# Numerical Linear Algebra

# 1 Matrices, Vectors, and Norms

# 1.1 Matrix-Vector Multiplication

Given a matrix  $A \in \mathbb{C}^{m \times n}$  and a vector  $x \in \mathbb{C}^n$ , the matrix-vector product  $Ax = b \in \mathbb{C}^m$  is defined as:

$$b_i = \sum_{j=1}^n a_{ij} x_j$$
, for  $i = 1, \dots, m$ 

where  $a_{ij}$  are the entries of the matrix A.

### Observation

The transformation  $x \mapsto Ax$  is a linear transformation from  $\mathbb{C}^n$  to  $\mathbb{C}^m$ , i.e., it satisfies:

$$A(x+y) = Ax + Ay$$
,  $A(\alpha x) = \alpha Ax$ , for all  $x, y \in \mathbb{C}^n$ ,  $\alpha \in \mathbb{C}$ 

### 1.1.1 A Matrix times a Vector

$$b = Ax = \sum_{j=1}^{n} x_j a_j$$

where  $a_j$  is the j-th column of A.

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

## **Example: Vandermonde Matrix**

A Vandermonde matrix is defined as:

$$A = \begin{bmatrix} 1 & x_1 & x_1^2 & \cdots & x_1^{n-1} \\ 1 & x_2 & x_2^2 & \cdots & x_2^{n-1} \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 1 & x_m & x_m^2 & \cdots & x_m^{n-1} \end{bmatrix} \in \mathbb{C}^{m \times n}$$

### Observation

Let fix the sequence  $\{x_1, x_2, \dots, x_m\}$ . If p and q are polynomials of degree at most n-1 and  $\alpha$  is a scalar, then:

- 1. (p+q) is a polynomial of degree at most n-1, and so are  $\alpha p, \alpha q$ .
- 2.  $(p+q)(x_i) = p(x_i) + q(x_i)$  for i = 1, ..., m.
- 3.  $(\alpha p)(x_i) = \alpha p(x_i)$  for  $i = 1, \dots, m$ .

### Observation

Suppose we have a vector  $c \in \mathbb{C}^n$  representing the coefficients of a polynomial  $p(x) = c_0 + c_1 x + c_2 x^2 + \cdots + c_{n-1} x^{n-1}$ . Then

$$p(x_i) = (Ac)_i = c_0 + c_1 x_i + c_2 x_i^2 + \dots + c_{n-1} x_i^{n-1}$$

Any polynomial of degree at most n-1 can be represented as

$$p(x) = \begin{bmatrix} 1 & x & x^2 & \cdots & x^{n-1} \end{bmatrix} \begin{bmatrix} c_0 \\ c_1 \\ c_2 \\ \vdots \\ c_{n-1} \end{bmatrix}$$

# 1.2 Matrix-Matrix Multiplication

Given matrices  $A \in \mathbb{C}^{l \times m}$  and  $C \in \mathbb{C}^{m \times n}$ , the matrix-matrix product  $B = AC \in \mathbb{C}^{l \times n}$  is defined as:

$$b_{ij} = \sum_{k=1}^{m} a_{ik} c_{kj}$$
, for  $i = 1, \dots, l, j = 1, \dots, n$ 

So,

$$B = \begin{bmatrix} b_1 & b_2 & \cdots & b_n \end{bmatrix} = \begin{bmatrix} a_1 & a_2 & \cdots & a_m \end{bmatrix} \begin{bmatrix} c_1 & c_2 & \cdots & c_n \end{bmatrix}$$
$$b_j = Ac_j = \sum_{k=1}^m c_{kj} a_k, \quad j = 1, \dots, n$$

# **Example: Outer Product**

Given two vectors  $u \in \mathbb{C}^{m \times 1}$  and  $v \in \mathbb{C}^{1 \times n}$ , the outer product  $uv^T \in \mathbb{C}^{m \times n}$  is defined as:

$$\begin{bmatrix} u_1 \\ u_2 \\ \vdots \\ u_m \end{bmatrix} \begin{bmatrix} v_1 & v_2 & \cdots & v_n \end{bmatrix} = \begin{bmatrix} u_1v_1 & u_1v_2 & \cdots & u_1v_n \\ u_2v_1 & u_2v_2 & \cdots & u_2v_n \\ \vdots & \vdots & \ddots & \vdots \\ u_mv_1 & u_mv_2 & \cdots & u_mv_n \end{bmatrix}$$

# Example

Let  $B \in \mathbb{C}^{m \times n}$ , let  $a_1, a_2, \dots, a_n$  be the columns of  $A \in \mathbb{C}^{m \times n}$  and let  $R \in \mathbb{C}^{n \times n}$  be an upper triangular matrix with all its superdiagonal entries equal to 1 such that

$$B = \begin{bmatrix} b_1 & b_2 & \dots & b_n \end{bmatrix} = \begin{bmatrix} a_1 & a_2 & \dots & a_n \end{bmatrix} \begin{bmatrix} 1 & 1 & \dots & 1 \\ 0 & 1 & \dots & 1 \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \dots & 1 \end{bmatrix}$$

Then,

$$b_j = Ar_j = \sum_{k=1}^{j} a_k, \quad j = 1, \dots, n$$

which is known as the indefinite integral operation.

# 1.3 Range and Null Space

### Theorem

The range of a matrix  $A \in \mathbb{C}^{m \times n}$  is the space spanned by its columns, i.e. it's column space.

"The set of vectors that can be written as Ax for some  $x \in \mathbb{C}^n$ ."

# 1.3.1 Null Space and Rank

The null space of a matrix  $A \in \mathbb{C}^{m \times n}$  is defined as:

$$\operatorname{span}\{x \in \mathbb{C}^n : Ax = 0\}$$

The rank of a matrix  $A \in \mathbb{C}^{m \times n}$  is defined as the dimension of its range:

$$rank(A) = dim(range(A))$$

### Theorem

For any matrix  $A \in \mathbb{C}^{m \times n}$ ,  $m \ge n$ , A has full rank if and only if it maps no two distinct vectors to the same vector:

$$Ax = Ay \implies x = y, \forall x, y \in \mathbb{C}^n$$
 and  $Ax \neq Ay \implies x \neq y, \forall x, y \in \mathbb{C}^n$ 

### 1.3.2 Non-singular Matrices

A non-singular matrix, or invertible matrix is a square matrix  $A \in \mathbb{C}^{n \times n}$  that has full rank.

### 1.3.3 Inverse of a Matrix

The matrix  $A \in \mathbb{C}^{n \times n}$  is the inverse of Z if and only if:

$$AZ = ZA = I = \begin{bmatrix} e_1 & e_2 & \cdots & e_n \end{bmatrix}$$

where  $e_i$  is the *i*-th standard basis vector. Furthermore, the inverse of a matrix is unique and

$$rank(A) = n = rank(Z)$$

### Theorem

For  $A \in \mathbb{C}^{n \times n}$ , the following statements are equivalent:

- 1. A has an inverse  $A^{-1}$ .
- 2. A has full rank, i.e. rank(A) = n.
- 3. The columns of A span  $\mathbb{C}^n$ , i.e. range $(A) = \mathbb{C}^n$ .
- 4. The columns of A are linearly independent, i.e.  $null(A) = \{0\}$ .
- 5. 0 is not an eigenvalue of A.
- 6. 0 is not a singular value of A.
- 7.  $|A| \neq 0$ .

# 1.4 Orthogonal Vectors and Matrices

## 1.4.1 Complex Conjugate and Conjugate Transpose

Given  $z \in \mathbb{C}$ , the complex conjugate of z is denoted by  $\bar{z}$  or  $z^*$ , where if z = x + iy, then  $\bar{z} = x - iy$ .

The conjugate transpose of a matrix  $A \in \mathbb{C}^{m \times n}$  is denoted by  $A^*$ , where  $A^* = \bar{A}^T$ .

$$A = \begin{bmatrix} a_{11} & a_{12} \\ a_{21} & a_{22} \\ a_{31} & a_{32} \end{bmatrix} \implies A^* = \begin{bmatrix} \bar{a}_{11} & \bar{a}_{21} & \bar{a}_{31} \\ \bar{a}_{12} & \bar{a}_{22} & \bar{a}_{32} \end{bmatrix}$$

### 1.4.2 Hermitian Matrices and Skew-Hermitian Matrices

A matrix  $A \in \mathbb{C}^{n \times n}$  is called Hermitian if  $A^* = A$ . (Or  $A = A^T$  if  $A \in \mathbb{R}^{n \times n}$ .) A matrix  $A \in \mathbb{C}^{n \times n}$  is called Skew-Hermitian if  $A^* = -A$ . (Or  $A = -A^T$  if  $A \in \mathbb{R}^{n \times n}$ .)

### 1.4.3 Inner Product and Norm

Given two vectors  $x, y \in \mathbb{C}^n$ , the inner product is defined as:

$$\langle x, y \rangle = y^* x = \sum_{i=1}^n \bar{y}_i x_i$$

The norm of a vector  $x \in \mathbb{C}^n$  is defined as:

$$||x|| = \sqrt{\langle x, x \rangle} = \sqrt{\sum_{i=1}^{n} |x_i|^2}$$

The modulus of a complex number  $z \in \mathbb{C}$  is defined as:

$$|z| = \sqrt{z\overline{z}} = \sqrt{x^2 + y^2}$$
 where  $z = x + iy$ 

The angle  $\theta$  between two vectors  $x, y \in \mathbb{C}^n$  is defined as:

$$\cos(\theta) = \frac{\langle x, y \rangle}{\|x\| \|y\|}$$

### Observation

The inner product is bilinear, which means:

1. 
$$\langle x_1 + x_2, y \rangle = \langle x_1, y \rangle + \langle x_2, y \rangle$$

2. 
$$\langle x, y_1 + y_2 \rangle = \langle x, y_1 \rangle + \langle x, y_2 \rangle$$

3. 
$$\langle \alpha x, \beta y \rangle = \alpha \bar{\beta} \langle x, y \rangle$$
 for all  $\alpha, \beta \in \mathbb{C}$ 

For all vectors or matrices A, B of compatible dimensions, we have:

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

### 1.4.4 Orthogonal Vectors

Two vectors  $x, y \in \mathbb{C}^n$  are orthogonal if  $\langle x, y \rangle = 0$ .

### Observation

If x and y are orthogonal, then they are perpendicular.

With  $x = \{x_1, x_2, \dots, x_n\}$  and  $y = \{y_1, y_2, \dots, y_n\}$ , x and y are orthogonal if and only if:

$$\langle x_i, y_i \rangle = 0, \quad \forall i, j = 1, 2, \dots, n$$

### 1.4.5 Orthonormal Vectors

A set of vectors  $\{x_1, x_2, \dots, x_n\} \subset \mathbb{C}^n$  is orthonormal if:

$$\langle x_i, x_j \rangle = 0$$
 for  $i \neq j$  and  $\langle x_i, x_i \rangle = ||x_i||^2 = 1$ 

# Observation

The vectors in an orthogonal set are linearly independent.

## 1.4.6 Orthogonal and Unitary Matrices

A matrix  $Q \in \mathbb{R}^{n \times n}$  is orthogonal if its columns form an orthonormal set, i.e.  $Q^TQ = QQ^T = I$ . A matrix  $U \in \mathbb{C}^{n \times n}$  is unitary if its columns form an orthonormal set, i.e.  $U^*U = UU^* = I$ .

#### Observation

If  $U \in \mathbb{C}^{n \times n}$  is unitary, then  $U^{-1} = U^*$  and

$$\begin{bmatrix} u_1^* \\ u_2^* \\ \vdots \\ u_n^* \end{bmatrix} \begin{bmatrix} u_1 & u_2 & \cdots & u_n \end{bmatrix} = \begin{bmatrix} 1 & 0 & \cdots & 0 \\ 0 & 1 & \cdots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & 1 \end{bmatrix}$$

which implies that  $\langle u_i, u_j \rangle = \delta_{ij}$  where  $\delta_{ij}$  is the Kronecker delta, i.e.  $\delta_{ij} = \begin{cases} 1 & \text{if } i = j \\ 0 & \text{if } i \neq j \end{cases}$ .

# 2 Vector Spaces

A vector space  $\mathcal V$  over a field  $\mathbb F$  (usually  $\mathbb R$  or  $\mathbb C$ ) is a set of objects called vectors, along with two operations: vector addition and scalar multiplication, satisfying the following axioms for all  $u,v,w\in\mathcal V$  and  $\alpha,\beta\in\mathbb F$ :

- 1. (Closure under addition)  $u + v \in \mathcal{V}$
- 2. (Commutativity) u + v = v + u
- 3. (Associativity) (u+v)+w=u+(v+w)
- 4. (Existence of additive identity) There exists an element  $0 \in \mathcal{V}$  such that u + 0 = u
- 5. (Existence of additive inverses) For each  $u \in \mathcal{V}$ , there exists an element  $-u \in \mathcal{V}$  such that u + (-u) = 0
- 6. (Closure under scalar multiplication)  $\alpha u \in \mathcal{V}$
- 7. (Distributivity of scalar multiplication with respect to vector addition)  $\alpha(u+v) = \alpha u + \alpha v$
- 8. (Distributivity of scalar multiplication with respect to field addition)  $(\alpha + \beta)u = \alpha u + \beta u$
- 9. (Associativity of scalar multiplication)  $\alpha(\beta u) = (\alpha \beta)u$
- 10. (Existence of multiplicative identity) 1u = u where 1 is the multiplicative identity in  $\mathbb{F}$

# 2.1 Linear Subspaces

A subset W of a vector space V is a subspace if W is itself a vector space under the operations of addition and scalar multiplication defined on V:

- 1. (Closure under addition)  $u + v \in \mathcal{W}$
- 2. (Closure under scalar multiplication)  $\alpha u \in \mathcal{W}$

# 2.2 Linear Independence

A set of vectors  $\{v_1, v_2, \dots, v_k\} \subset \mathcal{V}$  is linearly independent if the only solution to

$$\alpha_1 v_1 + \alpha_2 v_2 + \dots + \alpha_k v_k = 0$$

is  $\alpha_1 = \alpha_2 = \cdots = \alpha_k = 0$ . If there exists a non-trivial solution, the set is linearly dependent.

### 2.3 Basis and Dimension

A basis of a vector space  $\mathcal{V}$  is a set of linearly independent vectors that spans  $\mathcal{V}$ . The dimension of  $\mathcal{V}$ , denoted dim( $\mathcal{V}$ ), is the number of vectors in any basis of  $\mathcal{V}$ .

# 2.4 Inner Product and Norms

An inner product on a vector space  $\mathcal{V}$  over  $\mathbb{F}$  is a function  $\langle \cdot, \cdot \rangle : \mathcal{V} \times \mathcal{V} \to \mathbb{F}$ . The standard inner product on  $\mathbb{C}^n$  is defined as:

$$\langle x, y \rangle = y^* x = \sum_{i=1}^n \bar{y}_i x_i$$

A norm on a vector space  $\mathcal{V}$  is a function  $\|\cdot\|:\mathcal{V}\to\mathbb{R}$  satisfying:

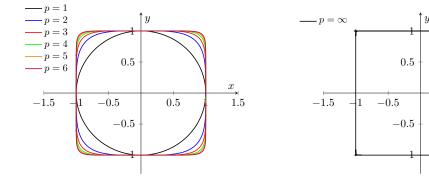
- 1. (Non-negativity)  $||v|| \ge 0$
- 2. (Definiteness) ||v|| = 0 if and only if v = 0
- 3. (Homogeneity)  $\|\alpha v\| = |\alpha| \|v\|$  for all  $\alpha \in \mathbb{F}$  and  $v \in \mathcal{V}$
- 4. (Triangle inequality)  $||u+v|| \le ||u|| + ||v||$  for all  $u, v \in \mathcal{V}$

The *p*-norm on  $\mathbb{C}^n$  is defined as:

$$||x||_p = \left(\sum_{i=1}^n |x_i|^p\right)^{1/p}$$

for  $1 \le p < \infty$ , and the  $\infty$ -norm is defined as:

$$||x||_{\infty} = \max_{1 \le i \le n} |x_i|$$



0.5

1.5

### 2.4.1 Lemma

Let  $\|\cdot\|_{\alpha}$  and  $\|\cdot\|_{\beta}$  be two norms on a finite-dimensional vector space  $\mathcal{V}$ . Then there exist positive constants  $c_1$  and  $c_2$  such that for all  $v \in \mathcal{V}$ :

$$c_1 \|v\|_{\alpha} \le \|v\|_{\beta} \le c_2 \|v\|_{\alpha}$$

 $\|\cdot\|_{\alpha}$  and  $\|\cdot\|_{\beta}$  are said to be equivalent norms. In  $\mathbb{R}^n$ :

$$||x||_{2} \le ||x||_{1} \le \sqrt{n} ||x||_{2}$$
$$||x||_{\infty} \le ||x||_{2} \le \sqrt{n} ||x||_{\infty}$$
$$||x||_{\infty} \le ||x||_{1} \le n ||x||_{\infty}$$

# 3 Matrix Norms

A matrix norm is a function  $\|\cdot\|:\mathbb{C}^{m\times n}\to\mathbb{R}$  satisfying:

- 1. (Non-negativity)  $||A|| \ge 0$
- 2. (Definiteness) ||A|| = 0 if and only if A = 0
- 3. (Homogeneity)  $\|\alpha A\| = |\alpha| \|A\|$  for all  $\alpha \in \mathbb{C}$  and  $A \in \mathbb{C}^{m \times n}$
- 4. (Triangle inequality)  $||A + B|| \le ||A|| + ||B||$  for all  $A, B \in \mathbb{C}^{m \times n}$
- 5. (Sub-multiplicativity)  $||AB|| \le ||A|| ||B||$  for all  $A \in \mathbb{C}^{m \times n}$  and  $B \in \mathbb{C}^{n \times p}$

### 3.1 Frobenius Norm

The Frobenius norm of a matrix  $A \in \mathbb{C}^{m \times n}$  is defined as:

$$||A||_F = \sqrt{\sum_{i=1}^m \sum_{j=1}^n |a_{ij}|^2} = \sqrt{\operatorname{trace}(A^*A)}$$

### 3.2 Induced Norms

The induced norm (or operator norm) of a matrix  $A \in \mathbb{C}^{m \times n}$  is defined as:

$$||A|| = \sup_{x \neq 0} \frac{||Ax||}{||x||} = \sup_{||x|| = 1} ||Ax||$$

# 3.3 Consistent Norms

Let  $\|\cdot\|_{m\times n}$ ,  $\|\cdot\|_{n\times p}$  and  $\|\cdot\|_{m\times p}$  be three matrix norms on  $\mathbb{C}^{m\times n}$ ,  $\mathbb{C}^{n\times p}$  and  $\mathbb{C}^{m\times p}$  respectively. They are said to be consistent if:

$$||AB||_{m \times p} \le ||A||_{m \times n} ||B||_{n \times p}, \quad \forall A \in \mathbb{C}^{m \times n}, B \in \mathbb{C}^{n \times p}$$

# 3.4 Properties of Induced Norms

- 1.  $||Ax|| \le ||A|| ||x||$  for all  $A \in \mathbb{C}^{m \times n}$  and  $x \in \mathbb{C}^n$
- 2.  $||AB|| \le ||A|| ||B||$  for all  $A \in \mathbb{C}^{m \times n}$  and  $B \in \mathbb{C}^{n \times p}$
- 3.  $||A||_{\infty} = \max_{1 \le i \le m} \sum_{j=1}^{n} |a_{ij}|$  (maximum absolute row sum)
- 4.  $||A||_1 = \max_{1 \le j \le n} \sum_{i=1}^m |a_{ij}|$  (maximum absolute column sum)
- 5.  $||A||_2 = \sqrt{\lambda_{\max}(A^*A)}$  (with  $\lambda_{\max}$  the largest eigenvalue)
- 6.  $||A||_2 = ||A^T||_2$
- 7.  $||A||_2 = \max |\lambda_i(A)|$  if A is normal (i.e.  $A^*A = AA^*$ )
- 8.  $||A||_F = \sqrt{\operatorname{trace}(A^*A)}$

So, in general,

$$||A||_{\alpha} = \max_{x \neq 0} \frac{||Ax||_{\alpha}}{||x||_{\alpha}} = \max_{||x||_{\alpha} = 1} ||Ax||_{\alpha}$$

The Frobenius norm is consistent with the euclidean vector norm, but it is not an induced norm.

## 3.5 Equivalence of Matrix Norms

Let  $\|\cdot\|_{\alpha}$  and  $\|\cdot\|_{\beta}$  be two matrix norms on  $\mathbb{C}^{m\times n}$ . Then there exist positive constants  $c_1$  and  $c_2$  such that for all  $A \in \mathbb{C}^{m\times n}$ :

$$c_1 ||A||_{\alpha} \le ||A||_{\beta} \le c_2 ||A||_{\alpha}$$

# 4 Conditioning and Stability

### 4.1 Conditioning

The conditioning is a measure of how the output value of a function changes with respect to small changes in the input value.

$$f: X \to Y$$

With a small perturbation in the input x to  $x + \delta x$ , the output changes from f(x) to  $f(x + \delta x)$ . We define

$$\delta f = f(x + \delta x) - f(x)$$

We want to measure how large  $\delta f$  is relative to f(x) when  $\delta x$  is small relative to x.

### 4.1.1 Absolute Condition Number

The normwise absolute condition number of a function f at a point x is defined as:

$$\hat{\kappa}_f(x) = \lim_{\delta x \to 0} \sup_{\|\delta x\| \le \delta} \frac{\|\delta f\|}{\|\delta x\|} \approx \sup_{\|\delta x\| < \delta} \frac{\|\delta f\|}{\|\delta x\|}$$

If f is differentiable at x, then

$$\hat{\kappa}_f(x) = \lim_{\delta x \to 0} \sup_{\|\delta x\| < \delta} \frac{\|\delta f(x)\|}{\|\delta x\|} = \|J_f(x)\|$$

where  $J_f(x)$  is the Jacobian of f at x.

### 4.1.2 Relative Condition Number

The normwise relative condition number of a function f at a point x is defined as:

$$\kappa_f(x) = \lim_{\delta x \to 0} \sup_{\frac{\|\delta x\|}{\|x\|} \le \delta} \frac{\|\delta f\| / \|f(x)\|}{\|\delta x\| / \|x\|} \xrightarrow{f \text{ differentiable}} \frac{\|J_f(x)\| \|x\|}{\|f(x)\|}$$

### Example

Let  $f: \mathbb{R} \setminus \{0\} \to \mathbb{R}$  be defined as  $f(x) = \frac{1}{x}$ . Then  $f'(x) = -\frac{1}{x^2}$  and

$$\hat{\kappa}_f(x) = |f'(x)| = \frac{1}{|x|^2}, \quad \kappa_f(x) = \frac{|f'(x)||x|}{|f(x)|} = 1$$

As  $x \to 0$ ,  $\hat{\kappa}_f(x) \to \infty$  but  $\kappa_f(x) = 1$ . This means that  $f(x) = \frac{1}{x}$  is well-conditioned for all  $x \neq 0$  in the relative sense, but it is ill-conditioned as  $x \to 0$  in the absolute sense.

## Example

Let  $f: \mathbb{R}^2 \to \mathbb{R}$  be defined as f(x,y) = x - y. Then, f is differentiable and

$$J_f(x,y) = \begin{bmatrix} 1 & -1 \end{bmatrix}$$

So,

$$\hat{\kappa}_f(x,y) = \|J_f(x,y)\|_2 = \sqrt{2}, \quad \|\kappa_f(x,y)\|_1 = \frac{\|J_f(x,y)\|_1 \|(x,y)\|_1}{|f(x,y)|} = \frac{2(|x|+|y|)}{|x-y|}$$

As  $x \to y$ ,  $\kappa_f(x,y) \to \infty$ . This means that f(x,y) = x - y is ill-conditioned when x is close to y.

#### Example

Let  $f: \mathbb{R}^2 \to \mathbb{R}$  be defined as  $f(x,y) = \frac{x}{y}$ . Then

$$\|\kappa_f(x,y)\|_2 = \left|\frac{x}{y}\right| + \left|\frac{y}{x}\right|$$

As  $y \to 0$  or  $x \to 0$ ,  $\kappa_f(x, y) \to \infty$ . This means that  $f(x, y) = \frac{x}{y}$  is ill-conditioned when y or x is close to 0.

### 4.2 Stability

An algorithm is stable if it produces an output that is close to the exact solution of a problem. Formally, an algorithm  $\hat{f}$  for the problem  $f: X \to Y$  is **numerically stable** if for every input  $x \in X$ , there exists  $\hat{x} = x + \delta x \in X$  such that

$$\frac{\|\hat{f}(x) - f(\hat{x})\|}{\|f(x)\|} = O(\epsilon_{\text{mach}}) \quad \text{and} \quad \frac{\|x - \hat{x}\|}{\|x\|} = O(\epsilon_{\text{mach}})$$

where  $\epsilon_{\text{mach}}$  is the machine precision.

An algorithm is **backward stable** if for every input  $x \in X$ , there exists  $\hat{x} = x + \delta x \in X$  such that

$$\hat{f}(x) = f(\hat{x})$$
 and  $\frac{\|x - \hat{x}\|}{\|x\|} = O(\epsilon_{\text{mach}})$ 

# 4.3 Accuracy

An algorithm  $\hat{f}$  is said to be **accurate** if it produces results that are close to the true solution f(x) for all inputs  $x \in X$ . Formally, this means that for every input  $x \in X$ , the following holds:

$$\frac{\|\hat{f}(x) - f(\hat{x})\|}{\|f(x)\|} = O(\epsilon_{\text{mach}})$$

## Observation

A numerically stable algorithm or a backward stable algorithm is accurate if the problem it solves is well-conditioned.

$$\frac{\|\hat{f}(x) - f(\hat{x})\|}{\|f(x)\|} = \frac{\|f(\hat{x}) + \Delta y - f(x)\|}{\|f(x)\|} \le \frac{\|f(x + \Delta x) - f(x)\|}{\|f(x)\|} + \frac{\|\Delta y\|}{\|f(x)\|} \cdot \|y\|$$

$$\le \frac{\|f(x + \Delta x) - f(x)\|/\|f(x)\|}{\|\Delta x\|/\|x\|} \cdot \frac{\|\Delta x\|}{\|x\|} + O(\epsilon_{\text{mach}}) = O(\kappa_f(x)\epsilon_{\text{mach}})$$

where  $\Delta x = \hat{x} - x$  and  $\Delta y = \hat{f}(x) - f(\hat{x})$ .

# 5 Solving Linear Systems

Let

$$A = \begin{bmatrix} 2 & 1 & 1 & 0 \\ 4 & 3 & 3 & 1 \\ 8 & 7 & 9 & 5 \\ 6 & 7 & 9 & 8 \end{bmatrix} = L \cdot U = \begin{bmatrix} 1 \\ 2 & 1 \\ 4 & 3 & 1 \\ 3 & 4 & 1 & 1 \end{bmatrix} \cdot \begin{bmatrix} 2 & 1 & 1 & 0 \\ 1 & 1 & 1 \\ & 2 & 2 \\ & & & 2 \end{bmatrix}$$

where L is a lower triangular matrix and U is an upper triangular matrix. Now, we define  $x_k$  as the k-th column of A:

$$x_{k} = \begin{bmatrix} x_{1k} \\ \vdots \\ x_{kk} \\ \vdots \\ x_{nk} \end{bmatrix} \longrightarrow L_{k} \cdot x_{k} = \begin{bmatrix} x_{1k} \\ \vdots \\ x_{kk} \\ 0 \\ \vdots \\ 0 \end{bmatrix} \quad \text{where} \quad L_{k} = \begin{bmatrix} 1 \\ & \ddots & \\ & & 1 \\ & & -l_{k+1,k} & 1 \\ & & \vdots \\ & & -l_{n,k} & & 1 \end{bmatrix}$$

with  $l_{ik} = \frac{x_{ik}}{x_{kk}}$  for  $k < i \le n$ . Thus, we can write

$$L_k = I - l_k e_k^T$$

where 
$$l_k = \begin{bmatrix} 0 \\ \vdots \\ 0 \\ l_{k+1,k} \\ \vdots \\ l_{n,k} \end{bmatrix}$$
 and  $e_k = \begin{bmatrix} 0 \\ \vdots \\ 1 \\ \vdots \\ 0 \end{bmatrix}$  (1 at the k-th position).

Therefore.

$$e_k^T l_k = 0$$
 and  $(I - l_k e_k^T)(I + l_k e_k^T) = I \implies L_k^{-1} = I + l_k e_k^T$ 

### Example

Let  $A = L \cdot U = (L_1^{-1}L_2^{-1}L_3^{-1}) \cdot U$  where

$$A = \begin{bmatrix} 1 & & & \\ 2 & 1 & & \\ 4 & 3 & 1 & \\ 3 & 4 & 1 & 1 \end{bmatrix} \cdot \begin{bmatrix} 2 & 1 & 1 & 0 \\ & 1 & 1 & 1 \\ & & 2 & 2 \\ & & & 2 \end{bmatrix}$$

Then,

$$L_1 = \begin{bmatrix} 1 & & & \\ -2 & 1 & & \\ -4 & & 1 & \\ -3 & & & 1 \end{bmatrix}, \quad L_2 = \begin{bmatrix} 1 & & & \\ & 1 & & \\ & -3 & 1 & \\ & -4 & & 1 \end{bmatrix}, \quad L_3 = \begin{bmatrix} 1 & & & \\ & 1 & & \\ & & 1 & \\ & & -1 & 1 \end{bmatrix}$$

Consider  $L_k^{-1} \cdot L_{k+1}^{-1} = (I + l_k e_k^T)(I + l_{k+1} e_{k+1}^T) = I + l_k e_k^T + l_{k+1} e_{k+1}^T$  because  $e_k^T l_{k+1} = 0$ . Thus,

$$L_1^{-1} \cdots L_{n-1}^{-1} = L = \begin{bmatrix} 1 & & & \\ l_{21} & 1 & & \\ l_{31} & l_{32} & 1 & & \\ \vdots & \vdots & \vdots & \ddots & \\ l_{n1} & l_{n2} & l_{n3} & \cdots & 1 \end{bmatrix}$$

## 5.1 Forward Elimination

Consider the Forward Elimination algorithm:

Input: 
$$A \in \mathbb{R}^{n \times n}, \bar{b} \in \mathbb{R}^n$$

Output:  $U \in \mathbb{R}^{n \times n}, \bar{b}^* \in \mathbb{R}^n$  such that LA = U and  $L\bar{b} = \bar{b}^*$ 

# **Algorithm 1:** Forward Elimination

Input:  $A \in \mathbb{R}^{n \times n}, \overline{b} \in \mathbb{R}^n$ 

**Output:**  $U \in \mathbb{R}^{n \times n}, \bar{b}^* \in \mathbb{R}^n$  such that LA = U and  $L\bar{b} = \bar{b}^*$ 

1 for k = 1 to n - 1 do

for 
$$i=k+1$$
 to  $n$  do

for  $i=k+1$  to  $n$  do

 $l_{ik}=\frac{a_{ik}}{a_{kk}};$ 

for  $j=k$  to  $n$  do

 $a_{ij}=a_{ij}-l_{ik}a_{kj};$ 
 $b_i=b_i-l_{ik}b_k;$ 

So, the output is

$$L\begin{bmatrix}A&|&\bar{b}\end{bmatrix}=\begin{bmatrix}U&|&\bar{b}^*\end{bmatrix}$$

### 5.2 Backward Substitution

Consider the Backward Substitution algorithm:

$$R\bar{x} = \bar{b}, \quad R = \begin{bmatrix} r_{11} & r_{12} & \cdots & r_{1n} \\ 0 & r_{22} & \cdots & r_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & r_{nn} \end{bmatrix}$$

$$r_{nn}x_n = b_n$$
 and  $r_{n-1,n}x_{n-1} + r_{n-1,n}x_n = b_{n-1}$ 

$$x_n = \frac{b_n}{r_{nn}}$$
 and  $x_i = \frac{b_i - \sum_{j=i+1}^n r_{ij} x_j}{r_{ii}}$ ,  $i = n - 1, n - 2, \dots, 1$ 

# Algorithm 2: Backward Substitution

Input:  $U \in \mathbb{R}^{n \times n}$  (upper triangular),  $\bar{b} \in \mathbb{R}^n$ 

**Output:**  $\bar{x} \in \mathbb{R}^n$  such that  $U\bar{x} = \bar{b}$ 

$$\mathbf{1} \ x_n = \frac{b_n}{r_{nn}};$$

2 for i = n - 1 down to 1 do

$$\mathbf{3} \quad | \quad x_i = b_i;$$

4 | for 
$$j = i + 1 \ to \ n \ do$$

$$\mathbf{6} \quad x_i = \frac{x_i}{r_{ii}};$$

# 5.3 Solving Triangular Systems

With  $D\overline{x} = \overline{b}$ , D diagonal  $(d_{11}, d_{22}, \dots, d_{nn}) \in \mathbb{R}^{n \times n}$ , we can solve for  $\overline{x}$  as follows:

$$x_i = \frac{b_i}{d_{ii}}, \quad i = 1, 2, \dots, n$$

With  $U\overline{x} = \overline{b}$ , L upper triangular, we can solve for  $\overline{x}$  using backward substitution:

$$\begin{bmatrix} u_{11} & u_{12} & \cdots & u_{1n} \\ 0 & u_{22} & \cdots & u_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & u_{nn} \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \\ \vdots \\ x_n \end{bmatrix} = \begin{bmatrix} b_1 \\ b_2 \\ \vdots \\ b_n \end{bmatrix}$$

 $u_{nn}x_n = b_n$  and  $u_{n-1,n}x_{n-1} + u_{n-1,n}x_n = b_{n-1}$ 

$$x_n = \frac{b_n}{u_{nn}}$$
 and  $x_i = \frac{b_i - \sum_{j=i+1}^n u_{ij} x_j}{u_{ii}}$ ,  $i = n - 1, n - 2, \dots, 1$ 

We can define the algorithm for solving  $U\overline{x} = \overline{b}$  where U is a upper triangular matrix as follows:

# Algorithm 3: Solve Upper Triangular System

**Input:**  $U \in \mathbb{R}^{n \times n}$  (upper triangular),  $\bar{b} \in \mathbb{R}^n$ 

**Output:**  $\bar{x} \in \mathbb{R}^n$  such that  $U\bar{x} = \bar{b}$ 

$$\mathbf{1} \ x_n = \frac{b_n}{u_{nn}};$$

2 for i = n - 1 down to 1 do

$$x_i = b_i$$
;

4 | for 
$$j = i + 1 \ to \ n \ do$$

$$\mathbf{6} \quad \boxed{ x_i = \frac{x_i}{u_{ii}};}$$

The cost of this algorithm is as follows:

Flops = 
$$\sum_{i=1}^{n-1} \left( 1 + 2 \sum_{j=i+1}^{n} 1 \right) = \dots = \frac{(n-2)n}{2} \sim \mathcal{O}(n^2)$$

### 5.4 Gaussian Elimination

The Gaussian Elimination algorithm can be defined as follows:

# Algorithm 4: Gaussian Elimination

```
Input: A \in \mathbb{R}^{n \times n}, \overline{b} \in \mathbb{R}^n

Output: U \in \mathbb{R}^{n \times n}, \overline{b}^* \in \mathbb{R}^n such that LA = U and L\overline{b} = \overline{b}^*

1 for k = 1 to n - 1 do

2 | for i = k + 1 to n do

3 | t = \frac{a_{ik}}{a_{kk}}; // t is a factor

for j = k to n do

5 | a_{ij} = a_{ij} - ta_{kj};

6 | a_{ij} = a_{ij} - ta_{kj};
```

The cost of this algorithm is as follows:

Flops = 
$$\mathcal{O}(n^3)$$

### 5.5 LU Factorization

With  $A \in \mathbb{R}^{n \times n}$  non-singular, we can factor A as  $A = L \cdot U$  where L is a lower triangular matrix with unit diagonal and U is an upper triangular matrix:

$$L = I + \sum_{k=1}^{n-1} l_k e_k^T$$

#### Observation

If an  $n \times n$  matrix A has an LU factorization, then it is unique. Furthermore, if we relax the condition that L has a unit diagonal (i.e., normalized), then there are infinitely many LU factorizations of A.

### Theorem

If all the leading principal minors of A are non-zero, i.e.,  $\det(A_k) \neq 0$  for k = 1, 2, ..., n-1 where  $A_k$  is the  $k \times k$  leading principal submatrix of A, then A has an LU factorization. In particular, if A is strictly diagonally dominant or symmetric positive definite, then A has an LU factorization.

*Proof.* We will prove this by induction on n. The base case n=1 is trivial since any non-zero scalar can be factored as  $1 \cdot a_{11}$ . A matrix with k row operations already done can be written as

$$A^{(k)} = \begin{bmatrix} a_{11} & a_{12} & \cdots & \cdots & a_{1n} \\ 0 & a_{22}^{(2)} & \cdots & \cdots & a_{2n}^{(2)} \\ 0 & 0 & a_{kk}^{(k)} & \cdots & a_{kn}^{(k)} \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & a_{n2}^{(k)} & \cdots & a_{nn}^{(k)} \end{bmatrix}$$

where every superindex (k) indicates that k row operations have been performed. Note that the leading principal submatrix of order k of  $A^{(k)}$  is the same as that of A. Thus,  $\det(A^{(k)}) = \det(A_k) \neq 0$ . Hence, the next step is possible.

### Theorem

If an invertible matrix  $A \in \mathbb{R}^{n \times n}$  has an LU factorization, then it is unique.

*Proof.* Suppose  $A = L_1U_1 = L_2U_2$  where  $L_1, L_2$  are lower triangular with unit diagonal and  $U_1, U_2$  are upper triangular. Then,

$$L_2^{-1}L_1 = U_2U_1^{-1}$$

The left-hand side is lower triangular with unit diagonal, and the right-hand side is upper triangular. Thus, both sides must be equal to the identity matrix. Therefore,  $L_1 = L_2$  and  $U_1 = U_2$ .

# 5.6 Pivoting

We use pivoting to avoid division by zero or small numbers during the elimination process. There are three types of pivoting:

1. Partial row pivoting: We interchange rows to ensure that the pivot element is the largest in its column.

$$|a_{lk}| = \max_{k \le i \le n} |a_{ik}^{(k)}|$$

2. Partial column pivoting: We interchange columns to ensure that the pivot element is the largest in its row.

$$|a_{kl}| = \max_{k \le j \le n} |a_{kj}^{(k)}|$$

3. Total pivoting: We interchange both rows and columns to ensure that the pivot element is the largest in the remaining submatrix.

$$|a_{lm}| = \max_{k \le i, j \le n} |a_{ij}^{(k)}|$$

Total pivoting is the most stable but also the most expensive. It yields the complete LU factorization. The cost of partial pivoting is  $\mathcal{O}(n^2)$ , while the cost of total pivoting is  $\mathcal{O}(n^3)$ .

The LU factorization with partial pivoting algorithm can be defined as follows:

## Algorithm 5: LU Factorization with Partial Pivoting

```
Input: A \in \mathbb{R}^{n \times n}, \overline{b \in \mathbb{R}^n}
  Output: P, L, U such that PA = LU
1 for k = 1 to n - 1 do
      l = \arg\max_{k \le i \le n} |a_{ik}^{(k)}|;
       Swap rows k and l of A and b;
3
       for i = k + 1 to n do
4
           t = \frac{a_{ik}}{a_{kk}};
5
           for j = k to n do
6
          7
          b_i = b_i - tb_k;
8
```

# 5.7 The role of L and U in Backward Stability

Let  $A \in \mathbb{R}^{n \times n}$  be a non-singular matrix with LU factorization A = LU. Then,

$$\hat{L}\hat{U} = A + \delta A, \quad \frac{\|\delta A\|}{\|L\|\|U\|} = \mathcal{O}(\varepsilon_{\text{mach}})$$

where  $\hat{L}$  and  $\hat{U}$  are the computed factors of A, and  $\delta A$  is the perturbation in A due to rounding errors.

Now, let us introduce partial pivoting. Let P be a permutation matrix such that PA = LU. Then,

$$L = \begin{bmatrix} \ddots & & & & \\ l_{ik} & \ddots & & & \\ \vdots & \vdots & \ddots & & \\ l_{nk} & l_{n,k+1} & \cdots & \ddots & \\ l_{n1} & l_{n2} & \cdots & l_{n,n-1} & 1 \end{bmatrix} \implies ||L|| = \mathcal{O}(1)$$

Thus, the stability of the LU factorization with partial pivoting depends on ||U|| relative to ||A||. This ratio is known as the growth factor:

$$\rho = \frac{\max |u_{ij}|}{\max |a_{ij}|}$$

With PA = LU, we have

$$\hat{L}\hat{U} = PA + \delta A, \quad \frac{\|\delta A\|}{\|P\|\|A\|} = \mathcal{O}(\varepsilon_{\text{mach}}) \implies \frac{\|\delta A\|}{\|A\|} = \mathcal{O}(\rho \varepsilon_{\text{mach}})$$

### Example

Let

Then,  $\rho = 16 = 2^{n-1}$ , which is not bounded. Then,

$$\frac{\|\delta A\|}{\|A\|} = \mathcal{O}(2^{n-1}\varepsilon_{\text{mach}})$$

# 6 Least Squares

To compute the QR factorization of a matrix  $A \in \mathbb{R}^{m \times n}$  with  $m \geq n$ , we can use different methods such as Gram-Schmidt, given rotations, or Householder reflections. The QR factorization decomposes A into an orthogonal matrix  $Q \in \mathbb{R}^{m \times m}$  and an upper triangular matrix  $R \in \mathbb{R}^{m \times n}$ :

$$A = QR$$

where

$$Q = \begin{bmatrix} q_1 & q_2 & \cdots & q_m \end{bmatrix}, \quad R = \begin{bmatrix} r_{11} & r_{12} & \cdots & r_{1n} \\ 0 & r_{22} & \cdots & r_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & r_{nn} \\ 0 & 0 & \cdots & 0 \\ \vdots & \vdots & & \vdots \\ 0 & 0 & \cdots & 0 \end{bmatrix}$$

## 6.1 Gram-Schmidt Process

Given a matrix  $A \in \mathbb{R}^{m \times n}$  with linearly independent columns, the Gram-Schmidt process constructs an orthonormal basis for the column space of A.

$$A = [a_1 \ a_2 \ \cdots \ a_n], \ \operatorname{span}\{a_1, a_2, \dots, a_n\} = \operatorname{span}\{q_1, q_2, \dots, q_n\}$$

The Gram-Schmidt process can be defined as follows:

$$v_{1} = a_{1} \qquad \longrightarrow \qquad q_{1} = \frac{v_{1}}{\|v_{1}\|}$$

$$v_{2} = a_{2} - (a_{2}^{T} \cdot q_{1}) \cdot q_{1} \qquad \longrightarrow \qquad q_{2} = \frac{v_{2}}{\|v_{2}\|}$$

$$v_{3} = a_{3} - (a_{3}^{T} \cdot q_{2}) \cdot q_{2} - (a_{3}^{T} \cdot q_{1}) \cdot q_{1} \qquad \longrightarrow \qquad q_{3} = \frac{v_{3}}{\|v_{3}\|}$$

$$\vdots$$

$$v_{n} = a_{n} - \sum_{i=1}^{n-1} (a_{n}^{T} \cdot q_{j}) \cdot q_{j} \qquad \longrightarrow \qquad q_{n} = \frac{v_{n}}{\|v_{n}\|}$$

The coefficients  $q_k$  satisfy the following:

$$q_i \cdot q_j = \delta_{ij} = \begin{cases} 1 & i = j \\ 0 & i \neq j \end{cases}$$

The matrix R can be constructed as follows:

$$R = \begin{bmatrix} r_{11} & r_{12} & \cdots & r_{1n} \\ 0 & r_{22} & \cdots & r_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & r_{nn} \end{bmatrix} = \begin{bmatrix} \|v_1\| & q_1^T a_2 & \cdots & q_1^T a_n \\ 0 & \|v_2\| & \cdots & q_2^T a_n \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \cdots & \|v_n\| \end{bmatrix}$$

where

$$r_{ij} = \begin{cases} q_i^T a_j & i < j \\ \|v_i\| & i = j \\ 0 & i > j \end{cases}$$

With this, we can express  $a_k$  as follows:

$$a_{1} = q_{1}r_{11} = q_{1}||v_{1}||$$

$$a_{2} = q_{1}r_{12} + q_{2}r_{22} = q_{1}(q_{1}^{T}a_{2}) + q_{2}|v_{2}||$$

$$a_{3} = q_{1}r_{13} + q_{2}r_{23} + q_{3}r_{33} = q_{1}(q_{1}^{T}a_{3}) + q_{2}(q_{2}^{T}a_{3}) + q_{3}||v_{3}||$$

$$\vdots$$

$$a_{n} = q_{1}r_{1n} + q_{2}r_{2n} + \dots + q_{n}r_{nn} = q_{1}(q_{1}^{T}a_{n}) + q_{2}(q_{2}^{T}a_{n}) + \dots + q_{n}||v_{n}||$$

The Gram-Schmidt algorithm can be summarized as follows:

# Algorithm 6: Gram-Schmidt Process

**Input:**  $A \in \mathbb{R}^{m \times n}$  with linearly independent columns

**Output:**  $Q \in \mathbb{R}^{m \times n}$  with orthonormal columns,  $R \in \mathbb{R}^{n \times n}$  upper triangular

1 for k = 1 to n do  $v_k = a_k;$ 3 for j = 1 to k - 1 do  $r_{jk} = q_j^T a_k;$  $v_k = v_k - r_{jk}q_j;$  $r_{kk} = ||v_k||;$  $q_k = \frac{v_k}{r_{kk}};$ 

This algorithm can be modified to improve numerical stability, resulting in the Modified Gram-Schmidt algorithm:

## Algorithm 7: Modified Gram-Schmidt Process

**Input:**  $A \in \mathbb{R}^{m \times n}$  with linearly independent columns

**Output:**  $Q \in \mathbb{R}^{m \times n}$  with orthonormal columns,  $R \in \mathbb{R}^{n \times n}$  upper triangular

1 for k = 1 to n do

 $\begin{array}{c|cccc} \mathbf{2} & & v_k = a_k; \\ \mathbf{3} & & r_{kk} = \|v_k\|; \\ \mathbf{4} & & q_k = \frac{v_k}{r_{kk}}; \\ \mathbf{5} & & \mathbf{for} \ j = k+1 \ \boldsymbol{to} \ n \ \mathbf{do} \\ \mathbf{6} & & & r_{kj} = q_k^T a_j; \\ \mathbf{7} & & & a_j = a_j - r_{kj} q_k; \end{array}$ 

### Remark

The matrix R is invertible if and only if the columns of A are linearly independent. The reduced QR factorization is unique up to a sign.

### 6.2 Householder Transformations

A Householder transformation is a linear transformation that reflects a vector about a plane or hyperplane. It is defined as follows:

Given a vector  $v \in \mathbb{R}^n$ , the Householder transformation H is defined as:

$$H = I - 2\frac{uu^T}{\|u\|}$$

where  $u = v - \alpha e_1$ ,  $\alpha = ||v||$ , and  $e_1$  is the first standard basis vector in  $\mathbb{R}^n$ .