

Advanced Engineering Mathematics Vectors, Matrices, and Vector Calculus by Dennis G. Zill

Notes

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1 Vectors

1.1 Vectors in 2-Space

- The zero vector can be assigned any direction
- The vectors \mathbf{i} and \mathbf{j} are known as the **standard basis vectors** for \mathbb{R}^2

1.2 Vectors in 3-Space

- In \mathbb{R}^3 the octant in which all coordinates are positive is known as the **first octant**. There is no agreement for naming the other seven octants.

1.3 Dot Product

- The **dot product** is also known as the **inner product** or the **scalar product** and is denoted $\mathbf{a} \cdot \mathbf{b}$
- Two non-zero vectors are orthogonal iff their dot product is 0
- The zero vector is considered orthogonal to all vectors
- The angles α , β , and γ between a vector and the unit vectors \mathbf{i} , \mathbf{j} , and \mathbf{k} , respectively are called the **direction angles** of the vector
- The cosines of a vectors direction angles (the **direction cosines**) can be calculated as

$$\begin{aligned}
\cos \alpha &= \frac{\mathbf{a} \cdot \mathbf{i}}{||\mathbf{a}|| ||\mathbf{i}||} \\
&= \frac{a_1}{||\mathbf{a}||} \\
\cos \beta &= \frac{\mathbf{a} \cdot \mathbf{j}}{||\mathbf{a}|| ||\mathbf{j}||} \\
&= \frac{a_2}{||\mathbf{a}||} \\
\cos \gamma &= \frac{\mathbf{a} \cdot \mathbf{k}}{||\mathbf{a}|| ||\mathbf{k}||} \\
&= \frac{a_3}{||\mathbf{a}||}
\end{aligned}$$

Equivalently, these can be calculated as the components of the unit vector $\mathbf{a}/||\mathbf{a}||$.

- To find the component of a vector \mathbf{a} in the direction of a vector \mathbf{b}

$$\text{comp}_{\mathbf{b}} \mathbf{a} = ||\mathbf{a}|| \cos \theta = \frac{\mathbf{a} \cdot \mathbf{b}}{||\mathbf{b}||}$$

- To project a vector \mathbf{a} onto a vector \mathbf{b}

$$\text{proj}_{\mathbf{b}} \mathbf{a} = (\text{comp}_{\mathbf{b}} \mathbf{a}) \frac{\mathbf{b}}{||\mathbf{b}||} = \left(\frac{\mathbf{a} \cdot \mathbf{b}}{\mathbf{b} \cdot \mathbf{b}} \right) \mathbf{b}$$

1.4 Cross Product

- The cross product is only defined in \mathbb{R}^3
- The **scalar triple product** of vectors \mathbf{a} , \mathbf{b} , and \mathbf{c} is defined as

$$\mathbf{a} \cdot (\mathbf{b} \times \mathbf{c}) = (\mathbf{a} \times \mathbf{b}) \cdot \mathbf{c} = \begin{vmatrix} a_1 & a_2 & a_3 \\ b_1 & b_2 & b_3 \\ c_1 & c_2 & c_3 \end{vmatrix}$$

- The area of a parallelogram with sides \mathbf{a} and \mathbf{b} is $||\mathbf{a} \times \mathbf{b}||$
- The area of a triangle with sides \mathbf{a} and \mathbf{b} is $\frac{1}{2} ||\mathbf{a} \times \mathbf{b}||$
- The volume of a parallelepiped with sides \mathbf{a} , \mathbf{b} , and \mathbf{c} is $|\mathbf{a} \cdot (\mathbf{b} \times \mathbf{c})|$
- $\mathbf{a} \cdot (\mathbf{b} \times \mathbf{c}) = 0$ iff \mathbf{a} , \mathbf{b} , and \mathbf{c} are coplanar

1.5 Lines and Planes in 3-Space

- There is a unique line between any two points \mathbf{r}_1 and \mathbf{r}_2 in 3-space. The equation for that line is

$$\mathbf{r} = \mathbf{r}_1 + t(\mathbf{r}_2 - \mathbf{r}_1) = \mathbf{r}_1 + t\mathbf{a}$$

where t is called a **parameter**, the nonzero vector \mathbf{a} is called a **direction vector**, and its components are called **direction numbers**.

- Equating the components of the equation above we find

$$\begin{aligned}x &= r_1 + ta_1 \\y &= r_2 + ta_2 \\z &= r_3 + ta_3.\end{aligned}$$

These are the **parametric equations** for the line through \mathbf{r}_1 and \mathbf{r}_2 .

- By solving the parametric equations for t and equating the results we find the **symmetric equations** for the line

$$t = \frac{x - r_1}{a_1} = \frac{y - r_2}{a_2} = \frac{z - r_3}{a_3}.$$

- Given a point P_1 and a vector \mathbf{n} , there exists only one plane containing P_1 with \mathbf{n} normal. The vector from P_1 to another point P on that plane will be perpendicular to \mathbf{n} , so the equation for the plane is

$$\mathbf{n} \cdot (\mathbf{r} - \mathbf{r}_1) = 0$$

where $\mathbf{r} = \overrightarrow{OP}$ and $\mathbf{r}_1 = \overrightarrow{OP_1}$. If

$$\mathbf{n} = a\hat{\mathbf{i}} + b\hat{\mathbf{j}} + c\hat{\mathbf{k}}$$

the cartesian form of this equation is

$$a(x - x_1) + b(y - y_1) + c(z - z_1) = 0$$

and is called the **point-normal form**.

- The graph of any equation $ax + by + cz + d = 0$, where a , b , and c are not all zero, is a plane with the normal vector $\mathbf{n} = a\hat{\mathbf{i}} + b\hat{\mathbf{j}} + c\hat{\mathbf{k}}$.
- Given three noncollinear points, a normal vector can be found by forming two vectors from two pairs of points and take their cross product.
- A line and a plane that aren't parallel intersect at a single point.
- Two planes that aren't parallel must intersect in a line.

1.6 Vector Spaces

- The length of a vector is called its **norm**
- The process of multiplying a vector by the reciprocal of its norm is called **normalizing** the vector
- Two nonzero vectors \mathbf{a} and \mathbf{b} in \mathbb{R}^n are said to be orthogonal if $\mathbf{a} \cdot \mathbf{b} = 0$

Definition 7.6.1 Vector Space

Let V be a set of elements on which two operations called **vector addition** and **scalar multiplication** are defined. Then V is said to be a **vector space** if the following 10 properties are satisfied.

Axioms for Vector Addition:

- (i) If \mathbf{x} and \mathbf{y} are in V , then $\mathbf{x} + \mathbf{y}$ is in V .
- (ii) For all \mathbf{x}, \mathbf{y} in V , $\mathbf{x} + \mathbf{y} = \mathbf{y} + \mathbf{x}$. ← commutative law
- (iii) For all $\mathbf{x}, \mathbf{y}, \mathbf{z}$ in V , $\mathbf{x} + (\mathbf{y} + \mathbf{z}) = (\mathbf{x} + \mathbf{y}) + \mathbf{z}$. ← associative law
- (iv) There is a unique vector $\mathbf{0}$ in V such that
 $\mathbf{0} + \mathbf{x} = \mathbf{x} + \mathbf{0} = \mathbf{x}$. ← zero vector
- (v) For each \mathbf{x} in V , there exists a vector $-\mathbf{x}$ such that
 $\mathbf{x} + (-\mathbf{x}) = (-\mathbf{x}) + \mathbf{x} = \mathbf{0}$. ← negative of a vector

Axioms for Scalar Multiplication:

- (vi) If k is any scalar and \mathbf{x} is in V , then $k\mathbf{x}$ is in V .
- (vii) $k(\mathbf{x} + \mathbf{y}) = k\mathbf{x} + k\mathbf{y}$ ← distributive law
- (viii) $(k_1 + k_2)\mathbf{x} = k_1\mathbf{x} + k_2\mathbf{x}$ ← distributive law
- (ix) $k_1(k_2\mathbf{x}) = (k_1k_2)\mathbf{x}$
- (x) $1\mathbf{x} = \mathbf{x}$

- If a subset W of a vector space V is itself a vector space under the operations of vector addition and scalar multiplication defined on V , then W is called a **subspace** of V
- Every vector space has at least two subspaces: itself and the zero subspace $\{\mathbf{0}\}$
- A set of vectors $\{\mathbf{x}_1, \mathbf{x}_2, \dots, \mathbf{x}_n\}$ is said to be **linearly independent** if the only constants satisfying the equation

$$k_1\mathbf{x}_1 + k_2\mathbf{x}_2 + \dots + k_n\mathbf{x}_n = \mathbf{0}$$

are $k_1 = k_2 = \dots = k_n = 0$. If the set of vectors is not linearly independent it is said to be **linearly dependent**.

- If a set of vectors $B = \{\mathbf{x}_1, \mathbf{x}_2, \dots, \mathbf{x}_n\}$ in a vector space V is linearly independent and every vector in V can be expressed as a linear combination of vectors in B then B is said to be a **basis** for V .
- The number of vectors in a basis B for a vector space V is said to be the **dimension** of the space.

- If the basis of a vector space contains a finite number of vectors, then the space is **finite dimensional**; otherwise it is **infinite dimensional**.
- If S denotes any set of vectors $\{\mathbf{x}_1, \mathbf{x}_2, \dots, \mathbf{x}_n\}$ in a vector space V , then the set of all linear combinations of the vectors in S

$$c_1\mathbf{x}_1 + c_2\mathbf{x}_2 + \dots + c_n\mathbf{x}_n$$

is called the **span** of the vectors and is denoted $\text{Span}(S)$.

- $\text{Span}(S)$ is a subspace of V and is said to be a subspace spanned by its vectors $\mathbf{x}_1, \mathbf{x}_2, \dots, \mathbf{x}_n$.
- If $V = \text{Span}(S)$ then S is said to be a **spanning set** for the vector space V or that S **spans** V .

1.7 Gram–Schmidt Orthogonalization Process

- An **orthonormal basis** is a basis whose vectors are mutually orthogonal and are unit vectors.
- If $B = \{\mathbf{w}_1, \mathbf{w}_2, \dots, \mathbf{w}_n\}$ is an orthonormal basis for \mathbb{R}^n then an arbitrary vector \mathbf{u} can be expressed as

$$\mathbf{u} = (\mathbf{u} \cdot \mathbf{w}_1)\mathbf{w}_1 + (\mathbf{u} \cdot \mathbf{w}_2)\mathbf{w}_2 + \dots + (\mathbf{u} \cdot \mathbf{w}_n)\mathbf{w}_n$$

- The **Gram-Schmidt Orthogonalization Process** is a process for converting any basis of a vector space into an orthonormal basis. First the basis vectors are made orthogonal to each other, then they are normalized. More specifically, to convert a basis $B = \{\mathbf{u}_1, \mathbf{u}_2, \dots, \mathbf{u}_n\}$ into an orthogonal basis $B' = \{\mathbf{v}_1, \mathbf{v}_2, \dots, \mathbf{v}_n\}$

1. Let $\mathbf{v}_1 = \mathbf{u}_1$
2. Let $\mathbf{v}_2 = \mathbf{u}_2 - \text{proj}_{\mathbf{v}_1} \mathbf{u}_2$
3. ...
4. Let $\mathbf{v}_n = \mathbf{u}_n - \text{proj}_{\mathbf{v}_1} \mathbf{u}_n - \text{proj}_{\mathbf{v}_2} \mathbf{u}_n - \dots - \text{proj}_{\mathbf{v}_{n-1}} \mathbf{u}_n$

and to convert B' into an orthonormal basis $B'' = \{\mathbf{w}_1, \mathbf{w}_2, \dots, \mathbf{w}_n\}$, normalize each \mathbf{v}_i , $i = 1, 2, \dots, n$.

2 Matrices

2.1 Matrix Algebra

- Vectors can be written as horizontal or vertical arrays of numbers

- A **matrix** is any rectangular array of numbers or functions

$$\begin{pmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ a_{21} & a_{22} & \cdots & a_{2n} \\ \vdots & & & \vdots \\ a_{m1} & a_{m2} & \cdots & a_{mn} \end{pmatrix}$$

- The numbers or functions in the array are called the **elements** or **entries** of the matrix
- If a matrix has m rows and n columns we say that its **size** is m by n or $m \times n$
- An $n \times n$ matrix is called a **square** matrix of **order** n
- The entry in the i th row and the j th column of an $m \times n$ matrix **A** is written a_{ij}
- An $m \times 1$ matrix

$$\begin{pmatrix} a_1 \\ a_2 \\ \vdots \\ a_n \end{pmatrix}$$

is called a **column vector**

- A $1 \times n$ matrix

$$(a_1 \quad a_2 \quad \cdots \quad a_n)$$

is called a **row vector**

Definition 8.1.6 Matrix Multiplication

Let **A** be a matrix having m rows and p columns, and let **B** be a matrix having p rows and n columns. The **product** **AB** is the $m \times n$ matrix

$$\begin{aligned} \mathbf{AB} &= \begin{pmatrix} a_{11} & a_{12} & \cdots & a_{1p} \\ a_{21} & a_{22} & \cdots & a_{2p} \\ \vdots & & & \vdots \\ a_{m1} & a_{m2} & \cdots & a_{mp} \end{pmatrix} \begin{pmatrix} b_{11} & b_{12} & \cdots & b_{1n} \\ b_{21} & b_{22} & \cdots & b_{2n} \\ \vdots & & & \vdots \\ b_{p1} & b_{p2} & \cdots & b_{pn} \end{pmatrix} \\ &= \begin{pmatrix} a_{11}b_{11} + a_{12}b_{21} + \cdots + a_{1p}b_{p1} & \cdots & a_{11}b_{1n} + a_{12}b_{2n} + \cdots + a_{1p}b_{pn} \\ a_{21}b_{11} + a_{22}b_{21} + \cdots + a_{2p}b_{p1} & \cdots & a_{21}b_{1n} + a_{22}b_{2n} + \cdots + a_{2p}b_{pn} \\ \vdots & & \vdots \\ a_{m1}b_{11} + a_{m2}b_{21} + \cdots + a_{mp}b_{p1} & \cdots & a_{m1}b_{1n} + a_{m2}b_{2n} + \cdots + a_{mp}b_{pn} \end{pmatrix} \\ &= \left(\sum_{k=1}^p a_{ik}b_{kj} \right)_{m \times n}. \end{aligned}$$

- Matrix multiplication is associative, i.e. $\mathbf{A}(\mathbf{BC}) = (\mathbf{AB})\mathbf{C}$
- Matrix multiplication is distributive, i.e. $\mathbf{A}(\mathbf{B} + \mathbf{C}) = \mathbf{AB} + \mathbf{AC}$ and $(\mathbf{B} + \mathbf{C})\mathbf{A} = \mathbf{BA} + \mathbf{CA}$
- The **transpose** of an $m \times n$ matrix \mathbf{A} is an $n \times m$ matrix \mathbf{A}^T

$$\begin{pmatrix} a_{11} & a_{21} & \cdots & a_{m1} \\ a_{12} & a_{22} & \cdots & a_{m2} \\ \vdots & & & \vdots \\ a_{1n} & a_{2n} & \cdots & a_{mn} \end{pmatrix}$$

i.e. the matrix is flipped along the main diagonal

Theorem 8.1.2 Properties of Transpose

Suppose \mathbf{A} and \mathbf{B} are matrices and k a scalar. Then

- | | |
|--|----------------------------------|
| (i) $(\mathbf{A}^T)^T = \mathbf{A}$ | ← transpose of a transpose |
| (ii) $(\mathbf{A} + \mathbf{B})^T = \mathbf{A}^T + \mathbf{B}^T$ | ← transpose of a sum |
| (iii) $(\mathbf{AB})^T = \mathbf{B}^T \mathbf{A}^T$ | ← transpose of a product |
| (iv) $(k\mathbf{A})^T = k\mathbf{A}^T$ | ← transpose of a scalar multiple |

- A matrix that consists of all zero entries is called a **zero matrix**
- A square matrix is said to be a **triangular matrix** if all of its entries above or below the main diagonal are zeroes. More specifically they are called **lower triangular** and **upper triangular** matrices, respectively.
- A square matrix is called a **diagonal matrix** if all entries not on the main diagonal are 0.
- A square matrix whose entries on the main diagonal are all equal is called a **scalar matrix**
- A square matrix that has the property $\mathbf{A} = \mathbf{A}^T$ is called a **symmetric matrix**

2.2 Systems of Linear Algebraic Equations

- In a linear system

$$\begin{aligned} a_{11}x_1 + a_{12}x_2 + \cdots + a_{1n}x_n &= b_1 \\ a_{21}x_1 + a_{22}x_2 + \cdots + a_{2n}x_n &= b_2 \\ &\vdots \\ a_{m1}x_1 + a_{m2}x_2 + \cdots + a_{mn}x_n &= b_n \end{aligned}$$

the values a_{ij} are called the **coefficients** and the values b_n are called the **constants**

- If all the constants are zero the system is said to be **homogeneous**, otherwise it is **nonhomogeneous**
- A linear system is said to be **consistent** if it has at least one solution, otherwise it's **inconsistent**
- A linear system can be transformed into an equivalent system (i.e. one that has the same solutions) via three elementary operations:
 1. Multiply an equation by a nonzero constant
 2. Interchange the positions of equations in the system
 3. Add a multiple of one equation to any other equation

- A linear system can be represented by an **augmented matrix**, e.g.

$$\left(\begin{array}{cc|c} a_{11} & a_{12} & b_1 \\ a_{21} & a_{22} & b_2 \end{array} \right)$$

- We say that two matrices are **row equivalent** if one can be obtained from the other via a series of elementary row operations
- **Gaussian elimination** is the process of applying elementary row operations to a matrix to put it into **row-echelon form** where:
 1. The first nonzero entry in a row is a 1
 2. In subsequent rows, the first 1 entry appears to the right of the 1 entry in earlier rows
 3. Rows consisting of all zeroes are at the bottom of the matrix
- **Gauss-Jordan elimination** is the same as Gaussian elimination with an additional constraint that puts the matrix into **reduced row-echelon form** where a column containing a first entry 1 has zeroes everywhere else
- A homogeneous linear system always has a trivial solution where all variables are equal to zero and will have an infinite number of nontrivial solutions if the number of equations m is less than the number of variables n , i.e. $m < n$
- If \mathbf{X}_1 is a solution to $\mathbf{AX} = \mathbf{0}$, then so is $c\mathbf{X}_1$ for any constant c
- If \mathbf{X}_1 and \mathbf{X}_2 are solutions of $\mathbf{AX} = \mathbf{0}$, then so is $\mathbf{X}_1 + \mathbf{X}_2$
- If a linear system contains more equations than variables it is said to be **overdetermined**; if it contains fewer equations than variables it is said to be **underdetermined**

2.3 Rank of a Matrix

- The **rank** of a matrix \mathbf{A} denoted $\text{rank}(\mathbf{A})$ is the number of linearly independent row vectors in \mathbf{A}
- The row vectors of an $m \times n$ matrix \mathbf{A} span a subspace of \mathbb{R}^n . This is called the **row space** of \mathbf{A} . The set of linearly independent row vectors in \mathbf{A} are a basis for that subspace

Theorem 8.3.1 Rank of a Matrix by Row Reduction

If a matrix \mathbf{A} is row equivalent to a row-echelon form \mathbf{B} , then

- (i) the row space of \mathbf{A} = the row space of \mathbf{B} ,
- (ii) the nonzero rows of \mathbf{B} form a basis for the row space of \mathbf{A} , and
- (iii) $\text{rank}(\mathbf{A})$ = the number of nonzero rows in \mathbf{B} .

- A linear system of equations $\mathbf{AX} = \mathbf{B}$ is consistent iff the rank of the coefficient matrix \mathbf{A} is equal to the rank of the augmented matrix of the system $(\mathbf{A}|\mathbf{B})$
- Suppose a linear system $\mathbf{AX} = \mathbf{B}$ with m equations and n variables is consistent. If $\text{rank}(\mathbf{A}) = r$ then the solution of the system contains $n - r$ variables

2.4 Determinants

- Suppose \mathbf{A} is an $n \times n$ matrix. Associated with \mathbf{A} is a number called the **determinant of \mathbf{A}** and is denoted by

$$\det \mathbf{A} = \begin{vmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ a_{21} & a_{22} & \cdots & a_{2n} \\ \vdots & & & \vdots \\ a_{n1} & a_{n2} & \cdots & a_{nn} \end{vmatrix}$$

- A determinant of an $n \times n$ matrix is called a **determinant of order n**
- The determinant of a 1×1 matrix is the element of the matrix
- Each element in an $n \times n$ matrix has an associated **cofactor** defined as

$$a_{ij} = (-1)^{i+j} M_{ij}$$

where M_{ij} is the determinant of the $(n-1) \times (n-1)$ matrix produced by deleting row i and column j from \mathbf{A}

- The determinant of an arbitrary $n \times n$ matrix \mathbf{A} can be calculated by choosing an arbitrary row or column and summing the products of each element in that column/row with their cofactors, e.g. if we choose the first row of a 3×3 matrix then

$$\begin{aligned}
\begin{vmatrix} a_{11} & a_{12} & a_{13} \\ a_{21} & a_{22} & a_{23} \\ a_{31} & a_{32} & a_{33} \end{vmatrix} &= a_{11}M_{11} + a_{12}M_{12} + a_{13}M_{13} \\
&= a_{11} \begin{vmatrix} a_{22} & a_{23} \\ a_{32} & a_{33} \end{vmatrix} - a_{12} \begin{vmatrix} a_{21} & a_{23} \\ a_{31} & a_{33} \end{vmatrix} + a_{13} \begin{vmatrix} a_{21} & a_{22} \\ a_{31} & a_{32} \end{vmatrix} \\
&= a_{11}(a_{22}|a_{33}| - a_{23}|a_{32}|) - a_{12}(a_{21}|a_{33}| - a_{23}|a_{31}|) \\
&\quad + a_{13}(a_{21}|a_{32}| - a_{22}|a_{31}|)
\end{aligned}$$

2.5 Properties of Determinants

- The determinant of a matrix and its transpose are the same
- If any two rows/columns of a matrix are the same its determinant is zero
- If all the entries in a row/column of a matrix are zero, then its determinant is zero
- Interchanging any two rows/columns of a matrix negates its determinant
- Multiplying a row/column of a matrix by a nonzero real number k also multiplies the determinant by k
- If \mathbf{A} and \mathbf{B} are both $n \times n$ matrices, then $\det \mathbf{AB} = \det \mathbf{A} \cdot \det \mathbf{B}$
- Adding a multiply of one row/column to another doesn't change the determinant
- The determinant of a triangular matrix is the product of the entries along the main diagonal
- Sometimes it's faster to calculate a matrix's determinant by reducing it to row-echelon form and multiplying the elements along the main diagonal than performing cofactor expansion
- Multiplying the entries of a row/column with the cofactors of another row/column and summing the results always equals zero

2.6 Inverse of a Matrix

- Given an $n \times n$ matrix \mathbf{A} , if there exists another $n \times n$ matrix \mathbf{B} such that $\mathbf{AB} = \mathbf{BA} = \mathbf{I}$ then \mathbf{A} is said to be **nonsingular** or **invertible** and \mathbf{B} is said to be the unique **inverse** of \mathbf{A} , i.e. $\mathbf{B} = \mathbf{A}^{-1}$
- Some $n \times n$ matrices don't have an inverse and are called **singular**

- The **adjoint** of an $n \times n$ matrix \mathbf{A} is the transpose of the matrix of cofactors corresponding to the entries of \mathbf{A}

$$\text{adj } \mathbf{A} = \begin{pmatrix} C_{11} & C_{12} & \cdots & C_{1n} \\ C_{21} & C_{22} & \cdots & C_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ C_{n1} & C_{n2} & \cdots & C_{nn} \end{pmatrix}^T = \begin{pmatrix} C_{11} & C_{21} & \cdots & C_{n1} \\ C_{12} & C_{22} & \cdots & C_{n2} \\ \vdots & \vdots & \ddots & \vdots \\ C_{1n} & C_{2n} & \cdots & C_{nn} \end{pmatrix}$$

- If \mathbf{A} is an $n \times n$ matrix and $\det \mathbf{A} \neq 0$ then

$$\mathbf{A}^{-1} = \left(\frac{1}{\det \mathbf{A}} \right) \text{adj } \mathbf{A}$$

- From the above, the inverse of a 2×2 matrix \mathbf{A} is

$$\mathbf{A}^{-1} = \frac{1}{\det \mathbf{A}} \begin{pmatrix} a_{22} & -a_{12} \\ -a_{21} & a_{11} \end{pmatrix}$$

- An $n \times n$ matrix \mathbf{A} is nonsingular (has an inverse) if $\det \mathbf{A} \neq 0$

Theorem 8.6.4 Finding the Inverse

If an $n \times n$ matrix \mathbf{A} can be transformed into the $n \times n$ identity \mathbf{I} by a sequence of elementary row operations, then \mathbf{A} is nonsingular. The same sequence of operations that transforms \mathbf{A} into the identity \mathbf{I} will also transform \mathbf{I} into \mathbf{A}^{-1} .

- Inverse matrices can be used to solve linear systems. If $\mathbf{A}\mathbf{X} = \mathbf{B}$ and \mathbf{A} is invertible, then

$$\mathbf{A}^{-1}\mathbf{A}\mathbf{X} = \mathbf{A}^{-1}\mathbf{B} \Rightarrow \mathbf{X} = \mathbf{A}^{-1}\mathbf{B}$$

- When $\det \mathbf{A} \neq 0$ the solution of the system $\mathbf{A}\mathbf{X} = \mathbf{B}$ is unique
- A homogeneous system of linear equations $\mathbf{A}\mathbf{X} = \mathbf{0}$ has only the trivial solution iff \mathbf{A} is nonsingular and an infinite number of solutions iff it is singular

2.7 Cramer's Rule

- If \mathbf{A} is the coefficient matrix of a linear system and $\det \mathbf{A} \neq 0$, then the solution of the system is given by

$$\begin{aligned} x_1 &= \frac{\det \mathbf{A}_1}{\det \mathbf{A}} \\ x_2 &= \frac{\det \mathbf{A}_2}{\det \mathbf{A}} \\ &\vdots \\ x_n &= \frac{\det \mathbf{A}_n}{\det \mathbf{A}} \end{aligned}$$

where \mathbf{A}_n is the matrix obtained by replacing column n of \mathbf{A} with the constants of the system.

2.8 The Eigenvalue Problem

- If \mathbf{A} is an $n \times n$ matrix, a number λ is said to be an **eigenvalue** of \mathbf{A} if there exists a nonzero solution vector \mathbf{K} of the linear system $\mathbf{A}\mathbf{K} = \lambda\mathbf{K}$. The solution vector \mathbf{K} is said to be an **eigenvector** corresponding to the eigenvalue λ .
- Rearranging the equation above we find

$$(\mathbf{A} - \lambda\mathbf{I})\mathbf{K} = \mathbf{0}$$

which only has nontrivial solutions if $\det(\mathbf{A} - \lambda\mathbf{I}) = 0$.

- Calculating $\det(\mathbf{A} - \lambda\mathbf{I})$ results in an n -th degree polynomial in λ called the **characteristic equation** of \mathbf{A} , the solutions to which are its eigenvalues.
- The eigenvector associated with a particular eigenvalue can be found by applying Gauss-Jordan elimination to the augmented matrix $(\mathbf{A} - \lambda\mathbf{I}|\mathbf{0})$.
- A nonzero constant multiple of an eigenvector is another eigenvector.
- If λ is a complex eigenvalue of a matrix, then its conjugate λ^* is also an eigenvalue. If \mathbf{K} is an eigenvector corresponding to λ then its conjugate \mathbf{K}^* is an eigenvector corresponding to λ^* .
- $\lambda = 0$ is an eigenvalue of a matrix iff the matrix isn't invertible
- The determinant of a matrix is the product of its eigenvalues
- If λ is an eigenvalue of a matrix \mathbf{A} with eigenvector \mathbf{K} , then $1/\lambda$ is an eigenvalue of \mathbf{A}^{-1} with the same eigenvector.
- The eigenvalues of a triangular matrix are the entries along the main diagonal.

2.9 Powers of Matrices

- Any $n \times n$ matrix \mathbf{A} satisfies its own characteristic equation, i.e. λ can be replaced with \mathbf{A} in the characteristic equation.
- This gives us an expression for \mathbf{A}^n as a linear combination

$$\mathbf{A}^n = c_0\mathbf{I} + c_1\mathbf{A} + c_2\mathbf{A}^2 + \cdots + c_{n-1}\mathbf{A}^{n-1}.$$

If we multiply this expression by \mathbf{A} we get an expression for \mathbf{A}^{n+1} and we can replace the \mathbf{A}^n term with the original expression. This can be repeated an arbitrary number of times to find expressions for any power of \mathbf{A} .

- The constants of the linear combination can be determined by substituting the matrix's eigenvalues into the characteristic equation, resulting in a linear system where the constants are the variables. Solving the system determines the constants.
- If \mathbf{A} is a nonsingular matrix, the fact that it satisfies its own characteristic equation can be used to determine its inverse. This can be achieved by replacing λ with \mathbf{A} in its characteristic equation, solving for \mathbf{I} , and multiplying both sides by \mathbf{A}^{-1} . This results in an expression for \mathbf{A}^{-1} as a linear combination of powers of \mathbf{A} .

2.10 Orthogonal Matrices

- If \mathbf{A} is a symmetric matrix with real entries, then the eigenvalues of \mathbf{A} are real.
- If \mathbf{A} is a symmetric matrix, then the eigenvectors corresponding to different eigenvalues are orthogonal.
- An $n \times n$ nonsingular matrix \mathbf{A} is **orthogonal** if $\mathbf{A}^{-1} = \mathbf{A}^T$.
- An $n \times n$ matrix \mathbf{A} is orthogonal iff its columns form an orthonormal set.
- If an $n \times n$ matrix \mathbf{A} has n distinct eigenvalues, an orthogonal matrix can be formed by normalizing its eigenvectors and using them as column vectors in a new matrix.

2.11 Approximation of Eigenvalues

- Let $\lambda_1, \lambda_2, \dots, \lambda_n$ denote the eigenvalues of an $n \times n$ matrix \mathbf{A} . The eigenvalue λ_k is said to be the **dominant eigenvalue** of \mathbf{A} if

$$|\lambda_k| > |\lambda_i|, i = 1, 2, \dots, n, i \neq k.$$

An eigenvector corresponding to λ_k is called the **dominant eigenvector** of \mathbf{A} .

- **Power iteration** is a method for approximating the dominant eigenvector of an $n \times n$ matrix \mathbf{A} .
 1. Choose an arbitrary starting vector \mathbf{X}_0
 2. An approximation of the dominant eigenvector is $\mathbf{X}_m = \mathbf{A}^m \mathbf{X}_0$
 3. An approximation of the dominant eigenvalue is

$$\lambda \approx \frac{\mathbf{A}\mathbf{X}_m \cdot \mathbf{X}_m}{\mathbf{X}_m \cdot \mathbf{X}_m}$$

- If \mathbf{X}_m is computed via repeated multiplications of \mathbf{A} rather than computing \mathbf{A}^m in advance the entries of the intermediary vectors can become quite large and pose problems for computers. This can be avoided by normalising or scaling down the vectors after each iteration.
- The **method of deflation** is a way to find nondominant eigenvalues of an $n \times n$ symmetric matrix \mathbf{A} that has eigenvalues $|\lambda_1| > |\lambda_2| > |\lambda_3| \geq \dots \geq |\lambda_n|$.
 1. Compute the dominant eigenvalue λ_1 and normalised eigenvector \mathbf{K}_1 of the matrix using power iteration.
 2. Compute the matrix $\mathbf{B} = \mathbf{A} - \lambda_1 \mathbf{K}_1 \mathbf{K}_1^T$ which has eigenvalues $0, \lambda_2, \lambda_3, \dots, \lambda_n$
 3. Apply power iteration to find λ_2 and \mathbf{K}_2
 4. Repeat steps 2 and 3 to compute subsequent eigenvalues
- The **inverse power method** is a way to find the eigenvalue with smallest absolute value. If \mathbf{A} is nonsingular then the eigenvalues of \mathbf{A}^{-1} are the reciprocals of the eigenvalues of \mathbf{A} . This means the eigenvalue of \mathbf{A} with smallest absolute value is the dominant eigenvalue of \mathbf{A}^{-1} and can be found via power iteration.

2.12 Diagonalization

- If an $n \times n$ nonsingular matrix \mathbf{P} can be found so that $\mathbf{P}^{-1}\mathbf{A}\mathbf{P} = \mathbf{D}$ is a diagonal matrix, then we say that the $n \times n$ matrix \mathbf{A} can be **diagonalised**, or is **diagonalisable**, and that \mathbf{P} **diagonalises** \mathbf{A} .
- An $n \times n$ matrix \mathbf{A} is diagonalisable iff \mathbf{A} has n linearly independent eigenvectors $\mathbf{K}_1, \mathbf{K}_2, \dots, \mathbf{K}_n$. If we let $\mathbf{P} = (\mathbf{K}_1 \ \mathbf{K}_2 \ \dots \ \mathbf{K}_n)$ then

$$\begin{aligned}
 \mathbf{A}\mathbf{P} &= (\mathbf{A}\mathbf{K}_1 \ \mathbf{A}\mathbf{K}_2 \ \dots \ \mathbf{A}\mathbf{K}_n) \\
 &= (\lambda_1 \mathbf{K}_1 \ \lambda_2 \mathbf{K}_2 \ \dots \ \lambda_n \mathbf{K}_n) \\
 &= (\mathbf{K}_1 \ \mathbf{K}_2 \ \dots \ \mathbf{K}_n) \begin{pmatrix} \lambda_1 & 0 & \dots & 0 \\ 0 & \lambda_2 & \dots & 0 \\ \vdots & & \ddots & \vdots \\ 0 & 0 & \dots & \lambda_n \end{pmatrix} \\
 &= \mathbf{P}\mathbf{D}
 \end{aligned}$$

- If an $n \times n$ matrix \mathbf{A} has n distinct eigenvalues, it is diagonalisable. If it has fewer than n distinct eigenvalues it may still be diagonalisable.
- Symmetric matrices with real entries are always diagonalisable.

2.13 LU-Factorisation

- If an $n \times n$ matrix \mathbf{A} can be written as a product $\mathbf{A} = \mathbf{L}\mathbf{U}$ where \mathbf{L} and \mathbf{U} are lower and upper triangular matrices, respectively, then we say that $\mathbf{A} = \mathbf{L}\mathbf{U}$ is an **LU-factorisation** of \mathbf{A} .
- An $n \times n$ matrix \mathbf{A} can have several LU-factorisations
- **Doolittle's method** is a method of performing LU-factorisation.
 1. Assume the diagonal entries of \mathbf{L} are 1, i.e. $l_{ii} = 1, i = 1, 2, \dots, n$
 2. Multiply \mathbf{L} and \mathbf{U} (with placeholder entries)
 3. Equate the resulting entries with those of the original matrix — this gives n^2 equations, but each equation only uses variables determined in previous equations allowing the system to be solved
- An alternative algorithm for Doolittle's method is
 1. Perform elementary row operations on \mathbf{A} until you have an upper triangular matrix \mathbf{U}
 2. Each time you add a c times row i to row j , record the $-c$ in the j -th row and i -th column of an identity matrix
 3. The matrix from step 2 is \mathbf{L}
- Given a linear system $\mathbf{A}\mathbf{X} = \mathbf{B}$, if \mathbf{A} has an LU-factorisation the system can be solved as follows:
 1. Rewrite the system $\mathbf{L}\mathbf{U}\mathbf{X} = \mathbf{B}$
 2. Let $\mathbf{U}\mathbf{X} = \mathbf{Y}$ where

$$\mathbf{Y} = \begin{pmatrix} y_1 \\ y_2 \\ \vdots \\ y_n \end{pmatrix}$$
 3. Solve $\mathbf{L}\mathbf{Y} = \mathbf{B}$ via forward substitution, i.e. find y_1 , use that to find y_2 , etc.
 4. Substitute the values of y_n into $\mathbf{U}\mathbf{X} = \mathbf{Y}$ and solve via back substitution, i.e. find x_n , use that to find x_{n-1} , etc.
- If a matrix \mathbf{A} has an LU-factorisation $\mathbf{A} = \mathbf{L}\mathbf{U}$ then the determinant of \mathbf{A} can be calculated as $\det \mathbf{A} = \det \mathbf{L} \cdot \det \mathbf{U}$ which is simply the product of the diagonal entries of \mathbf{L} and \mathbf{U}
- If row interchanges are required to arrive at \mathbf{U} then an LU-factorisation doesn't exist

2.14 Cryptography

- If you define a mapping between a set of characters allowed in messages and a list of integers, messages can be represented as an $n \times m$ matrix, a nonsingular $n \times n$ matrix \mathbf{A} can be used as an encryption key, and its inverse \mathbf{A}^{-1} can be used as a decryption key.

3 Vector Calculus

3.1 Vector Functions

- A curve C in the xy -plane is a set of ordered pairs (x, y) . We say C is a **parametric curve** if the x - and y -coordinates of a point on the curve are defined by a pair of functions $x = f(t)$ and $y = g(t)$ that are continuous on some interval $a \leq t \leq b$.
- If $\mathbf{r}(t) = \langle f(t), g(t), h(t) \rangle$, then $\mathbf{r}'(t) = \langle f'(t), g'(t), h'(t) \rangle$.
- If \mathbf{r} is a differentiable vector function and $s = u(t)$ is a differentiable scalar function, then the derivative of $\mathbf{r}(s)$ with respect to t is

$$\frac{d\mathbf{r}}{dt} = \frac{d\mathbf{r}}{ds} \frac{ds}{dt} = \mathbf{r}'(s)u'(t).$$

Theorem 9.1.4 Rules of Differentiation

Let \mathbf{r}_1 and \mathbf{r}_2 be differentiable vector functions and $u(t)$ a differentiable scalar function.

- (i) $\frac{d}{dt} [\mathbf{r}_1(t) + \mathbf{r}_2(t)] = \mathbf{r}_1'(t) + \mathbf{r}_2'(t)$
- (ii) $\frac{d}{dt} [u(t)\mathbf{r}_1(t)] = u(t)\mathbf{r}_1'(t) + u'(t)\mathbf{r}_1(t)$
- (iii) $\frac{d}{dt} [\mathbf{r}_1(t) \cdot \mathbf{r}_2(t)] = \mathbf{r}_1(t) \cdot \mathbf{r}_2'(t) + \mathbf{r}_1'(t) \cdot \mathbf{r}_2(t)$
- (iv) $\frac{d}{dt} [\mathbf{r}_1(t) \times \mathbf{r}_2(t)] = \mathbf{r}_1(t) \times \mathbf{r}_2'(t) + \mathbf{r}_1'(t) \times \mathbf{r}_2(t).$

- Because the cross product of two vectors isn't commutative, the order in which \mathbf{r}_1 and \mathbf{r}_2 appear above is important.
- The indefinite integral of a vector function is defined as

$$\int \mathbf{r}(t) dt = \left\langle \int f(t) dt, \int g(t) dt, \int h(t) dt \right\rangle = \mathbf{R}(t) + \mathbf{c}$$

- The definite integral of a vector function is defined as

$$\int_a^b \mathbf{r}(t) dt = \left\langle \int_a^b f(t) dt, \int_a^b g(t) dt, \int_a^b h(t) dt \right\rangle = \mathbf{R}$$

- The length of the curve traced out by a vector function from $t = a$ to $t = b$ is

$$s = \int_a^b \sqrt{f'(t)^2 + g'(t)^2 + h'(t)^2} dt = \int_a^b \|\mathbf{r}'(t)\| dt$$

3.3 Curvature and Components of Acceleration

- As $\mathbf{r}'(t)$ is always tangential to the curve a unit tangent vector is given by

$$\mathbf{T}(t) = \frac{\mathbf{r}'(t)}{\|\mathbf{r}'(t)\|}$$

- The **curvature** of a point on a curve is given by

$$\kappa = \left\| \frac{d\mathbf{T}}{ds} \right\|$$

where s is the arc length parameter or

$$\kappa = \frac{\|\mathbf{T}'(t)\|}{\|\mathbf{r}'(t)\|}$$

- By differentiating

$$\begin{aligned} \mathbf{T} \cdot \mathbf{T} &= 1 \\ \mathbf{T} \cdot \frac{d\mathbf{T}}{dt} + \frac{d\mathbf{T}}{dt} \cdot \mathbf{T} &= 0 \\ 2\mathbf{T} \cdot \frac{d\mathbf{T}}{dt} &= 0 \\ \mathbf{T} \cdot \frac{d\mathbf{T}}{dt} &= 0 \end{aligned}$$

we find that \mathbf{T} and $\frac{d\mathbf{T}}{dt}$ are orthogonal.

- If $\left\| \frac{d\mathbf{T}}{dt} \right\| \neq 0$ the vector

$$\mathbf{N} = \frac{d\mathbf{T}/dt}{\|d\mathbf{T}/dt\|}$$

is a unit normal to the curve and is called the **principal normal**.

- Since $\kappa = \frac{\|d\mathbf{T}/dt\|}{v}$, $d\mathbf{T}/dt = \kappa v \mathbf{N}$ and

$$\begin{aligned} \mathbf{a}(t) &= \frac{d}{dt} \mathbf{v}(t) \\ &= \frac{d}{dt} v \mathbf{T} \\ &= v \frac{d\mathbf{T}}{dt} + \frac{dv}{dt} \mathbf{T} \\ &= \kappa v^2 \mathbf{N} + \frac{dv}{dt} \mathbf{T} \\ &= a_N \mathbf{N} + a_T \mathbf{T} \end{aligned}$$

where a_N and a_T are the normal and tangential components of acceleration, respectively.

- The unit vector defined by

$$\mathbf{B}(t) = \mathbf{T}(t) \times \mathbf{N}(t)$$

is called the **binormal**.

- The three unit vectors \mathbf{T} , \mathbf{N} , and \mathbf{B} form a right-handed set of mutually orthogonal vectors called the **moving trihedral**. When used as a coordinate system they're called the **TNB-frame**.
- The plane of \mathbf{T} and \mathbf{N} is called the **osculating plane**.
- The plane of \mathbf{N} and \mathbf{B} is called the **normal plane**.
- The plane of \mathbf{T} and \mathbf{B} is called the **rectifying plane**.
- Explicit formulas for the tangential and normal components of acceleration are given by

$$a_T = \frac{\mathbf{r}'(t) \cdot \mathbf{r}''(t)}{\|\mathbf{r}'(t)\|}$$

$$a_N = \frac{\|\mathbf{r}'(t) \times \mathbf{r}''(t)\|}{\|\mathbf{r}'(t)\|^2}$$

and since $a_N = \kappa v^2$

$$\kappa = \frac{\|\mathbf{r}'(t) \times \mathbf{r}''(t)\|}{\|\mathbf{r}'(t)\|^3}$$

- The reciprocal of curvature $\rho = \frac{1}{\kappa}$ is called the **radius of curvature** and represents the radius of the circle that best “fits” the curve there.

3.4 Partial Derivatives

- The **level curves** of a function of two variables $z = f(x, y)$ are the curves resulting from the equation $c = f(x, y)$ for any real value of c .
- The **level surfaces** of a function of three variables $w = f(x, y, z)$ are the surfaces resulting from the equation $c = f(x, y, z)$ for any real value of c .
- The partial derivative of a function $f(x_1, x_2, \dots, x_n)$ with respect to a variable x_i is the derivative of that function with respect to x_i while holding all other variables constant.
- The partial derivative of f with respect to x can be denoted $\frac{\partial f}{\partial x}$ or f_x .

- Because partial derivatives are themselves multivariable functions you can take subsequent partial derivatives, including in other variables, e.g.

$$\frac{\partial^2 f}{\partial x^2} = \frac{\partial}{\partial x} \left(\frac{\partial f}{\partial x} \right) \text{ or } \frac{\partial f}{\partial x \partial y} = \frac{\partial}{\partial x} \left(\frac{\partial f}{\partial y} \right).$$

- When multiple derivatives are taken in different variables it's called a **mixed partial derivative**.
- The order in which a mixed partial derivative is computed doesn't matter, i.e.

$$\frac{\partial f}{\partial x \partial y} = \frac{\partial f}{\partial y \partial x}.$$

Theorem 9.4.1 Chain Rule

If $z = f(u, v)$ is differentiable and $u = g(x, y)$ and $v = h(x, y)$ have continuous first partial derivatives, then

$$\frac{\partial z}{\partial x} = \frac{\partial z}{\partial u} \frac{\partial u}{\partial x} + \frac{\partial z}{\partial v} \frac{\partial v}{\partial x}, \quad \frac{\partial z}{\partial y} = \frac{\partial z}{\partial u} \frac{\partial u}{\partial y} + \frac{\partial z}{\partial v} \frac{\partial v}{\partial y}. \quad (5)$$

3.5 Directional Derivative

- In n dimensions the **vector differential operator** is defined as

$$\nabla = \frac{\partial}{\partial x_1} \hat{\mathbf{e}}_1 + \frac{\partial}{\partial x_2} \hat{\mathbf{e}}_2 + \cdots + \frac{\partial}{\partial x_n} \hat{\mathbf{e}}_n.$$

- When the vector differential operator is applied to a scalar function the result is called the **gradient** of the function. The gradient of a function points in the direction in which the function increases most rapidly.
- The **directional derivative** of a function $f(x_1, x_2, \dots, x_n)$ in the direction of the unit vector \mathbf{u} is given by

$$D_{\mathbf{u}} f(x_1, x_2, \dots, x_n) = \nabla f(x_1, x_2, \dots, x_n) \cdot \mathbf{u}.$$

3.6 Tangent Planes and Normal Lines

- If $f(x, y)$ is a two-dimensional function, ∇f is always orthogonal to the level curves of $f(x, y)$.
- If $f(x, y, z)$ is a three-dimensional function, ∇f is always normal to the level surfaces of $f(x, y, z)$.

Theorem 9.6.1 Equation of Tangent Plane

Let $P(x_0, y_0, z_0)$ be a point on the graph of $F(x, y, z) = c$, where ∇F is not $\mathbf{0}$. Then an equation of the tangent plane at P is

$$F_x(x_0, y_0, z_0)(x - x_0) + F_y(x_0, y_0, z_0)(y - y_0) + F_z(x_0, y_0, z_0)(z - z_0) = 0. \quad (5)$$

3.7 Curl and Divergence

- The **curl** of a vector field $\mathbf{F}(x, y, z) = \langle P(x, y, z), Q(x, y, z), R(x, y, z) \rangle$ is defined as

$$\text{curl } \mathbf{F} = \nabla \times \mathbf{F} = \begin{vmatrix} \hat{\mathbf{i}} & \hat{\mathbf{j}} & \hat{\mathbf{k}} \\ \frac{\partial}{\partial x} & \frac{\partial}{\partial y} & \frac{\partial}{\partial z} \\ P & Q & R \end{vmatrix} = \left\langle \frac{\partial R}{\partial y} - \frac{\partial Q}{\partial z}, \frac{\partial P}{\partial z} - \frac{\partial R}{\partial x}, \frac{\partial Q}{\partial x} - \frac{\partial P}{\partial y} \right\rangle.$$

- The **divergence** of the same vector field is defined as

$$\nabla \cdot \mathbf{F} = \frac{\partial P}{\partial x} + \frac{\partial Q}{\partial y} + \frac{\partial R}{\partial z}.$$

- If f is a scalar function with continuous second partial derivatives then

$$\nabla \times (\nabla f) = \mathbf{0}.$$

- If \mathbf{F} is a vector function with continuous second partial derivatives then

$$\nabla \cdot (\nabla \times \mathbf{F}) = 0.$$

3.8 Line Integrals

- The line integral of a function $G(x, y)$ over a curve C defined parametrically by $x = f(t)$ and $y = g(t)$ is defined as

$$\begin{aligned} \int_C G(x, y) dx &= \int_a^b G(f(t), g(t)) f'(t) dt \\ \int_C G(x, y) dy &= \int_a^b G(f(t), g(t)) g'(t) dt \\ \int_C G(x, y) ds &= \int_a^b G(f(t), g(t)) \sqrt{[f'(t)]^2 + [g'(t)]^2} dt \end{aligned}$$

where a and b are the values of t at which the curve starts and ends, respectively.

- If the curve is defined by an explicit function $y = f(x)$ the integrals become

$$\begin{aligned} \int_C G(x, y) dx &= \int_a^b G(x, f(x)) dx \\ \int_C G(x, y) dy &= \int_a^b G(x, f(x)) f'(x) dx \\ \int_C G(x, y) ds &= \int_a^b G(x, f(x)) \sqrt{1 + [f'(x)]^2} dx \end{aligned}$$

- A line integral along a piecewise smooth curve C is defined as the sum of the integrals over the smooth segments that comprise C .
- The integral of a vector function $\mathbf{F}(x, y) = \langle P(x, y), Q(x, y) \rangle$ over a curve C described by $\mathbf{r}(t) = \langle f(t), g(t) \rangle$ can be denoted

$$\int_C \mathbf{F} \cdot d\mathbf{r} = \int_C P dx + Q dy.$$

3.9 Independence of the Path

- A vector function \mathbf{F} in 2- or 3-space is said to be **conservative** if \mathbf{F} can be written as the gradient of a scalar function ϕ . The function ϕ is called a **potential function** for \mathbf{F} .

Theorem 9.9.1 Fundamental Theorem

Suppose C is a path in an open region R of the xy -plane and is defined by $\mathbf{r}(t) = x(t)\mathbf{i} + y(t)\mathbf{j}$, $a \leq t \leq b$. If $\mathbf{F}(x, y) = P(x, y)\mathbf{i} + Q(x, y)\mathbf{j}$ is a conservative vector field in R and ϕ is a potential function for \mathbf{F} , then

$$\int_C \mathbf{F} \cdot d\mathbf{r} = \int_C \nabla \phi \cdot d\mathbf{r} = \phi(B) - \phi(A), \quad (2)$$

where $A = (x(a), y(a))$ and $B = (x(b), y(b))$.

- A region is said to be **connected** if every pair of points A and B in the region can be joined by a piecewise-smooth curve that lies entirely in the region.
- A region is said to be **simply connected** if it is connected and every simple closed curve C lying entirely within the region can be shrunk to a point without leaving the region. That is, the region contains no holes.
- A region is said to be **open** if it contains no boundary points.
- In an open and connected region R , $\int_C \mathbf{F} \cdot d\mathbf{r}$ is independent of the path C iff the vector field \mathbf{F} is conservative in R .

Theorem 9.9.4 Test for a Conservative Field

Suppose $\mathbf{F}(x, y) = P(x, y)\mathbf{i} + Q(x, y)\mathbf{j}$ is a conservative vector field in an open region R , and that P and Q are continuous and have continuous first partial derivatives in R . Then

$$\frac{\partial P}{\partial y} = \frac{\partial Q}{\partial x} \quad (6)$$

for all (x, y) in R . Conversely, if the equality (6) holds for all (x, y) in a simply connected region R , then $\mathbf{F} = P\mathbf{i} + Q\mathbf{j}$ is conservative in R .

- **Partial integration** is the integration counterpart to partial differentiation.

- The potential function $\phi(x, y)$ associated with a vector field $\mathbf{F}(x, y) = \langle P(x, y), Q(x, y) \rangle$ can be found by:
 1. Partially integrate $P(x, y)$ with respect to x to find an expression for $\phi(x, y)$ with a constant of integration $g(y)$.
 2. Take the partial derivative of the above with respect to y .
 3. Equate the result with $Q(x, y)$ to determine $g'(y)$.
 4. Integrate $g'(y)$ with respect to y to find $g(y)$.
 5. Substitute $g(y)$ into $\phi(x, y)$.

3.10 Double Integrals

- The double integral

$$V = \iint_R f(x, y) dA$$

is the volume above the region R and below the surface $z = f(x, y)$.

Theorem 9.10.1 Properties of Double Integrals

Let f and g be functions of two variables that are integrable over a region R . Then

- (i) $\iint_R kf(x, y) dA = k \iint_R f(x, y) dA$, where k is any constant
- (ii) $\iint_R [f(x, y) \pm g(x, y)] dA = \iint_R f(x, y) dA \pm \iint_R g(x, y) dA$
- (iii) $\iint_R f(x, y) dA = \iint_{R_1} f(x, y) dA + \iint_{R_2} f(x, y) dA$, where R_1 and R_2 are subregions of R that do not overlap and $R = R_1 \cup R_2$.

- A region defined as the area between two fixed x values $a \leq x \leq b$ and two variable y values $g_1(x) \leq y \leq g_2(x)$ is called a **region of Type I**.
- A region defined as the area between two fixed y values $c \leq y \leq d$ and two variable x values $h_1(y) \leq x \leq h_2(y)$ is called a **region of Type II**.

Theorem 9.10.2 Fubini's Theorem

Let f be continuous on a region R .

- (i) If R is of Type I, then

$$\iint_R f(x, y) dA = \int_a^b \int_{g_1(x)}^{g_2(x)} f(x, y) dy dx. \quad (6)$$

- (ii) If R is of Type II, then

$$\iint_R f(x, y) dA = \int_c^d \int_{h_1(y)}^{h_2(y)} f(x, y) dx dy. \quad (7)$$

- Sometimes it's difficult or impossible to evaluate the components of a double integral in a particular order. The order can be swapped to avoid this without changing the result.

3.11 Double Integrals in Polar Coordinates

- Double integrals can be evaluated in polar coordinates as

$$\iint_R f(r, \theta) dA = \int_{\alpha}^{\beta} \int_{g_1(\theta)}^{g_2(\theta)} f(r, \theta) r dr d\theta$$

or

$$\iint_R f(r, \theta) dA = \int_a^b \int_{h_1(r)}^{h_2(r)} f(r, \theta) r d\theta dr.$$

- An integral in cartesian coordinates can be converted to polar coordinates

$$\iint_R f(x, y) dA = \int_{\alpha}^{\beta} \int_{g_1(\theta)}^{g_2(\theta)} f(r \cos \theta, r \sin \theta) r dr d\theta$$

providing R can be described by the region bounded by $\alpha \leq \theta \leq \beta$ and $g_1(\theta) \leq r \leq g_2(\theta)$.