# CS 450: Numerical Anlaysis ${ }^{1}$ <br> Linear Systems 

University of Illinois at Urbana-Champaign

[^0]
## Vector Norms

- Properties of vector norms

$$
\begin{aligned}
\|\boldsymbol{x}\| & \geq 0 \\
\|\boldsymbol{x}\| & =0 \quad \Leftrightarrow \quad \boldsymbol{x}=\mathbf{0} \\
\|\alpha \boldsymbol{x}\| & =|\alpha| \cdot\|\boldsymbol{x}\| \\
\|\boldsymbol{x}+\boldsymbol{y}\| & \leq\|\boldsymbol{x}\|+\|\boldsymbol{y}\| \quad \text { (triangle inequality) implies continuity }
\end{aligned}
$$

- A norm is uniquely defined by its unit sphere: Surface defined by space of vectors $\mathbb{V} \subset \mathbb{R}^{n}$ such that $\forall \boldsymbol{x} \in \mathbb{V},\|\boldsymbol{x}\|=1$
- $p$-norms $\|\boldsymbol{x}\|_{p}=\left(\sum_{i}\left|x_{i}\right|^{p}\right)^{1 / p}$
- $p=1$ gives sum of absolute values of entry (unit sphere is diamond-like)
- $p=\infty$ gives maximum entry in absolute value (unit sphere is box-like)
- $p=2$ gives Euclidean distance metric (unit sphere is spherical)


## Inner-Product Spaces

- Properties of inner-product spaces: Inner products $\langle\boldsymbol{x}, \boldsymbol{y}\rangle$ must satisfy

$$
\begin{aligned}
\langle\boldsymbol{x}, \boldsymbol{x}\rangle & \geq 0 \\
\langle\boldsymbol{x}, \boldsymbol{x}\rangle & =0 \quad \Leftrightarrow \quad \boldsymbol{x}=\mathbf{0} \\
\langle\boldsymbol{x}, \boldsymbol{y}\rangle & =\langle\boldsymbol{y}, \boldsymbol{x}\rangle \\
\langle\boldsymbol{x}, \boldsymbol{y}+\boldsymbol{z}\rangle & =\langle\boldsymbol{x}, \boldsymbol{y}\rangle+\langle\boldsymbol{x}, \boldsymbol{z}\rangle \\
\langle\alpha \boldsymbol{x}, \boldsymbol{y}\rangle & =\alpha\langle\boldsymbol{x}, \boldsymbol{y}\rangle
\end{aligned}
$$

- Inner-product-based vector norms and Cauchy-Schwartz The $p=2$ vector norm is the Eucledian inner-product norm,

$$
\|\boldsymbol{x}\|_{2}=\sqrt{\boldsymbol{x}^{T} \boldsymbol{x}}
$$

and due to Cauchy-Schwartz inequality $|\langle\boldsymbol{x}, \boldsymbol{y}\rangle| \leq \sqrt{\langle\boldsymbol{x}, \boldsymbol{x}\rangle \cdot\langle\boldsymbol{y}, \boldsymbol{y}\rangle}$,

$$
\left|\boldsymbol{x}^{T} \boldsymbol{y}\right| \leq\|\boldsymbol{x}\|_{2}\|\boldsymbol{y}\|_{2} .
$$

Other inner-products can be expressed as $\langle\boldsymbol{x}, \boldsymbol{y}\rangle=\boldsymbol{x}^{T} \boldsymbol{A} \boldsymbol{y}$ where $\boldsymbol{A}$ is symmetric positive definite, yielding norms $\|\boldsymbol{x}\|_{\boldsymbol{A}}=\sqrt{\boldsymbol{x}^{T} \boldsymbol{A x}}$

## Matrix Norms

- Properties of matrix norms:

$$
\begin{aligned}
\|\boldsymbol{A}\| & \geq 0 \\
\|\boldsymbol{A}\| & =0 \quad \Leftrightarrow \quad \boldsymbol{A}=\mathbf{0} \\
\|\alpha \boldsymbol{A}\| & =|\alpha| \cdot\|\boldsymbol{A}\| \\
\|\boldsymbol{A}+\boldsymbol{B}\| & \leq\|\boldsymbol{A}\|+\|\boldsymbol{B}\| \quad \text { (triangle inequality) }
\end{aligned}
$$

- Frobenius norm:

$$
\|\mathbf{A}\|_{F}=\left(\sum_{i, j} a_{i j}^{2}\right)^{1 / 2}
$$

- Operator/induced/subordinate matrix norms:

For any vector norm $\|\cdot\|$, the induced matrix norm is

$$
\|\boldsymbol{A}\|=\max _{\boldsymbol{x} \neq \mathbf{0}}\|\boldsymbol{A} \boldsymbol{x}\| /\|\boldsymbol{x}\|=\max _{\|\boldsymbol{x}\|=1}\|\boldsymbol{A} \boldsymbol{x}\|
$$

## Induced Matrix Norms

- Interpreting induced matrix norms (amplification and reduction): A matrix is uniquely defined with respect to a norm by a unit-ball, which is the space of vectors $\boldsymbol{y}=\boldsymbol{A x}$ for all $\boldsymbol{x}$ on the unit-sphere of the norm.

$$
\|\boldsymbol{A}\|_{p}=\max _{\|\boldsymbol{x}\|_{p}=1}\|\boldsymbol{A} \boldsymbol{x}\|_{p}
$$

is the maximum possible p-norm amplification due to application of $\boldsymbol{A}$

$$
1 /\left\|\boldsymbol{A}^{-1}\right\|_{p}=\min _{\|\boldsymbol{x}\|_{p}=1}\|\boldsymbol{A} \boldsymbol{x}\|_{p}
$$

is the maximum possible p-norm reduction due to application of $\boldsymbol{A}$

## Matrix Condition Number

- Matrix condition number definition: $\kappa(\boldsymbol{A})=\|\boldsymbol{A}\| \cdot\left\|\boldsymbol{A}^{-1}\right\|$ is the ratio of the maximum $\boldsymbol{A}$ can amplify a vector and the minimum to which it can reduce the norm when applied to a unit-norm vector.
- Derivation from perturbations:

$$
\kappa(\boldsymbol{A})=\max _{\text {inputs }} \max _{\text {perturbations in input }}\left|\frac{\text { relative perturbation in output }}{\text { relative perturbation in input }}\right|
$$

since a matrix is a linear operator, we can decouple its action on the input $\boldsymbol{x}$ and the perturbation $\boldsymbol{\delta} \boldsymbol{x}$ since $\boldsymbol{A}(\boldsymbol{x}+\boldsymbol{\delta} \boldsymbol{x})=\boldsymbol{A x}+\boldsymbol{A} \boldsymbol{\delta} \boldsymbol{x}$, so


## Matrix Conditioning

- The matrix condition number $\kappa(\boldsymbol{A})$ is the ratio between the max and min distance from the surface to the center of the unit ball transformed by $\kappa(\boldsymbol{A})$ :
- The max distance to center is given by the vector maximizing $\max _{\|x\|=1}\|\boldsymbol{A x}\|_{2}$.
- The min distance to center is given by the vector minimizing $\min _{\|\boldsymbol{x}\|=1}\|\boldsymbol{A} \boldsymbol{x}\|_{2}=1 /\left(\max _{\|\boldsymbol{x}\|=1}\left\|\boldsymbol{A}^{-1} \boldsymbol{x}\right\|_{2}\right)$.
- Thus, we have that $\kappa(\boldsymbol{A})=\|\boldsymbol{A}\|_{2}\left\|\boldsymbol{A}^{-1}\right\|_{2}$
- The matrix condition number bounds the worst-case amplification of error in a matrix-vector product: Consider $\boldsymbol{y}+\boldsymbol{\delta} \boldsymbol{y}=\boldsymbol{A}(\boldsymbol{x}+\boldsymbol{\delta} \boldsymbol{x})$, assume $\|\boldsymbol{x}\|_{2}=1$
- In the worst case, $\|\boldsymbol{y}\|_{2}$ is minimized, that is $\|\boldsymbol{y}\|_{2}=1 /\left\|\boldsymbol{A}^{-1}\right\|_{2}$
- In the worst case, $\|\boldsymbol{\delta} \boldsymbol{y}\|_{2}$ is maximized, that is $\|\boldsymbol{\delta} \boldsymbol{y}\|_{2}=\|\boldsymbol{A}\|_{2}\|\boldsymbol{\delta} \boldsymbol{y}\|_{2}$
- So $\|\boldsymbol{\delta} \boldsymbol{y}\|_{2} /\|\boldsymbol{y}\|_{2}$ is at most $\kappa(\boldsymbol{A})\|\boldsymbol{\delta} \boldsymbol{x}\|_{2} /\|\boldsymbol{x}\|_{2}$


## Norms and Conditioning of Orthogonal Matrices

- Orthogonal matrices: A matrix $Q$ is orthogonal, if its square and its columns are orthonormal, or equivalently $Q^{T}=Q^{-1}$.
- Norm and condition number of orthogonal matrices: For any $\|\boldsymbol{v}\|_{2}=1$,

$$
\begin{aligned}
\|\boldsymbol{Q} \boldsymbol{v}\|_{2} & =\left(\left\langle\boldsymbol{v}^{T} \boldsymbol{Q}^{T}, \boldsymbol{Q} \boldsymbol{v}\right\rangle\right)^{1 / 2}=\left(\boldsymbol{v}^{T} \boldsymbol{Q}^{T} \boldsymbol{Q} \boldsymbol{v}\right)^{1 / 2}=\left(\boldsymbol{v}^{T} \boldsymbol{v}\right)^{1 / 2} \\
& =\|\boldsymbol{v}\|_{2}
\end{aligned}
$$

Consequently, $\|\boldsymbol{Q}\|_{2}=\left\|\boldsymbol{Q}^{-1}\right\|_{2}=\kappa(\boldsymbol{Q})=1$.
$\boldsymbol{Q} \boldsymbol{v}$ expresses $\boldsymbol{v}$ in a coordinate system whose axes are columns of $\boldsymbol{Q}^{T}$

## Singular Value Decomposition

- The singular value decomposition (SVD):

We can express any matrix $\boldsymbol{A}$ as

$$
\boldsymbol{A}=\boldsymbol{U} \boldsymbol{\Sigma} \boldsymbol{V}^{T}
$$

where $\boldsymbol{U}$ and $\boldsymbol{V}$ are orthogonal, and $\boldsymbol{\Sigma}$ is square nonnegative and diagonal,

$$
\boldsymbol{\Sigma}=\left[\begin{array}{lll}
\sigma_{\max } & & \\
& \ddots & \\
& & \sigma_{\min }
\end{array}\right]
$$

Any matrix is diagonal when expressed as an operator mapping vectors from a coordinate system given by $V$ to a coordinate system given by $U^{T}$.

## Norms and Conditioning via SVD

- Norm and condition number in terms of singular values:

When multiplying a vector by matrix $\boldsymbol{A}=\boldsymbol{U} \boldsymbol{\Sigma} \boldsymbol{V}^{T}$

- Multiplication by $\boldsymbol{V}^{T}$ changes coordinate systems, leaving the norm unchanged
- Multiplication by $\boldsymbol{U}$ changes coordinate systems, leaving the norm unchanged so, only multiplication by $\mathbf{\Sigma}$ has an effect on the vector norm
- Note that $\|\boldsymbol{\Sigma}\|_{2}=\sigma_{\max },\left\|\boldsymbol{\Sigma}^{-1}\right\|_{2}=1 / \sigma_{\text {min }}$, so

$$
\kappa(\boldsymbol{A})=\kappa(\boldsymbol{\Sigma})=\frac{\sigma_{\max }}{\sigma_{\min }}
$$

## Visualization of Matrix Conditioning



## Existence of SVD

- Consider any maximizer $\boldsymbol{x}_{1} \in \mathbb{R}^{n}$ with $\left\|\boldsymbol{x}_{1}\right\|_{2}=1$ to $\left\|\boldsymbol{A} \boldsymbol{x}_{1}\right\|_{2}$

Let $\boldsymbol{y}_{1}=\boldsymbol{A} \boldsymbol{x}_{1} /\left\|\boldsymbol{A} \boldsymbol{x}_{1}\right\|_{2}$ and $\sigma_{1}=\boldsymbol{y}_{1}^{T} \boldsymbol{A} \boldsymbol{x}_{1}=\left\|\boldsymbol{A} \boldsymbol{x}_{1}\right\|_{2}$, then consider any maximizer $\boldsymbol{x}_{2}$ of

$$
\left\|\left(\boldsymbol{A}-\sigma_{1} \boldsymbol{y}_{1} \boldsymbol{x}_{1}^{T}\right) \boldsymbol{x}_{2}\right\|_{2} .
$$

We can see that $\boldsymbol{x}_{1} \perp \boldsymbol{x}_{2}$ since, otherwise, we have $\boldsymbol{x}_{2}=\alpha \boldsymbol{x}_{1}+\tilde{\boldsymbol{x}}_{2}$ with $\tilde{\boldsymbol{x}}_{2} \perp \boldsymbol{x}_{1}$ and $\left\|\tilde{\boldsymbol{x}}_{2}\right\|_{2}<\left\|\boldsymbol{x}_{2}\right\|_{2}$ and

$$
\left\|\left(\boldsymbol{A}-\sigma_{1} \boldsymbol{y}_{1} \boldsymbol{x}_{1}^{T}\right)\left(\alpha \boldsymbol{x}_{1}+\tilde{\boldsymbol{x}}_{2}\right)\right\|_{2}=\left\|\left(\boldsymbol{A}-\sigma_{1} \boldsymbol{y}_{1} \boldsymbol{x}_{1}^{T}\right) \tilde{\boldsymbol{x}}_{2}\right\|_{2}
$$

Hence we have a contradiction, since

$$
\left\|\left(\boldsymbol{A}-\sigma_{1} \boldsymbol{y}_{1} \boldsymbol{x}_{1}^{T}\right) \boldsymbol{x}_{2}\right\|_{2}<\left(1 /\left\|\tilde{\boldsymbol{x}}_{2}\right\|_{2}\right)\left\|\left(\boldsymbol{A}-\sigma_{1} \boldsymbol{y}_{1} \boldsymbol{x}_{1}^{T}\right) \tilde{\boldsymbol{x}}_{2}\right\|_{2} .
$$

More generally, we can see that any maximizer $\boldsymbol{x}_{i+1}$ to

$$
\left\|\left(\boldsymbol{A}-\left[\begin{array}{lll}
\boldsymbol{y}_{1} & \cdots & \boldsymbol{y}_{i}
\end{array}\right]\left[\begin{array}{lll}
\sigma_{1} & & \\
& \ddots & \\
& & \sigma_{i}
\end{array}\right]\left[\begin{array}{lll}
\boldsymbol{x}_{1} & \cdots & \boldsymbol{x}_{i}
\end{array}\right]^{T}\right) \boldsymbol{x}_{i+1}\right\|_{2}
$$

is orthogonal to $\boldsymbol{x}_{1}, \ldots, \boldsymbol{x}_{i}$ and similar for $\boldsymbol{y}_{i+1}$.

## Conditioning of Linear Systems

- Lets now return to formally deriving the conditioning of solving $\boldsymbol{A x}=\boldsymbol{b}$ :

Consider a perturbation to the right-hand side (input) $\hat{\boldsymbol{b}}=\boldsymbol{b}+\boldsymbol{\delta b}$

$$
\begin{aligned}
\boldsymbol{A} \hat{\boldsymbol{x}} & =\hat{b} \\
\boldsymbol{A}(x+\boldsymbol{x} \boldsymbol{x}) & =b+\boldsymbol{\delta} b \\
\boldsymbol{A} \boldsymbol{\delta} \boldsymbol{x} & =\boldsymbol{\delta} b
\end{aligned}
$$

we wish to bound the size of the relative perturbation to the output $\|\boldsymbol{\delta} \boldsymbol{x}\| /\|\boldsymbol{x}\|$ with respect to the size of the relative perturbation the the input $\|\boldsymbol{\delta b}\| /\|\boldsymbol{b}\|$

$$
\begin{aligned}
\boldsymbol{\delta} \boldsymbol{x} & =\boldsymbol{A}^{-1} \boldsymbol{\delta} \boldsymbol{b} \\
\frac{\|\boldsymbol{\delta} \boldsymbol{x}\|}{\|x\|} & =\frac{\left\|\boldsymbol{A}^{-1} \boldsymbol{\delta} \boldsymbol{b}\right\|}{\|\boldsymbol{x}\|} \leq \frac{\left\|\boldsymbol{A}^{-1}\right\| \cdot\|\boldsymbol{\delta} \boldsymbol{\|}\|}{\|x\|}
\end{aligned}
$$

we can use that $\|\boldsymbol{x}\| \geq\|\boldsymbol{b}\| / \sigma_{\max }=\|\boldsymbol{b}\| /\|\boldsymbol{A}\|$ so

$$
\frac{\|\boldsymbol{\delta} \boldsymbol{x}\|}{\|\boldsymbol{x}\|} \leq \underbrace{\|\boldsymbol{A}\| \cdot\left\|\boldsymbol{A}^{-1}\right\|}_{\kappa(\boldsymbol{A})} \cdot \frac{\|\boldsymbol{\delta} \boldsymbol{b}\|}{\|\boldsymbol{b}\|}=\frac{\sigma_{\max }\|\boldsymbol{\delta} \boldsymbol{b}\|}{\sigma_{\min }\|\boldsymbol{b}\|}
$$

## Conditioning of Linear Systems II

- Consider perturbations to the input coefficients $\hat{A}=A+\delta A$ :

In this case, we solve the perturbed system

$$
\begin{aligned}
\hat{\boldsymbol{A}} \hat{\boldsymbol{x}} & =\boldsymbol{b} \\
(\boldsymbol{A}+\boldsymbol{\delta} \boldsymbol{A})(\boldsymbol{x}+\boldsymbol{\delta} \boldsymbol{x}) & =\boldsymbol{b} \\
\boldsymbol{\delta} \boldsymbol{A} \boldsymbol{x}+\boldsymbol{A} \boldsymbol{\delta} \boldsymbol{x}+\boldsymbol{\delta} \boldsymbol{A} \boldsymbol{\delta} \boldsymbol{x} & =\mathbf{0} \\
\|\boldsymbol{\delta} \boldsymbol{A} \boldsymbol{x}\| & =\|\hat{\boldsymbol{A}} \boldsymbol{\delta} \boldsymbol{x}\|+O\left(\|\boldsymbol{\delta} \boldsymbol{A}\|^{2}\right)
\end{aligned}
$$

we wish to bound the size of the relative perturbation to the output $\|\boldsymbol{\delta} \boldsymbol{x}\| /\|\boldsymbol{x}\|$ with respect to the size of the relative perturbation the the input $\|\boldsymbol{\delta} \boldsymbol{A}\| /\|\boldsymbol{A}\|$

$$
\begin{aligned}
\|\boldsymbol{A} \boldsymbol{\delta} \boldsymbol{x}\| & =\|\boldsymbol{\delta} \boldsymbol{A} \boldsymbol{x}\|+O\left(\|\boldsymbol{\delta} \boldsymbol{A}\|^{2}\right) \\
\|\boldsymbol{\delta} \boldsymbol{x}\| & \leq\left\|\boldsymbol{A}^{-1}\right\|\|\boldsymbol{\delta} \boldsymbol{A} \boldsymbol{x}\| \leq\left\|\boldsymbol{A}^{-1}\right\| \cdot\|\boldsymbol{\delta} \boldsymbol{A}\| \cdot\|\boldsymbol{x}\|+O\left(\|\boldsymbol{\delta} \boldsymbol{A}\|^{2}\right) \\
\frac{\|\boldsymbol{\delta} \boldsymbol{x}\|}{\|\boldsymbol{x}\|} & \leq \underbrace{\left\|\boldsymbol{A}^{-1}\right\| \cdot\|\boldsymbol{A}\|}_{\kappa(\boldsymbol{A})} \cdot \frac{\|\boldsymbol{\delta} \boldsymbol{A}\|}{\|\boldsymbol{A}\|}+O\left(\|\boldsymbol{\delta} \boldsymbol{A}\|^{2}\right)
\end{aligned}
$$

## Solving Basic Linear Systems

- Solve $\boldsymbol{D} \boldsymbol{x}=\boldsymbol{b}$ if $\boldsymbol{D}$ is diagonal

$$
x_{i}=b_{i} / d_{i i} \text { with total cost } O(n)
$$

- Solve $Q x=b$ if $Q$ is orthogonal $\boldsymbol{x}=\boldsymbol{Q}^{T} \boldsymbol{b}$ with total cost $O\left(n^{2}\right)$
- Given SVD $\boldsymbol{A}=\boldsymbol{U} \boldsymbol{\Sigma} \boldsymbol{V}^{T}$, solve $\boldsymbol{A} \boldsymbol{x}=\boldsymbol{b}$
- Compute $\boldsymbol{z}=\boldsymbol{U}^{T} \boldsymbol{b}$
- Solve $\Sigma \boldsymbol{y}=\boldsymbol{z}$ (diagonal)
- Compute $\boldsymbol{x}=\boldsymbol{V} \boldsymbol{x}$


## Solving Triangular Systems

- $L x=b$ if $L$ is lower-triangular is solved by forward substitution:

$$
\begin{aligned}
& l_{11} x_{1}=b_{1} \\
& l_{21} x_{1}+l_{22} x_{2}=b_{2} \\
& l_{31} x_{1}+l_{32} x_{2}+l_{33} x_{3}=b_{3}
\end{aligned} \quad \Rightarrow \quad x_{1}=b_{1} / l_{11}, \begin{aligned}
& x_{2}=\left(b_{2}-l_{21} x_{1}\right) / l_{22} \\
& x_{3}=\left(b_{3}-l_{31} x_{1}-l_{32} x_{2}\right) / l_{33}
\end{aligned}
$$

- Algorithm can also be formulated recursively by blocks:

$$
\left[\begin{array}{ll}
l_{11} & \\
\boldsymbol{l}_{21} & \boldsymbol{L}_{22}
\end{array}\right]\left[\begin{array}{l}
x_{1} \\
\boldsymbol{x}_{\mathbf{2}}
\end{array}\right]=\left[\begin{array}{l}
b_{1} \\
\boldsymbol{b}_{\mathbf{2}}
\end{array}\right]
$$

$x_{1}=b_{1} / l_{11}$, then solve recursively for $\boldsymbol{x}_{\mathbf{2}}$ in $\boldsymbol{L}_{\mathbf{2 2}} \boldsymbol{x}_{\mathbf{2}}=\boldsymbol{b}_{\mathbf{2}}-\boldsymbol{l}_{\mathbf{2 1}} x_{1}$.

## Solving Triangular Systems

- Existence of solution to $\boldsymbol{L} \boldsymbol{x}=\boldsymbol{b}$ :

If some $l_{i i}=0$, the solution may not exist, and $L^{-1}$ does not exist.

- Uniqueness of solution: Even if some $l_{i i}=0$ and $L^{-1}$ does not exist, the system may have a solution. The solution will not be unique since columns of $L$ are necessarily linearly dependent if a diagonal element is zero. May want to select solution minimizing norm of $x$.
- Computational complexity of forward/backward substitution: The recursive algorithm has the cost recurrence,

$$
T(n)=T(n-1)+n=\sum_{i=1}^{n} i=n(n+1) / 2
$$

The total cost is $n^{2} / 2$ multiplications and $n^{2} / 2$ additions to leading order.

## Properties of Triangular Matrices

- $Z=X Y$ is lower triangular is $X$ and $Y$ are both lower triangular:

$$
\left[\begin{array}{ll}
z_{11} & \boldsymbol{z}_{12} \\
\boldsymbol{z}_{21} & \boldsymbol{Z}_{22}
\end{array}\right]=\left[\begin{array}{ll}
x_{11} & \\
\boldsymbol{x}_{21} & \boldsymbol{X}_{22}
\end{array}\right]\left[\begin{array}{ll}
y_{11} & \\
\boldsymbol{y}_{21} & \boldsymbol{Y}_{22}
\end{array}\right]
$$

Clearly, $z_{11}=x_{11} y_{11}$ and $z_{12}=0$, then we proceed by the same argument for the triangular matrix product $\boldsymbol{Z}_{22}=\boldsymbol{X}_{22} \boldsymbol{Y}_{22}$.

- $L^{-1}$ is lower triangular if it exists:

We give a constructive proof by providing an algorithm for triangular matrix inversion. We need $\boldsymbol{Y}=\boldsymbol{X}^{-1}$ so

$$
\left[\begin{array}{ll}
\boldsymbol{Y}_{11} & \\
\boldsymbol{Y}_{21} & \boldsymbol{Y}_{22}
\end{array}\right]\left[\begin{array}{ll}
\boldsymbol{X}_{11} & \\
\boldsymbol{X}_{21} & \boldsymbol{X}_{22}
\end{array}\right]=\left[\begin{array}{ll}
\boldsymbol{I} & \\
& \boldsymbol{I}
\end{array}\right],
$$

from which we can deduce

$$
\boldsymbol{Y}_{11}=\boldsymbol{X}_{11}^{-1}, \quad \boldsymbol{Y}_{22}=\boldsymbol{X}_{22}^{-1}, \quad \boldsymbol{Y}_{21}=-\boldsymbol{Y}_{22} \boldsymbol{X}_{21} \boldsymbol{Y}_{11}
$$

## LU Factorization

- An LU factorization consists of a unit-diagonal lower-triangular factor $L$ and upper-triangular factor $U$ such that $A=L U$ :
- Unit-diagonal implies each $l_{i i}=1$, leaving $n(n-1) / 2$ unknowns in $\boldsymbol{L}$ and $n(n+1) / 2$ unknowns in $\boldsymbol{U}$, for a total of $n^{2}$, the same as the size of $\boldsymbol{A}$.
- For rectangular matrices $\boldsymbol{A} \in \mathbb{R}^{m \times n}$, one can consider a full $L U$ factorization, with $\boldsymbol{L} \in \mathbb{R}^{m \times \max (m, n)}$ and $\boldsymbol{U} \in \mathbb{R}^{\max (m, n) \times n}$, but it is fully described by a reduced $L U$ factorization, with lower-trapezoidal $L \in \mathbb{R}^{m \times \min (m, n)}$ and upper-trapezoidal $\boldsymbol{U} \in \mathbb{R}^{\min (m, n) \times n}$.
- Given an LU factorization of $A$, we can solve the linear system $A x=b$ :
- using forward substitution $L \boldsymbol{y}=\boldsymbol{b}$
- using backward substitution to solve $\boldsymbol{U} \boldsymbol{x}=\boldsymbol{y}$

Backward substitution is the same as forward substitution with a reversal of the ordering of the elements of the vectors and the ordering of the rows/columns of the matrix.

## Gaussian Elimination Algorithm

- Algorithm for factorization is derived from equations given by $A=L U$ :

$$
\left[\begin{array}{ll}
a_{11} & \boldsymbol{a}_{12} \\
\boldsymbol{a}_{21} & \boldsymbol{A}_{22}
\end{array}\right]=\left[\begin{array}{cc}
1 & \\
\boldsymbol{l}_{21} & \boldsymbol{L}_{22}
\end{array}\right]\left[\begin{array}{cc}
u_{11} & \boldsymbol{u}_{12} \\
& \boldsymbol{U}_{22}
\end{array}\right]=\left[\begin{array}{ll}
\boldsymbol{L}_{11} & \\
\boldsymbol{L}_{21} & \boldsymbol{L}_{22}
\end{array}\right]\left[\begin{array}{ll}
\boldsymbol{U}_{11} & \boldsymbol{U}_{12} \\
& \boldsymbol{U}_{22}
\end{array}\right]
$$

- First, observe $\left[\begin{array}{ll}u_{11} & \boldsymbol{u}_{12}\end{array}\right]=\left[\begin{array}{ll}a_{11} & \boldsymbol{a}_{12}\end{array}\right]$
- To obtain $\boldsymbol{l}_{21}$ compute $\boldsymbol{l}_{21}=\boldsymbol{a}_{21} / u_{11}$
- Obtain $\boldsymbol{L}_{22}$ and $\boldsymbol{U}_{22}$ by recursively computing LU of the Schur complement

$$
\boldsymbol{S}=\boldsymbol{A}_{22}-\boldsymbol{l}_{21} \boldsymbol{u}_{12}
$$

- The computational complexity of $\mathbf{L U}$ is $O\left(n^{3}\right)$ :

Computing $\boldsymbol{l}_{21}=\boldsymbol{a}_{21} / u_{11}$ requires $O(n)$ operations, finding $\boldsymbol{S}$ requires $2 n^{2}$, so to leading order the complexity of $L U$ is

$$
T(n)=T(n-1)+2 n^{2}=\sum_{i=1}^{n} 2 i^{2} \approx 2 n^{3} / 3
$$

## Existence of LU Factorization

- The LU factorization may not exist: Consider matrix $\left[\begin{array}{ll}3 & 2 \\ 6 & 4 \\ 0 & 3\end{array}\right]$.

Proceeding with Gaussian elimination we obtain

$$
\left[\begin{array}{ll}
3 & 2 \\
6 & 4 \\
0 & 3
\end{array}\right]=\left[\begin{array}{cc}
1 & 0 \\
2 & 1 \\
0 & l_{32}
\end{array}\right]\left[\begin{array}{cc}
3 & 2 \\
0 & u_{21}
\end{array}\right] .
$$

Then we need that $4=4+u_{21}$ so $u_{21}=0$, but at the same time $l_{32} u_{21}=3$. More generally, if and only if for any partitioning $\left[\begin{array}{ll}\boldsymbol{A}_{11} & \boldsymbol{A}_{12} \\ \boldsymbol{A}_{21} & \boldsymbol{A}_{22}\end{array}\right]$ the leading minor is singular $\left(\operatorname{det}\left(\boldsymbol{A}_{11}\right)=0\right), \boldsymbol{A}$ has no $L U$ factorization.

- Permutation of rows enables us to transform the matrix so the LU factorization does exist:
Gaussian elimination can only fail if dividing by zero. At every recursive step of Gaussian elimination, if the leading entry of the first row is zero, we permute it with a row with an leading nonzero (if $a_{21}=\mathbf{0}$, we set $u_{11}=0$ and $l_{21}=0$ ).


## Gaussian Elimination with Partial Pivoting

- Partial pivoting permutes rows to make divisor $u_{i i}$ maximal at each step:

Based on our argument above, for any matrix $\boldsymbol{A}$ there exists a permutation matrix $\boldsymbol{P}$ that can permute the rows of $\boldsymbol{A}$ to permit an LU factorization,

$$
P A=L U
$$

Partial pivoting finds such a permutation matrix $P$ one row at a time. The ith row is selected to maximize the magnitude of the leading element (over elements in the first column), which becomes the entry $u_{i i}$. This selection ensures that we are never forced to divide by zero during Gaussian elimination and that the magnitude of any element in $L$ is at most 1.

- A row permutation corresponds to an application of a row permutation matrix $\boldsymbol{P}_{j k}=\boldsymbol{I}-\left(\boldsymbol{e}_{j}-\boldsymbol{e}_{k}\right)\left(\boldsymbol{e}_{j}-\boldsymbol{e}_{k}\right)^{T}$ :
If we permute row $i_{j}$.o be the leading (ith) row at the ith step, the overall permutation matrix is given by $\boldsymbol{P}^{T}=\prod_{i=1}^{n-1} \boldsymbol{P}_{i i_{j}}$.


## Partial Pivoting Example

- Lets consider again the matrix $\boldsymbol{A}=\left[\begin{array}{ll}3 & 2 \\ 6 & 4 \\ 0 & 3\end{array}\right]$.
- The largest magnitude element in the first column is 6 , so we select this as our pivot and perform the first step of $L U$

$$
\underbrace{\left[\begin{array}{lll}
1 & 1 & \\
1 & & \\
& & 1
\end{array}\right]}_{P_{1}}\left[\begin{array}{ll}
6 & 4 \\
3 & 2 \\
0 & 3
\end{array}\right]=\left[\begin{array}{c}
1 \\
1 / 2 \\
0
\end{array}\right]\left[\begin{array}{ll}
6 & 4
\end{array}\right]+\left[\begin{array}{cc}
0 & 0 \\
0 & 2-(1 / 2) \cdot 4 \\
0 & 3-0 \cdot 4
\end{array}\right]
$$

- The Schur complement is $\left[\begin{array}{ll}0 & 3\end{array}\right]^{T}$ and we proceed with pivoted $L U$,

$$
\underbrace{\left[\begin{array}{ll} 
& 1 \\
1 &
\end{array}\right]}_{\boldsymbol{P}_{2}}\left[\begin{array}{l}
0 \\
3
\end{array}\right]=\left[\begin{array}{l}
1 \\
0
\end{array}\right][3]
$$

- The overall LU factorization is then given by $\boldsymbol{P}_{1}\left[\begin{array}{ll}1 & \\ & \boldsymbol{P}_{2}\end{array}\right] \boldsymbol{A}=\left[\begin{array}{cc}1 & \\ 0 & 1 \\ 1 / 2 & 0\end{array}\right]\left[\begin{array}{ll}6 & 4 \\ & 3\end{array}\right]$


## Complete Pivoting

- Complete pivoting permutes rows and columns to make divisor $u_{i i}$ is maximal at each step:
- Partial pivoting ensures that the magnitude of the multipliers satisfies $\left|\boldsymbol{l}_{21}\right|=\left|\boldsymbol{a}_{21}\right| /\left|u_{11}\right| \leq \mathbf{1}$
- Complete pivoting also gives $\left\|\boldsymbol{u}_{12}\right\|_{\infty} \leq\left|u_{11}\right|$ and consequently $\left|\boldsymbol{l}_{21}\right| \cdot\left|\left|\boldsymbol{u}_{12}\left\|_{\infty}=\left|\boldsymbol{a}_{21}\right| \cdot| | \boldsymbol{u}_{12}\left|\|_{\infty} /\left|u_{11}\right| \leq\left|\boldsymbol{a}_{21}\right|\right.\right.\right.\right.$
- Complete pivoting yields a factorization of the form $\mathbf{L U}=\boldsymbol{P A Q}$ where $\boldsymbol{P}$ and $Q$ are permutation matrices
- Complete pivoting is noticeably more expensive than partial pivoting:
- Partial pivoting requires just $O(n)$ comparison operations and a row permutation
- Complete pivoting requires $O\left(n^{2}\right)$ comparison operations, which somewhat increases the leading order cost of LU overall


## Round-off Error in LU

- Lets consider factorization of $\left[\begin{array}{ll}\epsilon & 1 \\ 1 & 1\end{array}\right]$ where $\epsilon<\epsilon_{\text {mach }}$ :
- Without pivoting we would compute $\boldsymbol{L}=\left[\begin{array}{cc}1 & 0 \\ 1 / \epsilon & 1\end{array}\right], \boldsymbol{U}=\left[\begin{array}{cc}\epsilon & 1 \\ 0 & 1-1 / \epsilon\end{array}\right]$
- Rounding yields $f l(\boldsymbol{U})=\left[\begin{array}{cc}\epsilon & 1 \\ 0 & -1 / \epsilon\end{array}\right]$
- This leads to $\boldsymbol{L} f l(\boldsymbol{U})=\left[\begin{array}{ll}\epsilon & 1 \\ 1 & 0\end{array}\right]$, a backward error of $\left[\begin{array}{ll}0 & 0 \\ 0 & 1\end{array}\right]$
- Permuting the rows of $\boldsymbol{A}$ in partial pivoting gives $\boldsymbol{P} \boldsymbol{A}=\left[\begin{array}{ll}1 & 1 \\ \epsilon & 1\end{array}\right]$
- We now compute $\boldsymbol{L}=\left[\begin{array}{ll}1 & 0 \\ \epsilon & 1\end{array}\right], \boldsymbol{U}=\left[\begin{array}{cc}1 & 1 \\ 0 & 1-\epsilon\end{array}\right]$, so $f l(\boldsymbol{U})=\left[\begin{array}{ll}1 & 1 \\ 0 & 1\end{array}\right]$
- This leads to $\boldsymbol{L} f l(\boldsymbol{U})=\left[\begin{array}{cc}1 & 1 \\ \epsilon & 1+\epsilon\end{array}\right]$, a backward error of $\left[\begin{array}{ll}0 & 0 \\ 0 & \epsilon\end{array}\right]$


## Error Analysis of LU

- The main source of round-off error in LU is in the computation of the Schur complement:
- Recall that division is well-conditioned, while addition can be ill-conditioned
- After $k$ steps of LU, we are working on Schur complement $\boldsymbol{A}_{22}-\boldsymbol{L}_{21} \boldsymbol{U}_{12}$ where $\boldsymbol{A}_{22}$ is $(n-k) \times(n-k), \boldsymbol{L}_{21}$ and $\boldsymbol{U}_{12}^{T}$ are $(n-k) \times k$
- Partial pivoting and complete pivoting improve stability by making sure $\boldsymbol{L}_{21} \boldsymbol{U}_{12}$ is small in norm
- When computed in floating point, absolute backward error $\delta A$ in LU (so $\hat{\boldsymbol{L}} \hat{\boldsymbol{U}}=\boldsymbol{A}+\boldsymbol{\delta} \boldsymbol{A})$ is $\left|\delta a_{i j}\right| \leq \epsilon_{\text {mach }}(|\hat{\boldsymbol{L}}| \cdot|\hat{\boldsymbol{U}}|)_{i j}$
For any $a_{i j}$ with $j \geq i$ (lower-triangle is similar), we compute

$$
a_{i j}-\sum_{k=1}^{i} \hat{l}_{i k} \hat{u}_{k j}=a_{i j}-\left\langle\hat{\boldsymbol{l}}_{i}, \hat{\boldsymbol{u}}_{j}\right\rangle,
$$

which in floating point incurs round-off error at most $\epsilon_{\text {mach }}\langle | \hat{\boldsymbol{l}}_{i}\left|,\left|\hat{\boldsymbol{u}}_{j}\right|\right\rangle$. Using this, for complete pivoting, we can show $\left|\delta a_{i j}\right| \leq \epsilon_{\text {mach }} n^{2}\|\boldsymbol{A}\|_{\infty}$.

## Helpful Matrix Properties

- Matrix is diagonally dominant, so $\sum_{i \neq j}\left|a_{i j}\right| \leq\left|a_{i i}\right|$ :

Pivoting is not required if matrix is strictly diagonally dominant
$\sum_{i \neq j}\left|a_{i j}\right|<\left|a_{i i}\right|$.

- Matrix is symmetric positive definite (SPD), so $\forall_{\boldsymbol{x} \neq 0}, \boldsymbol{x}^{T} \boldsymbol{A} \boldsymbol{x}>0$ :
$\boldsymbol{L}=\boldsymbol{U}$ and pivoting is not required, Cholesky algorithm $\boldsymbol{A}=\boldsymbol{L} \boldsymbol{L}^{T}$ can be used ( $L$ in Cholesky is not unit-diagonal).
- Matrix is symmetric but indefinite:

Compute pivoted $L D L$ factorization $\boldsymbol{P} \boldsymbol{A} \boldsymbol{P}^{T}=\boldsymbol{L} \boldsymbol{D} \boldsymbol{L}^{T}$ (where $\boldsymbol{L}$ is lower-triangular and unit-diagonal, while $\boldsymbol{D}$ is block-diagonal with 2-by-2 diagonal or antidiagonal blocks)

- Matrix is banded, $a_{i j}=0$ if $|i-j|>b$ :

LU without pivoting and Cholesky preserve banded structure and require only $O\left(n b^{2}\right)$ work.

## Solving Many Linear Systems

- Suppose we have computed $A=L U$ and want to solve $A X=B$ where $B$ is $n \times k$ with $k<n$ :
Cost is $O\left(n^{2} k\right)$ for solving the $k$ independent linear systems
- Suppose we have computed $\boldsymbol{A}=\boldsymbol{L} \boldsymbol{U}$ and now want to solve a perturbed system $\left(\boldsymbol{A}-\boldsymbol{u} \boldsymbol{v}^{T}\right) \boldsymbol{x}=\boldsymbol{b}$ :
Can use the Sherman-Morrison-Woodbury formula

$$
\left(\boldsymbol{A}-\boldsymbol{u} \boldsymbol{v}^{T}\right)^{-1}=\boldsymbol{A}^{-1}+\frac{\boldsymbol{A}^{-1} \boldsymbol{u} \boldsymbol{v}^{T} \boldsymbol{A}^{-1}}{1-\boldsymbol{v}^{T} \boldsymbol{A}^{-1} \boldsymbol{u}}
$$

- Consequently we have $\boldsymbol{A} \boldsymbol{x}=\boldsymbol{b}+\frac{\boldsymbol{u} \boldsymbol{v}^{T} \boldsymbol{A}^{-1}}{1-\boldsymbol{v}^{T} \boldsymbol{A}^{-1} \boldsymbol{u}} \boldsymbol{b}=\boldsymbol{b}+\frac{\boldsymbol{v}^{T} \boldsymbol{A}^{-1} \boldsymbol{b}}{1-\boldsymbol{v}^{T} A^{-1} \boldsymbol{u}} \boldsymbol{u}$
- Need not form $\boldsymbol{A}^{-1}$ or $\boldsymbol{L}^{-1}$ or $\boldsymbol{U}^{-1}$, suffices to use backward/forward substitution to solve $\boldsymbol{w}^{T} \boldsymbol{A}=\boldsymbol{v}^{T}$, i.e. solve $\boldsymbol{U}^{T} \boldsymbol{L}^{T} \boldsymbol{w}=\boldsymbol{v}$ and then solve

$$
\boldsymbol{L} \boldsymbol{U} \boldsymbol{x}=\boldsymbol{b}+\underbrace{\left(\frac{\boldsymbol{w}^{T} \boldsymbol{b}}{1-\boldsymbol{w}^{T} \boldsymbol{u}}\right)}_{\text {scalar }} \boldsymbol{u}
$$


[^0]:    ${ }^{1}$ These slides have been drafted by Edgar Solomonik as lecture templates and supplementary material for the book "Scientific Computing: An Introductory Survey" by Michael T. Heath (slides).

