

Computational Techniques: 233 – 2

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The following topics will be covered:

- Conditioning
- Convergence and fixed point problems
- Iterative solution of linear equations
- Laplace transforms
- Functions of several variables
- Introduction to continuous optimisation

Note that Sparse Matrix Techniques will not be covered:
they are discussed in the “Background Notes” however.

A polynomial equation

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$$(x - 1)^4 = 0$$

Four coincident roots, $x = 1$

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The change of 10^{-8} in a 'parameter' has caused a change of 10^{-2} in the solution: a ratio of 1000000 !

Linear equations

Now consider the equations

$$x + y = 1$$

$$x + \alpha y = 0$$

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If α increases from 0.999 to 0.9999, the solution changes from $(-999, 1000)$ to $(-9999, 10000)$: an increase by a factor of 10 from a change of about 0.1% in the parameter α .

Mathematical conditioning

The *condition number*, or just *condition*, of a problem P is the maximum size-ratio:

$$\kappa(P) = \max_{d_1, d_2} \frac{\|s(d_1) - s(d_2)\|}{\|d_1 - d_2\|}$$

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A relatively small number (near to 1) implies that even in the worst case, the solution will not be too sensitive to small changes in the input ... will not “blow up”.

Key point is that we're looking at the *worst case*.

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A lower bound on the norm is:

$$\|A\| \geq \frac{\|Ax\|}{\|x\|}$$

for any specific non-zero vector x .

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$$\begin{aligned}\|Ax\|_1 &= \|(\dots, \sum_j a_{ij}x_j, \dots)\|_1 \\ &\leq \sum_{i,j} |a_{ij}x_j| \\ &= \sum_{i,j} |a_{ij}| |x_j| \\ &= \sum_j \|a_j\|_1 |x_j| \\ &\leq \max_j \|a_j\|_1 \|x\|_1 \\ &= \|x\|_1 \|A\|_1\end{aligned}$$

with equality for some vector x .

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Relative perturbation is

$$\frac{\|\delta x_b\|}{\|x\|} \leq \|A^{-1}\| \|A\| \frac{\|\delta b\|}{\|b\|}$$

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Then

$$\text{cond}(A) \geq \max \left\{ \frac{\|\delta x_b\|/\|x\|}{\|\delta b\|/\|b\|}, \frac{\|\delta x_A\|/\|x\|}{\|\delta A\|/\|A\|} \right\}$$

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$\|A\| = \|A^{-1}\| = 10^4$ so Condition of A is 10^8

Application to Least Squares

- In least squares problems there are m equations in n variables, where $m > n$
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Application to Least Squares

- In least squares problems there are m equations in n variables, where $m > n$
- Gives a non-square ($m \times n$) matrix A , for which the condition *can* be calculated (see below)
- Condition of $A^T A$ is the square of the condition of A , for ℓ_2 -norm
- So unfortunately 4 norms get multiplied together, often giving a very big condition number
- For the normal equation, we calculate the condition of $A^T A$ directly:

$$\text{cond}(A^T A) = \|A^{-1}(A^T)^{-1}\| \|A^T A\|$$

Example

Consider:

$$A = \begin{bmatrix} 1 & 1 \\ 1 & 1 \\ 0 & 10^{-4} \end{bmatrix}$$

Normal equation matrix is

$$A^T A = \begin{bmatrix} 2 & 2 \\ 2 & 2 + 10^{-8} \end{bmatrix}$$

with inverse

$$(A^T A)^{-1} = 0.5 \times 10^8 \times \begin{bmatrix} 2 + 10^{-8} & -2 \\ -2 & 2 \end{bmatrix}$$

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So

$$\text{cond}(A^T A) = (4 + 10^{-8})(4 + 10^{-8}) \times 0.5 \times 10^8 \simeq 8 \times 10^8$$

Why are the notes different?

- Notes say $\text{cond}(A^T A) = 7.8 \times 10^8$
- They also say “condition of $A^T A$ is the square of the condition of A ” but this is *not true* for ℓ_1 - and ℓ_∞ -norms (try it!)
- So a big difference is using the ℓ_2 -norm.

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- So a big difference is using the ℓ_2 -norm.
- Singular value decomposition gives $\sigma_1(A) = \sigma_1(A^T) = 2.0$ and $\sigma_1(A^{-1}) = 14142.1$
- Thus, using these ℓ_2 -norms,

$$\text{cond}_2(A) = 28284.3$$

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- This agrees with the notes' 2.8×10^4 to 2 s.f. ... see page 78
- Squaring 2.8×10^4 does indeed give 7.8×10^8 to 2 s.f.
- However, squaring $\text{cond}_2(A)$ gives 8×10^8 !!

Definition: A *metric space* is a non-empty set S of points (or objects) together with a function $d : S \times S \rightarrow \mathbb{R}$ (the metric of the space) satisfying:

- ❶ $d(x, x) = 0$;
- ❷ $d(x, y) > 0$ if $x \neq y$;
- ❸ $d(x, y) = d(y, x)$
- ❹ $d(x, y) \leq d(x, z) + d(z, y)$

for all points $x, y, z \in S$.

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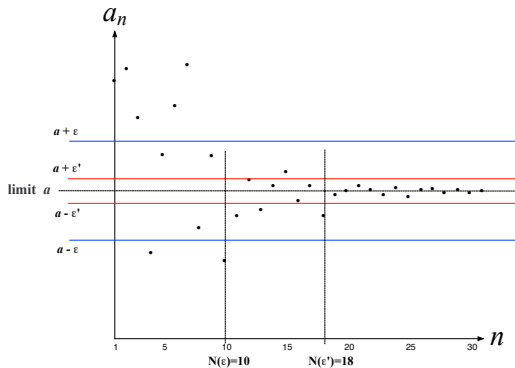
- Obviously true for distances in real spaces such as $\mathbb{R}, \mathbb{R}^2, \mathbb{R}^3$;
- Actually all we need to define notions of limits and convergence.

Definition: A sequence a_1, a_2, \dots converges to a limit $\ell \in \mathbb{R}$, written $a_n \rightarrow \ell$ as $n \rightarrow \infty$ or $\lim_{n \rightarrow \infty} a_n = \ell$, iff

$$\forall \epsilon > 0, \exists N \in \mathbb{N} \text{ such that } \forall n > N, |a_n - \ell| < \epsilon$$

- equivalently, $\ell - \epsilon < a_n < \ell + \epsilon$
- 'tramlines' ϵ away from the limit value ℓ

Illustration of convergence



Need a bigger N as ϵ decreases

Theorem (Cauchy): The sequence a_1, a_2, \dots is convergent if and only if $\forall \epsilon > 0, \exists N$ such that $|a_n - a_m| < \epsilon$ for all $n, m > N$.

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- Useful because you don't need to know what the limit is (when it exists), e.g. when a_n is defined by a recurrence relation or a recursive function
- Also a test for divergence

Example

$$a_n = \sum_{i=1}^n \frac{1}{i(i+1)}$$

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$$\begin{aligned} a_n - a_m &= \frac{1}{(m+1)(m+2)} + \dots + \frac{1}{n(n+1)} \\ &= \left(\frac{1}{m+1} - \frac{1}{m+2} \right) + \left(\frac{1}{m+2} - \frac{1}{m+3} \right) + \\ &\quad \dots + \left(\frac{1}{n} - \frac{1}{n+1} \right) \\ &= \frac{1}{m+1} - \frac{1}{n+1} \rightarrow 0 \quad \text{as } n > m \rightarrow \infty \end{aligned}$$

End of revision! More generally, in metric spaces ...

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Definition: A sequence $\{x_n\}$ in a metric space S is called a *Cauchy sequence* if for all $\epsilon > 0$ there exists an integer N such that for all $n, m \geq N$, $d(x_n, x_m) < \epsilon$.

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Definition: A metric space S in which every Cauchy sequence has a limit in S is called *complete*.

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Definition: A metric space S in which every Cauchy sequence has a limit in S is called *complete*.

Theorem: Suppose a sequence $\{x_n\}$ converges in a metric space S . Then $\{x_n\}$ is Cauchy.

Application and examples

- These theorems mean that in any complete metric space a sequence is convergent if and only if it is Cauchy, which is relatively easy to test.
- It can be shown that \mathbb{R}^k is complete for all $k \geq 1$.

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Proof hint:

- A cauchy sequence is bounded and so has a convergent sub-sequence (Fundamental axiom / Bolzano-Weierstrass theorem).
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Proof hint:
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 - It therefore has the same limit as the subsequence
- This is very useful computationally, for example in numerical iterations.
- Examples in \mathbb{R} given in the background notes.
- Used to prove convergence of certain fixed point iterations later.

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Proof: Suppose $x_n \rightarrow \ell_1$ and $x_n \rightarrow \ell_2$ as $n \rightarrow \infty$. Then

$$d(\ell_1, \ell_2) \leq d(\ell_1, x_n) + d(x_n, \ell_2)$$

for all n . But given any $\epsilon > 0$, for sufficiently large n , both terms on the RHS are less than $\epsilon/2$ so

$$d(\ell_1, \ell_2) < \epsilon$$

Hence $d(\ell_1, \ell_2) = 0$ since ϵ was arbitrary and so $\ell_1 = \ell_2$ by the definition of a metric.

Definition: Let $f : S \rightarrow S$ be a function from a metric space S to itself. A point p in S is called a *fixed point* (sometimes fixpoint) of f if $f(p) = p$. The function f is a *contraction* of S if there exists a real number α , $0 < \alpha < 1$, called a *contraction constant*, such that

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Theorem (Fixed point theorem): A continuous contraction f of a complete metric space S has a unique fixed point.

Proof of fixed point theorem

For any point $p \in S$, define the sequence $\{p_n\}$ by

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Then

$$d(p_{n+1}, p_n) = d(f(p_n), f(p_{n-1})) \leq \alpha d(p_n, p_{n-1}) \leq \dots \leq c\alpha^n$$

where $c = d(p_1, p_0)$.

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Hence, for $m > n$, by the triangle inequality,

$$d(p_m, p_n) \leq \sum_{k=n}^{m-1} d(p_{k+1}, p_k) \leq c \sum_{k=n}^{m-1} \alpha^k = c \frac{\alpha^n - \alpha^m}{1 - \alpha} < \frac{c\alpha^n}{1 - \alpha}$$

Since $\alpha < 1$, $d(p_m, p_n) \rightarrow 0$ as $m, n \rightarrow \infty$, and so $\{p_n\}$ is a Cauchy sequence.

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Finally, if p and p' are both fixed points,

$$d(p, p') = d(f(p), f(p')) \leq \alpha d(p, p')$$

where $0 < \alpha < 1$, so $d(p, p') = 0$ and the fixed points are the same and hence unique.

Iterative solutions of linear equations

A splitting of a square matrix A is defined by a nonsingular matrix $M = A - N$.

Suppose we are solving the equation $Ax = b$. Then we may write $Mx = b - Nx$ so that $x = M^{-1}b - M^{-1}Nx$.

If $G = -M^{-1}N$ and $c = M^{-1}b$, we need to solve

$$x = Gx + c$$

and we can define the iteration

$$x^{(k+1)} = Gx^{(k)} + c$$

together with some starting value $x^{(0)}$

Convergence of the iteration

Theorem : For any matrix norm, if $\|G\| < 1$, then $x^{(k+1)} = Gx^{(k)} + c$ converges for any starting point $x^{(0)}$.

Proof : Given that x is the correct solution, let $y^{(k)} = x^{(k)} - x$ for $k = 0, 1, 2, \dots$. Then $y^{(k+1)} = G(y^{(k)})$ and so

$$\|y^{(k+1)}\| \leq \|G\| \|y^{(k)}\| \leq \dots \leq \|G\|^{k+1} \|y^{(0)}\| \rightarrow 0$$

as $k \rightarrow \infty$ since $\|G\| < 1$

Proposition : A sufficient condition for convergence is any of:

- ① $\lim_{k \rightarrow \infty} G^k = 0$
- ② $\lim_{k \rightarrow \infty} G^k \vec{x} = \vec{0} \quad \forall \vec{x} \in \mathbb{R}^m$
- ③ $\rho(G) < 1$

where $\rho(G) = \max_i |\lambda_i|$ is the largest of the absolute values of the eigenvalues of G , called the *spectral radius* of G .

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where $\rho(G) = \max_i |\lambda_i|$ is the largest of the absolute values of the eigenvalues of G , called the *spectral radius* of G .

Proof : Looking at the proof of the Theorem,

$$y^{(k)} = G(y^{(k-1)}) = \dots = G^k y^{(0)}$$

Thus either of conditions 1 or 2 implies $y^{(k)} \rightarrow 0$ and so $x^{(k)} \rightarrow x$ as $k \rightarrow \infty$. For condition 3, diagonalising G (more generally, when there are multiple eigenvalues, putting G into Jordan Normal Form) we have

$$G^k = V^{-1} D^k V \rightarrow 0 \quad \text{as } k \rightarrow \infty$$

where $D = \text{diag}(\lambda_1, \dots, \lambda_m)$ and $|\lambda_i| < 1 \quad \forall i$.

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For example:

- $M = I$ is good for 1.
- $M = A$ is good for 2. but probably not for 1.

Common splitting of A

Assuming A has no zeros on the diagonal (can often relabel the variables x_i to permute the rows and columns, if necessary, to avoid this),

$$A = D - \tilde{L} - \tilde{U} = D(I - L - U)$$

where D is the diagonal of A , $-\tilde{L}$, \tilde{U} are the strict lower and upper triangular parts of A , $L = D^{-1}\tilde{L}$ and $U = D^{-1}\tilde{U}$

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Example :

$$A = \begin{bmatrix} 2 & -4 & 2 \\ -3 & 1 & -5 \\ 6 & -2 & 2 \end{bmatrix}$$

Using this splitting gives $N = \tilde{L} + \tilde{U}$ and the fixed point equation

$$x = D^{-1}b - D^{-1}Nx$$

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The iteration $x^{(k+1)} = Gx^{(k)} + c$ now yields

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- Update of x_i requires access to only row i of A
- Good for parallel computation (see Course 429)

Gauss-Seidel method

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$$x_i^{(k+1)} = \frac{1}{a_{ii}} \left(b_i - \sum_{j=1}^{i-1} a_{ij} x_j^{(k+1)} - \sum_{j=i+1}^m a_{ij} x_j^{(k)} \right)$$

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So now $G = (I - L)^{-1} U$ as opposed to $L + U$ for Jacobi.

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Now $G = (I - \omega L)^{-1}((1 - \omega)I + \omega U)$, so $\omega = 1$ gives Gauss-Seidel.

Definitions: A square matrix A is *strictly row diagonally dominant* if $|a_{ii}| > \sum_{j \neq i} |a_{ij}|$.

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Theorem: A *sufficient* condition for both Jacobi and Gauss-Seidel to converge is that A is strictly row diagonally dominant. G-S is faster.

Sketch of the proof

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- 2 Strict diagonal dominance implies all absolute row sums of G are less than one.
- 3 Hence the *maximum* absolute row sum is less than one, and so therefore is the ℓ_∞ -norm.
- 4 Any norm less than one leads to convergence, by the theorem above.

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where the block matrices A_{11} and A_{22} are square.

- 2 A necessary condition for $\text{SOR}(\omega)$ to converge is that $0 < \omega < 2$. This condition is also sufficient if A is positive definite (see later for definition).