Thus, if we apply $\prod_{k=1}^n (A - \lambda_k I) = p(A)$ to any vector in $E_n = V$, we will kill it piece by piece down to zero.
 - Let A be a square $n \times n$ matrix. Then p an arbitrary polynomial implies $\sigma(p(A)) = p(\sigma(A))$. (Any eigenvalue μ of $p(A)$ is $\mu = p(\lambda)$, where λ is an eigenvalue of A .)
 - Shows that polynomials of operators commute.

- Proof: Let λ be an eigenvalue of A . We want to show that $p(\lambda)$ is an eigenvalue of $p(A)$. This is obvious since $A\mathbf{x} = \lambda\mathbf{x}$ for some \mathbf{x} , so $A^k\mathbf{x} = \lambda^k\mathbf{x}$, so in particular, $p(A)\mathbf{x} = p(\lambda)\mathbf{x}$.
- On the other hand, if μ is an eigenvalue of $p(A)$, we want to show that there exists $\lambda \in \sigma(A)$ such that $\mu = p(\lambda)$.
- Consider $q(z) = p(z) - \mu$. Then $q(A) = p(A) - \mu I$. Since μ is an eigenvalue of $p(A)$, $q(A)$ is not invertible.
- Thus, $q(z) = (-1)^n(z - z_1) \cdots (z - z_n)$ and $q(A) = (-1)^n(A - z_1 I) \cdots (A - z_n I)$.
- But $q(A)$ is not invertible, so one of the $A - z_k I$ is not invertible. Take z_k such that $A - z_k I$ is not invertible. Then $z_k \in \sigma(A)$. It follows that $q(z_k) = p(z_k) - \mu = 0$.
- If A is $n \times n$, $\lambda_1, \dots, \lambda_n$ are its eigenvalues, p is a polynomial, then $p(A)$ is invertible if and only if $p(\lambda_k) \neq 0$ for each $k = 1, \dots, n$.
 - This is an immediate corollary to the previous result.
- We now build up to the **generalized eigenspace**, which is related to some “geometric” properties of the algebraic multiplicity of an eigenvalue.
- If $A : V \rightarrow V$ is a linear operator and $E \subset V$ is a subspace, E is A -invariant if $AE \subset E$.
- Facts:
 - If E is A -invariant, E is A^k -invariant.
 - Thus, E is $p(A)$ -invariant.
- Consider the restriction map $A|_E$.
- A has a block-diagonalized matrix where each block corresponds to the generalized eigenvectors of a generalized eigenvalue of A .
 - Let E_1, \dots, E_r be a **basis of invariant subspaces**.
 - Let $A_k = A|_{E_k}$. Then the A_k ’s act independently of each other.
- **Generalized eigenvector** (of A): A vector \mathbf{v} corresponding to an eigenvalue λ if there exists $k \geq 1$ such that $(A - \lambda I)^k \mathbf{v} = \mathbf{0}$.
- **Generalized eigenspace**: The set E_λ of all of the generalized eigenvectors of λ . *Given by*

$$E_\lambda = \bigcup_{k \geq 1} \ker(A - \lambda I)^k$$

- E_λ is a linear subspace of V .
- **Degree** (of λ): The smallest number k such that increasing k any more does not add further vectors to the generalized eigenspace. *Denoted by $d(\lambda)$. Also known as **depth**.*
 - Symbolically, $d(\lambda)$ is the smallest number such that

$$E_\lambda = \bigcup_{k=1}^{d(\lambda)} \ker(A - \lambda I)^k$$

- Start working through the first 25 problems of Rudin (1976) (his metric spaces problems).
- 10/27: • Jordan form.
- Reviews build up to generalized eigenvectors.

- Theorem: If $\sigma(A) = \{\lambda_1, \dots, \lambda_n\}$ and E_1, \dots, E_n are the corresponding generalized eigenspaces, then E_1, \dots, E_n is a basis of subspaces of U , i.e., $V = \oplus_k E_k$.
- Corollary: $A : V \rightarrow V$ can be represented as $A = D + N$ where D is diagonalizable and N is nilpotent and $ND = DN$.

– Proof: Consider the basis of generalized eigenspaces known to exist from the theorem. Then $A = \text{diag}\{A_1, \dots, A_r\}$.

– Let

$$N_k = A_k - \lambda_k I_{E_k}$$

This is nilpotent.

– Then let

$$D = \text{diag}\{\lambda_1 I_{E_1}, \dots, \lambda_n I_{E_n}\}$$

– These two matrices satisfy the necessary properties.

- Let $\dot{\mathbf{x}} = A\mathbf{x}$.

– Let $\mathbf{x}(t) = e^{tA}$, where

$$e^{tA} = \sum \frac{(tA)^k}{k!}$$

– $\|e^{tA}\| \leq \sum \frac{\|A^k\|}{k!} = \sum \frac{\|A\|^k}{k!}$.

– Let p be a polynomial of degree k . Then

$$p(a+x) = \sum_{k=0}^d \frac{p^{(k)}(a)}{k!} x^k$$

– If $A = D + N$, then...

- Nilpotent operators:

– Let $A = \text{diag}\{A_1, \dots, A_r\}$.

– We know that $A_k = \lambda_k I_{E_k} + N_k$ for each k .

– Every nilpotent N can be written in the form

$$\begin{pmatrix} 0 & 1 & & 0 \\ & \ddots & \ddots & \\ & & \ddots & 1 \\ 0 & & & 0 \end{pmatrix}$$

10/28:

- The exam is long but not that hard. The only question is will you do good or very good.
- Revise the previous two homeworks, especially the last two.
- No justification for any of the true/false questions. Just circle T or F.
 - There are four problems. One is T/F (with multiple subparts); the other three are problem problems (with subparts, too).
 - Some of the questions will take you 2 seconds. Some you've already seen in the PSets.
 - The exam is supposed to be boring.
- Calculators?
 - No calculators needed. Calculators are for chem/physics exams.

- Not a lot of computation.
- 50 minutes.
- Chloe will be proctoring.
- Remember the determinant of “special” matrices.
 - $|\det U| = 1$ if U is unitary.
 - $\det A = \pm 1$ if A is orthogonal.
 - Make a list of matrix types that are automatically diagonalizable.
 - Determinant is the product of the eigenvalues.
 - Determinant of A is equal to the conjugate of the determinant of A^* .
- Most of the exercises use the inner product.
 - Whenever you had something to prove about eigenvalues or eigenbasis, you went through diagonalization or SVD or the inner product or polar decomposition.
 - Proving eigenvalues of self-adjoint matrices are real w/ the inner product.
- Eigenvalues/eigenvectors of a projection.
 - It’s implied that it’s asking you the multiplicities!!!
- Know useful facts but have an idea how to prove them as well.
- Recommends against shorthanding in the exams.
- Not grading on clarity (since the exam is long).
- Max and min are for when you’re sure something will be attained. Otherwise use sup and inf.

9.2 Chapter 9: Advanced Spectral Theory

- 10/28:
- Theorem 9.1.1 (Cayley-Hamilton): If p is the characteristic polynomial of A , $p(A) = 0$.
 - Theorem 9.2.1 (Spectral Mapping Theorem): For a square matrix A and an arbitrary polynomial p , $\sigma(p(A)) = p(\sigma(A))$. In other words, μ is an eigenvalue of $p(A)$ if and only if $\mu = p(\lambda)$ for some eigenvalue λ of A .
 - Corollary 9.2.2: Let A be a square matrix with eigenvalues $\lambda_1, \dots, \lambda_n$ and let p be a polynomial. Then $p(A)$ is invertible iff $p(\lambda_k) \neq 0$ for all $k = 1, \dots, n$.
 - Algebraic multiplicity is the dimension of the corresponding generalized eigenspace.

Part II

Point Set Topology of Metric Spaces

Chapter 1

The Real and Complex Number Systems

1.1 Notes

- 11/1:
- Spent a lot of time trying to cheer us up regarding the midterm.
 - There may be some true/false on linear algebra on the final.
 - Facts:
 1. $\sqrt{2}$ is irrational.
 2. Archimedes principle: If $x > 0$ and $y \in \mathbb{R}$, then there exists n such that $nx > y$.
 3. If $x > y$, then there exists $q \in \mathbb{Q}$ such that $x > q > y$.

1.2 Chapter 1: The Real and Complex Number Systems

From Rudin (1976).

- 11/6:
- Rudin (1976) presents several interesting proofs throughout this section that may be of interest later by means of their divergence from the ones with which I am familiar.
 - **Least-upper-bound property:** The property pertaining to a set S that if $E \subset S$, $E \neq \emptyset$, and E is bounded above, then $\sup E \in S$.
 - For example, \mathbb{Q} does not have the least-upper-bound property.
 - The **greatest-lower-bound property** is analogously defined.
 - Theorem: Suppose S is an ordered set with the least-upper-bound property, $B \subset S$ is nonempty, and B is bounded below. Let L be the set of all lower bounds of B . Then $\alpha = \sup L$ exists in S , and $\alpha = \inf B$. In particular, $\inf B$ exists in S .
 - Essentially, this theorem states that any set that satisfies the least-upper-bound property satisfies the greatest lower bound property.
 - **Existence theorem:** There exists an ordered field \mathbb{R} which has the least-upper-bound property. Moreover, \mathbb{R} contains \mathbb{Q} as a subfield.
 - The second statement implies that the operations of addition and multiplication on \mathbb{R} , when applied to \mathbb{Q} , coincide with the operations of addition and multiplication on \mathbb{Q} .

- **Archimedean property** (of \mathbb{R}): If $x \in \mathbb{R}$, $y \in \mathbb{R}$, and $x > 0$, then there is a positive integer n such that $nx > y$.
- Rudin (1976) proves several theorems about the real numbers from the least-upper-bound property as opposed to the traditional construction of the real numbers.
- Introduces the decimal system.
- **Finite real number system**: That which has been defined thus far.
- **Extended real number system**: The set $\mathbb{R} \cup \{+\infty, -\infty\}$ where $+\infty, -\infty$ obey the expected properties (supremum [resp. infimum] of every set, $x + \infty = \infty$, etc.).
- Defines the complex field axiomatically with complex numbers in the form (a, b) for $a, b \in \mathbb{R}$.
 - Notes that the real numbers form a subfield of the complex field.
 - Defines $i = (0, 1)$, proves $i^2 = -1$, proves $a + bi = (a, b)$.
- **Schwarz inequality**: If a_1, \dots, a_n and b_1, \dots, b_n are complex numbers, then

$$\left| \sum_{j=1}^n a_j \bar{b}_j \right|^2 \leq \sum_{j=1}^n |a_j|^2 \sum_{j=1}^n |b_j|^2$$

- **Euclidean k -space**: The vector space \mathbb{R}^k over the real field.

Chapter 2

Basic Topology

2.1 Notes

11/1:

- Equivalence relationships are denoted $A \sim B$.
 - These are...
 - Reflexive ($A \sim A$).
 - Symmetric ($A \sim B \iff B \sim A$).
 - Transitive ($A \sim B \ \& \ B \sim C \implies A \sim C$).
 - Equivalence relations give rise to equivalence classes.
- **Countable** (set A): A set A such that $A \sim \mathbb{N}$, in the sense that there exists a one-to-one and onto map from $\mathbb{N} \rightarrow A$.
 - Alternatively, A can be written in the form $A = \{f(n) : n \in \mathbb{N}\}$.
- **Finite countable** vs. **infinite countable** (see Rudin (1976)).
- \mathbb{N} denotes the natural numbers.
- \mathbb{N}_0 denotes the natural numbers including 0.
- \mathbb{Z} denotes the integers.
- We know that $\mathbb{N} \sim \mathbb{Z}$: Let $f : \mathbb{N} \rightarrow \mathbb{Z}$ be defined by

$$f(n) = \begin{cases} \frac{n}{2} & n \text{ even} \\ \frac{n-1}{2} & n \text{ odd} \end{cases}$$

- More facts.
 1. Every subset of a countable set is countable.
 2. Unions of countable sets are countable.
 - If the sets E_n for some finite list of numbers are countable, then $\bigcup_n E_n$ is countable.
 - Soug goes over the diagonalization method of counting.
 3. n -fold Cartesian products of countable sets are countable (we induct on n).
 - If A is countable and B is countable, then $A \times B$ is countable.
 - If A is finite and to each $\alpha \in A$ we assign a countable set E_α , $\bigotimes_{\alpha \in A} E_\alpha$ is countable.
- **Metric space**: A space X along with a metric $d : X \times X \rightarrow [0, \infty)$ such that

- $d(x, y) > 0$ iff $x \neq y$, and $d(x, x) = 0$ iff $x = 0$.
- $d(x, y) = d(y, x)$.
- $d(x, y) \leq d(x, z) + d(z, y)$.

- Example (\mathbb{R}^n):

- We may define d by

$$d(x, y) = \sqrt{\sum (x_i - y_i)^2}$$

- We can also define the p -metrics (recall normed spaces) with p where 2 is.

- Example ($X_p = \{f : Y \rightarrow \mathbb{R} : 1 \leq p < \infty, \int_Y |f|^p dy < \infty\}$):

- This is ℓ_p .
- Define

$$\|f - g\|_p = \left[\int_Y |f - g|^p dy \right]^{1/p}$$

- Convergence: $x_n \rightarrow x \iff d(x_n, x) \rightarrow 0$.

- **Neighborhood**: The set of all points a distance less than r away from p . Denoted by $N_r(p)$. Given by

$$N_r(p) = \{q \in X : d(p, q) < r\}$$

- **Limit point** (of E): A point p such that every neighborhood of p intersects E at a point other than p . Also known as **accumulation point**.

- Symbolically,

$$N_r(p) \cap (E \setminus \{p\}) \neq \emptyset$$

for all $r > 0$.

- **Isolated point** (of E): A point p such that $p \in E$ and p is not a limit point of E .

- **Closed** (set E): A set E that contains all of its limit points.

- **Interior** (point p): A point p such that there exists $N_r(p) \subset E$.

- **Open** (set E): A set E , all points of which are interior points.

- **Perfect** (set E): A set E that is closed and every point of E is a limit point of E .

- **Bounded** (set E): There exists a number M and a $y \in X$ such that $E \subset \{p : d(p, y) \leq M\}$.

- **Dense** (set E in X): A set E such that every point of X is a limit point of E or a point of E , itself.

11/3:

- Every neighborhood is an open set.

- If p is a limit point of E , every neighborhood of p contains infinitely many points of E .

- Thus, a finite set cannot have a limit point.

- Prove by contradiction: Suppose there is a neighborhood that contains only finitely many points of E . Then the neighborhood with radius smaller than the distance to the closest point does not contain any points of E , a contradiction.

- E is open iff $E^{c[1]}$ is closed.

- Assume E^c closed. If $p \in E$, then p is not a limit point of E^c . It follows that there exists a neighborhood of p that is entirely contained within E , so p is interior, as desired.

¹The complement of E .

- Suppose E is open. Let p be any limit point of E^c . Then $p \in E^c$.
- F is closed iff F^c is open.
- If $(G_\alpha)_{\alpha \in A}$ is a family of open sets in X , then the union is open.
 - Let $p \in \bigcup_{\alpha \in A} G_\alpha$. Then $p \in G_\alpha$ for some $\alpha \in A$. It follows that p is an interior point of G_α , so thus an interior point of the union of G_α with everything else.
- Finite intersections of open sets are open.
 - In the infinite case $\bigcap_{n \in \mathbb{N}} (-1/n, 1/n) = \{0\}$, an intersection of infinitely many open sets is closed.
 - However, in the finite case, just consider the neighborhood with the smallest radius and take this one.
- The intersection of closed sets is closed.
- The union of finitely many closed sets is closed.
 - These follow from the previous two by De Morgan's rule.
- Let $\bar{E} = E \cup E'$ where E' is the set of limit points of E .
- Let X be a metric space and $E \subset X$. Then
 1. \bar{E} is closed.
 - WTS: \bar{E}^c is open. Let $p \in \bar{E}^c$. Then p is neither in E nor is it a limit point of E . Thus, there exists a neighborhood of \bar{E}^c containing entirely points of \bar{E}^c . Therefore, \bar{E}^c is open, so \bar{E} is closed.
 2. $E = \bar{E}$ iff E is closed.
 - Think $p \in \bigcap G_\alpha$?
 3. $\bar{E} \subset F$ for any closed $F \supset E$.
 - If $E \subset F$, then any limit point of E will be a limit point of F . Thus, $E' \subset F'$. Then $\bar{E} = E \cup E' \subset F \cup F' = \bar{F} = F$ where the last equality holds because F is closed.
- Types of sets.

	Closed	Open	Perfect	Bounded
$\{z \in \mathbb{Q} : z < 1\}$	N	Y	N	Y
$\{z \in \mathbb{Q} : z \leq 1\}$	Y	N	Y	Y
Nonempty finite set	Y	N	N	Y
\mathbb{Z}	Y	N	N	N
$\{1/n : n \in \mathbb{N}\}$	N	N	N	Y
\mathbb{R}^2	Y	Y	Y	N
(a, b)	N	?	N	Y

Table 2.1: Types of sets.

- **Relatively open** (set E to Y): A set $E \subset Y \subset X$ such that if $p \in E$, then there exists a Y -neighborhood of E contained in E .
- Let $N_r^X(p) = \{y \in X : d(y, p) < r\}$ be a neighborhood of p in X , and let $N_r^Y(p) = \{y \in Y : d(y, p) < r\}$ be a neighborhood of p in Y . Then $N_r^Y(p) = N_r^X(p) \cap Y$.

- E is open relative to Y iff $E = G \cap Y$ where G is open relative to X .
- Introduces the supremum.
- If $E \subset \mathbb{R}$, $E \neq \emptyset$, and E is bounded above, $\sup E < \infty$.
- Let $y = \sup E$. Then $y \in \bar{E}$.
- There exists a sequence $a_n \in A$ such that $a_n \rightarrow x = \sup A$.
- A is compact iff any open cover of the set has a finite subcover.
- Study and *know* all of these proofs.

11/5:

- Compactness: Defines compactness in terms of open covers.
- Finite sets are compact.
- Compactness is “absolute” (i.e., it is not a relative property like openness).
 - If $K \subset Y \subset X$, then K is compact relative to X iff K is compact relative to Y .
 - V is open relative to Y iff $V = G \cap Y$ where G is open relative to X .
- Compact implies closed.
 - We will show K compact implies K^c open.
 - WTS: For all $p \in K^c$, there exists $N_r(p) \subset K^c$ such that $N_r(p) \cap K = \emptyset$.
- A closed subset of a compact set is compact.
 - Let K be compact and let $F \subset K$ be closed.
 - Take any open cover of F . Extend it to an open cover of K . Take the finite subcover of K . Naturally, this finite subcover is also a finite cover of $F \subset K$.
- F closed, K compact implies $F \cap K$ compact.
- If $(K_\alpha)_{\alpha \in A}$ is compact in X with finite intersection property (every intersection of any finite number of these sets is nonempty), then $\bigcap_{\alpha \in A} K_\alpha \neq \emptyset$.
 - Argue by contradiction.
 - Let $G_\alpha = K_\alpha^c$.
 - Assume the intersection is empty. Assume WLOG that no point of K_1 is in any of the other K_α ’s.
 - Then $\{G_\alpha\}_{\alpha \in A}$ be an open cover of K_1 .
 - K_1 compact implies there is a finite subcover $G_{\alpha_1}, \dots, G_{\alpha_n}$. Then $K \subset G_{\alpha_1} \cup \dots \cup G_{\alpha_n}$. This implies that $K_1 \cap K_{\alpha_1} \cap \dots \cap K_{\alpha_n} = \emptyset$, a contradiction.
- Let E be an infinite subset of a compact K . Then E has a limit point in K .
 - Argue by contradiction.
 - Suppose for all $p \in K$, there exists $N_r(p)$ such that $N_r(p) \cap E = \{p\}$.
 - Consider the set $\{N_r(p) : p \in K\}$. This is an open cover of K . Thus, there exists a finite subcover of it. But since $E \subset K \subset N_{r_1}(p_1) \cup \dots \cup N_{r_n}(p_n) = \{p_1\} \cup \dots \cup \{p_n\}$, E is finite, a contradiction.
- **2-cell** (in \mathbb{R}^2): A set that is the Cartesian product of two closed intervals.
 - Generalizes to **k-cells**.
- Let $I_n = [a_n, b_n] \subset \mathbb{R}$ such that $I_{n+1} \subset I_n$. Then $\bigcap_{n=1}^{\infty} I_n \neq \emptyset$.

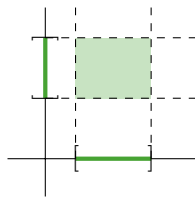
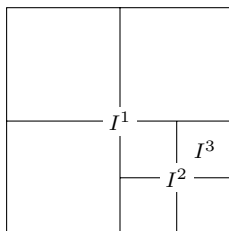


Figure 2.1: 2-cell.

- Let I_k be a k -cell in \mathbb{R}^k such that $I_k \supset I_{k+1}$. Then $\bigcap_k I_k \neq \emptyset$.
- We know that $a_m \leq a_{m+n} \leq b_{m+n} \leq b_m$, so $\sup a_n \in \bigcap I_n$.
- Every k -cell is compact.

Figure 2.2: k -cells are compact.

- Argue by contradiction.
- Consider an open cover of the k -cell I^1 . If it has a finite subcover, we're done. So suppose we have an open cover that doesn't have a finite subcover. Split the k -cell into 2^k chunks. At least one of the chunks I^2 must not have a finite subcover.
- Split that one into 2^k chunks. At least one of the chunks I^3 must not have a finite subcover.
- Continue.
- Thus, we have a decreasing family of k -cells, so by the previous result, their $\bigcap I^n \neq \emptyset$.
- Let $x \in \bigcap I^n$. Then the...
- Heine-Borel theorem: Let $E \subset \mathbb{R}^k$. Then TFAE^[2]
 1. E is closed and bounded.
 2. E is compact.
 3. Every infinite subset of E has a limit point in E .
 - $(1 \Rightarrow 2)$ E closed and bounded implies E is a closed subset of some I_k , so it's compact.
 - $(2 \Rightarrow 3)$ Already done.
 - $(3 \Rightarrow 1)$
 - Suppose E not bounded. Then there is an infinite sequence of points in E that never converges. Contradiction.
 - Suppose E is not closed. Then there exists a sequence of points in E which “converges” to an $x_0 \notin E$.

11/8: • Hewitt and Stromberg (1965) has harder analysis problems than Rudin (1976).

- Theorem: If P is a nonempty perfect subset of \mathbb{R}^k , then P is uncountable.

²The following are equivalent.

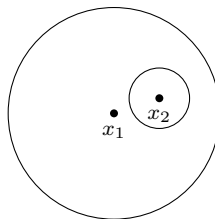


Figure 2.3: Nonempty perfect sets are uncountable.

- P perfect implies P infinite.
 - Suppose P is countable. Let $P = \{x_1, x_2, \dots\}$.
 - Start with x_1 . Take an open neighborhood V_1 of x_1 . Since x_1 is a limit point of P , there will be another point $x_2 \in P$ in V_1 . Choose V_2 to be a neighborhood of x_2 such that $\bar{V}_2 \subset V_1$.
 - Keep going — there is a point $x_3 \in P$ in V_2 , choose an appropriate neighborhood, etc.
 - Thus, we have a sequence of closed compact sets such that $\bar{V}_n \supset \bar{V}_{n+1}$ ($n \in \mathbb{N}$). It follows that $\bigcap \bar{V}_n \neq \emptyset$.
 - We also know that $V_n \cap P \neq \emptyset$ for each n .
 - Let $K_n = \bar{V}_n \cap P$. Each K_n is compact and $K_n \supset K_{n+1}$ for each n . Therefore, by compactness, $\bigcap K_n \neq \emptyset$. But the construction implies that $\bigcap K_n = \emptyset$ because we exhausted the whole sequence of possible points $x_i \in P$.
- Corollary: Any interval is uncountable.
 - The Cantor set:

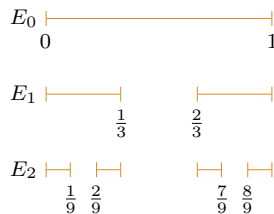


Figure 2.4: Constructing the Cantor set.

- Let $E_0 = [0, 1]$.
- Take out the middle third, so that $E_1 = [0, 1/3] \cup [2/3, 1]$.
- Take out the middle thirds of the remaining intervals and keep going.
- Thus, we are building a decreasing family of compact sets, so the overall intersection $E = \bigcap E_n$ of every set is nonempty.
- E^n is the union of 2^n closed intervals of length $n/3$. Thus, the overall length of E^n is $(2/3)^n$.
- Thus, we have a compact nonempty set with Lebesgue measure zero.
- E does not contain any segment of the form

$$\left(\frac{3k+1}{3^m}, \frac{3k+2}{3^m} \right)$$

for $k, m \in \mathbb{N}$.

- Therefore, no segment of the form (α, β) is contained in E (any segment of said form contains a segment of the above form).

- Moreover, E (the Cantor set) is perfect.
 - Let $x \in E$. WTS: For all segments S containing x , $S \cap (E \setminus \{x\}) \neq \emptyset$.
 - Let S be an arbitrary such segment...
- Consider the **Devil's staircase**.
 - $0 = \int_0^1 F'(x) dx = F(1) - F(0) = 1$. This function does not obey the fundamental theorem of calculus. A function satisfies the fundamental theorem of calculus if and only if it is absolutely continuous.
- Connected sets (motivation):
 - In a convex set, you can connect any two points with a straight line.
 - In a nonconvex connected set, there exist points that you must connect with a curve.
 - In a disconnected set, there exist points that cannot be connected via a line whose points lie wholly in the set.
- **Connected** (set E): A set E that is not the union of two **separated** sets.
- **Separated** (sets A, B): Two sets $A, B \subset X$ that are nonempty and such that $\bar{A} \cap B = \emptyset$, and $A \cap \bar{B} = \emptyset$.
- Theorem: $E \subset \mathbb{R}$ is connected iff $x, y \in E$ and $x < z < y$ implies $z \in E$.
 - If there is a $z \notin E$ between x, y , then $\{x \in E : x < z\}$ and $\{x \in E : z < y\}$ are separated sets, so E is not connected.

2.2 Chapter 2: Basic Topology

From Rudin (1976).

- 11/6:
- **Countable** (set A): A set A that is in bijective correspondence with the set of all positive integers. Also known as **enumerable**, **denumerable**.
 - **At most countable** (set A): A set A that is finite or countable.
 - An alternative definition of an **infinite** set would be a set that is equivalent to one of its proper subsets.
 - Theorem: Let A be the set of all sequences whose elements are the digits 0 and 1. This set A is uncountable.

Proof. Let $E = \{s_1, s_2, \dots\}$ be an arbitrary countable subset of A , where each s_j is a sequence whose elements are the digits 0 and 1. Let s be the sequence, the n^{th} term of which is the opposite of the n^{th} term of s_n (i.e., if the n^{th} term of s_n is 0, we set the n^{th} term of s equal to 1). This guarantees that s is distinct from each of the s_j , i.e., that $s \notin E$. It follows that $E \subsetneq A$, i.e., that every countable subset of A is a proper subset of A . Therefore, A must be uncountable (for otherwise A would be a proper subset of A , a contradiction). \square

- The idea of this proof is called **Cantor's diagonalization process**.
- Since every real number can be represented as a binary sequence of numbers, i.e., $A \sim \mathbb{R}$, the reals are uncountable.
- **Metric space**: A set X such that with any two points $p, q \in X$, there is associated a real number $d(p, q)$ such that
 1. $d(p, q) > 0$ if $p \neq q$; $d(p, p) = 0$.
 2. $d(p, q) = d(q, p)$.

3. $d(p, q) \leq d(p, r) + d(r, q)$ for any $r \in X$.

- **Distance** (from $p \in X$ to $q \in X$, X a metric space): The real number $d(p, q)$.
- **Distance function**: A function $d : X \times X \rightarrow \mathbb{R}$ that sends $(p, q) \mapsto d(p, q)$. Also known as **metric**.
- Every subset of a metric space is a metric space in its own right under the same distance function.
- **Segment** (from a to b): The set of all real numbers x such that $a < x < b$. Denoted by (a, b) .
- **Interval** (from a to b): The set of all real numbers x such that $a \leq x \leq b$. Denoted by $[a, b]$.
- **k -cell**: The set of all points $\mathbf{x} = (x_1, \dots, x_k) \in \mathbb{R}^k$ whose coordinates satisfy the inequalities $a_i \leq x_i \leq b_i$ where $a_i < b_i$ for each $1 \leq i \leq k$.

– Note that a 1-cell is an interval and a 2-cell is a rectangle.

- **Convex** (set E): A subset E of \mathbb{R}^k such that

$$\lambda \mathbf{x} + (1 - \lambda) \mathbf{y} \in E$$

for all $\mathbf{x}, \mathbf{y} \in E$ and $0 < \lambda < 1$.

– Balls and k -cells are both convex.

- The segment (a, b) is open as a subset of \mathbb{R}^1 , but not open as a subset of \mathbb{R}^2 .
- Since compactness is not relative, while it makes no sense to talk about *open* or *closed* metric spaces, it does make sense to talk about *compact* metric spaces.
- **Weierstrass theorem**: Every bounded infinite subset of \mathbb{R}^k has a limit point in \mathbb{R}^k .
- Theorem: Let P be a nonempty perfect set in \mathbb{R}^k . Then P is uncountable.

Proof. Since P is nonempty and perfect, there exists a limit point of P . It follows that P is infinite.

Now suppose for the sake of contradiction that P is countable, and denote the elements of P by $\mathbf{x}_1, \mathbf{x}_2, \dots$. We now construct a sequence $\{V_n\}$ of neighborhoods, as follows. Let $V_1 = N_r(\mathbf{x}_1)$. Clearly, $V_1 \subset P$ since $\mathbf{x}_1 \in P$. It follows that since V_1 is a neighborhood that V_1 contains infinitely many points of P . Now suppose inductively that V_n has been constructed. Thus, by analogous conditions to those on V_1 , we may let V_{n+1} be a neighborhood such that (i) $\bar{V}_{n+1} \subset V_n$, (ii) $\mathbf{x}_n \notin \bar{V}_{n+1}$, and (iii) $V_{n+1} \cap P$ is nonempty. By (iii), we can continue on to construct V_{n+2} , and so on and so forth.

Let $K_n = \bar{V}_n \cap P$. Since \bar{V}_n is closed and bounded, \bar{V}_n is compact. Additionally, since $\mathbf{x}_n \notin K_{n+1}$ for each n , no point of P lies in $\bigcap_1^\infty K_n$. Thus, since each $K_n \subset P$, $\bigcap_1^\infty K_n$ is empty. But this contradicts our previous result that since each K_n is nonempty, compact, and such that $K_n \supset K_{n+1}$, $\bigcap_1^\infty K_n$ is nonempty. \square

- Corollary: Every interval $[a, b]$ is uncountable. In particular, \mathbb{R} is uncountable.
- **Cantor set**: The set resulting from the following construction. Let $E_0 = [0, 1]$. Remove the segment $(1/3, 2/3)$, so that $E_1 = [0, 1/3] \cup [2/3, 1]$. Now remove the middle third of these two intervals to create E_2 . Continue on indefinitely.
 - This is a perfect set in \mathbb{R}^1 which contains no segment.
- **Separated** (sets A, B): Two subsets A, B of a metric space X such that $A \cap \bar{B}$ and $\bar{A} \cap B$ are empty.
- **Connected** (set E): A set E that is not the union of two nonempty separated sets.
- Separated sets are disjoint, but disjoint sets are not necessarily separated (consider $[0, 1]$ and $(1, 2)$).
- A subset E of \mathbb{R}^1 is connected if and only if it has the following property: If $x, y \in E$ and $x < z < y$, then $z \in E$.

Chapter 3

Numerical Sequences and Series

3.1 Notes

- 11/8: • Any bounded sequence in \mathbb{R}^k has a convergent subsequence.
- 11/10: • Read and understand the section about Cauchy sequences converging and the sup/inf.

3.2 Chapter 3: Numerical Sequences and Series

From Rudin (1976).

- 11/7: • Convergence of sequences is relative.
- For example, the sequence $1/n$ for $n = 1, 2, \dots$ converges in \mathbb{R} , but not in $(0, \infty)$.
- **Range** (of $\{p_n\}$): The set of all points p_n .
- This definition squares nicely with the formal definition of a sequence as a function p defined on \mathbb{N} .
- Theorem 3.6a: If $\{p_n\}$ is a sequence in a compact metric space X , then some subsequence of $\{p_n\}$ converges to a point of X .
- Theorem 3.7: The subsequential limits of a sequence $\{p_n\}$ in a metric space X form a closed subset of X .
- **Diameter** (of E): The supremum of the set

$$S = \{d(p, q) : p, q \in E\}$$

where E is a nonempty subset of a metric space X . Denoted by **diam** E .

- Theorem 3.10:
- (a) If \bar{E} is the closure of a set E in a metric space X , then
$$\text{diam } \bar{E} = \text{diam } E$$
 - (b) If K_n is a sequence of compact sets in X such that $K_n \supset K_{n+1}$ ($n = 1, 2, 3, \dots$) and if $\lim_{n \rightarrow \infty} \text{diam } K_n = 0$, then $\bigcap_1^\infty K_n$ consists of exactly one point.
- **Complete** (metric space): A metric space in which every Cauchy sequence converges.
- All compact metric spaces and all Euclidean spaces are complete.

- The metric space $(\mathbb{Q}, |x - y|)$ is not complete.
- **Monotonically increasing** (sequence $\{s_n\}$): A sequence $\{s_n\}$ of real numbers such that $s_n \leq s_{n+1}$ for each $n \in \mathbb{N}$.
- **Monotonically decreasing** (sequence $\{s_n\}$): A sequence $\{s_n\}$ of real numbers such that $s_n \geq s_{n+1}$ for each $n \in \mathbb{N}$.
- **Monotonic sequences**: The class of all sequences that are either monotonically increasing or monotonically decreasing.
- **Upper limit** (of $\{s_n\}$): The supremum of the set E of all subsequential limits of $\{s_n\}$. *Denoted by s^* , $\limsup_{n \rightarrow \infty} s_n$.*
- **Lower limit** (of $\{s_n\}$): The infimum of the set E of all subsequential limits of $\{s_n\}$. *Denoted by s_* , $\liminf_{n \rightarrow \infty} s_n$.*
- Theorem 3.17: Let $\{s_n\}$ be a sequence of real numbers. Then s^* has (and is the only number to have both of) the following two properties.
 - (a) $s^* \in E$.
 - (b) If $x > s^*$, then there is an integer N such that $n \geq N$ implies $s_n < x$.

An analogous result holds for s_* .

11/8:

- Series are defined in terms of sequences. Moreover, sequences can be defined in terms of series: Let $a_1 = s_1$, $a_n = s_n - s_{n-1}$ ($n \in \mathbb{N} + 1$). Thus, every theorem about sequences can be stated in terms of series and vice versa, but it is nevertheless useful to consider both concepts (Rudin, 1976, p. 59).
- Theorem 3.27: Suppose $\{a_n\}$ is a monotonically decreasing sequence of nonnegative terms. Then the series $\sum_{n=1}^{\infty} a_n$ converges if and only if the series

$$\sum_{k=0}^{\infty} 2^k a_{2^k} = a_1 + 2a_2 + 4a_4 + 8a_8 + \cdots$$

converges.

- Theorem 3.29: If $p > 1$,

$$\sum_{n=2}^{\infty} \frac{1}{n(\log n)^p}$$

converges; if $p \leq 1$, the series diverges.

- Note that $\log n = \ln n$.
- Note that we sum from $n = 2$ since $\log 1 = 0$.
- **e**: The number

$$e = \sum_{n=0}^{\infty} \frac{1}{n!}$$

- Theorem 3.31: $\lim_{n \rightarrow \infty} (1 + 1/n)^n = e$.
- Theorem 3.32: e is irrational.
- Theorem 3.39: Given the power series $\sum c_n z^n$, put

$$\alpha = \limsup_{n \rightarrow \infty} \sqrt[n]{|c_n|} \qquad R = \frac{1}{\alpha}$$

(If $\alpha = 0$, let $R = +\infty$; if $\alpha = +\infty$, let $R = 0$.) Then $\sum c_n z^n$ converges if $|z| < R$ and diverges if $|z| > R$.

- **Radius of convergence** (of a power series): The number R defined by Theorem 3.39.
- Theorem 3.41 (partial summation formula): Given two sequence $\{a_n\}, \{b_n\}$, put

$$A_n = \begin{cases} \sum_{k=0}^n a_k & n \geq 0 \\ 0 & n = -1 \end{cases}$$

Then if $0 \leq p \leq q$, we have

$$\sum_{n=p}^q a_n b_n = \sum_{n=p}^{q-1} A_n (b_n - b_{n+1}) + A_q b_q - A_{p-1} b_p$$

- **Product** (of $\sum a_n, \sum b_n$): The series $\sum c_n$ defined by

$$c_n = \sum_{k=0}^n a_k b_{n-k}$$

for each $n = 0, 1, 2, \dots$

- We motivate this definition by noting that if $\sum c_n$ is the product of $\sum a_n, \sum b_n$, then

$$\sum_{n=0}^{\infty} a_n z^n \cdot \sum_{n=0}^{\infty} b_n z^n = \sum_{n=0}^{\infty} c_n z^n$$

- Setting $z = 1$ then yields the given definition.

- The product of two convergent series may diverge. However...
- Theorem 3.50 (by Mertens): Suppose (a) $\sum_{n=0}^{\infty} a_n$ converges absolutely, (b) $\sum_{n=0}^{\infty} a_n = A$, (c) $\sum_{n=0}^{\infty} b_n = B$, and (d) $\sum_{n=0}^{\infty} c_n$ is the product of $\sum_{n=0}^{\infty} a_n$ and $\sum_{n=0}^{\infty} b_n$. Then

$$\sum_{n=0}^{\infty} c_n = AB$$

- Theorem 3.51 (by Abel): If $\sum a_n, \sum b_n, \sum c_n$ converge to A, B, C , respectively, and $\sum c_n$ is the product of $\sum a_n, \sum b_n$, then $C = AB$.
- **Rearrangement** (of $\sum a_n$): A series $\sum a'_n$ defined by $a'_n = a_{k_n}$ for each $n \in \mathbb{N}$, where $\{k_n\}$ is a sequence in which every positive integer appears once and only once (that is, $\{k_n\}$ is a 1-1 function from \mathbb{N} onto \mathbb{N}).
- Theorem 3.54: Let $\sum a_n$ be a series of real number which converges, but not absolutely. Suppose $-\infty \leq \alpha \leq \beta \leq \infty$. Then there exists a rearrangement $\sum a'_n$ with partial sums s'_n such that

$$\liminf_{n \rightarrow \infty} s'_n = \alpha \qquad \limsup_{n \rightarrow \infty} s'_n = \beta$$

- Theorem 3.55: If $\sum a_n$ is a series of complex numbers which converges absolutely, then every rearrangement of $\sum a_n$ converges, and they all converge to the same sum.

Chapter 4

Continuity

4.1 Notes

- 11/8:
- Consider a function $f : X \rightarrow Y$ whose domain and codomain are, respectively, the metric spaces (X, d_X) and (Y, d_Y) .
 - **Limit** (of f at p): A point $q \in Y$ such that for all $\epsilon > 0$, there exists δ such that $d_X(x, p) < \delta$ implies $d_Y(q, f(x)) < \epsilon$, where p is a limit point of X (otherwise, $x \nrightarrow p$).
 - **Continuous** (function f at p): A function f such that $\lim_{x \rightarrow p} f(x) = f(p)$.
 - f is continuous on X if it is continuous at every $p \in X$.
 - **Uniformly continuous** (function f): A function f such that for every $\epsilon > 0$, there exists a $\delta > 0$ such that $d_X(x, y) < \delta$ implies $d_Y(f(x), f(y)) < \epsilon$ for all $x, y \in X$.
- 11/10:
- f, g continuous implies $f + g, fg$, and f/g continuous, the latter where $g(x) \neq 0$.
 - If f, g continuous, then $h = g \circ f$ is continuous.
 - Theorem: $f : X \rightarrow Y$ is continuous iff $f^{-1}(V)$ is open in X for every $V \subset Y$ open.

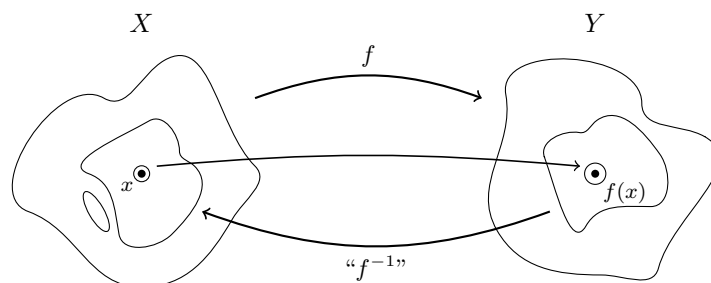


Figure 4.1: Set theoretic definition of continuity.

- This works in a general topological space, too, not just a metric space.
- Note that $f^{-1}(V)$ is not a function defined on V ; it's a specifically defined set $\{x \in X : f(x) \in V\}$.
- f being continuous means that open circular neighborhood of a point x in the domain maps to an area of the range encompassed by a circular neighborhood of $f(x)$.
- The other condition means that every open set surrounding $f(x)$ maps to an open set of the domain surrounding x . Indeed, going off of this definition, if an open set containing $f(x)$ maps to an open set containing x , then we can choose a neighborhood subset of the open set surrounding x and know that it will map into a neighborhood subset of the open set surrounding $f(x)$.

- Corollary: $f : X \rightarrow Y$ continuous iff $f^{-1}(C)$ closed for every $C \subset Y$ closed.
 - We use the property that $f^{-1}(X \subset C) = X \subset f^{-1}(C)$.
 - Let $f_1 : X_1 \rightarrow Y_1$ and $f_2 : X_2 \rightarrow Y_2$. Suppose $f : X_1 \times X_2 \rightarrow Y_1 \times Y_2$ is defined by $f(x_1, x_2) = (f_1(x_1), f_2(x_2))$. Then f is continuous iff f_1, f_2 are continuous, under appropriately defined metrics.
 - Continuity and compactness.
 - Theorem: $f : X \rightarrow Y$ continuous and X compact imply $f(X)$ compact.
 - Let $\{V_\alpha\}$ be an open cover of $f(X)$.
 - Then $\{f^{-1}(V_\alpha)\}$ is an open cover of X .
 - Choose a finite subcover of $\{f^{-1}(V_\alpha)\}$. Then the corresponding V_α 's form a finite subcover of $f(X)$.
 - If $f : X \rightarrow \mathbb{R}^k$ is continuous and X is compact, $f(X)$ is compact and closed/bounded.
 - If $f : X \rightarrow \mathbb{R}$ is continuous and X is compact, then $M = \sup_{x \in X} |f(x)| = |f(\bar{x})|$ and $m = \inf_{x \in X} |f(x)| = |f(\underline{x})|$ where $\bar{x}, \underline{x} \in X$.
 - There is a subsequence $\{x_m\}$ such that $|f(x_m)| \rightarrow M$. Since $f(X)$ is compact, the limit of this sequence is in $f(X)$.
 - If $f : X \rightarrow Y$ is continuous, bijective, and X, Y are compact, $f^{-1} : Y \rightarrow X$ is continuous.
 - Uniform continuity.
 - Examples.
 - Linear functions are uniformly continuous.
 - $f : (0, 1) \rightarrow \mathbb{R}$ defined by $f(x) = x^2$ is uniformly continuous.
 - $f : (a, \infty) \rightarrow \mathbb{R}$ defined by $f(x) = x^2$ is *not* uniformly continuous.
 - $f : (a, \infty) \rightarrow \mathbb{R}$ defined by $f(x) = 1/x$ is uniformly continuous if $a > 0$.
 - $f : (0, \infty) \rightarrow \mathbb{R}$ defined by $f(x) = 1/x$ is *not* uniformly continuous.
 - **Lipschitz continuous** (function f on $E \subset X$): A function such that $|f(x) - f(y)| \leq L|x - y|$ for each $x, y \in E$.
 - Theorem: $f : X \rightarrow Y$ continuous and X compact implies f is uniformly continuous.
 - Fix $\epsilon > 0$. There exists $\delta = \delta(p) > 0$.
 - Def. of continuity: $q \in N_{\delta(p)}(p)$ implies $f(q) \in N_\epsilon(f(p))$.
 - $\{N_{\delta(p)/2}(p) : p \in X\}$ is an open cover of X . Choose a finite subcover. Let $\delta = \min(\delta(p_1)/2, \dots, \delta(p_n)/2)$.
 - ...
- 11/12:
- $f : X \rightarrow Y$ continuous and $E \subset X$ connected implies $f(E)$ connected.
 - Suppose $f(E) = A \cup B$, A, B nonempty, separated subsets of Y .
 - Let $G = E \cap f^{-1}(A)$, $H = E \cap f^{-1}(B)$. It follows that $E = G \cup H$, where G, H nonempty.
 - $A \subset \bar{A}$ implies $G \subset f^{-1}(\bar{A})$ implies (inverse image def. of continuity) $\bar{G} \subset f^{-1}(A)$ implies $f(\bar{G}) \subset \bar{A}$.
 - $f(H) = B$ and $\bar{A} \cap B = \emptyset$ yields $\bar{G} \cap H = \emptyset$. Symmetrically, $G \cap \bar{H} = \emptyset$. This contradicts our assumption that E is connected.
 - Introduces monotone functions.
 - Theorem: If f is monotonic on (a, b) , then the set of points of (a, b) at which f is discontinuous is at most countable.
 - ...

4.2 Chapter 4: Continuity

From Rudin (1976).

- 11/8:
- **Limit** (of f at p): The point $q \in Y$, if it exists, such that for every $\epsilon > 0$, there exists a $\delta > 0$ such that $d_Y(f(x), q) < \epsilon$ for all points $x \in E$ for which $0 < d_X(x, p) < \delta$, where $(X, d_X), (Y, d_Y)$ are metric spaces, $E \subset X$, $f : E \rightarrow Y$, and $p \in E'$. Denoted by $\lim_{x \rightarrow p} f(x)$.
 - Note that we do not require that $p \in E$; only that some elements of the domain E approach p .
 - We also write $f(x) \rightarrow q$ as $x \rightarrow p$.
 - Theorem 4.2: Let X, Y, E, f , and p be as specified above. Then $\lim_{x \rightarrow p} f(x) = q$ iff $\lim_{n \rightarrow \infty} f(p_n) = q$ for every sequence $\{p_n\}$ in E such that $p_n \neq p$ for any n and $\lim_{n \rightarrow \infty} p_n = p$.
 - Rudin (1976) proves the sum, product, and quotient rules of limits from the analogous properties of series.
 - Continuity is defined.
 - Note that f *does* have to be defined at p to be continuous at p (in comparison to the fact that it can have a limit at a point p' at which it is not defined).
 - Thus, for proofs concerning continuity (as opposed to limits), we will consider functions f the domains of which are metric spaces, not *subsets* of metric spaces.
 - It follows from the definition that if $p \in E$ is isolated, then every possible f defined on E is continuous at p .
 - Theorem 4.7: Compositions of continuous functions are continuous.
 - Theorem 4.8: Preimage definition of continuity.
 - Theorem 4.9: If f, g are complex continuous functions on X , $f + g$, fg , and f/g are continuous on X .
 - Theorem 4.10: \mathbf{f} continuous implies f_1, \dots, f_k continuous. Also, $\mathbf{f}, \mathbf{g} : X \rightarrow \mathbb{R}^k$ continuous implies $\mathbf{f} + \mathbf{g}$ and $\mathbf{f} \cdot \mathbf{g}$ continuous.
- 11/9:
- Theorem 4.14: f continuous and X compact implies $f(X)$ compact.
 - Theorem 4.15: $\mathbf{f} : X \rightarrow \mathbb{R}^k$ continuous and X compact implies $f(X)$ closed and bounded.
 - Theorem 4.16: f continuous and X compact implies f attains its minimum and maximum.
 - Theorem 4.17: $f : X \rightarrow Y$ continuous, 1-1 for X, Y compact implies $f^{-1} : Y \rightarrow X$ continuous.
 - Theorem 4.19: f continuous and X compact implies f uniformly continuous.
 - Theorem 4.20: Compactness is a necessary condition in Theorems 4.14, 4.15, 4.16, and 4.19.
 - Theorem 4.22: $f : X \rightarrow Y$ continuous and $E \subset X$ connected implies $f(E)$ connected.
 - Theorem 4.23: Intermediate value theorem.
 - **Right-hand limit** (of f at x): Denoted by $f(x+)$.
 - **Left-hand limit** (of f at x): Denoted by $f(x-)$.
 - **Discontinuity of the first kind** (of f at x): A discontinuity of f at x such that $f(x+)$ and $f(x-)$ exist. Also known as **simple discontinuity**.
 - **Discontinuity of the second kind** (of f at x): A discontinuity of f at x that is not of the first kind (i.e., a discontinuity such that at least one of $f(x+)$ and $f(x-)$ does not exist).

- Theorem 4.29: If f is monotonic on (a, b) , then $f(x+), f(x-)$ exist at every $x \in (a, b)$.
- Corollary: Monotonic functions have no discontinuities of the second kind.
- Theorem 4.30: If f is monotonic on (a, b) , then the set of points of (a, b) at which f is discontinuous is at most countable.

11/20:

Proof. Suppose first that f is increasing. Let E be the set of points at which f is discontinuous. By Theorem 4.29, for every $x \in E$, $f(x-), f(x+)$ exist. Thus, we may pick a rational number $r(x)$ such that $f(x-) < r(x) < f(x+)$. Moreover, since $x_1 < x_2$ implies $f(x_1+) \leq f(x_2-)$, we have that $r(x_1) \neq r(x_2)$. Having established an injective function from E to the rationals \mathbb{Q} , we know that E is at most countable. The argument where f is decreasing is symmetric. \square

- Gives an example of a function with discontinuities that are not isolated.

Chapter 7

Sequences and Series of Functions

7.1 Notes

11/15:

- Soug will not test on differentiation/integration assuming that we know them already.
- **Pointwise convergent** (sequence $(f_n)_{n \in \mathbb{N}}$ of functions): A sequence of functions $f_n : E \rightarrow \mathbb{R}$ such that $\lim_{n \rightarrow \infty} f_n(x) = f(x)$ for all $x \in E$.
- Can we interchange “limit” in the above definition with continuity, convergence of series, integration, differentiation, etc.?
- Examples with negative answer:

1. Interchanging limits: Let $S_{mn} = \frac{m}{m+n}$. $S_{mn} \rightarrow 1$ as $m \rightarrow \infty$, and $S_{mn} \rightarrow 0$ as $n \rightarrow \infty$.
2. $f_n(x) = x^2/(1+x)^n$. $f(x) = \sum_{n=1}^{\infty} f_n(x)$. If $x = 0$, then $f_n(x) = 0$ for all n and $f(x) = 0$. If $x \neq 0$, we have

$$f(x) = \sum_{n=1}^{\infty} \frac{x^2}{(1+x^2)^n} = x^2 \sum_{n=1}^{\infty} X^n = \frac{x^2}{1-X} = \frac{x^2}{1-(1/(1+x^2))} = 1+x^2$$

3. Consider $f_m(x) = \lim_{n \rightarrow \infty} (\cos(m\pi x))^2 n$. $\lim_{m \rightarrow \infty} f_m(x)$ goes to 0 if $x \notin \mathbb{Q}$ and goes to 1 if $x \in \mathbb{Q}$. $f_m \rightarrow \chi_{\mathbb{Q}}$, where $\chi_{\mathbb{Q}}$ is the characteristic function of the rationals which is not Riemann integrable (partitions, upper and lower integrals, etc.).
 4. $f_n(x) = \sin nx / \sqrt{n} \rightarrow f(x) = 0$ for all x . However, $f'_n(x) = \sqrt{n} \cos nx \not\rightarrow 0$
 5. If $0 \leq x \leq 1$, define $f_n(x) = n^2 x(1-x^2)^n$. We know that $f_n(0) = 0$. $\lim_{n \rightarrow \infty} f_n(x) = 0$ for all $x \in (0, 1]$. We can show that $\int_0^1 x(1-x^2)^n dx = 1/(2n+2)$. Thus, $\int_0^1 f_n(x) dx = n^2/(2n+2)$. Limit of the functions is zero, but their integrals diverge to infinity.
- **Uniformly convergent** (sequence $(f_n)_{n \in \mathbb{N}}$ of functions on E): A sequence of functions $f_n : E \rightarrow \mathbb{R}$ such that for all $\epsilon > 0$, there exists N such that if $n \geq N$, then $|f_n(x) - f(x)| < \epsilon$ for all $x \in E$. Denoted by $f_n \rightrightarrows f$.
 - Theorem: $f_n \rightrightarrows f$ iff $(f_n)_{n \in \mathbb{N}}$ is uniformly Cauchy (i.e., for all $\epsilon > 0$, there exists N such that if $n, m \geq N$ then $|f_n(x) - f_m(x)| < \epsilon$ for all $x \in E$).
 - Let $M_n = \sup_{x \in E} |f_n(x) - f(x)|$. If $f_n \rightarrow f$ pointwise, then $f_n \rightrightarrows f$ if $M_n \rightarrow 0$.
 - Theorem: If $(f_n)_{n \in \mathbb{N}}$ and $|f_n(x)| \leq M_n$, then $\sum f_n \rightrightarrows f$ if $\sum M_n < \infty$.
 - Theorem: If E is a compact metric space, $f_n \rightrightarrows f$ in E , x is a limit point of E , and $\lim_{t \rightarrow x} f_n(t) = A_n$ exists, then $(A_n)_{n \in \mathbb{N}}$ converges and $\lim_{t \rightarrow x} f(t) = \lim_{n \rightarrow \infty} A_n$.

- Corollary: $\lim_{t \rightarrow x} \lim_{n \rightarrow \infty} f_n(t) = \lim_{n \rightarrow \infty} \lim_{t \rightarrow x} f_n(t)$.

11/16:

- Fix $\epsilon > 0$. Then $f_n \rightrightarrows f$ implies that there exists some N such that $n, m \geq N$ implies $|f_n(t) - f_m(t)| < \epsilon$ for all $t \in E$.
- x is a limit point of E and $t \rightarrow x$ implies $|A_n - A_m| < \epsilon$. Thus, $(A_n)_{n \in \mathbb{N}}$ is cauchy, so there exists A such that $A_n \rightarrow A$.
- WTS: $|f(t) - A| \leq |f(t) - f_n(t)| + |f_n(t) - A_n| + |A_n - A|$, so we WTS the three terms on the right are small.
- There exists n such that $|f(t) - f_n(t)| < \epsilon/3$ for all t since $f_n \rightrightarrows f$ by hypothesis.
- Since t is in a small neighborhood of x , there exists n such that $|A_n - A| < \epsilon/3$.
- We also have $|f_n(t) - A_n| < \epsilon/3$ by hypothesis.
- This is a very important proof to understand, because proofs like this pop up often.
- Corollary: f_n continuous and $f_n \rightrightarrows f$ implies f is continuous.
- Theorem: Let K be compact. Assume
 - (a) $(f_n)_{n \in \mathbb{N}} \subset C(K) = \{f : K \rightarrow \mathbb{R} \mid f \text{ continuous}\}$.
 - (b) $f_n \rightarrow f$ pointwise in K and $f \in C(K)$.
 - (c) $f_n(x) \geq f_{n+1}(x)$ for all $x \in K$.

Then $f_n \rightrightarrows f$.

- WLOG $f = 0$, $g_n = f_n - f \rightarrow 0$, $g_n \geq g_{n+1} \geq 0$.
- For all $\epsilon > 0$, there exists N such that $n \geq N$ and $0 \leq g_n(x) \leq \epsilon$ for all $x \in K$.
- $K_n = \{x \in K : g_n(x) \geq \epsilon\}$.
- g_n continuous implies K_n closed. This combined with K compact implies K_n is compact.
- g_n decreasing implies $K_n \supset K_{n+1}$. Thus, K_n is a nested family of compact sets, so $\bigcap K_n \neq \emptyset$.
- This implies that each K_n is nonempty, contradicting the fact that each $g_n \rightarrow 0$ for all x .
- Thus, there exists an N such that K_n is empty for all $n \geq N$. Thus $g_n(x) \leq \epsilon$ for all $x \in K$, $n \geq N$.
- Note that the compactness of K is important. If $f : (0, 1) \rightarrow \mathbb{R}$ is defined by $f(x) = 1/(nx + 1)$, then $f_n \rightarrow 0$, but $f_n \not\rightrightarrows f$.
- Let $C(X) = \{f : X \rightarrow \mathbb{R} \mid f \text{ continuous, bounded}\}$ for X a metric space.
- If we define $\|f\| = \sum_{x \in X} |f(x)|$, for $f, g \in C(X)$, we may define $d(f, g) = \|f - g\|$. This definition satisfies the properties of a distance function, and $\|\cdot\|$ is a norm.
 - Thus, $C(X)$ is a complete metric space, a normed space, or, specifically, a **Banach space**.
- Theorem: $(f_n)_{n \in \mathbb{N}} \subset C(X)$ such that $\|f_n - f_m\|_{n, m \rightarrow \infty} \rightarrow 0$. Then there exists $f \in C(X)$ such that $\|f_n - f\|_{n \rightarrow \infty} \rightarrow 0$.
 - We get such a strong statement using properties of the image, not properties of the domain.
 - For all $\epsilon > 0$, there exists N such that $n, m \geq N$.
 - $|f_n(x) - f_m(x)| \leq \|f_n - f_m\| < \epsilon$ for all x .
 - Then there exists f such that $f_m(x) \rightarrow f(x)$. It follows that $|f_n(x) - f_m(x)| < \epsilon$
- Uniform convergence and integration.
- Stieltjes integral.

- Define $\alpha : \mathbb{R} \rightarrow \mathbb{R}$ nondecreasing.
- If we sum over the minimums/maximums of a partition times $\alpha(x_{i+1}) - \alpha(x_i)$ instead of $x_{i+1} - x_i$, we obtain said integral as the upper/lower limits just like the Riemann integral.
- We write $\int_a^b f(x) d\alpha(x)$ where $d\alpha(x) = \alpha(x) dx$.
- Theorem: If α is nondecreasing on $[a, b]$, $f_n \in R(\alpha)$ such that $f_n \rightrightarrows f$ on $[a, b]$
 - We have

$$\begin{aligned}
 \left| \int f_n(x) d\alpha(x) - \int f(x) d\alpha(x) \right| &= \left| \int (f_n - f)(x) d\alpha(x) \right| \\
 &\leq \|f_n - f\|(\alpha(b) - \alpha(a)) \\
 &\leq \int |f_n - f| d\alpha(x) \\
 &\leq \int \|f_n - f\| d\alpha(x) \leq \|f_n - f\| \int_a^b d\alpha(x) = \|f_n - f\|(\alpha(b) - \alpha(a))
 \end{aligned}$$

11/19:

- Suppose $f_n \rightarrow f$ and $f'_n \rightarrow g$. When does $f' = g$?
- Theorem: If $f_n : [a, b] \rightarrow \mathbb{R}$ is differentiable, $f_n(x_0)$ converges for some $x_0 \in [a, b]$, and f'_n converges uniformly on $[a, b]$, then there exists f differentiable on $[a, b]$ such that $f_n \rightrightarrows f$ and $f'_n \rightrightarrows f'$.

- Assume the f'_n are continuous. Then $f_n(x) - f_n(x_0) = \int_{x_0}^x f'_n(y) dy$.
- Since $f'_n \rightrightarrows g$, $\int_{x_0}^x f'_n(y) dy \rightarrow \int_{x_0}^x g(y) dy$.
- It follows since $f_n(x_0) \rightarrow f(x_0)$ that $f_n \rightrightarrows f$.
- By the previous theorem, if

$$f'_n(x) = \lim_{h \rightarrow 0} \frac{f_n(x+h) - f_n(x)}{h}$$

then

$$\lim_{n \rightarrow \infty} f'_n(x) = \lim_{h \rightarrow 0} \lim_{n \rightarrow \infty} \frac{f_n(x+h) - f_n(x)}{h} = \lim_{h \rightarrow 0} \frac{f(x+h) - f(x)}{h}$$

- Fix $\epsilon > 0$. Then there exists N such that $n, m \geq N$ such that $|f_n(x_0) - f_m(x_0)| < \epsilon/2$ and $|f'_n(t) - f'_m(t)| < \epsilon/2$ for all $t \in [a, b]$.
- We know that $f_n(t) - f_n(x_0) = \int_{x_0}^t f'_n(y) dy$ and $f_m(t) - f_m(x_0) = \int_{x_0}^t f'_m(y) dy$.
- Thus,

$$|f_n(t) - f_n(x_0)| \leq |f_n(t) - f_m(t)| + |f_m(t) - f_m(x_0)| + |f_m(t) - f_m(x_0)|$$
- Let $f_n(t) - f_n(x_0) = c_n(t - x_0)$ and $f_m(t) - f_m(x_0) = c_m(t - x_0)$.
- ...
- Let $f : [a, b] \rightarrow \mathbb{R}$ be continuous. What conditions on f imply that f' exists?
- Suppose f is Lipschitz continuous (equivalent to saying there exists $L > 0$ such that $|f(x) - f(y)| \leq L|x - y|$); then f' exists **almost everywhere**.
 - If f differentiable, this is equivalent to saying f bounded.
- **Almost everywhere**: Something happens almost everywhere if the set of places where it doesn't happen has measure zero.
- Suppose f is **Hölder continuous**, then f' does not exist?
- **Hölder continuous** (function f): There exists $L > 0$ such that $|f(y) - f(x)| < L|x - y|^\alpha$ where $\alpha \in (0, 1)$

- Suppose f exists such that f is Hölder continuous in a neighborhood of every point in the domain. This function is not anywhere differentiable. Such a function does indeed exist (and it's Brownian motion). The construction of such a function is the essence of Stochastic analysis.
 - Probabilistically: Has mean zero, distributed as a normal function like the Gaussian, and the increments are independent of each other.
 - Analytically: It's a function that is Hölder continuous at half plus ϵ for every ϵ and it is nowhere differentiable.
- Theorem: There exists $f : \mathbb{R} \rightarrow \mathbb{R}$ continuous but nowhere differentiable.
 - This theorem is due to Weierstrass and as such, such functions are typically called Weierstrass functions.
- A general class of functions that are nowhere differentiable (not in Rudin (1976); we don't have to prove this).
 - Example 1:

$$f(x) = \sum_{n=0}^{\infty} a^n \cos(b^n \pi x)$$

where $0 < a < 1$, b positive odd integer greater than 1, and $ab > 1 + \frac{3}{2}\pi$.

■ This function at every point oscillates more and more and more.

- Rudin (1976)'s simple example.
 - $\phi : [-1, 1] \rightarrow \mathbb{R}$ defined by $\phi(x) = |x|$ is not differentiable at zero.
 - Takes ϕ extends it periodically with period 2, creating a sawtooth function.
 - Repeat the behavior so that the nondifferentiability becomes more and more frequent to get

$$f(x) = \sum_1^{\infty} \left(\frac{3}{4}\right)^n \phi(4^n x)$$

- This is continuous.
- Fix any $x \in \mathbb{R}$, $m \in \mathbb{N}$. Then $\delta_m = \pm \frac{1}{2} \cdot 4^{-m}$.
- Then consider $4^m x$, $4^m(x + \delta_m)$.
- Rudin asserts

$$\left| \frac{f(x + \delta_m) - f(x)}{\delta_m} \right| \rightarrow \infty$$

as $m \rightarrow \infty$ for all x .

11/29:

- Finding a uniformly convergent subsequence of a sequence of functions.
 - Pointwise, uniformly, bounded if there exist M_x such that $|f_n(x)| \leq M_x$ for all n, x . Uniformly bounded if there exists M such that $|f_n(x)| \leq M$ for all n, x .
- Theorem: If $(f_n)_{n \in \mathbb{N}}$ is pointwise bounded and $E \subset X$ is countable, then there exists a subsequence f_{n_k} which converges for every $x \in E$.
 - Let $E = \{x_i : i \in \mathbb{N}\}$. Consider $f_n(x_1)$. $f_{1,k}(x_1)$ converges.
 - $S_1 : f_{1,1}(x_1), f_{1,2}(x_1), f_{1,3}(x_1), f_{1,4}(x_1), \dots$
 - $S_2 : f_{2,1}(x_2), f_{2,2}(x_2), f_{2,3}(x_2), f_{2,4}(x_2), \dots$
 - Now consider $f_{2,k}(x_3)$.
 - $S_3 : f_{3,1}, f_{3,2}, f_{3,3}, f_{3,4}, \dots$

- Continue on and on to S_4, S_5, \dots . We know that each of these sequences converges pointwise by hypothesis.
 - Now consider the diagonal sequence $f_{1,1}, f_{2,2}, f_{3,3}, f_{4,4}$.
 - This subsequence of the original sequence we may call g_k .
 - We posit that g_k converges for every $x \in E$.
 - Theorem: There exists f_n which is uniformly bounded but does not converge uniformly.
 - Let $f_n(x) = \sin(2\pi x)$ for $0 \leq x \leq 2\pi$.
 - Let $f_n(x) = x^2/(x^2 + (1 - nx)^2)$ on $0 \leq x \leq 1$. This sequence is uniformly bounded, converges pointwise, but $f_n(1/n) = 1$ so f_n cannot converge uniformly to zero.
 - What does it mean that $f_n : [0, 1] \rightarrow \mathbb{R}$ does not converge uniformly?
 - It means that there exists a subsequence of the functions evaluated at certain points that is always greater than or equal to some fixed distance away from the limit.
 - Equicontinuity: $\mathcal{F}\{f : X \rightarrow \mathbb{R}\}$ for (X, d) a metric space is equicontinuous iff for all $\epsilon > 0$, there exists a $\delta > 0$ such that $d(x, y) < \delta$ implies $|f(x) - f(y)| < \epsilon$ for all $x, y \in X, f \in \mathcal{F}$.
 - Modulus of continuity: $f : X \rightarrow \mathbb{R}$ is continuous at x . A modulus of continuity is a function $\omega_X : [0, 1] \rightarrow [0, 1]$ such that $|f(y) - f(x)| \leq \omega_X|y - x|$.
 - The final result we'll prove: **Arzelà-Ascoli theorem**: If we have a family of functions on a compact set and we have a dense subset of that set, then if we have a sequence of functions that are equicontinuous, then they converge uniformly.
- 12/1:
- The final will be in this room.
 - The last PSet will not be graded, but there will be similar questions on the final.
 - No class on Friday.
 - Review from last time:
 - Equiboundedness and equicontinuity.
 - If E is a dense subset of X , then any pointwise bounded sequence has a subsequence that converges on E (diagonal argument.)
 - **Equicontinuous** (sequence $\{f_n\}$): For all $\epsilon > 0$, there exists a $\delta > 0$ such that $d(x, y) < \delta$ implies $|f_n(x) - f_n(y)| < \epsilon$ for all n .
 - Theorem: If K is a compact set and $\{f_n\} \in C(K)$ converges uniformly on K , then the f_n 's are equicontinuous on K .
 - The f_n are uniformly Cauchy: For all $\epsilon > 0$, there exists N such that $n, m \geq N$ imply $\|f_n - f_m\| < \epsilon$ where $\|f_n - f_m\| = \sup_{x \in K} (f_n - f_m)(x)$.
 - If $n \geq N$, then $|f_n(x) - f_n(y)| \leq |f_N(x) - f_N(y)| + 2\|f_n - f_N\|$ (since $f_n(x) - f_n(y) = f_n(x) - f_N(x) + f_N(x) - f_N(y) + f_N(y) - f_n(y)$).
 - Thus $|f_n(x) - f_n(y)| \leq |f_N(x) - f_N(y)| + 2\|f_n - f_N\| < 3\epsilon$ implies $|f_i(x) - f_i(y)| < \epsilon$ if $|x - y| < \delta$ for $i = 1, \dots, N$.
 - **Arzelà-Ascoli theorem**: If K is compact, $(f_n)_{n \geq 1} \subset C(K)$ which are pointwise bounded and equicontinuous, then
 - (a) $(f_n)_{n \geq 1}$ are uniformly bounded (equicontinuous).
 - (b) There exists $(f_{n_k})_{k \geq 1}$ which converges uniformly on K .

- Since K is compact, you can cover it by finitely many balls of radius δ .
- Thus $|f_n(p_k)| \leq M = \max(M_{p_1}, \dots, M_{p_k})$ where $K \subset \bigcup_{k \in K} B(p_k, \delta)$.
- K has a countable dense subset E (Exercise 2.25).
- $|f_n(x)| \leq M + \epsilon$ for all x .
- $\{B(x, \delta)\}_{x \in E}$ is an open cover of K .
- Thus has a finite subcover.
- ...
- If $f_n : \mathbb{R}^n \rightarrow \mathbb{R}$ are continuous, equibounded, equicontinuous, then there exists f_{n_k} which converges locally uniformly to some $f : \mathbb{R}^n \rightarrow \mathbb{R}$.
- How do you learn math?
 - In an ideal world, study by looking at theorems, thinking that you should be able to prove it, and etc.
 - Since we don't have the time to do everything ourselves, don't just get stuck in a place; move on and continue thinking if you have to.
- Let $\dot{\phi} = f(x, t)$ and $x(0) = c$. Let $\phi : \mathbb{R} \times [0, 1] \rightarrow \mathbb{R}$. Assume ϕ is bounded and continuous. Then there exists a solution of the differential equation and initial condition.
 - We need to find a function $x : [0, 1] \rightarrow \mathbb{R}$ continuous such that $x(t) = c + \int_0^t \phi(x(s), s) ds$.
 - Let $t_i = i/N$. Then $x_n(t) = \phi(x_i, t_i)$ on $t_i < t < t_{i+1}$.
 - $x_n(t) = x_n(t_i) + \phi(x_i, t_i)(t - t_i)$.
 - $\frac{x_{i+1} - x_i}{1/N} = \phi(x_i, t_i)$.
 - $\Delta_n(t) = x'_n(t) - \phi(x_n(t), t)$ for $\phi(x_i, t_i) - \phi(x_n(t), t)$ measures how close our solution is.
 - All of these things imply that our final formula is

$$x_n(t) = c + \int_0^t [\phi(x_n(s), s) + \Delta_n(s)] ds$$
 - If we know that $x_n \rightrightarrows x$, then $\Delta_n \rightrightarrows 0$.
 - We then use the A-A theorem to imply convergence.
- When we get to MATH 208, say we didn't do any multivariable calculus in MATH 207.
 - We didn't do how to integrate in \mathbb{R}^n , how to integrate by parts (Stoke's theorem), Lagrange multipliers (constraint minimization).
- Problem 4.23: Show the inequalities at the bottom first and then use those to show continuity.
 - Consider $\lim_{t \rightarrow u} f(t)$. Approach from two sides separately and show cancellation??? Chloe will write a solution.
 - This is a particular trick for convex functions; it's not exactly recyclable.
- The trick for Problem 4.26 is recyclable.
- Linear algebra questions on the final are easier than the midterm.
 - The last question will be the hard one this time.

References

- Hewitt, E., & Stromberg, K. (1965). *Real and abstract analysis: A modern treatment of the theory of functions of a real variable*. Springer.
- Rudin, W. (1976). *Principles of mathematical analysis* (A. A. Arthur & S. L. Langman, Eds.; Third). McGraw-Hill.
- Treil, S. (2017). *Linear algebra done wrong* [http://www.math.brown.edu/streil/papers/LADW/LADW_2017-09-04.pdf].