Numerical Analysis II

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Section 1

Initial Value Problems for ODEs

IVP of ODE

We study numerical solution for initial value problem (IVP) of ordinary differential equations (ODE).

A basic IVP:

$$rac{\mathsf{d} y}{\mathsf{d} t} = f(t,y), \quad ext{for } a \leq t \leq b$$

with initial value $y(a) = \alpha$.

Remark

f is given and called the defining function of IVP.
α is given and called the initial value.
y(t) is called the solution of the IVP if
y(a) = α;
y'(t) = f(t, y(t)) for all t ∈ [a, b].

IVP of ODE

Example

The following is a basic IVP:

$$y' = y - t^2 + 1$$
, $t \in [0, 2]$, and $y(0) = 0.5$

The defining function is f(t, y) = y - t² + 1.
Initial value is y(0) = 0.5.
The solution is y(t) = (t + 1)² - e^t/2 because:
y(0) = (0 + 1)² - e⁰/2 = 1 - 1/2 = 1/2;
We can check that y'(t) = f(t, y(t)):
y'(t) = 2(t + 1) - e^t/2
f(t, y(t)) = y(t) - t² + 1 = (t + 1)² - e^t/2 - t² + 1 = 2(t + 1) - e^t/2

IVP of ODE (cont.)

More general or complex cases:

► IVP of ODE system:

$$\begin{cases} \frac{dy_1}{dt} = f_1(t, y_1, y_2, \dots, y_n) \\ \frac{dy_2}{dt} = f_2(t, y_1, y_2, \dots, y_n) \\ \vdots \\ \frac{dy_n}{dt} = f_n(t, y_1, y_2, \dots, y_n) \end{cases} \text{ for } a \leq t \leq b$$

with initial value $y_1(a) = \alpha_1, \ldots, y_n(a) = \alpha_n$. • High-order ODE:

$$y^{(n)}=f(t,y,y',\ldots,y^{(n-1)}) \quad ext{for } a\leq t\leq b$$

with initial value $y(a) = \alpha_1, y'(a) = \alpha_2, \dots, y^{(n-1)}(a) = \alpha_n$.

Why numerical solutions for IVP?

- ODEs have extensive applications in real-world: science, engineering, economics, finance, public health, etc.
- Analytic solution? Not with almost all ODEs.
- ► Fast improvement of computers.

Definition (Lipschitz functions)

A function f(t, y) defined on $D = \{(t, y) : t \in \mathbb{R}_+, y \in \mathbb{R}\}$ is called Lipschitz with respect to y if there exists a constant L > 0

$$|f(t, y_1) - f(t, y_2)| \le L|y_1 - y_2|$$

for all $t \in \mathbb{R}_+$, and $y_1, y_2 \in \mathbb{R}$.

Remark

We also call f is Lipschitz with respect to y with constant L, or simply f is L-Lipschitz with respect to y.

Example

Function f(t, y) = t|y| is Lipschitz with respect to y on the set $D := \{(t, y)|t \in [1, 2], y \in [-3, 4]\}.$

Solution: For any $t \in [1, 2]$ and $y_1, y_2 \in [-3, 4]$, we have

$$|f(t,y_1) - f(t,y_2)| = |t|y_1| - t|y_2|| \le t|y_1 - y_2| \le 2|y_1 - y_2|.$$

So f(t, y) = t|y| is Lipschitz with respect to y with constant L = 2.

Definition (Convex sets)

A set $D \in \mathbb{R}^2$ is **convex** if whenever $(t_1, y_1), (t_2, y_2) \in D$ there is $(1 - \lambda)(t_1, y_1) + \lambda(t_2, y_2) \in D$ for all $\lambda \in [0, 1]$.



Theorem

If $D \subset \mathbb{R}^2$ is convex, and $\left|\frac{\partial f}{\partial y}(t, y)\right| \leq L$ for all $(t, y) \in D$, then f is Lipschitz with respect to y with constant L.

Remark

This is a sufficient (but not necessary) condition for f to be Lipschitz with respect to y.

Proof. For any $(t, y_1), (t, y_2) \in D$, define function g by $g(\lambda) = f(t, (1-\lambda)y_1 + \lambda y_2)$ for $\lambda \in [0, 1]$ (need convexity of D!). Then we have $g'(\lambda) = \partial_{y} f(t, (1-\lambda)y_1 + \lambda y_2) \cdot (y_2 - y_1)$ So $|g'(\lambda)| < L|y_2 - y_1|$. Then we have $|g(1) - g(0)| = \left| \int_{0}^{1} g'(\lambda) d\lambda \right| \le L|y_2 - y_1| \left| \int_{0}^{1} d\lambda \right| = L|y_2 - y_1|$

Note that $g(0) = f(t, y_1)$ and $g(1) = f(t, y_2)$. This completes the proof.

Theorem

Suppose $D = [a, b] \times \mathbb{R}$, a function f is continuous on D and Lipschitz with respect to y, then the initial value problem y' = f(t, y) for $t \in [a, b]$ with initial value $y(a) = \alpha$ has a unique solution y(t) for $t \in [a, b]$.

Remark

This theorem says that there must be one and only one solution of the IVP, provided that the defining f of the IVP is continuous and Lipschitz with respect to y on D.

Example

Show that $y' = 1 + t \sin(ty)$ for $t \in [0, 2]$ with y(0) = 0 has a unique solution.

Solution: First, we know $f(t, y) = 1 + t \sin(ty)$ is continuous on $[0, 2] \times \mathbb{R}$. Second, we can see

$$\left|\frac{\partial f}{\partial y}\right| = \left|t^2 \cos(ty)\right| \le |t^2| \le 4$$

So f(t, y) is Lipschitz with respect to y (with constant 4). From theorem above, we know the IVP has a unique solution y(t) on [0, 2].

Definition (Well-posedness)

An IVP y' = f(t, y) for $t \in [a, b]$ with $y(a) = \alpha$ is called **well-posed** if

- lt has a unique solution y(t);
- There exist ε₀ > 0 and k > 0, such that ∀ε ∈ (0, ε₀) and function δ(t), which is continuous and satisfies |δ(t)| < ε for all t ∈ [a, b], the perturbed problem z' = f(t, z) + δ(t) with initial value z(a) = α + δ₀ (where |δ₀| ≤ ε) satisfies

$$|z(t) - y(t)| < k\epsilon, \quad \forall t \in [a, b].$$

Remark

This theorem says that a small perturbation on defining function f by $\delta(t)$ and initial value y(a) by δ_0 will only cause small change to original solution y(t).

Theorem

Let $D = [a, b] \times \mathbb{R}$. If f is continuous on D and Lipschitz with respect to y, then the IVP is well-posed.

Remark

Again, a sufficient but not necessary condition for well-posedness of IVP.

Given an IVP y' = f(t, y) for $t \in [a, b]$ and $y(a) = \alpha$, we want to compute y(t) on **mesh points** $\{t_0, t_1, \ldots, t_N\}$ on [a, b].

To this end, we partition [a, b] into N equal segments: set $h = \frac{b-a}{N}$, and define $t_i = a + ih$ for i = 0, 1, ..., N. Here h is called the **step size**.



From Taylor's theorem, we have

$$y(t_{i+1}) = y(t_i) + y'(t_i)(t_{i+1} - t_i) + \frac{1}{2}y''(\xi_i)(t_{i+1} - t_i)^2$$

for some $\xi_i \in (t_i, t_{i+1})$. Note that $t_{i+1} - t_i = h$ and $y'(t_i) = f(t_i, y(t_i))$, we get

$$y(t_{i+1}) \approx y(t_i) + hf(t, y(t_i))$$

Denote $w_i = y(t_i)$ for all i = 0, 1, ..., N, we get the **Euler's** method:

$$\begin{cases} w_0 = \alpha \\ w_{i+1} = w_i + hf(t_i, w_i), \quad i = 0, 1, \dots, N-1 \end{cases}$$



Example

Use Euler's method with h = 0.5 for IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ with initial value y(0) = 0.5.

Solution: We follow Euler's method step-by-step:

$$t_{0} = 0: \quad w_{0} = y(0) = 0.5$$

$$t_{1} = 0.5: \quad w_{1} = w_{0} + hf(t_{0}, w_{0}) = 0.5 + 0.5 \times (0.5 - 0^{2} + 1) = 1.25$$

$$t_{2} = 1.0: \quad w_{2} = w_{1} + hf(t_{1}, w_{1}) = 1.25 + 0.5 \times (1.25 - 0.5^{2} + 1) = 2.25$$

$$t_{3} = 1.5: \quad w_{3} = w_{2} + hf(t_{2}, w_{2}) = 2.25 + 0.5 \times (2.25 - 1^{2} + 1) = 3.375$$

$$t_{4} = 2.0: \quad w_{4} = w_{3} + hf(t_{3}, w_{3}) = 3.375 + 0.5 \times (3.375 - 1.5^{2} + 1) = 4.4375$$

Theorem

Suppose f(t, y) in an IVP is continuous on $D = [a, b] \times \mathbb{R}$ and Lipschitz with respect to y with constant L. If $\exists M > 0$ such that $|y''(t)| \leq M$ (y(t) is the unique solution of the IVP), then for all i = 0, 1, ..., N there is

$$|y(t_i) - w_i| \leq \frac{hM}{2L} \left(e^{L(t_i-a)} - 1\right)$$

Remark

- Numerical error depends on h (also called O(h) error).
- Also depends on M, L of f.
- Error increases for larger t_i.

Proof. Taking the difference of

$$y(t_{i+1}) = y(t_i) + hf(t_i, y_i) + \frac{1}{2}y''(\xi_i)(t_{i+1} - t_i)^2$$

$$w_{i+1} = w_i + hf(t_i, w_i)$$

we get

$$\begin{split} |y(t_{i+1}) - w_{i+1}| &\leq |y(t_i) - w_i| + h|f(t_i, y_i) - f(t_i, w_i)| + \frac{Mh^2}{2} \\ &\leq |y(t_i) - w_i| + hL|y_i - w_i| + \frac{Mh^2}{2} \\ &= (1 + hL)|y_i - w_i| + \frac{Mh^2}{2} \end{split}$$

Proof (cont). Denote $d_i = |y(t_i) - w_i|$, then we have

$$d_{i+1} \leq (1+hL)d_i + \frac{Mh^2}{2} = (1+hL)\left(d_i + \frac{hM}{2L}\right) - \frac{hM}{2L}$$

for all $i = 0, 1, \dots, N - 1$. So we obtain

$$\begin{aligned} d_{i+1} + \frac{hM}{2L} &\leq (1+hL) \left(d_i + \frac{hM}{2L} \right) \\ &\leq (1+hL)^2 \left(d_{i-1} + \frac{hM}{2L} \right) \\ &\leq \cdots \\ &\leq (1+hL)^{i+1} \left(d_0 + \frac{hM}{2L} \right) \end{aligned}$$

and hence $d_i \leq (1 + hL)^i \cdot \frac{hM}{2L} - \frac{hM}{2L}$ (since $d_0 = 0$).

Proof (cont).

Note that $1 + x \le e^x$ for all x > -1, and hence $(1 + x)^a \le e^{ax}$ if a > 0. Based on this, we know $(1 + hL)^i \le e^{ihL} = e^{L(t_i - a)}$ since $ih = t_i - a$. Therefore we get

$$d_i \leq e^{L(t_i-a)} \cdot \frac{hM}{2L} - \frac{hM}{2L} = \frac{hM}{2L}(e^{L(t_i-a)} - 1)$$

This completes the proof.

Example

Estimate the error of Euler's method with h = 0.2 for IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ with initial value y(0) = 0.5.

Solution: We first note that $\frac{\partial f}{\partial y} = 1$, so f is Lipschitz with respect to y with constant L = 1. The IVP has solution $y(t) = (t-1)^2 - \frac{e^t}{2}$ so $|y''(t)| = |\frac{e^t}{2} - 2| \le \frac{e^2}{2} - 2 =: M$. By theorem above, the error of Euler's method is

$$|y(t_i) - w_i| \leq \frac{hM}{2L} \left(e^{L(t_i - a)} - 1 \right) = \frac{0.2(0.5e^2 - 2)}{2} \left(e^{t_i} - 1 \right)$$

Example

Estimate the error of Euler's method with h = 0.2 for IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ with initial value y(0) = 0.5. Solution: (cont)

w_i	$y_i = y(t_i)$	$ y_i - w_i $
0.5000000	0.5000000	0.0000000
0.8000000	0.8292986	0.0292986
1.1520000	1.2140877	0.0620877
1.5504000	1.6489406	0.0985406
1.9884800	2.1272295	0.1387495
2.4581760	2.6408591	0.1826831
2.9498112	3.1799415	0.2301303
3.4517734	3.7324000	0.2806266
3.9501281	4.2834838	0.3333557
4.4281538	4.8151763	0.3870225
4.8657845	5.3054720	0.4396874
	$\begin{array}{c} w_i \\ \hline 0.5000000 \\ 0.8000000 \\ 1.1520000 \\ 1.5504000 \\ 1.984800 \\ 2.4581760 \\ 2.9498112 \\ 3.4517734 \\ 3.9501281 \\ 4.4281538 \\ 4.8657845 \end{array}$	$\begin{array}{c cccc} w_i & y_i = y(t_i) \\ \hline 0.5000000 & 0.5000000 \\ 0.8000000 & 0.8292986 \\ 1.1520000 & 1.2140877 \\ 1.5504000 & 1.6489406 \\ 1.9884800 & 2.1272295 \\ 2.4581760 & 2.6408591 \\ 2.9498112 & 3.1799415 \\ 3.4517734 & 3.7324000 \\ 3.9501281 & 4.2834838 \\ 4.4281538 & 4.8151763 \\ 4.8657845 & 5.3054720 \\ \hline \end{array}$

Round-off error of Euler's method

Due to round-off errors in computer, we instead obtain

$$\begin{cases} u_0 = \alpha + \delta_0 \\ u_{i+1} = u_i + hf(t_i, u_i) + \delta_i, \quad i = 0, 1, \dots, N-1 \end{cases}$$

Suppose $\exists \delta > 0$ such that $|\delta_i| \leq \delta$ for all *i*, then we can show

$$|y(t_i) - u_i| \leq \frac{1}{L} \left(\frac{hM}{2} + \frac{\delta}{h}\right) \left(e^{L(t_i - a)} - 1\right) + \delta e^{L(t_i - a)}.$$

Note that $\frac{hM}{2} + \frac{\delta}{h}$ does not approach 0 as $h \to 0$. $\frac{hM}{2} + \frac{\delta}{h}$ reaches minimum at $h = \sqrt{\frac{2\delta}{M}}$ (often much smaller than what we choose in practice).

Definition (Local truncation error) We call the difference method

$$\begin{cases} w_0 = \alpha \\ w_{i+1} = w_i + h\phi(t_i, w_i), \quad i = 0, 1, \dots, N-1 \end{cases}$$

to have local truncation error

$$au_{i+1}(h) = rac{y_{i+1} - (y_i + h\phi(t_i, y_i))}{h}$$

where $y_i := y(t_i)$.

Example

Euler's method has local truncation error

$$\tau_{i+1}(h) = \frac{y_{i+1} - (y_i + hf(t_i, y_i))}{h} = \frac{y_{i+1} - y_i}{h} - f(t_i, y_i)$$

Note that Euler's method has local truncation error $\tau_{i+1}(h) = \frac{y_{i+1}-y_i}{h} - f(t_i, y_i) = \frac{hy''(\xi_i)}{2}$ for some $\xi_i \in (t_i, t_{i+1})$. If $|y''| \le M$ we know $|\tau_{i+1}(h)| \le \frac{hM}{2} = O(h)$.

Question: What if we use higher-order Taylor's approximation?

$$y(t_{i+1}) = y(t_i) + hy'(t_i) + \frac{h^2}{2}y''(t_i) + \dots + \frac{h^n}{n!}y^{(n)}(t_i) + R$$

where $R = \frac{h^{n+1}}{(n+1)!} y^{(n+1)}(\xi_i)$ for some $\xi_i \in (t_i, t_{i+1})$.

First note that we can always write $y^{(n)}$ using f(t, y(t)):

$$y'(t) = f$$

$$y''(t) = f' = \partial_t f + (\partial_y f) f$$

$$y'''(t) = f'' = \partial_t^2 f + (\partial_t \partial_y f + (\partial_y^2 f) f) f + \partial_y f (\partial_t f + (\partial_y f) f)$$

...

$$y^{(n)}(t) = f^{(n-1)} = \cdots$$

albeit it's quickly getting very complicated.

Now substitute them back to high-order Taylor's approximation (ignore residual R)

$$y(t_{i+1}) = y(t_i) + hy'(t_i) + \frac{h^2}{2}y''(t_i) + \dots + \frac{h^n}{n!}y^{(n)}(t_i)$$

= $y(t_i) + hf + \frac{h^2}{2}f' + \dots + \frac{h^n}{n!}f^{(n-1)}$

We can get the *n*-th order Taylor's method:

$$\begin{cases} w_0 = \alpha \\ w_{i+1} = w_i + hT^{(n)}(t_i, w_i), \quad i = 0, 1, \dots, N-1 \end{cases}$$

where

$$T^{(n)}(t_i, w_i) = f(t_i, w_i) + \frac{h}{2}f'(t_i, w_i) + \cdots + \frac{h^{n-1}}{n!}f^{(n-1)}(t_i, w_i)$$

- Euler's method is the first order Taylor's method.
- High-order Taylor's method is more accurate than Euler's method, but at much higher computational cost.
- Together with Hermite interpolating polynomials, it can be used to interpolate values not on mesh points more accurately.

Theorem If $y(t) \in C^{n+1}[a, b]$, then the n-th order Taylor method has local truncation error $O(h^n)$.

Runge-Kutta (RK) method attains high-order local truncation error **without** expensive evaluations of derivatives of f.

To derive RK method, first recall Taylor's formula for two variables (t, y):

$$f(t,y) = P_n(t,y) + R_n(t,y)$$

where $\partial_t^{n-k} \partial_y^k f = \frac{\partial^n f(t_0, y_0)}{\partial t^{n-k} \partial y^k}$ and

$$\begin{aligned} P_n(t,y) &= f(t_0,y_0) + (\partial_t f \cdot (t-t_0) + \partial_y f \cdot (y-y_0)) \\ &+ \frac{1}{2} \left(\partial_t^2 f \cdot (t-t_0)^2 + 2\partial_y \partial_t f \cdot (t-t_0)(y-y_0) + \partial_y^2 f \cdot (y-y_0)^2 \right) \\ &+ \dots + \frac{1}{n!} \sum_{k=0}^n \binom{n}{k} \partial_t^{n-k} \partial_y^k f \cdot (t-t_0)^{n-k} (y-y_0)^k \\ R_n(t,y) &= \frac{1}{(n+1)!} \sum_{k=0}^{n+1} \binom{n+1}{k} \partial_t^{n+1-k} \partial_y^k f(\xi,\mu) \cdot (t-t_0)^{n+1-k} (y-y_0)^k \end{aligned}$$

The second order Taylor's method uses

$$T^{(2)}(t,y) = f(t,y) + \frac{h}{2}f'(t,y) = f(t,y) + \frac{h}{2}(\partial_t f + \partial_y f \cdot f)$$

to get $O(h^2)$ error. Suppose we use $af(t + \alpha, y + \beta)$ (with some

 a, α, β to be determined) to reach the same order of error. To that end, we first have

$$af(t + \alpha, y + \beta) = a\left(f + \partial_t f \cdot \alpha + \partial_y f \cdot \beta + R\right)$$

where $R = \frac{1}{2}(\partial_t^2 f(\xi, \mu) \cdot \alpha^2 + 2\partial_y \partial_t f(\xi, \mu) \cdot \alpha\beta + \partial_y^2 f(\xi, \mu) \cdot \beta^2).$

Suppose we try to match the terms of these two formulas (ignore R):

$$T^{(2)}(t,y) = f + \frac{h}{2}\partial_t f + \frac{hf}{2}\partial_y f$$
$$af(t+\alpha, y+\beta) = af + a\alpha\partial_t f + a\beta\partial_y f$$

then we have

$$a = 1, \quad \alpha = \frac{h}{2}, \quad \beta = \frac{h}{2}f(t, y)$$

So instead of $T^{(2)}(t, y)$, we use

$$af(t+lpha,y+eta)=f\left(t+rac{h}{2},y+rac{h}{2}f(t,y)
ight)$$
Note that R we ignored is

$$R = \frac{1}{2} \left(\partial_t^2 f(\xi, \mu) \cdot \left(\frac{h}{2}\right)^2 + 2 \partial_y \partial_t f(\xi, \mu) \cdot \left(\frac{h}{2}\right)^2 f(\xi, \mu) \cdot \left(\frac{h}{2}\right)^2 f^2 \right)$$

which means $R = O(h^2)$.

Also note that

$$R = T^{(2)}(t, y) - f\left(t + \frac{h}{2}, y + \frac{h}{2}f(t, y)\right) = O(h^2)$$

and the error of $T^{(2)}(t, y)$ is of $O(h^2)$, we know $f\left(t + \frac{h}{2}, y + \frac{h}{2}f(t, y)\right)$ has error of $O(h^2)$.

This is the **RK2 method (Midpoint method)**:

$$\begin{cases} w_0 = \alpha \\ w_{i+1} = w_i + h \ f\left(t_i + \frac{h}{2}, w_i + \frac{h}{2}f(t_i, w_i)\right), \quad i = 0, 1, \dots, N-1. \end{cases}$$

Remark

If we have (t_i, w_i) , we only need to evaluate f twice (i.e., compute $k_1 = f(t_i, w_i)$ and $k_2 = f(t_i + \frac{h}{2}, w_i + \frac{h}{2}k_1)$) to get w_{i+1} .

We can also consider higher-order RK method by fitting

$$T^{(3)}(t,y) = f(t,y) + rac{h}{2}f'(t,y) + rac{h}{6}f''(t,y)$$

with $af(t, y) + bf(t + \alpha, y + \beta)$ (has 4 parameters a, b, α, β).

Unfortunately we can't make match to the $\frac{hf''}{6}$ term of $T^{(3)}$, which contains $\frac{h^2}{6}f \cdot (\partial_y f)^2$, by this way. But it leaves us open choices if we're OK with $O(h^2)$ error: let a = b = 1, $\alpha = h$, $\beta = hf(t, y)$, then we get the **modified Euler's method**:

$$\begin{cases} w_0 = \alpha \\ w_{i+1} = w_i + \frac{h}{2} \left(f(t_i, w_i) + f(t_{i+1}, w_i + hf(t_i, w_i)) \right), \ i = 0, 1, \dots, N-1. \end{cases}$$

Also need evaluation of f twice in each step.

Example

Use Midpoint method (RK2) and Modified Euler's method with h = 0.2 to solve IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ and y(0) = 0.5.

Solution: Apply the main steps in the two methods:

Midpoint:
$$w_{i+1} = w_i + h f\left(t_i + \frac{h}{2}, w_i + \frac{h}{2}f(t_i, w_i)\right)$$

Modified Euler's: $w_{i+1} = w_i + \frac{h}{2}\left(f(t_i, w_i) + f(t_{i+1}, w_i + hf(t_i, w_i))\right)$

Example

Use Midpoint method (RK2) and Modified Euler's method with h = 0.2 to solve IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ and y(0) = 0.5. Solution: (cont)

rror
32986
71677
16982
69938
31715
03627
87138
83866
95577
24173
2

Midpoint (RK2) method is better than modified Euler's method.

We can also consider higher-order RK method by fitting

$$T^{(3)}(t,y) = f(t,y) + \frac{h}{2}f'(t,y) + \frac{h^2}{6}f''(t,y)$$

with $af(t, y) + bf(t + \alpha_1, y + \delta_1(f(t + \alpha_2, y + \delta_2 f(t, y)))$ (has 6 parameters $a, b, \alpha_1, \alpha_2, \delta_1, \delta_2$) to reach $O(h^3)$ error.

For example, Heun's choice is $a = \frac{1}{4}$, $b = \frac{3}{4}$, $\alpha_1 = \frac{2h}{3}$, $\alpha_2 = \frac{h}{3}$, $\delta_1 = \frac{2h}{3}f$, $\delta_2 = \frac{h}{3}f$.

Nevertheless, methods of order $O(h^3)$ are rarely used in practice.

4-th Order Runge-Kutta (RK4) method

Most commonly used is the **4-th order Runge-Kutta method** (**RK4**): start with $w_0 = \alpha$, and iteratively do

$$\begin{cases} k_1 = f(t_i, w_i) \\ k_2 = f(t_i + \frac{h}{2}, w_i + \frac{h}{2}k_1) \\ k_3 = f(t_i + \frac{h}{2}, w_i + \frac{h}{2}k_2) \\ k_4 = f(t_{i+1}, w_i + hk_3) \\ w_{i+1} = w_i + \frac{h}{6}(k_1 + 2k_2 + 2k_3 + k_4) \end{cases}$$

Need to evaluate f for 4 times in each step. Reach error $O(h^4)$.

4-th Order Runge-Kutta (RK4) method

Example

Use RK4 (with h = 0.2) to solve IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ and y(0) = 0.5.

Solution: With h = 0.2, we have N = 10 and $t_i = 0.2i$ for i = 0, 1, ..., 10. First set $w_0 = 0.5$, then the first iteration is

$$k_{1} = f(t_{0}, w_{0}) = f(0, 0.5) = 0.5 - 0^{2} + 1 = 1.5$$

$$k_{2} = f(t_{0} + \frac{h}{2}, w_{0} + \frac{h}{2}k_{1}) = f(0.1, 0.5 + 0.1 \times 1.5) = 1.64$$

$$k_{3} = f(t_{0} + \frac{h}{2}, w_{0} + \frac{h}{2}k_{2}) = f(0.1, 0.5 + 0.1 \times 1.64) = 1.654$$

$$k_{4} = f(t_{1}, w_{0} + hk_{3}) = f(0.2, 0.5 + 0.2 \times 1.654) = 1.7908$$

$$w_{1} = w_{0} + \frac{h}{6}(k_{1} + 2k_{2} + 2k_{3} + k_{4}) = 0.8292933$$

So w_1 is our RK4 approximation of $y(t_1) = y(0.2)$. Numerical Analysis II – Xiaojing Ye, Math & Stat, Georgia State University

4-th Order Runge-Kutta (RK4) method

Example

Use RK4 (with h = 0.2) to solve IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ and y(0) = 0.5.

Solution: (cont) Continue with $i = 1, 2, \dots, 9$:

	Runge-Kutta					
	Exact	Order Four	Error			
t_i	$y_i = y(t_i)$	w_i	$ y_i - w_i $			
0.0	0.5000000	0.5000000	0			
0.2	0.8292986	0.8292933	0.0000053			
0.4	1.2140877	1.2140762	0.0000114			
0.6	1.6489406	1.6489220	0.0000186			
0.8	2.1272295	2.1272027	0.0000269			
1.0	2.6408591	2.6408227	0.0000364			
1.2	3.1799415	3.1798942	0.0000474			
1.4	3.7324000	3.7323401	0.0000599			
1.6	4.2834838	4.2834095	0.0000743			
1.8	4.8151763	4.8150857	0.0000906			
2.0	5.3054720	5.3053630	0.0001089			

High-order Runge-Kutta method

Can we use even higher-order method to improve accuracy?

#f eval	2	3	4	$5 \le n \le 7$	$8 \le n \le 9$	$n \ge 10$	So
Best error	$O(h^2)$	$O(h^3)$	$O(h^4)$	$O(h^{n-1})$	$O(h^{n-2})$	$O(h^{n-3})$	50

RK4 is the sweet spot.

Remark

Note that RK4 requires 4 evaluations of f each step. So it would make sense only if it's accuracy with step size 4h is higher than Midpoint with 2h or Euler's with h!

High-order Runge-Kutta method

Example

Use RK4 (with h = 0.1), Midpoint (with h = 0.05), and Euler's method (with h = 0.025) to solve IVP $y' = y - t^2 + 1$ for $t \in [0, 0.5]$ and y(0) = 0.5.

Solution:

t_i	Exact	Euler $h = 0.025$	Modified Euler h = 0.05	Runge-Kutta Order Four h = 0.1
0.0	0.5000000	0.5000000	0.5000000	0.5000000
0.1	0.6574145	0.6554982	0.6573085	0.6574144
0.2	0.8292986	0.8253385	0.8290778	0.8292983
0.3	1.0150706	1.0089334	1.0147254	1.0150701
0.4	1.2140877	1.2056345	1.2136079	1.2140869
0.5	1.4256394	1.4147264	1.4250141	1.4256384

RK4 is better with same computation cost!

Error control

Can we control the error of Runge-Kutta method by using variable step sizes?

Let's compare two difference methods with errors $O(h^n)$ and $O(h^{n+1})$ (say, RK4 and RK5) for fixed step size h, which have schemes below:

$$w_{i+1} = w_i + h\phi(t_i, w_i, h) \qquad O(h^n)$$

$$\tilde{w}_{i+1} = \tilde{w}_i + h\tilde{\phi}(t_i, \tilde{w}_i, h) \qquad O(h^{n+1})$$

Suppose $w_i \approx \tilde{w}_i \approx y(t_i) =: y_i$. Then for any given $\epsilon > 0$, we want to see how small h should be for the $O(h^n)$ method so that its error $|\tau_{i+1}(h)| \leq \epsilon$?

Error control

We recall that the local truncation errors of these two methods are:

$$\tau_{i+1}(h) = \frac{y_{i+1} - y_i}{h} - \phi(t_i, y_i, h) \approx O(h^n)$$

$$\tilde{\tau}_{i+1}(h) = \frac{y_{i+1} - y_i}{h} - \tilde{\phi}(t_i, y_i, h) \approx O(h^{n+1})$$

Given that $w_i \approx \tilde{w}_i \approx y_i$ and $O(h^{n+1}) \ll O(h^n)$ for small h, we see

$$egin{aligned} & au_{i+1}(h) pprox au_{i+1}(h) = ilde{\phi}(t_i, y_i, h) - \phi(t_i, y_i, h) \ &pprox ilde{\phi}(t_i, ilde{w}_i, h) - \phi(t_i, w_i, h) = rac{ ilde{w}_{i+1} - ilde{w}_i}{h} - rac{w_{i+1} - w_i}{h} \ &pprox rac{ ilde{w}_{i+1} - w_{i+1}}{h} pprox extsf{Kh}^n \end{aligned}$$

for some K > 0 independent of h, since $\tau_{i+1}(h) \approx O(h^n)$.

Error control

Suppose that we can scale h by q > 0, such that

$$| au_{i+1}(qh)| pprox \mathcal{K}(qh)^n = q^n \mathcal{K}h^n pprox q^n rac{| ilde w_{i+1} - w_{i+1}|}{h} \leq \epsilon$$

So we need q to satisfy

$$q \leq \left(\frac{\epsilon h}{|\tilde{w}_{i+1} - w_{i+1}|}\right)^{1/n}$$

q < 1: reject the initial h and recalculate using qh.
q ≥ 1: accept computed value and use qh for next step.

The **Runge-Kutta-Fehlberg (RKF) method** uses specific 4th-order and 5th-order RK schemes, which share some computed values and together only need 6 evaluation of f, to estimate

$$q = \left(rac{\epsilon h}{2| ilde{w}_{i+1} - w_{i+1}|}
ight)^{1/4} = 0.84 \left(rac{\epsilon h}{| ilde{w}_{i+1} - w_{i+1}|}
ight)^{1/4}$$

This q is used to tune step size so that error is always bounded by the prescribed ϵ .

Multistep method

Definition

Let m > 1 be an integer, then an m-step multistep method is given by the form of

$$w_{i+1} = a_{m-1}w_i + a_{m-2}w_{i-1} + \dots + a_0w_{i-m+1} + h \left[b_m f(t_{i+1}, w_{i+1}) + b_{m-1}f(t_i, w_i) + \dots + b_0 f(t_{i-m+1}, w_{i-m+1}) \right]$$

for
$$i = m - 1, m, \dots, N - 1$$
.

Here
$$a_0, \ldots, a_{m-1}, b_0, \ldots, b_m$$
 are constants. Also
 $w_0 = \alpha, w_1 = \alpha_1, \ldots, w_{m-1} = \alpha_{m-1}$ need to be given.
 $b_m = 0$: Explicit m-step method.
 $b_m \neq 0$: Implicit m-step method.

Definition

The **local truncation error** of the *m*-step multistep method above is defined by

$$\tau_{i+1}(h) = \frac{y_{i+1} - (a_{m-1}y_i + \dots + a_0y_{i-m+1})}{h} - [b_m f(t_{i+1}, y_{i+1}) + b_{m-1}f(t_i, y_i) + \dots + b_0 f(t_{i-m+1}, y_{i-m+1})]$$
where $y_i := y(t_i)$.

Adams-Bashforth Two-Step Explicit method:

$$\begin{cases} w_0 = \alpha, & w_1 = \alpha_1, \\ w_{i+1} = w_i + \frac{h}{2} \left[3f(t_i, w_i) - f(t_{i-1}, w_{i-1}) \right] \end{cases}$$

for i = 1, ..., N - 1.

The local truncation error is

$$au_{i+1}(h) = rac{5}{12} y'''(\mu_i) h^2$$

for some $\mu_i \in (t_{i-1}, t_{i+1})$.

Adams-Bashforth Explicit method: local truncation error

We denote $y_i^{(k)} := y^{(k)}(t_i)$ for short. If $w_j = y_j$ for $j \leq i$, then

$$y_{i+1} = y_i + hy'_i + \frac{h^2}{2}y''_i + \frac{h^3}{6}y'''(\xi_i), \quad \xi_i \in (t_i, t_{i+1})$$
$$w_{i+1} = y_i + hy'_i + \frac{h}{2}(y'_i - y'_{i-1}),$$
$$y'_{i-1} = y'_i - hy''_i + \frac{h^2}{2}y'''(\eta_i), \quad \eta_i \in (t_{i-1}, t_i)$$

Plugging the equations above into the formula of local truncation error:

$$\tau_{i+1}(h) = \frac{y_{i+1} - w_{i+1}}{h} = \left(\frac{1}{6}y^{\prime\prime\prime}(\xi_i) + \frac{1}{4}y^{\prime\prime\prime}(\eta_i)\right)h^2 = \frac{5}{12}y^{\prime\prime\prime}(\mu_i)h^2$$

for some $\mu_i \in (t_{i-1}, t_{i+1})$, where in the last equality we used the intermediate value theorem and $y \in C^3$ (so y''' is continuous) to obtain $y'''(\mu_i) = \frac{\frac{1}{6}y'''(\xi_i) + \frac{1}{4}y'''(\eta_i)}{\frac{1}{6} + \frac{1}{4}}$ which is between $y'''(\xi_i)$ and $y'''(\eta_i)$.

Adams-Bashforth Three-Step Explicit method:

$$\begin{cases} w_0 = \alpha, \quad w_1 = \alpha_1, \quad w_2 = \alpha_2, \\ w_{i+1} = w_i + \frac{h}{12} \left[23f(t_i, w_i) - 16f(t_{i-1}, w_{i-1}) + 5f(t_{i-2}, w_{i-2}) \right] \end{cases}$$

for i = 2, ..., N - 1.

The local truncation error is

$$au_{i+1}(h) = rac{3}{8} y^{(4)}(\mu_i) h^3$$

for some $\mu_i \in (t_{i-2}, t_{i+1})$.

Adams-Bashforth Four-Step Explicit method:

$$\begin{cases} w_0 = \alpha, & w_1 = \alpha_1, & w_2 = \alpha_2, & w_3 = \alpha_3 \\ w_{i+1} = w_i + \frac{h}{24} \left[55f(t_i, w_i) - 59f(t_{i-1}, w_{i-1}) + 37f(t_{i-2}, w_{i-2}) - 9f(t_{i-3}, w_{i-3}) \right] \end{cases}$$

for
$$i = 3, ..., N - 1$$
.

The local truncation error is

$$\tau_{i+1}(h) = \frac{251}{720} y^{(5)}(\mu_i) h^4$$

for some $\mu_i \in (t_{i-3}, t_{i+1})$.

Adams-Bashforth Five-Step Explicit method:

$$\begin{cases} w_0 = \alpha, \quad w_1 = \alpha_1, \quad w_2 = \alpha_2, \quad w_3 = \alpha_3, \quad w_4 = \alpha_4 \\ w_{i+1} = w_i + \frac{h}{720} [1901f(t_i, w_i) - 2774f(t_{i-1}, w_{i-1}) + 2616f(t_{i-2}, w_{i-2}) \\ - 1274f(t_{i-3}, w_{i-3}) + 251f(t_{i-4}, w_{i-4})] \end{cases}$$

for i = 4, ..., N - 1.

The local truncation error is

$$\tau_{i+1}(h) = \frac{95}{288} y^{(6)}(\mu_i) h^5$$

for some $\mu_i \in (t_{i-4}, t_{i+1})$.

Adams-Moulton Implicit method

Adams-Moulton Two-Step Implicit method:

$$\begin{cases} w_0 = \alpha, & w_1 = \alpha_1, \\ w_{i+1} = w_i + \frac{h}{12} [5f(t_{i+1}, w_{i+1}) + 8f(t_i, w_i) - f(t_{i-1}, w_{i-1})] \end{cases}$$

for i = 1, ..., N - 1.

The local truncation error is

$$au_{i+1}(h) = -rac{1}{24} y^{(4)}(\mu_i) h^3$$

for some $\mu_i \in (t_{i-1}, t_{i+1})$.

Adams-Moulton Implicit method

Adams-Moulton Three-Step Implicit method:

$$\begin{cases} w_0 = \alpha, \quad w_1 = \alpha_1, \quad w_2 = \alpha_2 \\ w_{i+1} = w_i + \frac{h}{24} [9f(t_{i+1}, w_{i+1}) + 19f(t_i, w_i) - 5f(t_{i-1}, w_{i-1}) + f(t_{i-2}, w_{i-2})] \end{cases}$$

for
$$i = 2, ..., N - 1$$
.

The local truncation error is

$$au_{i+1}(h) = -rac{19}{720}y^{(5)}(\mu_i)h^4$$

for some $\mu_i \in (t_{i-2}, t_{i+1})$.

Adams-Moulton Implicit method

Adams-Moulton Four-Step Implicit method:

$$\begin{cases} w_0 = \alpha, & w_1 = \alpha_1, & w_2 = \alpha_2, & w_3 = \alpha_3 \\ w_{i+1} = w_i + \frac{h}{720} [251f(t_{i+1}, w_{i+1}) + 646f(t_i, w_i) - 264f(t_{i-1}, w_{i-1}) \\ &+ 106f(t_{i-2}, w_{i-2}) - 19f(t_{i-3}, w_{i-3})] \end{cases}$$

for i = 3, ..., N - 1.

The local truncation error is

$$au_{i+1}(h) = -rac{3}{160}y^{(6)}(\mu_i)h^5$$

for some $\mu_i \in (t_{i-3}, t_{i+1})$.

Steps to develop multistep methods

Construct interpolating polynomial P(t) (e.g., Newton's backward difference method) using previously computed (t_{i-m+1}, w_{i-m+1}), ..., (t_i, w_i).

• Approximate $y(t_{i+1})$ based on

$$y(t_{i+1}) = y(t_i) + \int_{t_i}^{t_{i+1}} y'(t) dt = y(t_i) + \int_{t_i}^{t_{i+1}} f(t, y(t)) dt$$
$$\approx y(t_i) + \int_{t_i}^{t_{i+1}} f(t, P(t)) dt$$

and construct difference method:

$$w_{i+1} = w_i + h\phi(t_i, \ldots, t_{i-m+1}, w_i, \ldots, w_{i-m+1})$$

Explicit vs. Implicit

- Implicit methods are generally more accurate than the explicit ones (e.g., Adams-Moulton three-step implicit method is even more accurate than Adams-Bashforth four-step explicit method).
- Implicit methods require solving for w_{i+1} from

$$w_{i+1} = \cdots + \frac{h}{xxx}f(t_{i+1}, w_{i+1}) + \cdots$$

which can be difficult or even impossible.

There could be multiple solutions of w_{i+1} when solving the equation above in implicit methods.

Due to the aforementioned issues, implicit methods are often cast in "predictor-corrector" form in practice.

In each step *i*:

Prediction: Compute w_{i+1} using an explicit method \u03c6 to get w_{i+1,p} using

$$w_{i+1,p} = w_i + h\phi(t_i, w_i, \dots, t_{i-m+1}, w_{i-m+1})$$

• **Correction:** Substitute w_{i+1} by $w_{i+1,p}$ in the implicit method $\tilde{\phi}$ and compute w_{i+1} using

$$w_{i+1} = w_i + h\tilde{\phi}(t_{i+1}, w_{i+1,p}, t_i, w_i, \dots, t_{i-m+1}, w_{i-m+1})$$

Example

Use the Adams-Bashforth 4-step explicit method and Adams-Moulton 3-step implicit method to form the Adams 4th-order Predictor-Corrector method.

With initial value $w_0 = \alpha$, suppose we first generate w_1, w_2, w_3 using RK4 method. Then for i = 3, 4, ..., N - 1:

Use Adams-Bashforth 4-step explicit method to get a predictor w_{i+1,p}:

$$\mathbf{w}_{i+1,p} = w_i + \frac{h}{24} \left[55f(t_i, w_i) - 59f(t_{i-1}, w_{i-1}) + 37f(t_{i-2}, w_{i-2}) - 9f(t_{i-3}, w_{i-3}) \right]$$

Use Adams-Moulton 3-step implicit method to get a corrector w_{i+1}:

$$w_{i+1} = w_i + \frac{h}{24} [9f(t_{i+1}, w_{i+1,p}) + 19f(t_i, w_i) - 5f(t_{i-1}, w_{i-1}) + f(t_{i-2}, w_{i-2})]$$

Example

Use Adams Predictor-Corrector Method with h = 0.2 to solve IVP $y' = y - t^2 + 1$ for $t \in [0, 2]$ and y(0) = 0.5.

			Error
t_i	$y_i = y(t_i)$	w_i	$ y_i - w_i $
0.0	0.5000000	0.5000000	0
0.2	0.8292986	0.8292933	0.0000053
0.4	1.2140877	1.2140762	0.0000114
0.6	1.6489406	1.6489220	0.0000186
0.8	2.1272295	2.1272056	0.0000239
1.0	2.6408591	2.6408286	0.0000305
1.2	3.1799415	3.1799026	0.0000389
1.4	3.7324000	3.7323505	0.0000495
1.6	4.2834838	4.2834208	0.0000630
1.8	4.8151763	4.8150964	0.0000799
2.0	5.3054720	5.3053707	0.0001013

We can also use Milne's 3-step explicit method and Simpson's 2-step implicit method below:

$$w_{i+1,p} = w_{i-3} + \frac{4h}{3} \left[2f(t_i, w_i) - f(t_{i-1}, w_{i-1}) + 2f(t_{i-2}, w_{i-2}) \right]$$

$$w_{i+1} = w_{i-1} + \frac{h}{3} [f(t_{i+1}, w_{i+1,p}) + 4f(t_i, w_i) + f(t_{i-1}, w_{i-1})]$$

This method is $O(h^4)$ and generally has better accuracy than Adams PC method. However it is more likely to be vulnerable to round-off error.

- PC methods have comparable accuracy as RK4, but often require only 2 evaluations of f in each step.
- Need to store values of f for several previous steps.
- Sometimes are more restrictive on step size h, e.g., in the stiff differential equation case later.

Variable step-size multistep method

Now let's take a closer look at the errors of the multistep methods. Denote $y_i := y(t_i)$.

The Adams-Bashforth 4-step explicit method has error

$$\tau_{i+1}(h) = \frac{251}{720} y^{(5)}(\mu_i) h^4$$

The Adams-Moulton 3-step implicit method has error

$$ilde{ au}_{i+1}(h) = -rac{19}{720} y^{(5)}(ilde{\mu}_i) h^4$$

where $\mu_i \in (t_{i-3}, t_{i+1})$ and $\tilde{\mu}_i \in (t_{i-2}, t_{i+1})$.

Question: Can we find a way to scale step size h so the error is under control?

Variable step-size multistep method

Consider the their local truncation errors:

$$y_{i+1} - w_{i+1,p} = rac{251}{720} y^{(5)}(\mu_i) h^5$$

 $y_{i+1} - w_{i+1} = -rac{19}{720} y^{(5)}(\tilde{\mu}_i) h^5$

Assume $y^{(5)}(\mu_i) \approx y^{(5)}(\tilde{\mu}_i)$, we take their difference to get

$$w_{i+1} - w_{i+1,p} = \frac{1}{720}(19 + 251)y^{(5)}(\mu_i)h^5 \approx \frac{3}{8}y^{(5)}(\mu_i)h^5$$

So the error of Adams-Moulton (corrector step) is

$$ilde{ au}_{i+1}(h) = rac{|y_{i+1} - w_{i+1}|}{h} pprox rac{19|w_{i+1} - w_{i+1,p}|}{270h} = Kh^4$$

where K is independent of h since $\tilde{\tau}_{i+1}(h) = O(h^4)$.

Variable step-size multistep method

If we want to keep error under a prescribed ϵ , then we need to find q > 0 such that with step size qh, there is

$$ilde{ au}_{i+1}(qh) = rac{|y(t_i+qh)-w_{i+1}|}{qh} pprox rac{19q^4|w_{i+1}-w_{i+1,p}|}{270h} < \epsilon$$

This implies that

$$q < \left(\frac{270h\epsilon}{19|w_{i+1}-w_{i+1,p}|}\right)^{1/4} \approx 2\left(\frac{h\epsilon}{|w_{i+1}-w_{i+1,p}|}\right)^{1/4}$$

To be conservative, we may replace 2 by 1.5 above.

In practice, we tune q (as less as possible) such that the estimated error is between $(\epsilon/10,\epsilon)$

System of differential equations

The IVP for a system of ODE has form

$$\begin{cases} \frac{\mathrm{d}u_1}{\mathrm{d}t} = f_1(t, u_1, u_2, \dots, u_m) \\ \frac{\mathrm{d}u_2}{\mathrm{d}t} = f_2(t, u_1, u_2, \dots, u_m) \\ \vdots \\ \frac{\mathrm{d}u_m}{\mathrm{d}t} = f_m(t, u_1, u_2, \dots, u_m) \end{cases} \text{ for } a \leq t \leq b \end{cases}$$

with initial value $u_1(a) = \alpha_1, \ldots, u_m(a) = \alpha_m$.

Definition

A set of functions $u_1(t), \ldots, u_m(t)$ is a solution of the IVP above if they satisfy both the system of ODEs and the initial values.
In this case, we will solve for $u_1(t), \ldots, u_m(t)$ which are interdependent according to the ODE system.



Definition

A function f is called **Lipschitz** with respect to u_1, \ldots, u_m on $D := [a, b] \times \mathbb{R}^m$ if there exists L > 0 s.t.

$$|f(t, u_1, \ldots, u_m) - f(t, z_1, \ldots, z_m)| \le L \sum_{j=1}^m |u_j - z_j|$$

for all $(t, u_1, ..., u_m), (t, z_1, ..., z_m) \in D$.

Theorem If $f \in C^1(D)$ and $|\frac{\partial f}{\partial u_j}| \leq L$ for all j, then f is Lipschitz with respect to $u = (u_1, \dots, u_m)$ on D.

Proof.

Note that D is convex. For any $(t, u_1, \ldots, u_m), (t, z_1, \ldots, z_m) \in D$, define

$$g(\lambda) = f(t, (1 - \lambda)u_1 + \lambda z_1, \dots, (1 - \lambda)u_m + \lambda z_m)$$

for all $\lambda \in [0,1]$. Then from $|g(1) - g(0)| \leq \int_0^1 |g'(\lambda)| d\lambda$ and the definition of g, the conclusion follows.

Theorem If $f \in C^1(D)$ and is Lipschitz with respect to $u = (u_1, ..., u_m)$, then the IVP with f as defining function has a unique solution.

Now let's use vector notations below

$$\mathbf{a} = (\alpha_1, \dots, \alpha_m)$$
$$\mathbf{y} = (y_1, \dots, y_m)$$
$$\mathbf{w} = (w_1, \dots, w_m)$$
$$\mathbf{f}(t, \mathbf{w}) = (f_1(t, \mathbf{w}), \dots, f_m(t, \mathbf{w}))$$

Then the IVP of ODE system can be written as

$$\mathbf{y}' = \mathbf{f}(t, \mathbf{y}), \quad t \in [a, b]$$

with initial value $\mathbf{y}(a) = \mathbf{a}$. So the difference methods developed

above, such as RK4, still apply.

Example

Use RK4 (with h = 0.1) to solve IVP for ODE system

$$\begin{cases} l'_1(t) = f_1(t, l_1, l_2) = -4l_1 + 3l_2 + 6\\ l'_2(t) = f_2(t, l_1, l_2) = -2.4l_1 + 1.6l_2 + 3.6 \end{cases}$$

with initial value $I_1(0) = I_2(0) = 0$.

Solution: The exact solution is

$$\begin{cases} l_1(t) = -3.375e^{-2t} + 1.875e^{-0.4t} + 1.5\\ l_2(t) = -2.25e^{-2t} + 2.25e^{-0.4t} \end{cases}$$

for all $t \geq 0$.

Example

Use RK4 (with h = 0.1) to solve IVP for ODE system

$$\begin{cases} l_1'(t) = f_1(t, l_1, l_2) = -4l_1 + 3l_2 + 6\\ l_2'(t) = f_2(t, l_1, l_2) = -2.4l_1 + 1.6l_2 + 3.6 \end{cases}$$

with initial value $l_1(0) = l_2(0) = 0$. Solution: (cont) The result by RK4 is

t_j	$w_{1,j}$	$w_{2,j}$	$ I_1(t_j) - w_{1,j} $	$ I_2(t_j) - w_{2,j} $
0.0	0	0	0	0
0.1	0.5382550	0.3196263	0.8285×10^{-5}	0.5803×10^{-5}
0.2	0.9684983	0.5687817	0.1514×10^{-4}	0.9596×10^{-5}
0.3	1.310717	0.7607328	0.1907×10^{-4}	0.1216×10^{-4}
0.4	1.581263	0.9063208	0.2098×10^{-4}	0.1311×10^{-4}
0.5	1.793505	1.014402	0.2193×10^{-4}	0.1240×10^{-4}

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A general IVP for mth-order ODE is

$$y^{(m)} = f(t, y, y', \dots, y^{(m-1)}), \quad t \in [a, b]$$

with initial value $y(a) = \alpha_1, y'(a) = \alpha_2, \dots, y^{(m-1)}(a) = \alpha_m$.

Definition

A function y(t) is a solution of IVP for the *m*th-order ODE above if y(t) satisfies the differential equation for $t \in [a, b]$ and all initial value conditions at t = a.

We can define a set of functions u_1, \ldots, u_m s.t.

$$u_1(t) = y(t), \quad u_2(t) = y'(t), \quad \dots, \quad u_m(t) = y^{(m-1)}(t)$$

Then we can convert the mth-order ODE to a system of first-order ODEs (total of m coupled ODEs):

$$\begin{cases} u'_{1} = u_{2} \\ u'_{2} = u_{3} \\ \vdots \\ u'_{m} = f(t, u_{1}, u_{2}, \dots, u_{m}) \end{cases} \text{ for } a \leq t \leq b$$

with initial values $u_1(a) = \alpha_1, \ldots, u_m(a) = \alpha_m$.

Example

Use RK4 (with h = 0.1) to solve IVP for ODE system

$$y'' - 2y' + 2y = e^{2t} \sin t, \quad t \in [0, 1]$$

with initial value y(0) = -0.4, y'(0) = -0.6.

Solution: The exact solution is $y(t) = u_1(t) = 0.2e^{2t}(\sin t - 2\cos t)$. Also $u_2(t) = y'(t) = u'_1(t)$ but we don't need it.

Example

Use RK4 (with h = 0.1) to solve IVP for ODE system

$$y'' - 2y' + 2y = e^{2t} \sin t, \quad t \in [0, 1]$$

with initial value y(0) = -0.4, y'(0) = -0.6. Solution: (cont) The result by RK4 is

tj	$y(t_j) = u_1(t_j)$	$w_{1,j}$	$y'(t_j) = u_2(t_j)$	$w_{2,j}$	$ y(t_j) - w_{1,j} $	$ y'(t_j) - w_{2,j} $
0.0	-0.40000000	-0.40000000	-0.6000000	-0.60000000	0	0
0.1	-0.46173297	-0.46173334	-0.6316304	-0.63163124	3.7×10^{-7}	7.75×10^{-7}
0.2	-0.52555905	-0.52555988	-0.6401478	-0.64014895	8.3×10^{-7}	1.01×10^{-6}
0.3	-0.58860005	-0.58860144	-0.6136630	-0.61366381	1.39×10^{-6}	8.34×10^{-7}
0.4	-0.64661028	-0.64661231	-0.5365821	-0.53658203	2.03×10^{-6}	1.79×10^{-7}
0.5	-0.69356395	-0.69356666	-0.3887395	-0.38873810	2.71×10^{-6}	5.96×10^{-7}
0.6	-0.72114849	-0.72115190	-0.1443834	-0.14438087	3.41×10^{-6}	7.75×10^{-7}
0.7	-0.71814890	-0.71815295	0.2289917	0.22899702	4.05×10^{-6}	2.03×10^{-6}
0.8	-0.66970677	-0.66971133	0.7719815	0.77199180	4.56×10^{-6}	5.30×10^{-6}
0.9	-0.55643814	-0.55644290	1.534764	1.5347815	4.76×10^{-6}	9.54×10^{-6}
1.0	-0.35339436	-0.35339886	2.578741	2.5787663	4.50×10^{-6}	1.34×10^{-5}

The difference methods we developed above, e.g., Euler's, midpoints, RK4, multistep explicit/implicit, predictor-corrector methods, are

- based on step-by-step derivation and easy to understand;
- widely used in many practical problems;
- fundamental to more advanced and complex techniques.

Definition (Consistency)

A difference method is called consistent if

$$\lim_{h\to 0} \left(\max_{1\leq i\leq N} \tau_i(h) \right) = 0$$

where $\tau_i(h)$ is the local truncation error of the method.

Remark

Since local truncation error $\tau_i(h)$ is defined assuming previous $w_i = y_i$, it does not take error accumulation into account. So the consistency definition above only considers how good $\phi(t, w_i, h)$ in the difference method is.

For any step size h > 0, the difference method $w_{i+1} = w_i + h\phi(t_i, w_i, h)$ can generate a sequence of w_i which depend on h. We call them $\{w_i(h)\}_i$. Note that w_i gradually accumulate errors as i = 1, 2, ..., N.

Definition (Convergent)

A difference method is called convergent if

$$\lim_{h\to 0} \left(\max_{1\leq i\leq N} |y_i - w_i(h)| \right) = 0$$

Example

Show that Euler's method is convergent.

Solution: We have showed before that for fixed h > 0 there is

$$|y(t_i) - w_i| \leq \frac{hM}{2L} \left(e^{L(t_i - a)} - 1 \right) \leq \frac{hM}{2L} \left(e^{L(b-a)} - 1 \right)$$

for all $i = 0, \ldots, N$. Therefore we have

$$\max_{1\leq i\leq N} \left| y(t_i) - w_i \right| \leq \frac{hM}{2L} \left(e^{L(b-a)} - 1 \right) \to 0$$

as $h \to 0$. Therefore $\lim_{h \to 0} (\max_{1 \le i \le N} |y(t_i) - w_i|) = 0$.

Definition

A numerical method is called **stable** if its results depend on the initial data continuously.

Theorem

For a given IVP y' = f(t, y), $t \in [a, b]$ with $y(a) = \alpha$, consider a difference method $w_{i+1} = w_i + h\phi(t_i, w_i, h)$ with $w_0 = \alpha$. If there exists $h_0 > 0$ such that ϕ is continuous on $[a, b] \times \mathbb{R} \times [0, h_0]$, and ϕ is L-Lipschitz with respect to w, then

- The difference method is stable.
- ► The difference method is convergent if and only if it is consistent (i.e., φ(t, y, 0) = f(t, y)).
- ▶ If there exists bound $\tau(h)$ such that $|\tau_i(h)| \le \tau(h)$ for all i = 1, ..., N, then $|y(t_i) w_i| \le \tau(h)e^{L(t_i a)}/L$.

Proof.

Let *h* be fixed, then $w_i(\alpha)$ generated by the difference method are functions of α . For any two values $\alpha, \hat{\alpha}$, there is

$$\begin{split} |w_{i+1}(\alpha) - w_{i+1}(\hat{\alpha})| &= |(w_i(\alpha) - h\phi(t_i, w_i(\alpha), h)) - (w_i(\hat{\alpha}) - h\phi(t_i, w_i(\hat{\alpha}), h))| \\ &\leq |w_i(\alpha) - w_i(\hat{\alpha})| + h|\phi(t_i, w_i(\alpha), h) - \phi(t_i, w_i(\hat{\alpha}), h)| \\ &\leq |w_i(\alpha) - w_i(\hat{\alpha})| + hL|w_i(\alpha) - w_i(\hat{\alpha})| \\ &= (1 + hL)|w_i(\alpha) - w_i(\hat{\alpha})| \\ &\leq \cdots \\ &\leq (1 + hL)^{i+1}|w_0(\alpha) - w_0(\hat{\alpha})| \\ &= (1 + hL)^{i+1}|\alpha - \hat{\alpha}| \\ &\leq (1 + hL)^N |\alpha - \hat{\alpha}| \end{split}$$

Therefore $w_i(\alpha)$ is Lipschitz with respect to α (with constant at most $(1 + hL)^N$), and hence is continuous with respect to α . We omit the proofs for the other two assertions here.

Example

Use the result of Theorem above to show that the Modified Euler's method is stable.

Solution: Recall the Modified Euler's method is given by

$$w_{i+1} = w_i + \frac{h}{2} \left(f(t_i, w_i) + f(t_{i+1}, w_i + hf(t_i, w_i)) \right)$$

So we have $\phi(t, w, h) = \frac{1}{2}(f(t, w) + f(t + h, w + hf(t, w)))$. Now we want to show ϕ is continuous in (t, w, h), and Lipschitz with respect to w.

Solution: (cont) It is obvious that ϕ is continuous in (t, w, h) since f(t, w) is continuous. Fix t and h. For any $w, \bar{w} \in \mathbb{R}$, there is

$$\begin{split} |\phi(t, w, h) - \phi(t, \bar{w}, h)| &\leq \frac{1}{2} |f(t, w) - f(t, \bar{w})| \\ &+ \frac{1}{2} |f(t + h, w + hf(t, w)) - f(t + h, \bar{w} + hf(t, \bar{w}))| \\ &\leq \frac{L}{2} |w - \bar{w}| + \frac{L}{2} |(w + hf(t, w)) - (\bar{w} + hf(t, \bar{w}))| \\ &\leq L |w - \bar{w}| + \frac{Lh}{2} |f(t, w) - f(t, \bar{w})| \\ &\leq L |w - \bar{w}| + \frac{L^2h}{2} |w - \bar{w}| \\ &= (L + \frac{L^2h}{2}) |w - \bar{w}| \end{split}$$

So ϕ is Lipschitz with respect to w. By first part of Theorem above, the Modified Euler's method is stable.

Definition Suppose a multistep difference method given by

 $w_{i+1} = a_{m-1}w_i + a_{m-2}w_{i-1} + \dots + a_0w_{i-m+1} + hF(t_i, h, w_{i+1}, \dots, w_{i-m+1})$

Then we call the following the **characteristic polynomial** of the method:

$$\lambda^m - (a_{m-1}\lambda^{m-1} + \cdots + a_1\lambda + a_0)$$

Definition

A difference method is said to satisfy the **root condition** if all the m roots $\lambda_1, \ldots, \lambda_m$ of its characteristic polynomial have magnitudes ≤ 1 , and all of those which have magnitude =1 are single roots.

Definition

- A difference method that satisfies root condition is called strongly stable if the only root with magnitude 1 is λ = 1.
- A difference method that satisfies root condition is called weakly stable if there are multiple roots with magnitude 1.
- A difference method that does not satisfy root condition is called unstable.

Theorem

- A difference method is stable if and only if it satisfies the root condition.
- If a difference method is consistent, then it is stable if and only if it is convergent.

Example

Show that the Adams-Bashforth 4-step explicit method is strongly stable.

Solution: Recall that the method is given by

$$w_{i+1} = w_i + \frac{h}{24} \left[55f(t_i, w_i) - 59f(t_{i-1}, w_{i-1}) + 37f(t_{i-2}, w_{i-2}) - 9f(t_{i-3}, w_{i-3}) \right]$$

So the characteristic polynomial is simply $\lambda^4 - \lambda^3 = \lambda^3(\lambda - 1)$, which only has one root $\lambda = 1$ with magnitude 1. So the method is strongly stable.

Example

Show that the Milne's 4-step explicit method is weakly stable but not strongly stable.

Solution: Recall that the method is given by

$$w_{i+1} = w_{i-3} + \frac{4h}{3} \left[2f(t_i, w_i) - f(t_{i-1}, w_{i-1}) + 2f(t_{i-2}, w_{i-2}) \right]$$

So the characteristic polynomial is simply $\lambda^4 - 1$, which have roots $\lambda = \pm 1, \pm i$. So the method is weakly stable but not strongly stable.

Remark

This is the reason we chose Adams-Bashforth-Moulton PC rather than Milne-Simpsons PC since the former is strongly stable and likely to be more robust. Stiff differential equations have e^{-ct} terms (c > 0 large) in their solutions. These terms $\rightarrow 0$ quickly, but their derivatives (of form $c^n e^{-ct}$) do not, especially at small t.

Recall that difference methods have errors proportional to the derivatives, and hence they may be inaccurate for stiff ODEs.

Example

Use RK4 to solve the IVP for a system of two ODEs:

$$\begin{cases} u_1' = 9u_1 + 24u_2 + 5\cos t - \frac{1}{3}\sin t \\ u_2' = -24u_1 - 51u_2 - 9\cos t + \frac{1}{3}\sin t \end{cases}$$

with initial values $u_1(0) = 4/3$ and $u_2(0) = 2/3$. Solution: The exact solution is

$$\begin{cases} u_1(t) = 2e^{-3t} - e^{-39t} + \frac{1}{3}\cos t \\ u_2(t) = -e^{-3t} + 2e^{-39t} - \frac{1}{3}\cos t \end{cases}$$

for all $t \geq 0$.

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Solution: (cont) When we apply RK4 to this stiff ODE, we obtain

	<i>u</i> (f)	$w_1(t) = 0.05$	$w_1(t)$ h = 0.1	<i>u</i> (t)	$w_2(t) = 0.05$	$w_2(t) = 0.1$
ı	$u_1(t)$	n = 0.03	n = 0.1	$u_2(i)$	n = 0.05	n = 0.1
0.1	1.793061	1.712219	-2.645169	-1.032001	-0.8703152	7.844527
0.2	1.423901	1.414070	-18.45158	-0.8746809	-0.8550148	38.87631
0.3	1.131575	1.130523	-87.47221	-0.7249984	-0.7228910	176.4828
0.4	0.9094086	0.9092763	-934.0722	-0.6082141	-0.6079475	789.3540
0.5	0.7387877	9.7387506	-1760.016	-0.5156575	-0.5155810	3520.00
0.6	0.6057094	0.6056833	-7848.550	-0.4404108	-0.4403558	15697.84
0.7	0.4998603	0.4998361	-34989.63	-0.3774038	-0.3773540	69979.87
0.8	0.4136714	0.4136490	-155979.4	-0.3229535	-0.3229078	311959.5
0.9	0.3416143	0.3415939	-695332.0	-0.2744088	-0.2743673	1390664.
1.0	0.2796748	0.2796568	-3099671.	-0.2298877	-0.2298511	6199352.

which can blow up for larger step size h.

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Now let's use a simple example to see why this happens: consider an IVP $y' = \lambda y$, $t \ge 0$, and $y(0) = \alpha$. Here $\lambda < 0$. We know the problem has solution $y(t) = \alpha e^{\lambda t}$.

Suppose we apply Euler's method, which is

$$w_{i+1} = w_i + hf(t_i, w_i) = w_i + h\lambda w_i = (1 + \lambda h)w_i$$

 $= \cdots = (1 + \lambda h)^{i+1}w_0 = (1 + \lambda h)^{i+1}lpha$

Therefore we simply have $w_i = (1 + \lambda h)^i \alpha$. So the error is

$$|y(t_i) - w_i| = |lpha e^{\lambda i h} - (1 + \lambda h)^i lpha| = |e^{\lambda i h} - (1 + \lambda h)^i||lpha|$$

In order for the error not to blow up, we need at least $|1 + \lambda h| < 1$, which yields $h < \frac{2}{|\lambda|}$. So h needs to be sufficiently small for large λ .

Similar issue occurs for other one-step methods, which for this IVP can be written as $w_{i+1} = Q(\lambda h)w_i = \cdots = (Q(\lambda h))^{i+1}\alpha$. For the solution not to blow up, we need $|Q(\lambda h)| < 1$.

For example, in *n*th-order Taylor's method, we need

$$|Q(\lambda h)| = \left|1 + \lambda h + \frac{\lambda^2 h^2}{2} + \dots + \frac{\lambda^n h^n}{n!}\right| < 1$$

which requires h to be very small.

The same issue occurs for multistep methods too.

A remedy of stiff ODE is using implicit method, e.g., the implicit Trapezoid method:

$$\mathbf{w}_{i+1} = w_i + \frac{h}{2}(f(t_{i+1}, \mathbf{w}_{i+1}) + f(t_i, w_i))$$

In each step, we need to solve for w_{i+1} from the equation above. Namely, we need to solve for the root of F(w):

$$F(w) := w - w_i - \frac{h}{2}(f(t_{i+1}, w) + f(t_i, w_i)) = 0$$

We can use fixed point iteration or Newton's method to solve F(x) = 0.

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Section 2

Direct Methods for Linear Systems

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Linear system of equations

In many real-world applications, we need to solve linear system of n equations with n variables x_1, \ldots, x_n :

$$E_1: \qquad a_{11}x_1 + a_{12}x_2 + \cdots + a_{1n}x_n = b_1$$

$$E_2: \qquad a_{21}x_1 + a_{22}x_2 + \cdots + a_{2n}x_n = b_2$$

÷

$$E_n: \qquad a_{n1}x_1 + a_{n2}x_2 + \cdots + a_{nn}x_n = b_n$$

We're given a_{ij} , $1 \le i, j \le n$ and b_i , $(1 \le i \le n)$, and want to find x_1, \ldots, x_n that satisfy the *n* equations E_1, \ldots, E_n .

Linear system of equations

General approach: Gauss elimination.

We use three operations to simplify the linear system:

- Equation E_i can be multiplied by λE_i for any $\lambda \neq 0$: $\lambda E_i \rightarrow E_i$
- E_j is multiplied by λ and added to E_i : $\lambda E_j + E_i \rightarrow E_i$
- Switch E_i and E_j : $E_i \leftrightarrow E_j$

The goal is to simply the linear system into a triangular form, and apply backward substitution to get x_1, \ldots, x_n .

Linear system of equations

Generally, we form the augmented matrix

$$\tilde{A} = [A \mathbf{b}], \text{ where } A = \begin{bmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ a_{21} & a_{22} & \cdots & a_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ a_{n1} & a_{n2} & \cdots & a_{nn} \end{bmatrix} \text{ and } \mathbf{b} = \begin{bmatrix} b_1 \\ b_2 \\ \vdots \\ b_n \end{bmatrix}$$

and apply Gaussian elimination to get a triangular form of \tilde{A} then apply backward substitution. Total cost is $O(n^3)$.

Pivoting strategies

Standard Gauss elimination may not work properly in numerical computations.

Example

Apply Gauss elimination to the system

with four digits for arithmetic rounding. Compare the result to exact solution $x_1 = 10.00$ and $x_2 = 1.000$.
Pivoting strategies

Solution: We need to multiply E_1 by $\frac{5.291}{0.003000} = 1763.66\overline{6} \approx 1764$, then subtract it from E_2 and get:

 $\begin{array}{l} 0.003000 x_1 + 59.14 x_2 \approx 59.17 \\ -104300 x_2 \approx -104400 \end{array}$

On the other hand, the exact system without rounding error:

 $\begin{array}{l} 0.003000 x_1 + 59.14 x_2 \approx 59.17 \\ -104309.37 \bar{6} x_2 \approx -104309.37 \bar{6} \end{array}$

Solving the former yields $x_2 = 1.001$ (still close to exact solution 1.000), but $x_1 = \frac{59.17-59.14x_2}{0.003000} = -10.00$ (far from exact solution 10.00).

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Pivoting strategies



Partial pivoting

- The issue above comes up because the pivot a_{kk} is smaller than the remaining a_{ij} (i, j > k).
- One remedy, called **partial pivoting**, is interchanging rows k and p (where |a_{ik}| = max{|a_{ik}| : i = k,...,n}).

Sometimes interchange columns can also be performed. For example, when we are about to do pivoting for the *k*-th time (i.e., $a_{kk}x_k$ term), we switch row *p* and current row *k* so that

$$p = \operatorname*{argmax}_{k \le i \le n} |a_{ik}|$$

Redo the example above, we will get exact solution.

Scaled partial pivoting

Consider the following example:

 $E_1: \quad 30.00x_1 + 591400x_2 = 591700$

$$E_2$$
: 5.291 $x_1 - 6.130x_2 = 46.78$

This is equivalent to example above, except that E_1 is multiplied by 10^4 .

If we apply partial pivoting above, we will not exchange E_1 and E_2 since 30.00 > 5.291, and will end up with the same *incorrect* answer $x_2 = 1.001$ and $x_1 = -10.00$.

To overcome this issue, we can scale the coefficients of each row i by $1/s_i$ where $s_i = \max_{1 \le j \le n} |a_{ij}|$. Then apply partial pivoting based on the scaled values.

Scaled partial pivoting

Applying scaled partial pivoting to the example above, we first have

 $s_1 = \max\{30.00, 519400\} = 519400, s_2 = \max\{5.291, 6.130\} = 6.130$

Hence we get $\frac{a_{11}}{s_1} = \frac{30.00}{519400} \approx 0.5073 \times 10^{-4}$, and $\frac{a_{21}}{s_2} = \frac{5.291}{6.130} = 0.8631$, the others are ± 1 . By comparing $\frac{a_{11}}{s_1}$ and $\frac{a_{21}}{s_2}$, we will exchange E_1 and E_2 , and hence obtain

$$E_1: \quad 5.291x_1 - 6.130x_2 = 46.78 \\ E_2: \quad 30.00x_1 + 591400x_2 = 591700 \\$$

and apply Gauss elimination to obtain correct answer $x_2 = 1.000$ and $x_1 = 10.00$. For each of the *n* steps, find the largest magnitude among all coefficients a_{ij} for $k \le i, j \le n$. Then switch rows and/or columns so that the one with largest magnitude is in the pivot position.

This requires $O(n^3)$ comparisons. Only worth it if the accuracy improvement justifies the cost.

We call A an $m \times n$ (*m*-by-*n*) matrix if A is an array of *mn* numbers with *m* rows and *n* columns

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

We may simply denote it by $A = [a_{ij}]$, when its size is clear in the context.

- We call two matrices equal, i.e., A = B, if $a_{ij} = b_{ij}$ for all i, j.
- The sum of two matrices of same size is: $A \pm B = [a_{ij} \pm b_{ij}]$.
- Scalar multiplication of A by $\lambda \in \mathbb{R}$ is $\lambda A = [\lambda a_{ij}]$.
- We denote the matrix of all zeros by 0, and $-A = [-a_{ij}]$.

The set of all $m \times n$ matrices forms a **vector space**:

•
$$A + B = B + A$$

• $(A + B) + C = A + (B + C)$
• $A + 0 = 0 + A$
• $A + (-A) = 0$
• $\lambda(A + B) = \lambda A + \lambda B$
• $(\lambda + \mu)A = \lambda A + \mu A$
• $\lambda(\mu A) = (\lambda \mu)A$
• $1A = A$

For matrix A of size $m \times n$ and (column) vector b of dimension n, we define the matrix-vector multiplication (product) by

$$Ab = \begin{bmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ a_{21} & a_{22} & \cdots & a_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ a_{m1} & a_{m2} & \cdots & a_{mn} \end{bmatrix} \begin{bmatrix} b_1 \\ b_2 \\ \vdots \\ b_n \end{bmatrix} = \begin{bmatrix} \sum_{j=1}^n a_{1j} b_j \\ \sum_{j=1}^n a_{2j} b_j \\ \vdots \\ \sum_{j=1}^n a_{mj} b_j \end{bmatrix}$$

For matrix A of size $m \times n$ and matrix B of size $n \times k$, we define the matrix-matrix multiplication (product) by

$$AB = \begin{bmatrix} a_{11} & a_{12} & \cdots & a_{1n} \\ a_{21} & a_{22} & \cdots & a_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ a_{m1} & a_{m2} & \cdots & a_{mn} \end{bmatrix} \begin{bmatrix} b_{11} & b_{12} & \cdots & b_{1k} \\ b_{21} & b_{22} & \cdots & b_{2k} \\ \vdots & \vdots & \ddots & \vdots \\ b_{n1} & b_{n2} & \cdots & b_{nk} \end{bmatrix}$$
$$= \begin{bmatrix} \sum_{j=1}^{n} a_{1j}b_{j1} & \sum_{j=1}^{n} a_{1j}b_{j1} & \cdots & \sum_{j=1}^{n} a_{1j}b_{jk} \\ \sum_{j=1}^{n} a_{2j}b_{j1} & \sum_{j=1}^{n} a_{2j}b_{j1} & \cdots & \sum_{j=1}^{n} a_{2j}b_{jk} \\ \vdots & \vdots & \ddots & \vdots \\ \sum_{j=1}^{n} a_{mj}b_{j1} & \sum_{j=1}^{n} a_{mj}b_{j1} & \cdots & \sum_{j=1}^{n} a_{mj}b_{jk} \end{bmatrix} \in \mathbb{R}^{m \times k}$$

That is, if C = AB, then $[c_{ij}] = [\sum_{r} a_{ir}b_{rj}]$ for all i, j.

Some properties of matrix product

$$\blacktriangleright A(BC) = (AB)C$$

$$\blacktriangleright A(B+D) = AB + AD$$

$$\blacktriangleright \ \lambda(AB) = (\lambda A)B = A(\lambda B)$$

Remark

Note that $AB \neq BA$ in general, even if both exists.

Some special matrices

- Square matrix: A is of size $n \times n$
- Diagonal matrix: $a_{ij} = 0$ if $i \neq j$.
- Identity matrix of order *n*: $I = [\delta_{ij}]$ where $\delta_{ij} = 1$ if i = j and = 0 otherwise.
- Upper triangle matrix: $a_{ij} = 0$ if i > j.
- Lower triangle matrix: $a_{ij} = 0$ if i < j.

Definition (Inverse of matrix)

An $n \times n$ matrix A is said to be **nonsingular** (or **invertible**) if there exists an $n \times n$ matrix, denoted by A^{-1} , such that $A(A^{-1}) = (A^{-1})A = I$. Here A^{-1} is called the **inverse** of matrix A.

Definition

An $n \times n$ matrix A without an inverse is called singular (or noninvertible)

Several properties of inverse matrix:

► A^{-1} is unique.

•
$$(A^{-1})^{-1} = A$$
.

• If B is also nonsingular, then $(AB)^{-1} = B^{-1}A^{-1}$.

Definition (Transpose)

The transpose of an $m \times n$ matrix $A = [a_{ij}]$ is the $n \times m$ matrix $A^{\top} = [a_{ji}]$.

Sometimes A^{\top} is also written as A^t, A', A^{\top} .

$$(A^{\top})^{\top} = A$$

$$(AB)^{\top} = B^{\top}A^{\top}$$

$$(A+B)^{\top} = A^{\top} + B^{\top}$$

• If A is nonsingular, then $(A^{-1})^{\top} = (A^{\top})^{-1}$.

Definition (Determinant)

- If A = [a] is a 1×1 matrix, then det(A) = a.
- If A is n × n where n > 1, then the minor M_{ij} is the determinant of the (n − 1) × (n − 1) submatrix of A by deleting its ith row and jth column.
- The cofactor A_{ij} associated with the minor M_{ij} is defined by $A_{ij} = (-1)^{i+j} M_{ij}$.
- The determinant of the n × n matrix A, denoted by det(A) (or |A|), is given by either of the followings:

$$det(A) = \sum_{j=1}^{n} a_{ij}A_{ij} = \sum_{j=1}^{n} (-1)^{i+j}a_{ij}M_{ij}, \text{ for any } i = 1, \dots, n$$
$$det(A) = \sum_{i=1}^{n} a_{ij}A_{ij} = \sum_{j=1}^{n} (-1)^{i+j}a_{ij}M_{ij}, \text{ for any } j = 1, \dots, n$$

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Some properties of determinant

- If A has any zero row or column, then det(A) = 0.
- If two rows (or columns) of A are the same, or one is a multiple of the other, then det(A) = 0.
- Switching two rows (or columns) of A results in a matrix with determinant – det(A).
- Multiplying a row (or column) of A by λ results in a matrix with determinant λ det(A).
- $(E_i + \lambda E_j) \rightarrow E_i$ results in a matrix of the same determinant.

Some properties of determinant

- det(AB) = det(A) det(B) if A and B are square matrices of same size.
- ► $det(A^{\top}) = det(A)$
- A is singular if any only if det(A) = 0.
- ▶ If A is nonsingular, then $det(A) \neq 0$ and $det(A^{-1}) = det(A)^{-1}$.
- If A is an upper or lower triangular matrix, then det(A) = ∏ⁿ_{i=1} a_{ii}.

The following statements are equivalent:

- Ax = 0 has unique solution x = 0.
- Ax = b has a unique solution for every b.
- A is nonsingular, i.e., A^{-1} exists.
- det(A) \neq 0.

Gauss elimination can be used to compute **LU factorization** of a square matrix *A*:

$$A = LU$$

where L is a lower triangular matrix, and U is an upper triangular matrix.

If we have **LU factorization** of *A*, then

$$Ax = LUx = L(Ux) = b$$

so we solve x easily:

1. Solve y from Ly = b by forward substitution;

2. Solve x from Ux = y by backward substitution. Total cost is $O(2n^2)$.

Matrix factorization

The cost reduction from $O(n^3/3)$ to $O(2n^2)$ is huge, especially for large *n*:

п	n ³ /3	2 <i>n</i> ²	% Reduction
10	$3.\overline{3} imes10^2$	$2 imes 10^2$	40
100	$3.ar{3} imes10^5$	$2 imes 10^4$	94
1000	$3.ar{3} imes10^8$	$2 imes 10^6$	99.4

Unfortunately, LU factorization itself requires $O(n^3)$ in general.

Now let's see how to obtain LU factorization by Gauss elimination.

Suppose we can perform Gauss elimination without any row exchange. In first round, we use a_{11} as the pivot and cancel each of a_{21}, \ldots, a_{n1} by

$$(E_j-m_{j1}E_1)
ightarrow E_j$$
 where $m_{j1}=rac{a_{j1}}{a_{11}}, \quad j=2,\ldots,4$

This is equivalent to multiplying $M^{(1)}$ to A and get $A^{(2)} := M^{(1)}A$ where

$$M^{(1)} = \begin{bmatrix} 1 & 0 & \cdots & 0 \\ -m_{21} & 1 & \cdots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ -m_{n1} & 0 & \cdots & 1 \end{bmatrix} \text{ and } A^{(2)} = \begin{bmatrix} a_{11} & * & \cdots & * \\ 0 & * & \cdots & * \\ \vdots & \vdots & \ddots & \vdots \\ 0 & * & \cdots & * \end{bmatrix}$$

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In second round, we use <u>current</u> a_{22} as the pivot and cancel each of a_{32}, \ldots, a_{n2} by

$$(E_j - m_{j2}E_2) \to E_j$$
 where $m_{j2} = \frac{a_{j2}}{a_{22}}, j = 3, \dots, 4$

This is equivalent to multiplying $M^{(2)}$ to $A^{(2)}$ and get $A^{(3)} := M^{(2)}A^{(2)}$ where

$$M^{(2)} = \begin{bmatrix} 1 & 0 & 0 & \cdots & 0 \\ 0 & 1 & 0 & \cdots & 0 \\ 0 & -m_{32} & 1 & \cdots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & -m_{n2} & 0 & \cdots & 1 \end{bmatrix} \text{ and } A^{(3)} = \begin{bmatrix} a_{11} & * & * & \cdots & * \\ 0 & * & * & \cdots & * \\ 0 & 0 & * & \cdots & * \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & * & \cdots & * \end{bmatrix}$$

When Gauss elimination finishes (total n-1 rounds), we will get an upper triangular matrix U:

$$U := M^{(n-1)} M^{(n-2)} \cdots M^{(1)} A$$

Define matrix L

$$L = (M^{(n-1)}M^{(n-2)}\cdots M^{(1)})^{-1} = (M^{(1)})^{-1}\cdots (M^{(n-2)})^{-1}(M^{(n-1)})^{-1}$$

Note that L is lower triangular (because each M is lower triangular, and inverse and product of lower triangular matrices are still lower triangular). So we get the LU factorization of A:

$$LU = (M^{(1)})^{-1} \cdots (M^{(n-2)})^{-1} (M^{(n-1)})^{-1} M^{(n-1)} M^{(n-2)} \cdots M^{(1)} A = A$$

It is easy to check that:

$$M^{(1)} = \begin{bmatrix} 1 & 0 & \cdots & 0 \\ -m_{21} & 1 & \cdots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ -m_{n1} & 0 & \cdots & 1 \end{bmatrix} \text{ and } (M^{(1)})^{-1} = \begin{bmatrix} 1 & 0 & \cdots & 0 \\ m_{21} & 1 & \cdots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ m_{n1} & 0 & \cdots & 1 \end{bmatrix}$$
$$M^{(2)} = \begin{bmatrix} 1 & 0 & 0 & \cdots & 0 \\ 0 & 1 & 0 & \cdots & 0 \\ 0 & -m_{32} & 1 & \cdots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & -m_{n2} & 0 & \cdots & 1 \end{bmatrix} \text{ and } (M^{(2)})^{-1} = \begin{bmatrix} 1 & 0 & 0 & \cdots & 0 \\ 0 & 1 & 0 & \cdots & 0 \\ 0 & m_{32} & 1 & \cdots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & m_{n2} & 0 & \cdots & 1 \end{bmatrix}$$

and finally there is

$$L = (M^{(1)})^{-1} \cdots (M^{(n-2)})^{-1} (M^{(n-1)})^{-1} = \begin{bmatrix} 1 & 0 & 0 & \cdots & 0 \\ m_{21} & 1 & 0 & \cdots & 0 \\ m_{31} & m_{32} & 1 & \cdots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ m_{n1} & m_{n2} & m_{n3} & \cdots & 1 \end{bmatrix}$$

To summarize, the LU factorization of A gives L as above, and U as the result of Gauss elimination of A.

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Gauss elimination row exchange

If Gauss elimination is done with row exchanges, then we will get LU factorization of PA where P is some row permutation matrix.

For example, to switch rows 2 and 4 of a 4×4 matrix *A*, the permutation matrix *P* is

$$P = \begin{bmatrix} 1 & 0 & 0 & 0 \\ 0 & 0 & 0 & 1 \\ 0 & 0 & 1 & 0 \\ 0 & 1 & 0 & 0 \end{bmatrix}$$

Some properties of permutation matrices:

If P₁, P₂ are permutations, then P₂P₁ is still permutation.
 P⁻¹ = P[⊤].

Now we consider two types of matrices for which Gauss elimination can be used effectively without row interchanges.

Definition (Diagonally dominate matrices) An $n \times n$ matrix A is called **diagonally dominate** if

$$|a_{ii}| \ge \sum_{j \ne i} |a_{ij}|, \quad \forall i = 1, 2, \dots, n$$

An $n \times n$ matrix A is called strictly diagonally dominate if

$$|a_{ii}| > \sum_{j \neq i} |a_{ij}|, \quad \forall i = 1, 2, \dots, n$$

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Example

Consider the following matrices:

$$A = \begin{bmatrix} 1 & -1 & 0 & 0 \\ -1 & 2 & -1 & 0 \\ 0 & -1 & 2 & -1 \\ 0 & 0 & -1 & 1 \end{bmatrix}$$
$$B = \begin{bmatrix} 7 & 2 & 0 \\ 3 & 5 & -1 \\ 0 & 5 & -6 \end{bmatrix} \quad C = \begin{bmatrix} 6 & 4 & -3 \\ 4 & -2 & 0 \\ -3 & 0 & 1 \end{bmatrix}$$

A (and A^{\top}) is diagonally dominate, B is strictly diagonally dominate, B^{\top} , C, C^{\top} are not diagonally dominate.

Theorem

If A is strictly diagonally dominant, then A is nonsingular. Moreover, Gauss elimination can be performed without row interchange to obtain the unique solution of Ax = b.

Proof.

If A is singular, then Ax = 0 has nonzero solution x. Suppose x_k is the component of x with largest magnitude:

$$|x_k| > 0$$
 and $|x_k| \ge |x_j|, \ \forall j \ne k$

Then the product of x and the k-th row of A gives

$$a_{kk}x_k + \sum_{j \neq k} a_{kj}x_j = 0$$

From this we obtain

$$|\mathbf{a}_{kk}| = \left|-\sum_{j \neq k} \frac{a_{kj} x_j}{x_k}\right| \le \sum_{j \neq k} \frac{|x_j|}{|x_k|} |\mathbf{a}_{kj}| \le \sum_{j \neq k} |\mathbf{a}_{kj}|$$

Contradiction. So A is nonsingular.

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Proof (cont.) Now let's see how Gauss elimination works when A is strictly diagonally dominant. Consider 1st and *i*th $(i \ge 2)$ rows of A:

$$|\mathsf{a}_{11}| > \sum_{j
eq 1} |\mathsf{a}_{1j}|, \quad |\mathsf{a}_{ii}| > \sum_{j
eq i} |\mathsf{a}_{ij}|$$

If we perform $E_i - \frac{a_{i1}}{a_{11}}E_1 \rightarrow E_i$, the new values in row *i* are $a_{i1}^{(2)} = 0$ and $a_{ij}^{(2)} = a_{ij} - \frac{a_{i1}}{a_{11}}a_{1j}$ for $j \ge 2$. Therefore

$$\begin{split} \sum_{\substack{j=2\\j\neq i}}^{n} |a_{ij}^{(2)}| &\leq \sum_{\substack{j=2\\j\neq i}}^{n} |a_{ij}| + \sum_{\substack{j=2\\j\neq i}}^{n} \left| \frac{a_{1j}}{a_{11}} \right| |a_{i1}| < |a_{ii}| - |a_{i1}| + \frac{|a_{11}| - |a_{1i}|}{|a_{11}|} |a_{i1}| \\ &= |a_{ii}| - \frac{|a_{1i}|}{|a_{11}|} |a_{i1}| \leq \left| |a_{ii}| - \frac{|a_{1i}|}{|a_{11}|} |a_{i1}| \right| = |a_{ii}^{(2)}| \end{split}$$

As *i* is arbitrary, we know *A* remains strictly diagonally dominant after first round. By induction we know *A* stays as strictly diagonally dominant and Gauss elimination can be performed without row interexchange.

Definition (Positive definite matrix)

A matrix A is called **positive definite** (PD) if it is symmetric and $x^{\top}Ax > 0$ for any $x \neq 0$

Remark

In some texts, A is called positive definite as long as $x^{\top}Ax > 0$ for any $x \neq 0$, so A is not necessarily symmetric. In these texts, the matrix in our definition above is called **symmetric positive definite** (SPD).

Positive definite matrices

We first have the following formula: if $x = (x_1, \ldots, x_n)^{\top}$ and $A = [a_{ij}]$, then

$$x^{ op} A x = \sum_{i,j} \mathsf{a}_{ij} x_i x_j$$
Example

Show that the matrix A below is PD:

$$A = \begin{bmatrix} 2 & -1 & 0 \\ -1 & 2 & -1 \\ 0 & -1 & 2 \end{bmatrix}$$

Solution: First A is symmetric. For any $x \in \mathbb{R}^3$, we have

$$x^{\top}Ax = 2x_1^2 - 2x_1x_2 + 2x_2^2 - 2x_2x_3 + 2x_3^2$$

= $x_1^2 + (x_1^2 - 2x_1x_2 + x_2^2) + (x_2^2 - 2x_2x_3 + x_3^2) + x_3^2$
= $x_1^2 + (x_1 - x_2)^2 + (x_2 - x_3)^2 + x_3^2$

Therefore $x^{\top}Ax = 0$ if and only if $x_1 = x_2 = x_3 = 0$. So A is PD.

Theorem

If A is an $n \times n$ positive definite matrix, then

• $a_{ii} > 0$ for all i;

•
$$(a_{ij})^2 < a_{ii}a_{jj}$$
 for any $i \neq j$.

Proof.

- If Ax = 0, then x^TAx = 0 and hence x = 0 since A is PD. So A is nonsingular.
- Set $x = e_i$, where $e_i \in \mathbb{R}^n$ has 1 as the *i*-th component and zeros elsewhere. Then $x^{\top}Ax = e_i^{\top}Ae_i = a_{ii} > 0$.
- For any k, j, define $x, z \in \mathbb{R}^n$ such that $x_j = z_k = z_j = 1$ and $x_k = -1$, and $x_i = z_i = 0$ if $i \neq k, j$. Then we can show

$$0 < x^{ op} A x = a_{jj} + a_{kk} - a_{kj} - a_{jk}$$

 $0 < z^{ op} A z = a_{jj} + a_{kk} + a_{kj} + a_{jk}$

Note that $a_{kj} = a_{jk}$, so we get $|a_{kj}| < \frac{a_{ij} + a_{kk}}{2} \le \max_i a_{ii}$.

For any i ≠ j, set x ∈ ℝⁿ such that x_i = α and x_j = 1, and 0 elsewhere. Therefore 0 < x^TAx = a_{ii}α² + 2a_{ij}α + a²_{jj} for any α. This implies that 4a²_{ij} - 4a_{ii}a_{jj} < 0.</p>

Definition (Leading principal submatrix)

A leading principal submatrix of A is the $k \times k$ upper left submatrix

$$A_{k} = \begin{bmatrix} a_{11} & a_{12} & \cdots & a_{1k} \\ a_{21} & a_{22} & \cdots & a_{2k} \\ \vdots & \vdots & \ddots & \vdots \\ a_{k1} & a_{k2} & \cdots & a_{kk} \end{bmatrix}$$

Theorem

A symmetric matrix A is PD if and only if every leading principal submatrix has a positive determinant.

Example

Use the Theorem above to check A is PD:

$$A = \begin{bmatrix} 2 & -1 & 0 \\ -1 & 2 & -1 \\ 0 & -1 & 2 \end{bmatrix}$$

Theorem

A matrix A is PD if and only either of the followings is true:

- ► There exist a lower triangular matrix L with all 1 on its diagonal and a diagonal matrix D with all diagonal entries positive, such that A = LDL^T.
- ► There exists a lower triangular matrix L with all diagonal entries positive such that A = LL^T (Cholesky factorization).
- Gauss elimination of A without row interchanges can be performed and all pivot elements are positive.

Definition (Band matrix)

An $n \times n$ matrix A is called **band matrix** if there exist p, q such that a_{ij} can be nonzero only if $i - q \le j \le i + p$. The band width is defined by w = p + q + 1.

Definition (Tridiagonal matrix)

A band matrix with p = q = 1 is called tridiagonal matrix.

Crout factorization

The **Crout factorization** of a tridiagonal matrix is A = LU where L is lower triangle, U is upper triangle, and both L, U are tridiagonal:

$$L = \begin{bmatrix} l_{11} & 0 & \cdots & 0 & 0 \\ l_{21} & l_{22} & \cdots & 0 & 0 \\ \vdots & \vdots & \ddots & \vdots & \vdots \\ 0 & 0 & \cdots & l_{n-1,n-1} & 0 \\ 0 & 0 & \cdots & l_{n,n-1} & l_{nn} \end{bmatrix} \quad U = \begin{bmatrix} 1 & u_{12} & \cdots & 0 & 0 \\ 0 & 1 & \cdots & 0 & 0 \\ \vdots & \vdots & \ddots & \vdots & \vdots \\ 0 & 0 & \cdots & 1 & u_{n-1,n} \\ 0 & 0 & \cdots & 0 & 1 \end{bmatrix}$$

Note that a tridiagonal matrix A has 3n - 2 unknowns, and the L and U together also have 3n - 2 unknowns.

Crout factorization

Theorem

A tridiagonal matrix A has a Crout factorization if either of the following statements is true:

- A is positive definite;
- A is strictly diagonally dominant;
- ▶ A is diagonally dominant, $|a_{11}| > |a_{12}|$, $|a_{nn}| > |a_{n,n-1}|$, and $a_{i,i-1}, a_{i,i+1} \neq 0$ for all i = 2, ..., n 1.

Crout factorization

With the special form of A, L and U, we can obtain the Crout factorization A = LU by solving I_{ij} (i = 1, ..., n and j = i - 1, i) and $u_{i,i+1}$ (i = 1, ..., n - 1) from

$$\begin{aligned} a_{11} = I_{11} \\ a_{i,i-1} = I_{i,i-1}, & \text{for } i = 2, \dots, n \\ a_{i,i} = I_{i,i-1} u_{i-1,i} + I_{ii}, & \text{for } i = 2, \dots, n \\ a_{i,i+1} = I_{ii} u_{i,i+1}, & \text{for } i = 1, \dots, n-1 \end{aligned}$$

When we use Crout factorization to solve Ax = b, the cost is only 5n - 4 multiplications/divisions and 3n - 3 additions/subtractions.

Section 3

Iterative Methods in Matrix Algebra

Vector norm

Definition

A vector norm on \mathbb{R}^n , denoted by $\|\cdot\|$, is a mapping from \mathbb{R}^n to \mathbb{R} such that

Vector norm

Definition $(I_p \text{ norms})$

The I_p (sometimes L_p or ℓ_p) norm of a vector is defined by

$$1 \le p < \infty : \qquad \|x\|_p = \left(\sum_{i=1}^n |x_i|^p\right)^{1/p}$$
$$p = \infty : \qquad \|x\|_\infty = \max_{1 \le i \le n} |x_i|$$

In particular, the l_2 norm is also called the **Euclidean norm**. Note that when $0 \le p < 1$, $\|\cdot\|_p$ is not norm, strictly speaking, but have some usages in specific applications.

I_2 norm



I_{∞} norm



Vector norms

Example

Compute the l_2 and l_{∞} norms of vector $x = (1, -1, 2) \in \mathbb{R}^3$. Solution:

$$\|x\|_{2} = \sqrt{|1|^{2} + |-1|^{2} + |2|^{2}} = \sqrt{6}$$
$$\|x\|_{\infty} = \max_{1 \le i \le 3} |x_{i}| = \max\{|1|, |-1|, |2|\} = 2$$

Theorem (Cauchy-Schwarz inequality) For any vectors $x = (x_1, ..., x_n)^\top \in \mathbb{R}^n$ and $y = (y_1, ..., y_n)^\top \in \mathbb{R}^n$, there is

$$|x^{\top}y| = \left|\sum_{i=1}^{n} x_{i}y_{i}\right| \le \left(\sum_{i=1}^{n} |x_{i}|^{2}\right)^{1/2} \left(\sum_{i=1}^{n} |y_{i}|^{2}\right)^{1/2} = \|x\|_{2}\|y\|_{2}$$

Proof.

It is obviously true for x = 0 or y = 0. If $x, y \neq 0$, then for any $\lambda \in \mathbb{R}$, there is

$$0 \le \|x - \lambda y\|_2^2 = \|x\|_2^2 - 2\lambda x^\top y + \lambda^2 \|y\|_2^2$$

and the equality holds when $\lambda = ||x||_2/||y||_2$.

Definition (Distance between two vectors)

The l_p distance $(1 \le p \le \infty)$ between two vectors $x, y \in \mathbb{R}^n$ is defined by $||x - y||_p$.

Definition (Convergence of a sequence of vectors) A sequence $\{x^{(k)}\}$ is said to converge with respect to the l_p norm if for any given $\epsilon > 0$, there exists an integer $N(\epsilon)$ such that

$$||x^{(k)} - x|| < \epsilon$$
, for all $k \ge N(\epsilon)$

Convergence of a sequence of vectors

Theorem A sequence of vectors $\{x^{(k)}\}$ converges to x if and only if $x_i^{(k)} \rightarrow x_i$ for every i = 1, 2, ..., n.

Theorem

For any vector $x \in \mathbb{R}^n$, there is

$$\|x\|_{\infty} \le \|x\|_2 \le \sqrt{n} \|x\|_{\infty}$$

Proof.

$$\|x\|_{\infty} = \max_{i} |x_{i}| = \sqrt{\max_{i} |x_{i}|^{2}} \le \sqrt{|x_{1}|^{2} + \dots + |x_{n}|^{2}} = \|x\|_{2}$$
$$\|x\|_{2} = \sqrt{|x_{1}|^{2} + \dots + |x_{n}|^{2}} \le \sqrt{n\max_{i} |x_{i}|^{2}}$$
$$= \sqrt{n}\sqrt{\max_{i} |x_{i}|^{2}} = \sqrt{n}\max_{i} |x_{i}| = \sqrt{n}\|x\|_{\infty}$$

Compare I_2 and I_∞ norms in \mathbb{R}^2



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Definition

A matrix norm on the set of $n \times n$ matrices is a real-valued function, denoted by $\|\cdot\|$, that satisfies the follows for all $A, B \in \mathbb{R}^{n \times n}$ and $\alpha \in \mathbb{R}$:

- $\blacktriangleright ||A|| \ge 0$
- ||A|| = 0 if and only if A = 0 the zero matrix,
- $\blacktriangleright \|\alpha A\| = |\alpha| \|A\|$
- ► $||A + B|| \le ||A|| + ||B||$
- $\blacktriangleright \|AB\| \le \|A\| \|B\|$

Distance between matrices

Definition

Suppose $\|\cdot\|$ is a norm defined on $\mathbb{R}^{n \times n}$. Then the **distance between two** $n \times n$ **matrices** A **and** B with respect to $\|\cdot\|$ is $\|A - B\|$ (check that it's a distance)

Matrix norm can be induced by vector norms, and hence there are many choices. Here we focus on those induced by l_2 and l_∞ vector norms.

Definition If $\|\cdot\|$ is a vector norm on \mathbb{R}^n , then the norm defined below

$$||A|| = \max_{||x||=1} ||Ax||$$

is called the matrix norm induced by vector norm $\|\cdot\|$.

Remark

- Induced norms are also called natural norms of matrices.
- Unless otherwise specified, by matrix norms most books/papers refer to induced norms.
- The induced norm can be written equivalently as

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

• It can be easily extended to case $A \in \mathbb{R}^{m \times n}$.

Corollary

For any vector $x \in \mathbb{R}^n$, there is $||Ax|| \le ||A|| ||x||$.

Proof.

It is obvious for x = 0. If $x \neq 0$, then

$$\frac{\|Ax\|}{\|x\|} \le \max_{x' \ne 0} \frac{\|Ax'\|}{\|x'\|} = \|A\|$$

Induced *I*₂ matrix norm



Induced I_{∞} matrix norm



Theorem Suppose $A = [a_{ij}] \in \mathbb{R}^{n \times n}$, then $||A||_{\infty} = \max_{1 \le i \le n} \sum_{j=1}^{n} |a_{ij}|$.

Proof. For any x with $||x||_{\infty} = 1$, i.e., $\max_i |x_i| = 1$, there is

$$\begin{aligned} \|Ax\|_{\infty} &= \max\left\{\left|\sum_{j} a_{1j}x_{j}\right|, \dots, \left|\sum_{j} a_{nj}x_{j}\right|\right\} \\ &\leq \max\left\{\sum_{j} |a_{1j}||x_{j}|, \dots, \sum_{j} |a_{nj}||x_{j}|\right\} \\ &\leq \max\left\{\sum_{j} |a_{1j}|, \dots, \sum_{j} |a_{nj}|\right\} = \max_{1 \leq i \leq n} \sum_{j=1}^{n} |a_{ij}| \end{aligned}$$

Suppose *i'* is such that $\sum_{j=1}^{n} |a_{i'j}| = \max_{1 \le i \le n} \sum_{j=1}^{n} |a_{ij}|$, then by choosing \hat{x} such that $\hat{x}_j = 1$ if $a_{i'j} \ge 0$ and -1 otherwise, we have $\sum_{j=1}^{n} a_{i'j} \hat{x}_j = \sum_{j=1}^{n} |a_{i'j}|$. So $||A\hat{x}||_{\infty} = \max_{1 \le i \le n} \sum_{j=1}^{n} |a_{ij}|$. Note that $||\hat{x}||_{\infty} = 1$. Therefore $||A||_{\infty} = \max_{1 \le i \le n} \sum_{j=1}^{n} |a_{ij}|$.

Definition

The characteristic polynomial of a square matrix $A \in \mathbb{R}^{n \times n}$ is defined by

 $p(\lambda) = \det(A - \lambda I)$

We call λ an **eigenvalue** of A if λ is a root of p, i.e., det $(A - \lambda I) = 0$. Moreover, any nonzero solution $x \in \mathbb{R}^n$ of $(A - \lambda I)x = 0$ is called an **eigenvector** of A corresponding to the eigenvalue λ .

Remark

- $p(\lambda)$ is a polynomial of degree n, and hence has n roots.
- x is an eigenvector of A corresponding to eigenvalue λ iff (A − λI)x = 0, i.e., Ax = λx. This also means A applied to x is stretching x by λ.
- If x is an eigenvector of A corresponding to λ, so is αx for any α ≠ 0:

$$A(\alpha x) = \alpha A x = \alpha \lambda x = \lambda(\alpha x)$$

Definition

Let $\lambda_1, \ldots, \lambda_n$ be the eigenvalues of $A \in \mathbb{R}^{n \times n}$, then the **spectral** radius $\rho(A)$ is defined by $\rho(A) = \max_i |\lambda_i|$ where $|\cdot|$ is the absolute value (aka magnitude) of complex numbers.

Some properties

Theorem

For a matrix $A \in \mathbb{R}^{n \times n}$, there are

•
$$||A||_2 = \sqrt{\rho(A^{\top}A)}$$

• $\rho(A) \le ||A||$ for any norm $||\cdot||$ of A

Proof.

- We later will show that both sides = σ₁², where σ₁ is the largest singular value of A.
- Let λ := ρ(A) be the eigenvalue with largest magnitude.
 Then there exists eigenvector x such that

$$(\|A\| \ge) \frac{\|Ax\|}{\|x\|} = \frac{\|\lambda x\|}{\|x\|} = \frac{|\lambda| \|x\|}{\|x\|} = |\lambda|$$

Convergent matrix

Definition A matrix $A \in \mathbb{R}^{n \times n}$ is said to be **convergent** if

$$\lim_{k\to\infty}A^k=0$$

Theorem

The following statements are equivalent:

- 1. A is convergent.
- 2. $\lim_{k\to\infty} ||A^k|| = 0$ for any norm $||\cdot||$.
- **3**. $\rho(A) < 1$.

4.
$$\lim_{k\to\infty} A^k x = 0$$
 for any $x \in \mathbb{R}$.

Jacobi iterative method

To solve x from Ax = b where $A \in \mathbb{R}^{n \times n}$ and $b \in \mathbb{R}^n$, the **Jacobi** iterative method is

▶ Initialize $x^{(0)} \in \mathbb{R}^n$. Set D = diag(A), R = A - D.

• Repeat the following for k = 0, 1, ... until convergence:

$$x^{(k+1)} = D^{-1}(b - Rx^{(k)})$$

Remark

- Needs nonzero diagonal entries, i.e., a_{ii} ≠ 0 for all i.
- ► Usually faster convergence with larger |a_{ii}|.
- Stopping criterion can be ^{||x(k)−x(k-1)||}/_{||x(k)||} ≤ ε for some prescribed ε > 0.
Gauss-Seidel iterative method

To solve x from Ax = b where $A \in \mathbb{R}^{n \times n}$ and $b \in \mathbb{R}^n$, the **Gauss-Seidel iterative method** is

- Initialize x⁽⁰⁾ ∈ ℝⁿ. Set L to the lower triangular part (including diagonal) of A and U = A − L.
- Repeat the following for k = 0, 1, ... until convergence:

$$x^{(k+1)} = L^{-1}(b - Ux^{(k)})$$

Remark

- Inverse of L requires forward substitution.
- Again needs nonzero diagonal entries, i.e., $a_{ii} \neq 0$ for all i.
- Stopping criterion can be $\frac{\|x^{(k)}-x^{(k-1)}\|}{\|x^{(k)}\|} \leq \epsilon$ for some prescribed $\epsilon > 0$.
- Faster than Jacobi iterative method most of times.

Lemma (
$$\rho(T) < 1 \Rightarrow I - T$$
 invertible)
If $\rho(T) < 1$, then $(I - T)^{-1}$ exists and

$$(I - T)^{-1} = I + T + T^2 + \dots = \sum_{j=0}^{\infty} T^j$$

Proof.

We first show that I - T is invertible, i.e., (I - T)x = 0 has unique solution x = 0. If not, then $\exists x \neq 0$ such that (I - T)x = 0, i.e., Tx = x, or x is an e.v. corresponding to e.w. 1, contradiction to $\rho(T) < 1$.

Define $S_m = I + T + \dots + T^m$. Then $(I - T)S_m = I - T^{m+1}$. Note $\rho(T) < 1$ implies $\lim_{m\to\infty} T^m = 0$, and hence

$$(I - T) \lim_{m \to \infty} S_m = \lim_{m \to \infty} (I - T) S_m = \lim_{m \to \infty} (I - T^{m+1}) = I$$

That is, $\sum_{m=0}^{\infty} T^m = \lim_{m \to \infty} S_m = (I - T)^{-1}$.

General iterative method has form $x^{(k)} = Tx^{(k-1)} + c$ for k = 1, 2, ...

Example (Jacobi and GS are iterative methods)

Jacobi iterative method:

$$x^{(k)} = D^{-1}(b - Rx^{(k-1)}) = -(D^{-1}R)x^{(k-1)} + D^{-1}b^{(k-1)}$$

So
$$T = -D^{-1}R$$
 and $c = D^{-1}b$.

Gauss-Seidel iterative method:

$$x^{(k)} = L^{-1}(b - Ux^{(k-1)}) = -(L^{-1}U)x^{(k-1)} + L^{-1}b$$

So $T = -L^{-1}U$ and $c = L^{-1}b$.

Theorem (Sufficient and necessary condition of convergence) For any initial $x^{(0)}$, the sequence $\{x^{(k)}\}_k$ defined by

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

converges to the unique solution of x = Tx + c iff $\rho(T) < 1$.

Proof.
(
$$\Leftarrow$$
) Suppose $\rho(T) < 1$. Then
 $x^{(k)} = Tx^{(k-1)} + c = T(Tx^{(k-2)} + c) + c = T^2x^{(k-2)} + (I+T)c$
 $= \cdots = T^k x^{(0)} + (I+T+\cdots+T^k)c$

Note $\rho(T) < 1 \Rightarrow T^k \to 0$ and $(I + T + \dots + T^k) \to (I - T)^{-1}$, so $x^{(k)} \to (I - T)^{-1}c$, the unique solution of x = Tx + c.

Proof.

(⇒) Let x^* be the unique solution of x = Tx + c. Then for any $z \in \mathbb{R}^n$, we set initial $x^{(0)} = x^* - z$. Then

$$\begin{aligned} x^* - x^{(k)} &= (Tx^* + c) - (Tx^{(k-1)} + c) = T(x^* - x^{(k-1)}) \\ &= \cdots = T^k(x^* - x^{(0)}) = T^k z \to 0 \end{aligned}$$

This implies $\rho(T) < 1$.

Corollary (Linear convergence rate) If ||T|| < 1 for any matrix norm $||\cdot||$, and c is given, then $\{x^{(k)}\}$ generated by $x^{(k)} = Tx^{(k-1)} + c$ converges to the unique solution x^* of x = Tx + c. Moreover 1. $||x^* - x^{(k)}|| \le ||T||^k ||x^* - x^{(0)}||$. 2. $||x^* - x^{(k)}|| \le \frac{||T||^k}{1 - ||T||} ||x^{(1)} - x^{(0)}||$.

Proof.

1. Note $\rho(T) \leq ||T|| < 1$. Follow (\Rightarrow) part of the theorem above.

2. Note that
$$||x^* - x^{(1)}|| \le ||T|| ||x^* - x^{(0)}||$$
 and hence
 $||x^{(1)} - x^{(0)}|| \ge ||x^* - x^{(0)}|| - ||x^* - x^{(1)}|| \ge (1 - ||T||) ||x^* - x^{(0)}||.$

Theorem (Jacobi and GS are convergent)

If A is strictly diagonally dominant, then from any initial $x^{(0)}$ both Jacobi and Gauss-Seidel iterative methods generate sequences that converge to the unique solution of Ax = b.

Proof.

For Jacobi, we can show $\rho(D^{-1}R) < 1$: if not, then exists ew λ such that $|\lambda| = \rho(D^{-1}R) \ge 1$, and ev $x \ne 0$ such that $D^{-1}Rx = \lambda x$, i.e., $(R + \lambda D)x = 0$ or $R + \lambda D$ invertible, contradiction to A = D + R strictly diagonally dominant given $|\lambda| \ge 1$. Similar for GS.

The theory of general iterative methods suggest using a matrix T with smaller spectrum $\rho(T)$. To this end, we can use the relaxation technique to modify the iterative scheme.

Original Gauss-Seidel iterative method:

$$x^{(k)} = -(L^{-1}U)x^{(k-1)} + L^{-1}b$$

Successive Over-Relaxation¹ (SOR) for Gauss-Seidel iterative method (ω > 1):

$$x^{(k)} = (D - \omega L)^{-1} [(1 - \omega)D + \omega U] x^{(k-1)} + \omega (D - \omega L)^{-1} b$$

where D, -L, -U are the diagonal, strict lower, and strict upper triangular parts of A, respectively.

$${}^{1}Ax = b \Leftrightarrow \omega(-L+D-U)x = \omega b \Leftrightarrow (D-\omega L)x = ((1-\omega)D+\omega U)x + \omega b.$$

Example

Compare Gauss-Seidel and SOR with $\omega = 1.25$, both using $x^{(0)} = (1, 1, 1)^{\top}$ as initial, to solve the system:

$$4x_1 + 3x_2 = 24$$

$$3x_1 + 4x_2 - x_3 = 30$$

$$-x_2 + 4x_3 = -24$$

Solution: Compare with true solution $(3, 4, -5)^{\top}$, we get:

Gauss-Seidel:												
k	0	1	2	3	4	5	6	7				
x ₁ ⁽²⁾	1	5.250000	3.1406250	3.0878906	3.0549316	3.0343323	3.0214577	3.0134110				
$x_{2}^{(2)}$	1	3.812500	3.8828125	3.9667578	3.9542236	3.9713898	3.9821186	3.9888241				
$x_{3}^{(2)}$	1	-5.046875	-5.0292969	-5.0183105	-5.0114441	-5.0071526	-5.0044703	-5.0027940				

Successive Over-Relaxation:

k	0	1	2	3	4	5	6	7
$x_{1}^{(k)}$	1	6.312500	2.6223145	3.1333027	2.9570512	3.0037211	2.9963276	3.0000498
$x_{2}^{(k)}$	1	3.5195313	3.9585266	4.0102646	4.0074838	4.0029250	4.0009262	4.0002586
$x_{3}^{(k)}$	1	-6.6501465	-4.6004238	-5.0966863	-4.9734897	-5.0057135	-4.9982822	-5.0003486

The 5th iteration of SOR is better than 7th of GS.

Theorem (Kahan's theorem)

If all diagonal entries of A are nonzero, then $\rho(T_{\omega}) \ge |\omega - 1|$, where $T_{\omega} = (D - \omega L)^{-1}[(1 - \omega)D + \omega U]$.

Proof. Let $\lambda_1, \ldots, \lambda_n$ be the ew of T_{ω} , then

$$\prod_{i=1}^n \lambda_i = \det(T_\omega) = \det(D)^{-1} \det((1-\omega)D) = (1-\omega)^n$$

since $D - \omega L$ and $(1 - \omega)D + \omega U$ are lower/upper triangular matrices. Hence $\rho(T_{\omega})^n \ge \prod_{i=1}^n |\lambda_i| = |1 - \omega|^n$.

This result says that SOR can converge only if $|\omega - 1| < 1$.

Theorem (Ostrowski-Reich theorem)

If A is positive definite and $|\omega - 1| < 1$, then the SOR converges starting from any initial $x^{(0)}$.

Theorem

If A is positive definite and tridiagonal, then $\rho(T_g) = [\rho(T_j)]^2 < 1$, where T_g and T_j are the T matrices of GS and Jacobi methods respectively, and the optimal ω for SOR is

$$\omega = \frac{2}{1 + \sqrt{1 - (\rho(T_j))^2}}$$

With this choice of ω , the spectrum $\rho(T_{\omega}) = \omega - 1$.

Definition (Residual)

Let \tilde{x} be an approximation to the solution x of linear system Ax = b. Then $r = b - A\tilde{x}$ is called the **residual** of approximation \tilde{x} .

Remark

It seems intuitive that a small residual r implies a close approximation \tilde{x} to x. However, it is not always true.

Example (small residual \Rightarrow small approximation error) The linear system Ax = b is given by

$$\begin{bmatrix} 1 & 2 \\ 1.0001 & 2 \end{bmatrix} \begin{bmatrix} x_1 \\ x_2 \end{bmatrix} = \begin{bmatrix} 3 \\ 3.0001 \end{bmatrix}$$

has a unique solution $x = (1, 1)^{\top}$. Determine the residual vector r of a poor approximation $\tilde{x} = (3, -0.0001)^{\top}$.

Solution: The residual is

$$r = b - A\tilde{x} = \begin{bmatrix} 3\\3.0001 \end{bmatrix} - \begin{bmatrix} 1 & 2\\1.0001 & 2 \end{bmatrix} \begin{bmatrix} 3\\-0.0001 \end{bmatrix} = \begin{bmatrix} 0.0002\\0 \end{bmatrix}$$

So $\|r\|_{\infty} = 0.0002$ is small but $\|\tilde{x} - x\|_{\infty} = 2$ is large.

Theorem (Relation between residual and error)

Suppose A is nonsingular, and \tilde{x} is an approximation to the solution x of Ax = b, and $r = b - A\tilde{x}$ is the residual vector of \tilde{x} , then for any norm, there is

$$||x - \tilde{x}|| \le ||r|| \cdot ||A^{-1}||$$

Moreover, if $x \neq 0$ and $b \neq 0$, then there is

$$\frac{\|x - \tilde{x}\|}{\|x\|} \le \|A\| \cdot \|A^{-1}\| \cdot \frac{\|r\|}{\|b\|}$$

If $||A|| ||A^{-1}||$ is large, then small ||r|| does not guarantee small $||x - \tilde{x}||$.

Proof.

Since x is a solution, we have Ax = b, we have $r = b - A\tilde{x} = Ax - A\tilde{x} = A(x - \tilde{x})$. Since A is nonsingular, we have $x - \tilde{x} = A^{-1}r$, and hence

$$||x - \tilde{x}|| = ||A^{-1}r|| \le ||r|| \cdot ||A^{-1}||$$

If $x \neq 0$ and $b \neq 0$, from $||b|| = ||Ax|| \le ||A|| \cdot ||x||$ we have $1/||x|| \le ||A||/||b||$. Multiplying this to the inequality above, we get

$$\frac{\|x - \tilde{x}\|}{\|x\|} \le \|A\| \cdot \|A^{-1}\| \cdot \frac{\|r\|}{\|b\|}$$

The number $||A|| \cdot ||A^{-1}||$ provide an indication between the error of approximation $||x - \tilde{x}||$ and size of residual *r*. So the larger $||A|| \cdot ||A^{-1}||$ is, the less power we have to control error using residual.

Definition (Condition number)

The condition number of a nonsingular matrix A relative to a norm $\|\cdot\|_p$ is

$$K_p(A) = \|A\|_p \cdot \|A^{-1}\|_p$$

The subscript p is often omitted if it's clear from context or it's not important.

Condition number

Remark

• The condition number $K(A) \ge 1$:

$$1 = \|I\| = \|AA^{-1}\| \le \|A\| \cdot \|A^{-1}\| = K(A)$$

A matrix A is called well-conditioned if K(A) is close to 1.
A matrix A is called ill-conditioned if K(A) >> 1.

Condition number

Example (Condition number)

Determine the condition number of matrix

$$A = \begin{bmatrix} 1 & 2\\ 1.0001 & 2 \end{bmatrix}$$

Condition number

Solution: Let's use I_{∞} norm. Then

$$||A||_{\infty} = \max\{|1| + |2|, |1.0001| + |2|\} = 3.0001$$

Furthermore, there is

$$A^{-1} = \begin{bmatrix} -10000 & 10000\\ 5000.5 & -5000 \end{bmatrix}$$

and hence $\|A^{-1}\|_{\infty} = 20000$. Therefore

 $K(A) = \|A\| \cdot \|A^{-1}\| = 3.0001 \times 20000 = 60002$

Suppose \tilde{x} is our current approximation to x. Let $\tilde{y} = x - \tilde{x}$, then $A\tilde{y} = A(x - \tilde{x}) = Ax - A\tilde{x} = b - A\tilde{x} = r$. If we can solve for \tilde{y} here, we would get a new approximation $\tilde{x} + \tilde{y}$, expectedly to approximate x better.

This procedure is called iterative refinement.

Given A and b, Iterative Refinement first applies Gauss eliminations to Ax = b and obtains approximation x.

Then, for each iteration k = 1, 2, ..., N, do the following:

• Compute residual
$$r = b - Ax$$
;

Solve y from Ay = r using the same Gauss elimination steps.

Set
$$x \leftarrow x + y$$

The actual Iterative Refinement algorithm can also find approximation of condition number $K_{\infty}(A)$ (See textbook).

In reality, A and b may be perturbed by noise or rounding errors δA and δb . Therefore, we are actually solving

$$(A + \delta A)x = b + \delta b$$

rather than Ax = b. This won't cause much issue if A is well-conditioned, but could be a problem otherwise.

Perturbed linear system

Theorem

Suppose A is nonsingular and $\|\delta A\| < \frac{1}{\|A^{-1}\|}$, then the solution \tilde{x} of perturbed linear system $(A + \delta A)x = b + \delta b$ has an error estimate given by

$$\frac{\|x - \tilde{x}\|}{\|x\|} \leq \frac{\mathcal{K}(A)\|A\|}{\|A\| - \mathcal{K}(A)\|\delta A\|} \left(\frac{\|\delta b\|}{\|b\|} + \frac{\|\delta A\|}{\|A\|}\right)$$

where x is the solution of the original linear system Ax = b.

Note that $K(A) \|\delta A\| = \|A\| \|A^{-1}\| \|\delta A\| < \|A\|$ so the denominator is positive.

Conjugate gradient (CG) method is particularly efficient for solving linear systems with large, sparse, and positive definite matrix A.

Equipped with proper preconditioning, CG can often reach very good result in \sqrt{n} iterations (*n* the size of system).

The per-iteration cost is also low when A is sparse.

An alternate perspective of linear system

Theorem

Let A be positive definite, then x^* is the solution of Ax = b iff x^* is the minimizer of

$$g(x) = \frac{1}{2}x^{\top}Ax - b^{\top}x$$

Proof. Note that $\nabla g(x) = Ax - b$ and $\nabla^2 g(x) = A \succ 0$, so $g(x^*) = Ax^* - b = 0$ iff x^* is a minimizer of g(x).

An alternate perspective of linear system

We have following observations:

- ▶ $r = b Ax = -\nabla g(x)$ is the residual and also the steepest descent direction of g(x) (recall that $\nabla g(x)$ is the steepest ascent direction).
- It seems intuitive to update x ← x + t · r = x − t∇g(x) with proper step size t.
- It turns out that we can find such t that makes the most progress.
- This method is called the "steepest descent method".
- However, it converges slowly and exhibits "zigzag" path for ill-conditioned A.

A-orthogonal

Conjugate gradient method amends this issue of steepest descent. To derive CG, we first present the following concept:

Definition

Two vectors v and w are called A-orthogonal if $\langle v, Aw \rangle = 0$.

Theorem

If A is positive definite, then there exists a set of independent vectors $\{v^{(1)}, \dots, v^{(n)}\}$ such that $\langle v^{(i)}, Av^{(j)} \rangle = 0$ for all $i \neq j$.

Key idea of CG

Given previous estimate $x^{(k-1)}$ and a "search direction" $v^{(k)}$, CG will find scalars t_k and s_k to update x and v:

$$x^{(k)} = x^{(k-1)} + t_k v^{(k)}$$
$$v^{(k+1)} = r^{(k)} + s_k v^{(k)}$$

(where $r^{(k)} = b - Ax^{(k)}$), such that: $\langle v^{(k+1)}, Av^{(j)} \rangle = 0, \quad \forall j \le k$ $\langle r^{(k)}, v^{(j)} \rangle = 0, \quad \forall j \le k$

If this can be done, then $\{v^{(1)}, \cdots, v^{(n)}\}$ is A-orthogonal.

The main tool is mathematical induction: given $x^{(0)}$, first set $v^{(0)} = 0$, $r^{(0)} = b - Ax^{(0)}$, $v^{(1)} = r^{(0)}$. So

$$egin{aligned} &\langle m{v}^{(k+1)}, Am{v}^{(j)}
angle &= 0, \quad \forall j \leq k \ &\langle m{r}^{(k)}, m{v}^{(j)}
angle &= 0, \quad \forall j \leq k \end{aligned}$$

is true for k = 0. Assume they hold for k - 1, we need to find t_k and s_k such that they also hold for k.

We first find t_k : note that

$$r^{(k)} = b - Ax^{(k)} = b - A(x^{(k-1)} + t_k v^{(k)}) = r^{(k-1)} - t_k Av^{(k)}$$

Therefore, by induction hypothesis, there is

So we just need

$$t_k = \frac{\langle r^{(k-1)}, v^{(k)} \rangle}{\langle v^{(k)}, A v^{(k)} \rangle}$$

to make $\langle r^{(k)}, v^{(j)} \rangle = 0$.

Then we find s_k : by the update of $v^{(k+1)}$, we have

Note that $Av^{(j)} = \frac{Ax^{(j)} - Ax^{(j-1)}}{t_j} = \frac{r^{(j-1)} - r^{(j)}}{t_j}$, and $r^{(j-1)} - r^{(j)}$ is linear combination of $v^{(j-1)}, v^{(j)}, v^{(j+1)}$, so $\langle r^{(k)}, Av^{(j)} \rangle = 0$ for $j \leq k-1$ due to induction hypothesis. Hence we just need

$$s_k = -rac{\langle r^{(k)}, A v^{(k)}
angle}{\langle v^{(k)}, A v^{(k)}
angle}$$

to make $\langle r^{(k)}, Av^{(j)} \rangle = 0$ for all $j \leq k$.

We can further simply t_k and s_k :

Since that $v^{(k)} = r^{(k-1)} + s_{k-1}v^{(k-1)}$ and $\langle r^{(k-1)}, v^{(k-1)} \rangle = 0$, we have

$$t_{k} = \frac{\langle r^{(k-1)}, v^{(k)} \rangle}{\langle v^{(k)}, Av^{(k)} \rangle} = \frac{\langle r^{(k-1)}, r^{(k-1)} \rangle}{\langle v^{(k)}, Av^{(k)} \rangle}$$

Since $r^{(k-1)} = v^{(k)} - s_{k-1}v^{(k-1)}$, we have $\langle r^{(k)}, r^{(k-1)} \rangle = 0$. Since $Av^{(k)} = \frac{Ax^{(k)} - Ax^{(k-1)}}{t_k} = \frac{r^{(k-1)} - r^{(k)}}{t_k}$, we have $\langle r^{(k)}, Av^{(k)} \rangle = -\frac{\langle r^{(k)}, r^{(k)} \rangle}{t_k}$. Combining t_k expression above, we have

$$s_k = -\frac{\langle r^{(k)}, Av^{(k)} \rangle}{\langle v^{(k)}, Av^{(k)} \rangle} = -\frac{\frac{-\langle r^{(k)}, r^{(k)} \rangle}{t_k}}{\frac{\langle r^{(k-1)}, r^{(k-1)} \rangle}{t_k}} = \frac{\langle r^{(k)}, r^{(k)} \rangle}{\langle r^{(k-1)}, r^{(k-1)} \rangle}$$

Conjugate gradient method

Since $\langle r^{(n)}, v^{(k)} \rangle = 0$ for all k = 1, ..., n and the *A*-orthogonal set $\{v^{(1)}, \dots, v^{(n)}\}$ is independent when *A* is positive definite, we know $r^{(n)} = b - Ax^{(n)} = 0$, i.e., $x^{(n)}$ is the solution.

This shows that CG converges in at most n steps, assuming all arithmetics are exact.

Conjugate gradient method

• Input:
$$x^{(0)}$$
, $r^{(0)} = b - Ax^{(0)}$, $v^{(1)} = r^{(0)}$.

• Repeat the following for k = 1, ..., n until $r^{(k)} = 0$:

$$t_{k} = \frac{\langle r^{(k-1)}, r^{(k-1)} \rangle}{\langle v^{(k)}, Av^{(k)} \rangle}$$
$$x^{(k)} = x^{(k-1)} + t_{k}v^{(k)}$$
$$r^{(k)} = r^{(k-1)} - t_{k}Av^{(k)}$$
$$s_{k} = \frac{\langle r^{(k)}, r^{(k)} \rangle}{\langle r^{(k-1)}, r^{(k-1)} \rangle}$$
$$v^{(k+1)} = r^{(k)} + s_{k}v^{(k)}$$

• Output:
$$x^{(k)}$$
.
Preconditioning

The convergence rate of CG can be greatly improved by **preconditioning**. Preconditioning reduces condition number of A first if A is ill-conditioned. With preconditioning, CG usually converges in \sqrt{n} steps.

The preconditioning is done by using some nonsingular matrix C, we can get $\tilde{A} = C^{-1}A(C^{-1})^{\top}$ such that $K(\tilde{A}) \ll K(A)$.

Now by defining $\tilde{x} = C^{\top}x$ and $\tilde{b} = C^{-1}b$, we obtain a new linear system $\tilde{A}\tilde{x} = \tilde{b}$, which is equivalent to Ax = b. Then we can apply CG to the new system $\tilde{A}\tilde{x} = \tilde{b}$.

Preconditioner

There are various methods to choose the preconditioner C.

- Choose $C = \text{diag}(\sqrt{a_{11}}, \dots, \sqrt{a_{nn}}).$
- Approximate Cholesky's factorization LL^T ≈ A (by ignoring small values in A) and set C = L (then C⁻¹A(C⁻¹)^T ≈ L⁻¹(LL^T)L^{-T} = I).
- Many others...

Preconditioned conjugate gradient method

▶ Input: Preconditioner *C*,
$$x^{(0)}$$
, $r^{(0)} = b - Ax^{(0)}$,
 $w^{(0)} = C^{-1}r^{(0)}$, $v^{(1)} = C^{-T}w^{(0)}$.

• Repeat the following for k = 1, ..., n until $r^{(k)} = 0$:

$$\begin{split} \tilde{t}_{k} &= \frac{\langle w^{(k-1)}, w^{(k-1)} \rangle}{\langle v^{(k)}, Av^{(k)} \rangle} \\ x^{(k)} &= x^{(k-1)} + \tilde{t}_{k} v^{(k)} \\ r^{(k)} &= r^{(k-1)} - \tilde{t}_{k} Av^{(k)} \\ w^{(k)} &= C^{-1} r^{(k)} \\ \tilde{s}_{k} &= \frac{\langle w^{(k)}, w^{(k)} \rangle}{\langle w^{(k-1)}, w^{(k-1)} \rangle} \\ v^{(k+1)} &= C^{-\top} w^{(k)} + \tilde{s}_{k} v^{(k)} \end{split}$$

• Output: $x^{(k)}$.

A comparison

Example

Given A and b below, we use the methods above to solve Ax = b.

$$A = \begin{bmatrix} 0.2 & 0.1 & 1 & 1 & 0 \\ 0.1 & 4 & -1 & 1 & -1 \\ 1 & -1 & 60 & 0 & -2 \\ 1 & 1 & 0 & 8 & 4 \\ 0 & -1 & -2 & 4 & 700 \end{bmatrix}, \quad b = \begin{bmatrix} 1 \\ 2 \\ 3 \\ 4 \\ 5 \end{bmatrix}$$

True solution is

$$x^* = \begin{bmatrix} 7.859713071\\ 0.4229264082\\ -0.07359223906\\ -0.5406430164\\ 0.01062616286 \end{bmatrix}$$

A comparison

A comparison of Jacobi, Gauss-Seidel, SOR, CG, and PCG on the problem above.

Method	Number of Iterations	$\mathbf{x}^{(k)}$	$\ \mathbf{x}^* - \mathbf{x}^{(k)}\ _\infty$
Jacobi	49	$(7.86277141, 0.42320802, -0.07348669, -0.53975964, 0.01062847)^{t}$	0.00305834
Gauss-Seidel	15	(7.83525748, 0.42257868, -0.07319124, -0.53753055, 0.01060903)'	0.02445559
SOR ($\omega = 1.25$)	7	$(7.85152706, 0.42277371, -0.07348303, -0.53978369, 0.01062286)^t$	0.00818607
Conjugate Gradient	5	$(7.85341523, 0.42298677, -0.07347963, -0.53987920, 0.008628916)^{t}$	0.00629785
Conjugate Gradient (Preconditioned)	4	$(7.85968827, 0.42288329, -0.07359878, -0.54063200, 0.01064344)^t$	0.00009312

Section 4

Boundary Value Problems for ODEs

BVP for ODE

We study numerical solution for boundary value problem (BVP).

If the BVP involves first-order ODE, then

$$y'(x) = f(x, y(x)), \quad a \le x \le b, \quad y(a) = \alpha.$$

This reduces to an initial value problem we learned before.

So we start by considering second-order ODE:

$$\begin{cases} y''(x) = f(x, y(x), y'(x)), & a \le x \le b \\ y(a) = \alpha, & y(b) = \beta \end{cases}$$

Existence of solutions

Consider the BVP with second-order ODE:

$$\begin{cases} y''(x) = f(x, y(x), y'(x)), & a \le x \le b \\ y(a) = \alpha, & y(b) = \beta \end{cases}$$

Theorem (Existence and uniqueness of solution) Let $D = [a, b] \times \mathbb{R} \times \mathbb{R}$. Suppose f(x, y, y') satisfies:

- 1. f is continuous on D,
- 2. $\frac{\partial f}{\partial y} > 0$ in D,
- 3. $\exists M > 0$ such that $\left|\frac{\partial f}{\partial y'}\right| \leq M$ in D.

Then the BVP has unique solution.

Example (Existence and uniqueness of solution) Show that the BVP below has unique solution:

$$\begin{cases} y''(x) = -e^{-xy} + \sin(y'), & 1 \le x \le 2\\ y(a) = 0, & y(b) = 0 \end{cases}$$

Solution: We have $f(x, y, y') = -e^{-xy} - \sin(y')$. It is obvious that f is continuous. Moreover $\partial_y f = xe^{-xy} > 0$, and $|\partial_{y'}f| = |-\cos(y')| \le 1$. So the BVP has unique solution by the theorem above.

BVP with linear ODE

Now we first consider a linear second-order ODE:

$$\begin{cases} y'' = p(x)y' + q(x)y + r(x), & a \le x \le b \\ y(a) = \alpha, & y(b) = \beta \end{cases}$$

where $p, q, r : [a, b] \to \mathbb{R}$ are given functions.

Corollary

If p, q, r are continuous on [a, b], q > 0 for all x, then the BVP with linear ODE above has a unique solution.

Proof.

Set f = py' + qy + r. Note that p is bounded since it is continuous on [a, b]. Hence the theorem (check the 3 conditions) above applies.

Now we consider how to solve BVP with linear ODE:

$$\begin{cases} y'' = py' + qy + r, & a \le x \le b\\ y(a) = \alpha, & y(b) = \beta \end{cases}$$

We consider two associated initial value problems:

$$\begin{cases} y_1'' = py_1' + qy_1 + r, & a \le x \le b \\ y_1(a) = \alpha, & y_1'(a) = 0 \\ \end{cases}$$
$$\begin{cases} y_2'' = py_2' + qy_2, & a \le x \le b \\ y_2(a) = 0, & y_2'(a) = 1 \end{cases}$$

Suppose the solution y to the BVP can be written as $y = y_1 + cy_2$ for some constant c (to be determined soon), where y_1, y_2 are the solutions to the two IVPs. Then y satisfies the ODE:

$$y'' = (y_1 + cy_2)'' = y_1'' + cy_2''$$

= $(py_1' + qy_1 + r) + c(py_2' + qy_2)$
= $p(y_1 + cy_2)' + q(y_1 + cy_2) + r$
= $py' + qy + r$

To make y satisfy the boundary conditions, we need c such that

$$y(a) = y_1(a) + cy_2(a) = y_1(a) = \alpha$$

 $y(b) = y_1(b) + cy_2(b) = \beta$

So we just need to set $c = \frac{\beta - y_1(b)}{y_2(b)}$.



Here y_1, y_2 are two shot trajectories based on their initial height and angle. Their linear combination $y_1 + \frac{\beta - y_1(b)}{y_2(b)}y_2$ is the solution y.

Steps of the linear shooting method:

- 1. Partition [a, b] into N equal subintervals.
- 2. Solve y_1 and y_2 from their own IVPs (e.g., using RK4) $(u_1 = y_1, u_2 = y'_1, v_1 = y_2, v_2 = y'_2)$, and get $\{u_{1,i}, v_{1,i} : 0 \le i \le N\}$
- 3. Set $c = (\beta u_{1,N})/v_{1,N}$, and set $w_{1,i} = u_{1,i} + cv_{1,i}$ for $0 \le i \le N$.

Example (Linear shooting method) Solve the BVP with N = 10.

$$\begin{cases} y'' = -\frac{2}{x}y' + \frac{2}{x^2}y + \frac{\sin(\ln x)}{x^2}, & 1 \le x \le 2\\ y(1) = 1, \ y(2) = 2 \end{cases}$$

Solution: Partition [1, 2] into N = 10 subintervals, and solve

$$\begin{cases} y_1'' = -\frac{2}{x}y_1' + \frac{2}{x^2}y_1 + \frac{\sin(\ln x)}{x^2}, & 1 \le x \le 2\\ y_1(1) = 1, & y_1'(1) = 0\\ \begin{cases} y_2'' = -\frac{2}{x}y_2' + \frac{2}{x^2}y_2, & 1 \le x \le 2\\ y_2(1) = 0, & y_2'(1) = 1 \end{cases}$$

Then set
$$w_i = u_{1,i} + \frac{2-u_{1,N}}{v_{1,N}}v_{1,i}$$
 for $i = 0, \dots, 10$.

Numerical result:

xi	$u_{1,i} \approx y_1(x_i)$	$v_{1,i} \approx y_2(x_i)$	$w_i \approx y(x_i)$	$y(x_i)$	$ y(x_i) - w_i $
1.0	1.00000000	0.00000000	1.00000000	1.00000000	
1.1	1.00896058	0.09117986	1.09262917	1.09262930	$1.43 imes10^{-7}$
1.2	1.03245472	0.16851175	1.18708471	1.18708484	$1.34 imes10^{-7}$
1.3	1.06674375	0.23608704	1.28338227	1.28338236	$9.78 imes10^{-8}$
1.4	1.10928795	0.29659067	1.38144589	1.38144595	$6.02 imes10^{-8}$
1.5	1.15830000	0.35184379	1.48115939	1.48115942	$3.06 imes10^{-8}$
1.6	1.21248372	0.40311695	1.58239245	1.58239246	$1.08 imes10^{-8}$
1.7	1.27087454	0.45131840	1.68501396	1.68501396	$5.43 imes10^{-10}$
1.8	1.33273851	0.49711137	1.78889854	1.78889853	$5.05 imes10^{-9}$
1.9	1.39750618	0.54098928	1.89392951	1.89392951	$4.41 imes10^{-9}$
2.0	1.46472815	0.58332538	2.00000000	2.00000000	

This accurate result is due to $O(h^4)$ of RK4 used for the two IVPs.

Round-off error in linear shooting method

If $y_1(x)$ grows too fast such that $y_1(b) \gg \beta$, then

$$rac{eta-y_1(b)}{y_2(b)}pprox -rac{y_1(b)}{y_2(b)}$$

which is prone to round-off error.

In this case, we can solve the two IVPs backward in x:

$$\begin{cases} y_1'' = py_1' + qy_1 + r, & a \le x \le b \\ y_1(b) = \beta, & y_1'(b) = 0 \\ \end{cases}$$
$$\begin{cases} y_2'' = py_2' + qy_2, & a \le x \le b \\ y_2(b) = 0, & y_2'(b) = 1 \end{cases}$$

and set $y(x) = y_1(x) + \frac{\alpha - y_1(a)}{y_2(a)}y_2(x)$ for $a \le x \le b$

Nonlinear shooting method

Consider the BVP with nonlinear ODE (f is a nonlinear function):

$$\begin{cases} y'' = f(x, y, y'), & a \le x \le b \\ y(a) = \alpha, & y(b) = \beta \end{cases}$$

Suppose we try to solve the IVP with some given *t*:

$$\begin{cases} y'' = f(x, y, y'), & a \le x \le b \\ y(a) = \alpha, & y'(a) = t \end{cases}$$

and obtain solution y(x, t) (since the solution depends on t) for $a \le x \le b$.

Then we hope to find t such that $y(b, t) = \beta$.

Secant method for nonlinear shooting

Suppose we have two initials t_0, t_1 , then we use the secant method to solve $y(b, t) - \beta = 0$ by iterating

$$t_k = t_{k-1} - rac{(y(b,t_{k-1})-eta)(t_{k-1}-t_{k-2})}{y(b,t_{k-1})-y(b,t_{k-2})}$$

For each k, we need to compute $y(b, t_k)$ by solving the IVP:

$$\begin{cases} y'' = f(x, y, y'), & a \le x \le k \\ y(a) = \alpha, & y'(a) = t_k \end{cases}$$

Nonlinear shooting method



Here $y(x, t_k)$ is "shooting" at an angle (with slope t_k) and try to "hit" β at x = b.

We can also consider Newton's method to $y(b, t) - \beta = 0$ for fewer iterations:

$$t_k = t_{k-1} - \frac{y(b, t_{k-1}) - \beta}{\partial_t y(b, t_{k-1})}$$

However, we need to know $\partial_t y(b, t)$...

We denote the solution of IVP below by y(x, t):

$$\begin{cases} y''(x,t) = f(x,y(x,t),y'(x,t)), & a \le x \le b \\ y(a,t) = \alpha, & y'(a,t) = t \end{cases}$$

where $y' = \partial_x y$ and $y'' = \partial_x^2 y$ (i.e., the ' is on x).

Taking partial derivatives with respect to t above yields:

$$\begin{cases} \partial_t y'' = \partial_y f \cdot \partial_t y + \partial_{y'} f \cdot \partial_t y', & a \le x \le b \\ \partial_t y(a, t) = 0, & \partial_t y'(a, t) = 1 \end{cases}$$

Denote $z(x,t) = \partial_t y(x,t)$. Suppose ∂_x and ∂_t can exchange, then

$$\begin{cases} z''(x,t) = \partial_y f \cdot z(x,t) + \partial_{y'} f \cdot z'(x,t), & a \le x \le b \\ z(a,t) = 0, & z'(a,t) = 1 \end{cases}$$

and set $\partial_t y(b, t) = z(b, t)$.

Steps of Newton's method for nonlinear shooting:

- 1. Initialize t_0 (e.g. $t_0 = \frac{\beta \alpha}{b a}$). Set k = 1.
- 2. For $t = t_{k-1}$, solve y(x, t) and z(x, t) from

$$\begin{cases} y''(x,t) = f(x,y(x,t),y'(x,t)), & a \le x \le b \\ y(a,t) = \alpha, & y'(a,t) = t \\ \begin{cases} z''(x,t) = \partial_y f \cdot z(x,t) + \partial_{y'} f \cdot z'(x,t), & a \le x \le b \\ z(a,t) = 0, & z'(a,t) = 1 \end{cases}$$

and set
$$t_k = t_{k-1} - \frac{y(b, t_{k-1}) - \beta}{z(b, t_{k-1})}$$
.
3. Set $k \leftarrow k + 1$ and go to Step 2.

Example (Newton's method for nonlinear BVP)

Solve the BVP with nonlinear ODE using Newton's method with N = 20 for maximal of 10 iterations or $|w_N(t_k) - y(3)| \le 10^{-5}$:

$$\begin{cases} y'' = \frac{1}{8} \left(32 + 2x^3 - yy' \right), & 1 \le x \le 3\\ y(1) = 17, \ y(3) = \frac{43}{3} \end{cases}$$

Solution: Note that $\partial_y f = -\frac{1}{8}y'$ and $\partial_{y'} f = -\frac{1}{8}y$. For every *t*, the two IVPs are (note *z* depends on *y* but not vice versa):

$$\begin{cases} y'' = \frac{1}{8} \left(32 + 2x^3 - yy' \right), & 1 \le x \le 3 \\ y(1) = 17, & y'(1) = t \\ \\ z'' = -\frac{1}{8} (y'z + yz'), & 1 \le x \le 3 \\ z(1) = 0, & z'(1) = 1 \end{cases}$$

Nonlinear shooting using Newton's method

xi	w _{1,i}	$y(x_i)$	$ w_{1,i} - y(x_i) $
1.0	17.000000	17.000000	
1.1	15.755495	15.755455	4.06×10^{-5}
1.2	14.773389	14.773333	5.60×10^{-5}
1.3	13.997752	13.997692	5.94×10^{-5}
1.4	13.388629	13.388571	5.71×10^{-5}
1.5	12.916719	12.916667	5.23×10^{-5}
1.6	12.560046	12.560000	$4.64 imes 10^{-5}$
1.7	12.301805	12.301765	4.02×10^{-5}
1.8	12.128923	12.128889	3.14×10^{-5}
1.9	12.031081	12.031053	2.84×10^{-5}
2.0	12.000023	12.000000	2.32×10^{-5}
2.1	12.029066	12.029048	1.84×10^{-5}
2.2	12.112741	12.112727	1.40×10^{-5}
2.3	12.246532	12.246522	1.01×10^{-5}
2.4	12.426673	12.426667	$6.68 imes 10^{-6}$
2.5	12.650004	12.650000	3.61×10^{-6}
2.6	12.913847	12.913845	9.17×10^{-7}
2.7	13.215924	13.215926	1.43×10^{-6}
2.8	13.554282	13.554286	$3.46 imes 10^{-6}$
2.9	13.927236	13.927241	$5.21 imes 10^{-6}$
3.0	14.333327	14.333333	$6.69 imes10^{-6}$

Netwon's method requires solving two IVPs in each iteration, but converges much faster than secant method. Still sensitive to round-off errors if y or z increases rapidly.

Idea: Partition [a, b] into N + 1 subintervals with nodes $a = x_0 < \cdots < x_{N+1} = b$ and step size $h = \frac{b-a}{N+1}$. Then approximate y', y'' by finite differences, and solve $w_i = y(x_i)$ for $0 \le i \le N + 1$.

Recall the centered-difference approximation of $y'(x_i)$:

$$y(x_{i+1}) = y(x_i + h) = y(x_i) + hy'(x_i) + \frac{h^2}{2}y''(x_i) + \frac{h^3}{6}y'''(\eta_i^+)$$

$$y(x_{i-1}) = y(x_i - h) = y(x_i) - hy'(x_i) + \frac{h^2}{2}y''(x_i) - \frac{h^3}{6}y'''(\eta_i^-)$$

where η_i^{\pm} is between x_i and $x_{i\pm 1}$. Then subtracting the two above:

$$y'(x_i) = \frac{y(x_{i+1}) - y(x_{i-1})}{2h} - \frac{h^2}{6}y'''(\eta_i)$$

for some $\eta_i \in (x_{i-1}, x_{i+1})$ due to IVT and $y \in C^3$.

Similarly, we have the centered-difference approximation of $y''(x_i)$:

$$y(x_{i+1}) = y(x_i + h) = y(x_i) + hy'(x_i) + \frac{h^2}{2}y''(x_i) + \frac{h^3}{6}y'''(x_i) + \frac{h^4}{24}y^{(4)}(\xi_i^+)$$

$$y(x_{i-1}) = y(x_i - h) = y(x_i) - hy'(x_i) + \frac{h^2}{2}y''(x_i) - \frac{h^3}{6}y'''(x_i) + \frac{h^4}{24}y^{(4)}(\xi_i^-)$$

where ξ_i^{\pm} is between x_i and $x_{i\pm 1}$. Then adding the two above:

$$y''(x_i) = \frac{y(x_{i+1}) - 2y(x_i) + y(x_{i-1})}{h^2} - \frac{h^2}{12}y^{(4)}(\xi_i)$$

for some $\xi_i \in (x_{i-1}, x_{i+1})$ due to IVT and $y \in C^4$.

Plugging the two identities about $y'(x_i)$ and $y''(x_i)$ above into y'' = py' + qy + r:

$$\frac{y(x_{i+1}) - 2y(x_i) + y(x_{i-1})}{h^2} = p(x_i) \left[\frac{y(x_{i+1}) - y(x_{i-1})}{2h} \right] + q(x_i) y(x_i)$$
$$+ r(x_i) - \frac{h^2}{12} \left[2p(x_i) y'''(\eta_i) - y^{(4)}(\xi_i) \right]$$

which has truncation error $O(h^2)$.

Now we approximate $y(x_i)$ by w_i for $0 \le i \le N + 1$. Note that $w_0 = y(a) = \alpha$ and $w_{N+1} = y(b) = \beta$, and for $i = 1, \dots, N$:

$$\left(\frac{-w_{i+1}+2w_{i}-w_{i-1}}{h^{2}}\right)+p(x_{i})\left(\frac{w_{i+1}-w_{i-1}}{2h}\right)+q(x_{i})w_{i}=-r(x_{i})$$

The equation above can be rearranged into

$$-\left(1+\frac{h}{2}p(x_{i})\right)w_{i-1}+\left(2+h^{2}q(x_{i})\right)w_{i}-\left(1-\frac{h}{2}p(x_{i})\right)w_{i+1}=-h^{2}r(x_{i})$$

This is a linear system Aw = b where $w = (w_1, \ldots, w_N)^{\top}$, A is tridiagonal, and b is known.

Theorem

Suppose that p, q, r are continuous on [a, b] and $q \ge 0$, then the tridiagonal linear system Aw = b has a unique solution provided that h < 2/L where $L = \max_{a \le x \le b} |p(x)|$.

Example (Finite-difference method for linear problems) Solve the BVP below using finite difference method with N = 9:

$$\begin{cases} y'' = -\frac{2}{x}y' + \frac{2}{x^2}y + \frac{\sin(\ln x)}{x^2}, & 1 \le x \le 2\\ y(1) = 1, \ y(2) = 2 \end{cases}$$

Solution: Note that p(x) = -2/x, $q(x) = 2/x^2$, and $r(x) = \sin(\ln x)/x^2$. Step size h = (b - a)/(N + 1) = 0.1.

xi	Wi	$y(x_i)$	$ w_i - y(x_i) $
1.0	1.00000000	1.00000000	
1.1	1.09260052	1.09262930	$2.88 imes10^{-5}$
1.2	1.18704313	1.18708484	$4.17 imes10^{-5}$
1.3	1.28333687	1.28338236	$4.55 imes10^{-5}$
1.4	1.38140205	1.38144595	$4.39 imes10^{-5}$
1.5	1.48112026	1.48115942	$3.92 imes10^{-5}$
1.6	1.58235990	1.58239246	$3.26 imes10^{-5}$
1.7	1.68498902	1.68501396	$2.49 imes10^{-5}$
1.8	1.78888175	1.78889853	$1.68 imes10^{-5}$
1.9	1.89392110	1.89392951	$8.41 imes10^{-6}$
2.0	2.00000000	2.00000000	

The error is $O(h^2)$, which is worse than the linear shooting method.

We can improve the error order by Richardson's extrapolation since the truncation errors are in even orders of h.

Consider the same example above, we use step sizes h = 0.1, 0.05, and 0.025 to compute w(h = 0.1), w(h = 0.05) and w(h = 0.025), and compute

$$Ext_{1i} = \frac{4w_i(h = 0.05) - w_i(h = 0.1)}{3}$$

$$Ext_{2i} = \frac{4w_i(h = 0.025) - w_i(h = 0.05)}{3}$$

$$Ext_{3i} = \frac{16Ext_{2i} - Ext_{1i}}{15}$$

xi	$w_i(h = 0.05)$	$w_i(h = 0.025)$	Ext _{1i}	Ext _{2i}	Ext _{3i}
1.0	1.00000000	1.00000000	1.00000000	1.00000000	1.00000000
1.1	1.09262207	1.09262749	1.09262925	1.09262930	1.09262930
1.2	1.18707436	1.18708222	1.18708477	1.18708484	1.18708484
1.3	1.28337094	1.28337950	1.28338230	1.28338236	1.28338236
1.4	1.38143493	1.38144319	1.38144589	1.38144595	1.38144595
1.5	1.48114959	1.48115696	1.48115937	1.48115941	1.48115942
1.6	1.58238429	1.58239042	1.58239242	1.58239246	1.58239246
1.7	1.68500770	1.68501240	1.68501393	1.68501396	1.68501396
1.8	1.78889432	1.78889748	1.78889852	1.78889853	1.78889853
1.9	1.89392740	1.89392898	1.89392950	1.89392951	1.89392951
2.0	2.00000000	2.00000000	2.00000000	2.00000000	2.00000000

The error reduces to 6.3×10^{-11} which significantly improves the case with h = 0.1 (about 10^{-5}).

Consider the BVP with nonlinear ODE:

$$\begin{cases} y'' = f(x, y, y'), & a \le x \le b \\ y(a) = \alpha, & y(b) = \beta \end{cases}$$

Theorem

Let $D = [a, b] \times \mathbb{R} \times \mathbb{R}$. If f satisfies the following conditions:

- 1. f is continuous on D,
- 2. $\exists \delta > 0$ such that $\partial_y f(x, y, y') \ge \delta$ on D,
- 3. $\exists L > 0$ such that $|\partial_y f|, |\partial_{y'} f| \leq L$ on D.

Then the BVP has a unique solution.

We apply the same partition of [a, b] into N + 1 subintervals and centered-difference approximations for $y'(x_i)$ and $y''(x_i)$: $w_0 = \alpha$, $w_{N+1} = \beta$, and for i = 1, ..., N

$$-\frac{w_{i+1}-2w_i+w_{i-1}}{h^2}+f\left(x_i,w_i,\frac{w_{i+1}-w_{i-1}}{2h}\right)=0$$

This is a system of N nonlinear equations of (w_1, \ldots, w_N) :

$$2w_{1} - w_{2} + h^{2}f\left(x_{1}, w_{1}, \frac{w_{2} - \alpha}{2h}\right) - \alpha = 0$$
$$-w_{1} + 2w_{2} - w_{3} + h^{2}f\left(x_{2}, w_{2}, \frac{w_{3} - w_{1}}{2h}\right) = 0$$
$$\vdots$$
$$w_{N-2} + 2w_{N-1} - w_{N} + h^{2}f\left(x_{N-1}, w_{N-1}, \frac{w_{N} - w_{N-2}}{2h}\right) = 0$$
$$-w_{N-1} + 2w_{N} + h^{2}f\left(x_{N}, w_{N}, \frac{\beta - w_{N-1}}{2h}\right) - \beta = 0$$

We can write the system as F(w) = 0 (note that $F : \mathbb{R}^N \to \mathbb{R}^N$).

To solve this system, we can apply the Newton's method:

$$w^{(k)} = w^{(k-1)} - J(w^{(k-1)})^{-1}F(w^{(k-1)})$$

starting from some initial value $w^{(0)}$. Here $J(w) \in \mathbb{R}^{N \times N}$ is the Jacobian of F(w).

The key is to solve $v = J(w)^{-1}F(w)$ from J(w)v = F(w) for given w.
Finite-difference method for nonlinear problems

Recall that
$$F(w) = (F_1(w), \dots, F_N(w))^\top \in \mathbb{R}^N$$
 where
 $F_i(w) = -w_{i-1} + 2w_i - w_{i+1} + h^2 f\left(x_i, w_i, \frac{w_{i+1} - w_{i-1}}{2h}\right)$

Jacobian $J(w) = [\frac{\partial F_i(w)}{\partial w_j}] \in \mathbb{R}^{N \times N}$ is tridiagonal:

$$J(w_1, \dots, w_N)_{ij} = \frac{\partial F_i(w)}{\partial w_j}$$

$$= \begin{cases}
-1 + \frac{h}{2} f_{y'}\left(x_i, w_i, \frac{w_{i+1} - w_{i-1}}{2h}\right), & \text{for } i = j - 1 \text{ and } j = 2, \dots, N \\
2 + h^2 f_y\left(x_i, w_i, \frac{w_{i+1} - w_{i-1}}{2h}\right), & \text{for } i = j \text{ and } j = 1, \dots, N \\
-1 - \frac{h}{2} f_{y'}\left(x_i, w_i, \frac{w_{i+1} - w_{i-1}}{2h}\right), & \text{for } i = j + 1 \text{ and } j = 1, \dots, N - 1 \\
0, & \text{for } |i - j| > 1
\end{cases}$$

Example (Finite-difference method for nonlinear BVP) Solve the BVP with nonlinear ODE using finite difference method with h = 0.1:

$$\begin{cases} y'' = \frac{1}{8} \left(32 + 2x^3 - yy' \right), & 1 \le x \le 3\\ y(1) = 17, \ y(3) = \frac{43}{3} \end{cases}$$

Finite-difference method for nonlinear problems

xi	Wi	$y(x_i)$	$ w_i - y(x_i) $
1.0	17.000000	17.000000	
1.1	15.754503	15.755455	$9.520 imes10^{-4}$
1.2	14.771740	14.773333	$1.594 imes10^{-3}$
1.3	13.995677	13.997692	$2.015 imes10^{-3}$
1.4	13.386297	13.388571	$2.275 imes10^{-3}$
1.5	12.914252	12.916667	$2.414 imes10^{-3}$
1.6	12.557538	12.560000	$2.462 imes10^{-3}$
1.7	12.299326	12.301765	$2.438 imes 10^{-3}$
1.8	12.126529	12.128889	$2.360 imes10^{-3}$
1.9	12.028814	12.031053	$2.239 imes10^{-3}$
2.0	11.997915	12.000000	$2.085 imes10^{-3}$
2.1	12.027142	12.029048	$1.905 imes10^{-3}$
2.2	12.111020	12.112727	$1.707 imes10^{-3}$
2.3	12.245025	12.246522	$1.497 imes10^{-3}$
2.4	12.425388	12.426667	$1.278 imes10^{-3}$
2.5	12.648944	12.650000	$1.056 imes10^{-3}$
2.6	12.913013	12.913846	$8.335 imes10^{-4}$
2.7	13.215312	13.215926	$6.142 imes10^{-4}$
2.8	13.553885	13.554286	$4.006 imes10^{-4}$
2.9	13.927046	13.927241	$1.953 imes10^{-4}$
3.0	14.333333	14.333333	

Finite-difference method for nonlinear problems

The error order again can be improved by Richardson's extrapolation: solve the problem with h = 0.1, 0.05, and 0.025, and then use extrapolation as before. Accuracy can be improved from 10^{-3} to 10^{-10} .

Idea: Convert the BVP to an integral minimization problem, and then find the minimizer from the function space spanned by a set of basis functions.

We consider a standard BVP with second-order ODE:

$$\begin{cases} -\frac{\mathrm{d}}{\mathrm{d}x} \left(p(x) \frac{\mathrm{d}y}{\mathrm{d}x} \right) + q(x)y = f, & 0 \le x \le 1\\ y(0) = 0, & y(1) = 0 \end{cases}$$

Problems with general interval [a, b] and boundary conditions $y(a) = \alpha$, $y(b) = \beta$ can be converted into the standard one above. For example, if $y(0) = \alpha$, $y(1) = \beta$, then set $z(x) = y(x) - ((1 - x)\alpha + x\beta)$ and derive the ODE of z with boundary value z(0) = z(1) = 0.

Theorem (Variational form of BVP)

Suppose $p \in C^1$, $q, f \in C$, $p \ge \delta$ for some $\delta > 0$ and $q \ge 0$ on [0, 1], and $y \in C_0^2$, then y is the unique solution to

$$-\frac{\mathsf{d}}{\mathsf{d}x}\left(p(x)\frac{\mathsf{d}y}{\mathsf{d}x}\right) + q(x)y = f, \quad 0 \le x \le 1 \quad \underline{ODE}$$

if and only if y is the unique function that minimizes I[·] where

$$I[u] = \int_0^1 \left(p(x)[u'(x)]^2 + q(x)[u(x)]^2 - 2f(x)u(x) \right) dx \quad \underline{Energy}$$

Proof.

1. A solution y to (ODE) satisfies:

$$\int_0^1 f(x)u(x)dx = \int_0^1 p(x)\frac{dy}{dx}(x)\frac{du}{dx}(x) + q(x)y(x)u(x)dx, \quad \forall u \in C_0^1[0,1] \quad \underline{Weak}$$

This can be verified by multiplying u on both sides of <u>ODE</u>, taking integral, and integrating by part.

- 2. *y* minimizes Energy iff *y* satisfies Weak: For any *y*, $u \in C_0^1[0, 1]$, define $g(\epsilon) = I[y + \epsilon u]$, then $g''(\epsilon) \ge 0$, so *I* is a convex functional. Therefore *y* minimizes Energy iff g'(0) = 0 for all *u* (i.e., *y* satisfies Weak).
- 3. Weak admits at most one solution: if y_1, y_2 both satisfies Weak, then $y = y_1 - y_2$ satisfies Weak with f = 0, i.e., y minimizes $J[u] = \int_0^1 (p(u')^2 + qu^2) dx$. Hence $y \equiv 0$ (since $J[u] \ge 0$ and = 0 only if $u \equiv 0$).

Now we know BVP is equivalent to an energy minimization problem:

$$I[u] = \int_0^1 \left(p(x)[u'(x)]^2 + q(x)[u(x)]^2 - 2f(x)u(x) \right) dx$$

Steps of Rayleigh-Ritz method:

- 1. Create a set of basis functions $\{\phi_i \mid 1 \le i \le n\}$, and set approximation $\phi = \sum_i c_i \phi_i$ to $y = \operatorname{argmin}_u I[u]$.
- 2. Find *c* by minimizing $I[\phi] = I[\sum_i c_i \phi_i]$, i.e., $\partial_{c_i} I[\sum_i c_i \phi_i] = 0$ for all *i*.

Step 2 above yields a linear **normal equation** of *c*, denoted by Ac = b, where $A = [a_{ij}] \in \mathbb{R}^{n \times n}$ and $b \in \mathbb{R}^n$ with

$$\begin{aligned} \mathsf{a}_{ij} &= \int_0^1 \left[\mathsf{p}(x)\phi_i'(x)\phi_j'(x) + \mathsf{q}(x)\phi_i(x)\phi_j(x) \right] dx \\ \mathsf{b}_i &= \int_0^1 f(x)\phi_i(x)dx \end{aligned}$$

Once c is solved, the minimizer of I can be set to $\phi = \sum_i c_i \phi_i$.

Now the key is the design of basis functions in Step 1. If properly designed, A will be a band matrix (and even tridiagonal matrix).

Steps to create a piecewise linear basis:

1. Partition [0, 1] into n + 1 subintervals:

$$0 = x_0 < x_1 < \cdots < x_{n+1} = 1$$

Step size
$$h_i = x_{i+1} - x_i$$
 for $i = 0, ..., n$.
2. Set $\{\phi_i\}$ for $i = 1, ..., n$ by:

$$\phi_i(x) = \begin{cases} 0, & \text{if } 0 \le x \le x_{i-1} \\ \frac{1}{h_{i-1}} (x - x_{i-1}), & \text{if } x_{i-1} < x \le x_i \\ \frac{1}{h_i} (x_{i+1} - x), & \text{if } x_i < x \le x_{i+1} \\ 0, & \text{if } x_{i+1} < x \le 1 \end{cases}$$

Namely, $\phi_i(x)$ is 1 at $x = x_i$ and linearly decays to 0 at $x = x_{i\pm 1}$, then stays as 0 outside of $[x_{i-1}, x_{i+1}]$.

Example of piecewise linear basis functions:



Several properties about piecewise linear basis:

1. ϕ_i is differentiable except at x_{i-1}, x_i, x_{i+1} :

$$\phi_i'(x) = \begin{cases} 0, & \text{if } 0 < x < x_{i-1} \\ \frac{1}{h_{i-1}}, & \text{if } x_{i-1} < x < x_i \\ -\frac{1}{h_i}, & \text{if } x_i < x < x_{i+1} \\ 0, & \text{if } x_{i+1} < x < 1 \end{cases}$$

2. ϕ_i and ϕ_j do not interfere if |i - j| > 1:

$$\phi_i(x)\phi_j(x)\equiv 0$$
 and $\phi_i'(x)\phi_j'(x)\equiv 0$

Hence $A = [a_{ij}]$ in the normal equation Ac = b is a tridiagonal matrix.

$$\begin{aligned} \mathbf{a}_{ii} &= \int_{0}^{1} \left\{ p(\mathbf{x}) \left[\phi_{i}'(\mathbf{x}) \right]^{2} + q(\mathbf{x}) \left[\phi_{i}(\mathbf{x}) \right]^{2} \right\} d\mathbf{x} \\ &= \left(\frac{1}{h_{i-1}} \right)^{2} \int_{x_{i-1}}^{x_{i}} p(\mathbf{x}) d\mathbf{x} + \left(\frac{-1}{h_{i}} \right)^{2} \int_{x_{i}}^{x_{i+1}} p(\mathbf{x}) d\mathbf{x} \\ &+ \left(\frac{1}{h_{i-1}} \right)^{2} \int_{x_{i-1}}^{x_{i}} (\mathbf{x} - \mathbf{x}_{i-1})^{2} q(\mathbf{x}) d\mathbf{x} + \left(\frac{1}{h_{i}} \right)^{2} \int_{x_{i}}^{x_{i+1}} (\mathbf{x}_{i+1} - \mathbf{x})^{2} q(\mathbf{x}) d\mathbf{x} \\ \mathbf{a}_{i,i+1} &= \int_{0}^{1} \left\{ p(\mathbf{x}) \phi_{i}'(\mathbf{x}) \phi_{i+1}'(\mathbf{x}) + q(\mathbf{x}) \phi_{i}(\mathbf{x}) \phi_{i+1}(\mathbf{x}) \right\} d\mathbf{x} \\ &= - \left(\frac{1}{h_{i}} \right)^{2} \int_{x_{i}}^{x_{i+1}} p(\mathbf{x}) d\mathbf{x} + \left(\frac{1}{h_{i}} \right)^{2} \int_{x_{i}}^{x_{i+1}} (\mathbf{x}_{i+1} - \mathbf{x}) (\mathbf{x} - \mathbf{x}_{i}) q(\mathbf{x}) d\mathbf{x} \\ \mathbf{a}_{i,i-1} &= \int_{0}^{1} \left\{ p(\mathbf{x}) \phi_{i}'(\mathbf{x}) \phi_{i-1}'(\mathbf{x}) + q(\mathbf{x}) \phi_{i}(\mathbf{x}) \phi_{i-1}(\mathbf{x}) \right\} d\mathbf{x} \\ &= - \left(\frac{1}{h_{i-1}} \right)^{2} \int_{x_{i-1}}^{x_{i}} p(\mathbf{x}) d\mathbf{x} + \left(\frac{1}{h_{i-1}} \right)^{2} \int_{x_{i-1}}^{x_{i}} (\mathbf{x}_{i} - \mathbf{x}) (\mathbf{x} - \mathbf{x}_{i-1}) q(\mathbf{x}) d\mathbf{x} \\ &= \int_{0}^{1} f(\mathbf{x}) \phi_{i}(\mathbf{x}) d\mathbf{x} = \frac{1}{h_{i-1}} \int_{x_{i-1}}^{x_{i}} (\mathbf{x} - \mathbf{x}_{i-1}) f(\mathbf{x}) d\mathbf{x} + \frac{1}{h_{i}} \int_{x_{i}}^{x_{i+1}} (\mathbf{x}_{i+1} - \mathbf{x}) f(\mathbf{x}) d\mathbf{x} \end{aligned}$$

There are 6n integrals to evaluate:

$$\begin{aligned} Q_{1,i} &= \left(\frac{1}{h_i}\right)^2 \int_{x_i}^{x_{i+1}} (x_{i+1} - x) (x - x_i) q(x) dx, & \text{ for each } i = 1, 2, \dots, n-1 \\ Q_{2,i} &= \left(\frac{1}{h_{i-1}}\right)^2 \int_{x_{i-1}}^{x_i} (x - x_{i-1})^2 q(x) dx, & \text{ for each } i = 1, 2, \dots, n \\ Q_{3,i} &= \left(\frac{1}{h_i}\right)^2 \int_{x_i}^{x_{i+1}} (x_{i+1} - x)^2 q(x) dx, & \text{ for each } i = 1, 2, \dots, n \\ Q_{4,i} &= \left(\frac{1}{h_{i-1}}\right)^2 \int_{x_{i-1}}^{x_i} p(x) dx, & \text{ for each } i = 1, 2, \dots, n+1 \\ Q_{5,i} &= \frac{1}{h_{i-1}} \int_{x_{i-1}}^{x_i} (x - x_{i-1}) f(x) dx, & \text{ for each } i = 1, 2, \dots, n \\ Q_{6,i} &= \frac{1}{h_i} \int_{x_i}^{x_{i+1}} (x_{i+1} - x) f(x) dx, & \text{ for each } i = 1, 2, \dots, n \end{aligned}$$

Then A and b are computed as

$$\begin{aligned} a_{i,i} &= Q_{4,i} + Q_{4,i+1} + Q_{2,i} + Q_{3,i}, & \text{for each } i = 1, 2, \dots, n \\ a_{i,i+1} &= -Q_{4,i+1} + Q_{1,i}, & \text{for each } i = 1, 2, \dots, n-1 \\ a_{i,i-1} &= -Q_{4,i} + Q_{1,i-1}, & \text{for each } i = 2, 3, \dots, n \\ b_i &= Q_{5,i} + Q_{6,i}, & \text{for each } i = 1, 2, \dots, n \end{aligned}$$

We can show that A is positive definite.

Two ways to approximate the 6n integrals Q's:

- 1. Quadratures such as Simpson's rule.
- 2. Approximate p, q, r by piecewise linear functions and compute integrals. For example, $p(x) \approx \sum_{i} p(x_i)\phi_i(x)$ etc., then

$$\begin{aligned} &Q_{1,i} \approx \frac{h_i}{12} [q(x_i) + q(x_{i+1})] \\ &Q_{2,i} \approx \frac{h_{i-1}}{12} \left[3q(x_i) + q(x_{i-1}) \right] \\ &Q_{3,i} \approx \frac{h_i}{12} \left[3q(x_i) + q(x_{i+1}) \right] \\ &Q_{4,i} \approx \frac{h_{i-1}}{2} \left[p(x_i) + p(x_{i-1}) \right] \\ &Q_{5,i} \approx \frac{h_{i-1}}{6} \left[2f(x_i) + f(x_{i-1}) \right] \\ &Q_{6,i} \approx \frac{h_i}{6} \left[2f(x_i) + f(x_{i+1}) \right] \end{aligned}$$

Each approximation has error order $O(h_i^3)$.

Example (Rayleigh-Ritz method with piecewise linear basis) Solve the BVP below using Rayleigh-Ritz method and piecewise linear basis with $h_i = h = 0.1$:

$$-y'' + \pi^2 y = 2\pi^2 \sin(\pi x), \quad 0 \le x \le 1, \quad y(0) = 0, \ y(1) = 0$$

Solution: We have $p(x) \equiv 1$, $q(x) \equiv \pi^2$, $f(x) = 2\pi^2 \sin(\pi x)$. Then apply the formula above to obtain $Q_{1,i}, \ldots, Q_{6,i}$ for $i = 0, \ldots, 9$, and then A and b. Then solve c from Ac = b, and obtain $\phi(x) = \sum_i c_i \phi_i(x)$ (note that $\phi(x)$ is piecewise linear function and $\phi(x_i) = c_i$).

i	Xi	$\phi(\mathbf{x}_i)$	$y(x_i)$	$ \phi(x_i) - y(x_i) $
1	0.1	0.3102866742	0.3090169943	0.00127
2	0.2	0.5902003271	0.5877852522	0.00241
3	0.3	0.8123410598	0.8090169943	0.00332
4	0.4	0.9549641896	0.9510565162	0.00390
5	0.5	1.0041087710	1.0000000000	0.00411
6	0.6	0.9549641893	0.9510565162	0.00390
7	0.7	0.8123410598	0.8090169943	0.00332
8	0.8	0.5902003271	0.5877852522	0.00241
9	0.9	0.3102866742	0.3090169943	0.00127

The error order is $O(h^2)$ due to the nature of linear (first-order) approximation of the integand.

We can create C^2 basis functions using the idea of cubic splines. These are called the **B**-splines (basis splines).

We start from the cubic spline function S:

$$S(x) = \begin{cases} 0, & \text{if } x \leq -2 \\ \frac{1}{4}(2+x)^3, & \text{if } -2 \leq x \leq -1 \\ \frac{1}{4}\left[(2+x)^3 - 4(1+x)^3\right], & \text{if } -1 < x \leq 0 \\ \frac{1}{4}\left[(2-x)^3 - 4(1-x)^3\right], & \text{if } 0 < x \leq 1 \\ \frac{1}{4}(2-x)^3, & \text{if } 1 < x \leq 2 \\ 0, & \text{if } 2 < x \end{cases}$$

Then construct B-spline basis functions $\{\phi_i \mid 0 \le i \le n+1\}$:

$$\phi_i(x) = \begin{cases} S\left(\frac{x}{h}\right) - 4S\left(\frac{x+h}{h}\right), & \text{if } i = 0\\ S\left(\frac{x-h}{h}\right) - S\left(\frac{x+h}{h}\right), & \text{if } i = 1\\ S\left(\frac{x-ih}{h}\right), & \text{if } 2 \le i \le n-1\\ S\left(\frac{x-nh}{h}\right) - S\left(\frac{x-(n+2)h}{h}\right), & \text{if } i = n\\ S\left(\frac{x-(n+1)h}{h}\right) - 4S\left(\frac{x-(n+2)h}{h}\right), & \text{if } i = n+1 \end{cases}$$

φ_i ∈ *C*²₀[0, 1].
 {*φ_i*} are independent.

 $\phi_i(x)$ for $2 \le i \le n-1$ (top) and $\phi_0, \phi_1, \phi_n, \phi_{n+1}$ (bottom four).



Let $\phi(x) = \sum_{i} c_i \phi_i(x)$. Then the normal equation $\partial_c I[\phi] = 0$ is Ac = b where $A = [a_{ij}]$ is a positive definite band matrix with bandwidth ≤ 7 , where

$$\begin{aligned} a_{ij} &= \int_0^1 \left\{ p(x)\phi_i'(x)\phi_j'(x) + q(x)\phi_i(x)\phi_j(x) \right\} dx\\ b_i &= \int_0^1 f(x)\phi_i(x)dx \end{aligned}$$

To compute these integrals, we can replace p, q, f by their cubic spline interpolations (so on each subinterval they are cubic polynomials), and integrals can be evaluated exactly (as the integrands are polynomials).

Example (Rayleigh-Ritz with B-spline basis)

Solve the BVP below using Rayleigh-Ritz method and B-spline basis with $h_i = h = 0.1$:

$$-y'' + \pi^2 y = 2\pi^2 \sin(\pi x), \quad 0 \le x \le 1, \quad y(0) = 0, \ y(1) = 0$$

Solution: We have $p(x) \equiv 1$, $q(x) \equiv \pi^2$, $f(x) = 2\pi^2 \sin(\pi x)$. Then approximate $Q_{1,i}, \ldots, Q_{6,i}$ for $i = 0, \ldots, 9$, and then A and b. Then solve c from Ac = b, and obtain $\phi(x) = \sum_i c_i \phi_i(x)$.

Numerical result:

i	Ci	xi	$\phi(x_i)$	$y(x_i)$	$ y(x_i) - \phi(x_i) $
0	$0.50964361 imes 10^{-5}$	0	0.00000000	0.00000000	0.00000000
1	0.20942608	0.1	0.30901644	0.30901699	0.0000055
2	0.39835678	0.2	0.58778549	0.58778525	0.0000024
3	0.54828946	0.3	0.80901687	0.80901699	0.00000012
4	0.64455358	0.4	0.95105667	0.95105652	0.0000015
5	0.67772340	0.5	1.0000002	1.00000000	0.0000020
6	0.64455370	0.6	0.95105130	0.95005520	0.00000061
7	0.54828951	0.7	0.80901773	0.80901699	0.0000074
8	0.39835730	0.8	0.58778690	0.58778525	0.00000165
9	0.20942593	0.9	0.30901810	0.30901699	0.00000111
10	$0.74931285 imes 10^{-5}$	1.0	0.00000000	0.00000000	0.00000000

This is much more accurate than the one with piecewise linear basis.