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

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Contents

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 **a** in the direction of a vector **b**

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

- To project a vector **a** onto a vector **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 **a**, **b**, and **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 **a** and **b** is $||\mathbf{a} \times \mathbf{b}||$
- The area of a triangle with sides **a** and **b** is $\frac{1}{2}||\mathbf{a} \times \mathbf{b}||$
- The volume of a paralleleipied with sides **a**, **b**, and **c** is $|\mathbf{a} \cdot (\mathbf{b} \times \mathbf{c})|$
- $\mathbf{a} \cdot (\mathbf{b} \times \mathbf{c}) = 0$ iff **a**, **b**, and **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 **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 parellel 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 multipying 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 $\{\mathbf{x_1}, \mathbf{x_2}, \ldots, \mathbf{x_n}\}$ is said to be **linearly independent** if the only constants satisfying the equation

$$k_1\mathbf{x_1} + k_2\mathbf{x_2} + \cdots + k_n\mathbf{x_n} = \mathbf{0}$$

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

- If a set of vectors $B = \{\mathbf{x}_1, \mathbf{x}_2, \ldots, \mathbf{x}_n\}$ 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 $\{\mathbf{x}_1, \mathbf{x}_2, \ldots, \mathbf{x}_n\}$ in a vector space V, then the set of all linear combinations of the vectors in S

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

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

- $\mathrm{Span}(S)$ is a subspace of V and is said to be a subspace spanned by its vectors $\mathbf{x}_1, \mathbf{x}_2, \ldots, \mathbf{x}_n$.

- If $V = \mathrm{Span}(S)$ 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 $B = \{\mathbf{w}_1, \mathbf{w}_2, \ldots, \mathbf{w}_n\}$ is an orthonormal basis for $\mathbb{R}^n$ then an arbitrary vector $\mathbf{u}$ can be expressed as

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

- 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 $B = \{\mathbf{u}_1, \mathbf{u}_2, \ldots, \mathbf{u}_n\}$ into an orthogonal basis $B' = \{\mathbf{v}_1, \mathbf{v}_2, \ldots, \mathbf{v}_n\}$

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

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

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 ith row and the jth column of an $m \times n$ matrix $\mathbf{A}$ 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

$$\begin{pmatrix} a_1 & a_2 & \cdots & a_n \end{pmatrix}$$

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

$$\mathbf{AB} = \begin{pmatrix} a_{11} & a_{12} & \cdots & a_{1p} \\ a_{21} & a_{22} & \cdots & a_{2p} \\ \vdots & & & \vdots \\ a_{m1} & a_{m2} & \cdots & a_{mp} \end{pmatrix} \begin{pmatrix} b_{11} & b_{12} & \cdots & b_{1n} \\ b_{21} & b_{22} & \cdots & b_{2n} \\ \vdots & & & \vdots \\ b_{p1} & b_{p2} & \cdots & b_{pn} \end{pmatrix}$$

$$= \begin{pmatrix} a_{11}b_{11} + a_{12}b_{21} + \cdots + a_{1p}b_{p1} & \cdots & a_{11}b_{1n} + a_{12}b_{2n} + \cdots + a_{1p}b_{pn} \\ a_{21}b_{11} + a_{22}b_{21} + \cdots + a_{2p}b_{p1} & \cdots & a_{21}b_{1n} + a_{22}b_{2n} + \cdots + a_{2p}b_{pn} \\ \vdots & & \vdots \\ a_{m1}b_{11} + a_{m2}b_{21} + \cdots + a_{mp}b_{p1} & \cdots & a_{m1}b_{1n} + a_{m2}b_{2n} + \cdots + a_{mp}b_{pn} \end{pmatrix}$$

$$= \left(\sum_{k=1}^{p} a_{ik}b_{kj} \right)_{m \times n}.$$

- 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 **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 **A** and **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 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*) 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 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 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/colum 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 $\mathbf{A}$

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

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

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

- 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 $\mathbf{AX} = \mathbf{B}$ and $\mathbf{A}$ is invertible, then

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

- When $\det \mathbf{A} \neq 0$ the solution of the system $\mathbf{AX} = \mathbf{B}$ is unique
- A homogeneous system of linear equations $\mathbf{AX} = \mathbf{0}$ 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 $\mathbf{AK} = \lambda\mathbf{K}$. 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 $\lambda*$ is also an eigenvalue. If $\mathbf{K}$ is an eigenvector corresponding to λ then its conjugate $\mathbf{K}*$ is an eigenvector corresponding to $\lambda*$.

- $\lambda = 0$ 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 $\lambda_1, \lambda_2, \ldots, \lambda_n$ 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

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

  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 $|\lambda_1| > |\lambda_2| > |\lambda_3| \geq \cdots \geq |\lambda_n|$.

  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 $0, \lambda_2, \lambda_3, \cdots, \lambda_n$
  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 $\mathbf{P}^{-1}\mathbf{AP} = \mathbf{D}$ is a diagonal matrix, then we say that the $n \times n$ matrix $\mathbf{A}$ can be **diagonalised**, or is **diagonalisable**, and that **P diagonalises A**.

- An $n \times n$ matrix $\mathbf{A}$ is diagonalisable iff $\mathbf{A}$ has n linearly independent eigenvectors $\mathbf{K}_1, \mathbf{K}_2, \ldots, \mathbf{K}_n$. If we let $\mathbf{P} = \begin{pmatrix} \mathbf{K}_1 & \mathbf{K}_2 & \cdots & \mathbf{K}_n \end{pmatrix}$ then

$$\begin{aligned}
\mathbf{AP} &= \begin{pmatrix} \mathbf{AK_1} & \mathbf{AK_2} & \cdots & \mathbf{AK_n} \end{pmatrix} \\
&= \begin{pmatrix} \lambda_1 \mathbf{K}_1 & \lambda_2 \mathbf{K}_2 & \cdots & \lambda_n \mathbf{K}_n \end{pmatrix} \\
&= \begin{pmatrix} \mathbf{K}_1 & \mathbf{K}_2 & \cdots & \mathbf{K}_n \end{pmatrix} \begin{pmatrix} \lambda_1 & 0 & \cdots & 0 \\ 0 & \lambda_2 & \cdots & 0 \\ \vdots & & \ddots & \vdots \\ 0 & 0 & \cdots & \lambda_n \end{pmatrix} \\
&= \mathbf{PD}
\end{aligned}$$

- 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 $\mathbf{A} = \mathbf{LU}$ where $\mathbf{L}$ and $\mathbf{U}$ are lower and upper triangular matrices, respectively, then we say that $\mathbf{A} = \mathbf{LU}$ 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. $l_{ii} = 1, i = 1, 2, \ldots, n$
  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 $\mathbf{AX} = \mathbf{B}$, if $\mathbf{A}$ has an LU-factorisation the system can be solved as follows:

  1. Rewrite the system $\mathbf{LUX} = B$
  2. Let $\mathbf{UX} = \mathbf{Y}$ where
  $$\mathbf{Y} = \begin{pmatrix} y_1 \\ y_2 \\ \vdots \\ y_n \end{pmatrix}$$
  3. Solve $\mathbf{LY} = \mathbf{B}$ via forward substitution, i.e. find y_1, use that to find y_2, etc.
  4. Substitute the values of y_n into $\mathbf{UX} = \mathbf{Y}$ 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 $\mathbf{A} = \mathbf{LU}$ then the determinant of $\mathbf{A}$ can be calculated as $\det \mathbf{A} = \det \mathbf{L} \cdot \det \mathbf{L}$ 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

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

- The definite integral of a vector function is defined as

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

- 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

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

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 $\left|\left|\frac{d\mathbf{T}}{dt}\right|\right| \neq 0$ 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

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

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 $f(x_1, x_2, \ldots, x_n)$ 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.

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

- When multiple derivates 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

$$\frac{\partial z}{\partial x} = \frac{\partial z}{\partial u}\frac{\partial u}{\partial x} + \frac{\partial z}{\partial v}\frac{\partial v}{\partial x}, \qquad \frac{\partial z}{\partial y} = \frac{\partial z}{\partial u}\frac{\partial u}{\partial y} + \frac{\partial z}{\partial v}\frac{\partial v}{\partial y}. \tag{5}$$

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 $f(x_1, x_2, \ldots, x_n)$ in the direction of the unit vector $\mathbf{u}$ is given by

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

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

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