

**Analysis and development of finite
volume methods for the new generation of
cubed sphere dynamical cores for the
atmosphere**

Luan da Fonseca Santos

REPORT PRESENTED TO THE
INSTITUTE OF MATHEMATICS AND STATISTICS
OF THE UNIVERSITY OF SÃO PAULO
FOR THE DOCTOR OF SCIENCE
QUALIFYING EXAMINATION

Program: Applied Mathematics

Advisor: Prof. Pedro da Silva Peixoto

During the development of this work the author was supported by CAPES and FAPESP (grant number 20/10280-4)

São Paulo
November, 2022

**Analysis and development of finite
volume methods for the new generation of
cubed sphere dynamical cores for the
atmosphere**

Luan da Fonseca Santos

This is the original version of the
qualifying text prepared by candidate
Luan da Fonseca Santos, as submitted
to the Examining Committee.

Resumo

Luan da Fonseca Santos. **Análise e desenvolvimento de métodos de volumes finitos para modelos da nova geração da dinâmica atmosférica baseados na esfera cubada.** Exame de Qualificação (Doutorado). Instituto de Matemática e Estatística, Universidade de São Paulo, São Paulo, 2022.

O modelo atmosférico global FV3 do GFDL-NOAA-USA, inicialmente desenvolvido para malhas do tipo latitude-longitude, foi adaptado para a esfera cubada visando atingir melhor escalabilidade em super-computadores massivamente paralelos. Entretanto, neste tipo de malhas estamos mais sujeitos à problemas como o grid imprinting. Além disso, o modelo carece de algumas propriedades miméticas, que são altamente desejáveis. Este projeto de doutorado propõe-se a analisar as propriedades das discretizações de volumes finitos utilizadas no modelo FV3 na esfera cubada. Iremos investigar como propriedades das células da esfera cubada interferem na precisão dos esquemas numéricos. O estudo irá começar com a implementação de um código para gerar a esfera cubada e calcular os operados discretos do FV3. Então, iremos analisar como a malha interfere nos modelos de advecção e de águas rasas na esfera.

Palavras-chave: Núcleo dinâmico da atmosfera, esfera cubada, volumes finitos.

Abstract

Luan da Fonseca Santos. **Analysis and development of finite volume methods for the new generation of cubed sphere dynamical cores for the atmosphere.**

Qualifying Exam (Doctorate). Institute of Mathematics and Statistics, University of São Paulo, São Paulo, 2022.

The global atmospheric model FV3 from GFDL-NOAA-USA, which was originally designed for latitude-longitude grids, was adapted to the cubed sphere aiming to improve its scalability in massively parallel supercomputers. However, in this kind of grid, we are more likely to have grid imprinting problems. Besides that, the FV3 model lacks some highly desirable mimetic properties. This work aims to analyze the properties of the finite volume discretizations employed in the global atmospheric model FV3 on the cubed-sphere. We will investigate how the properties of the cells may impact on the accuracy of the numerical schemes. This study will firstly implement a cubed-sphere grid generator and the FV3 discrete operators on this grid. Then, we will analyze how the cubed-sphere grid properties influence in the numerical schemes by assessing it using the advection and shallow-water equations on the sphere. We will study the numerical dispersion and conservations properties of the scheme aiming to propose modifications in the numerical schemes to develop a mimetic finite volume version of the model.

Keywords: Dynamical core, cubed-sphere, finite-volume.

Contents

1	Introduction	1
1.1	Background	1
1.2	Motivations	4
1.3	Goals	5
1.4	Outline	6
2	One-dimensional finite-volume methods	7
2.1	One-dimensional advection equation in integral form	8
2.1.1	Notation	8
2.1.2	The 1D advection equation	10
2.2	The finite-volume Semi-Lagrangian approach	15
2.3	Departure point computation	16
2.3.1	RK1 scheme	16
2.3.2	RK2 scheme	17
2.4	Reconstruction	18
2.4.1	The Piecewise-Parabolic Method	20
2.4.2	Monotonization	21
2.5	Flux	23
2.6	Numerical experiments	24
2.7	Concluding remarks	27
3	Two-dimensional finite-volume methods	29
3.1	Two-dimensional advection equation in integral form	30
3.2	The finite-volume approach	31
3.2.1	Discretization of the problem	32
3.2.2	Convergence, consistency and stability	35
3.3	Dimension splitting	37
3.4	Numerical experiments	42

3.4.1	Linear advection equation with constant velocity simulations	42
3.4.2	Linear advection equation with variable velocity simulations	43
3.5	Concluding remarks	46
4	Cubed-sphere grids	49
4.1	Cubed-sphere mappings	50
4.1.1	Equidistant cubed-sphere	50
4.1.2	Equiangular cubed-sphere	52
4.1.3	Examples	53
4.2	Edges treatment	53
4.2.1	Ghost cells interpolation	53
4.2.2	Edges reconstruction	55
5	Cubed-sphere finite-volume methods	61
5.1	Advection finite-volume scheme	61
Appendixes		
A	Numerical Analysis	63
A.1	Finite-difference estimates	63
A.2	Lagrange interpolation	67
A.3	Numerical integration	67
A.3.1	Midpoint rule	68
A.3.2	Multi-step schemes	69
A.4	PPM reconstruction accuracy analysis	71
A.5	Convergence of 1D FV-SL schemes	75
A.5.1	Consistency and convergence	75
A.5.2	Stability	77
A.5.3	Flux accuracy analysis	80
B	Spherical coordinates and geometry	81
B.1	Conversions between latitude-longitude and contravariant coordinates	81
B.2	Covariant/contravariant conversion	83
C	Code availability	85
References		87

Chapter 1

Introduction

1.1 Background

Weather and climate predictions are recognized as a good for mankind, due to the information they yield for diverse activities. For instance, short-range forecasts are useful for public use, while medium-range forecasts are helpful for industrial activities and agriculture. Seasonal forecasts (one up to three months) are important to energy planning and agriculture. At last, longer-range forecasts (one century, for instance) are useful for climate change projections that are important for government planning.

The first global Numerical Weather Prediction models emerged in the 1960s with applications to weather, seasonal and climate forecasts. All these applications are essentially based on the same set of Partial Differential Equations (PDEs) but with distinct time scales (D. L. Williamson, 2007). These PDEs are defined on the sphere and model the evolution of the atmospheric fluid given the initial conditions. One important component of global models is the dynamical core, which is responsible for solving the PDEs that governs the atmosphere dynamics on grid-scale. The development of numerical methods for dynamical cores has been an active research area since the 1960s.

Global models use the sphere as the computational domain and therefore they require a discretization of the sphere. The first global models used the latitude-longitude grid (Figure 1.1a), which is very suitable for finite-differences schemes due to its orthogonality. The major drawback of the latitude-longitude grid is the clustering of points at the poles, known as the “pole problem”, which leads to extremely small time steps for explicit-in-time schemes due to the Courant-Friedrichs-Lowy (CFL) condition, making these schemes computationally very expensive.

The most successful method adopted in global atmospheric dynamical cores that overcomes the CFL restriction is the Semi-Implicit Semi-Lagrangian (SI-SL) scheme (Randall et al., 2018), which emerged in the 1980s and consists of the Lagrangian advection scheme applied at each time-step and the solution of fast gravity waves implicitly, allowing very large time steps despite the pole problem. The SI-SL approach combined with finite differences is still used nowadays, for instance in the UK Met Office global model ENDGame (Benacchio & Wood, 2016; Wood et al., 2014). The expensive part of the SI-SL approach is to

solve an elliptic equation at each time step, that comes from the semi-implicit discretization, which requires global data communication, being inefficient to run in massive parallel supercomputers. Besides that, Semi-Lagrangian schemes are inherently non-conservatives for mass, which is critical for climate forecasts (D. L. Williamson, 2007).



Figure 1.1: Examples of spherical grids: latitude-longitude grid (a) and grids based on Platonic solids (b)-(d).

The emergence of the Fast Fourier Transform (FFT) in the 1960s with the work from Cooley and Tukey (1965) allowed the computation of discrete Fourier transforms with $N \log(N)$ complexity. The viability of the usage of FFTs for solving atmospheric flows was shown by Orszag (1970), using the barotropic vorticity equation on the sphere, and by Eliasen et al. (1970), using the primitive equations. The spectral transform method expresses latitude-longitude grid values, that represent some scalar field, using truncated spherical harmonics expansions, which consists of Fourier expansions in latitude circles and Legendre functions expansions in longitude circles. The coefficients in the spectral expansions are known as spectral coefficients and are usually thought to live in the so-called spectral space. Given the grid values, the spectral coefficients are obtained by performing a FFT followed by a Legendre Transform (LT). Conversely, given the spectral coefficients, the grid values are obtained by performing an inverse LT followed by an inverse FFT. The main idea of the spectral method is to apply the spectral transform, in order to go the spectral space, and evaluate spatial derivatives in the spectral space, which consists of multiplying the spectral coefficients by constants. Then, the method performs the inverse spectral transform in order to get back to grid space, and the nonlinear terms are treated on the grid space (Krishnamurti et al., 2006).

The spectral transform makes the use of SI-SL methods computationally cheap, since the solution to elliptic problems becomes easy, once the spherical harmonics are eigenfunctions of the Laplacian operator on the sphere. Therefore, the spectral transform method gets faster when combined with the SI-SL approach due to the larger times-steps allowed in this case. Due to these enhancements, the spectral transform dominated global atmospheric modeling (Randall et al., 2018) since the 1980s. Indeed, the spectral method is still used in many current operational Weather Forecasting models such as the Integrated Forecast System (IFS) from European Centre for Medium-Range Weather Forecasts (ECMWF), Global Forecast System (GFS) from National Centers for Environmental Prediction (NCEP) and the Brazilian Global Atmospheric Model (BAM) (Figueroa et al., 2016) from Center for Weather Forecasting and Climate Research [Centro de Previsão de Tempo e Estudos Climáticos (CPTEC)].

With the beginning of the multicore era in the 1990s, the global atmospheric models started to move towards parallel efficiency aiming to run at very high resolutions. Even though the spectral transform expansions have a global data dependency, some parallelization is feasible among all the computations of FFTs, LTs and their inverses (Barros et al., 1995). However, the parallelization of the spectral method requires data transpositions in order to compute FFTs and LTs in parallel. These transpositions demand a lot of global communication using, for instance, the Message Passing Interface (MPI) (Zheng & Marguinaud, 2018). Indeed, the spectral transform becomes the most expensive component of global spectral models when the resolution is increased due to the amount of MPI communications (Müller et al., 2019).

The adiabatic and frictionless continuous equations that govern the atmospheric flow have conserved quantities. Among them, some of the most important are mass, total energy, angular momentum and potential vorticity (Thuburn, 2011). Numerical schemes that are known for having discrete analogous of these conservative properties are known as mimetic schemes. As we pointed out, Semi-Lagrangian schemes lack mass conservation. Nevertheless, these schemes have been employed in dynamical cores for better computational performance. However, dynamical cores should have discrete analogous of the continuous conserved quantities, especially concerning for longer simulation runs.

Aiming for better performance in massively parallel computers and conservation properties, new dynamical cores have been developed since the beginning of the 2000s. Novel spherical grids have been proposed, in order to avoid the pole problem. A popular choice are grids based on Platonic solids (Staniforth & Thuburn, 2012). The construction of these grids relies on a Platonic circumscribed on the sphere and the projection of its faces onto the sphere, which leads to quasi-uniform and more isotropic spherical grids. Some examples of spherical grids based on Platonic solids employed in the new generation of dynamical cores are the cubed-sphere (Figure 1.1b), icosahedral grid (Figure 1.1c), the pentagonal/hexagonal or Voronoi grid (Figure 1.1d) and octahedral grid (Figure 1.1e), which are based on the cube, icosahedron, dodecahedron and octahedron, respectively (Ullrich et al., 2017).

1.2 Motivations

The cubed-sphere became a popular quasi-uniform grid for the new generation of dynamical cores. It was originally proposed by Sadourny (1972) and it was revisited by Ronchi et al. (1996). Some of the cubed-sphere advantages are: uniformity; quadrilateral structure, making the grid indexing trivial; no overlappings; it is cheap to generate. However, the major drawbacks of the cubed-sphere are: non-orthogonal coordinate system, which leads to metric terms on the differential operator; discontinuity of the coordinate system at the cube edges, which may generate numerical noise and demands special treatment of discrete operators at the cube edges.

Despite of its drawbacks, the cubed-sphere has been adopted in some of the new generation dynamical cores. For instance, the cubed-sphere is used in the Community Atmosphere Model (CAM-SE) from the NCAR using spectral elements (Dennis et al., 2012) and in the Nonhydrostatic Unified Model of the Atmosphere (NUMA) from the US Navy using Discontinuous Galerkin methods (Giraldo et al., 2013). The cubed-sphere was also chosen to be used in the next UK Met Office global model using mixed finite elements (Kent et al., 2022). At last, the Finite Volume Cubed-Sphere dynamical core (FV3) from the Geophysical Fluid Dynamics Laboratory (GFDL) and the National Oceanic and Atmospheric Administration (NOAA) (L. M. Harris & Lin, 2013; Putman & Lin, 2007) is another example of new generation dynamical core based on the cubed-sphere.

The FV3 model is an extension of the Finite-Volume dynamical core (FVcore) from latitude-longitude grids to the cubed-sphere. The numerical methods from FVcore started to be developed with the transport scheme from the work Lin et al. (1994), which is based on the piecewise linear scheme from Van Leer (1977). This scheme was later improved, using the Piecewise Parabolic Method (PPM) (Carpenter et al., 1990; Colella & Woodward, 1984) using dimension splitting techniques that guarantee monotonicity and mass conservation, for the transport equation (Lin & Rood, 1996) and the shallow-water equations (Lin & Rood, 1997). An important feature is that the FVcore combines the Arakawa C- and D-grids (Arakawa & Lamb, 1977), where the C-grid values are computed in an intermediate time step. The full global model was then presented by Lin (2004).

The FVcore was adapted to the cubed-sphere grid (Putman, 2007; Putman & Lin, 2007), to reach better performance in parallel computers, leading to the FV3 model. Later, the FV3 also was improved to allow locally refinement grids through grid-nesting or grid-stretching (L. M. Harris & Lin, 2013). Currently, the FV3 model is capable of performing hydrostatic and non-hydrostatic atmospheric simulations and it was chosen as the new US global weather prediction model, indeed, it replaced the spectral transform Global Forecast System (GFS) in June, 2019 (Samenow, 2019).

However, a well-known problem that occurs on cubed-sphere models that use low-order numerical methods is the grid imprinting visible due to the coordinate system discontinuity, especially at larger scales, leading to the emergence of a wavenumber 4 pattern. This was reported in the paper of Rančić et al. (2017), where the authors employ a finite-difference numerical scheme on the Uniform Jacobian cubed-sphere using a Arakawa B-grid. The unpublished report from Whitaker (2015) shows grid imprinting in other models, including the FV3. Generally speaking, grid imprinting is the presence of artificial behaviors on

the numerical solution that is associated with the grid employed. It is important to stress out that other quasi-uniform grids may also suffer from grid imprinting. For instance, a popular mimetic method, known as TRiSK, was proposed in the literature by Thuburn et al. (2009) and Ringler et al. (2010) using finite difference and finite volume schemes. This scheme is designed for general orthogonal grids, such as the Voronoi and icosahedral grids, and ensures mass and total energy conservation. This method has been employed in the dynamical core of the Model for Prediction Across Scales (MPAS) from National Center for Atmospheric Research (NCAR) (Skamarock et al., 2012), which intended to work on general Voronoi grids, including locally refined Voronoi grids. However, the TRiSK scheme is a low-order scheme and also suffers from grid imprinting, *i.e.*, geometric properties of the grid, such as cell alignment, interfere with the method accuracy (Peixoto, 2016; Peixoto & Barros, 2013; Weller, 2012). Furthermore, in locally refined Voronoi grids, the scheme may become unstable due to ill-aligned cells and numerical dissipation is needed (Santos & Peixoto, 2021), breaking the total energy conservation of the method.

Despite being chosen as the new US global weather prediction model, there is a lack of numerical studies of the FV3 discretizations in the literature, especially regarding the grid imprinting problem and its mimetic properties. Numerical results for the advection equation on the cubed-sphere using the FV3 dynamical core was presented in Putman and Lin (2007) and some shallow-water simulations were presented in L. M. Harris and Lin (2013), considering cubed-spheres with local refinement through grid nesting. From the work L. M. Harris and Lin (2013) we can notice that the FV3 dynamical lack convergence on the maximum norm for the shallow-water model considering the classical balanced geostrophic flow test case from D. Williamson et al. (1992). The authors attribute these errors to the abrupt change in the grid resolution near the nested grid, but no quantitative results are shown considering the quasi-uniform grid. Many other papers available in the literature use the complete FV3 model in three-dimensional frameworks which make it harder to perform a numerical analysis study due not only to its computational cost but also due to the complexity of three-dimensional atmospheric models. There are no detailed works published in intermediate two-dimensional frameworks, using, for instance, the shallow-water equations on the sphere. Even though the advection equation on the sphere plays a key role in the dynamical core development, since it models the transport of scalar fields on the sphere, important features captured by the shallow-water equations on the sphere, such as the Coriolis effect, inertia-gravity waves, geostrophic adjustment, Rossby waves, among others, are not captured by a simple advection model. Hence, shallow-water equations provide an excellent benchmark to assess dynamical cores in general, since it is only two-dimensional but is a complex enough geophysical model for atmosphere dynamics.

1.3 Goals

The aim of this work is to fill the gap in the literature regarding numerical studies of the FV3 discrete operators that we pointed out before. More explicitly, the goals of this work are:

- Investigate the occurrence of grid imprinting on the cubed-sphere using the advection equations and the shallow-water equations on the sphere;

- Propose improvements on the FV3 discrete operators and modifications on the cubed-sphere that alleviate grid imprinting;
- Investigate how we can add more mimetic properties to the FV3 discretizations.

1.4 Outline

This report is outlined as follows. Chapter 2 is dedicated to review the Piecewise Parabolic Method (PPM) for the one-dimensional advection equation. Chapter 3 reviews the dimension splitting method, which allow us to use one-dimensional methods, such as the PPM, to solve the two-dimensional advection equation. Chapter 4 introduces the cubed-sphere grid and shows some of its geometric properties. Chapter 5 extends the ideas of Chapter 3 to the cubed-sphere grid. The dimension-splitting method on each cubed-sphere panel works as in the plane, with the addition of metric terms, due to non-orthogonality of the grid, and interpolation between panels to obtain ghost cells values needed for stencil computations.

Chapter 2

One-dimensional finite-volume methods

The aim of this chapter is to provide a detailed description of one-dimensional (1D) finite-volume (FV) schemes within a Semi-Lagrangian (SL) framework, specifically applied to the 1D advection equation with periodic boundary conditions. These schemes are also known as flux-form Semi-Lagrangian schemes, and they allow for time steps beyond the Courant-Friedrichs-Lowy (CFL) condition while preserving the total mass. FV-SL schemes have been explored in the literature since the work of LeVeque (1985), which extended the finite-volume schemes from Godunov (1959) to accommodate larger time steps. This approach has been further investigated in the literature (c.f, e.g. . Leonard et al. (1996) and Lin and Rood (1996)).

To introduce the FV-SL schemes, we begin by discretizing the spatial and temporal domains into uniform grids. Subsequently, the FV-SL schemes involve three steps. The first step involves computing the departure points of the spatial grid edges. The second step, known as reconstruction, utilizes the grid cell average values to determine a piecewise function within each cell. This piecewise function approximates the values of the advected quantity and ensures the preservation of its local mass within each grid cell. The third step entails updating the fluxes at the grid edges by integrating the reconstruction function over a domain that extends from the departure point of the grid edge to the grid edge itself.

The first step of FV-SL schemes can be accomplished by integrating an ordinary differential equation backward in time. The second step is performed using the Piecewise-Parabolic Method (PPM) proposed by Colella and Woodward (1984). As the name suggests, PPM employs piecewise-parabolic functions. The third and final step is computed easily, as the reconstruction functions consist of parabolas that preserve the local mass.

It is worth noting that the reconstruction function can be constructed using functions other than parabolas. In fact, the Piecewise-Parabolic Method (PPM) can be seen as an extension of the Piecewise-Linear method proposed by Van Leer (1977), which, in turn, was inspired by the Piecewise-Constant method introduced by Godunov (1959). Additionally, other schemes inspired by PPM have been proposed in the literature utilizing higher-order

polynomials, such as quartic polynomials (White & Adcroft, 2008). For a comprehensive review of general piecewise-polynomial reconstruction, we recommend referring to the technical report by Engwirda and Kelley (2016), Lauritzen et al. (2011), and the references therein.

The PPM approach has become popular in the literature for gas dynamics simulations, astrophysical phenomena modeling (Woodward, 1986), and later on atmospheric simulations (Carpenter et al., 1990). Indeed, PPM has been implemented in the FV3 dynamical core on its latitude-longitude grid (Lin, 2004) and cubed-sphere (Putman & Lin, 2007) versions. Although many other shapes for the basis functions and higher-order schemes are available in the literature, L. Harris et al. (2021) points out that the PPM scheme suits the needs of FV3 well. It is a flexible method that can be modified to ensure low diffusivity or shape preservation, for example. Additionally, a finite-volume numerical method usually requires monotonicity constraints, which, according to Godunov's theorem, limit the order of convergence to at most 1. Therefore, a higher-order scheme needs to strike a well-balanced trade-off between increasing computational cost and potential benefits.

This chapter begins with a basic review of one-dimensional advection equation in the integral form in Section 2.1. In Section 2.2, we establish the framework for general one-dimensional finite-volume Semi-Lagrangian schemes. Section 2.3 presents methods for computing the departure point. The PPM reconstruction is described in Section 2.4, while Subsection 2.4.2 introduces different approaches to ensure the monotonicity of parabolas. Subsection 2.5 focuses on the description and investigation of the PPM flux computation. Section 2.6 presents numerical results using the PPM scheme for the advection equation. Finally, Section 2.7 presents some concluding remarks. The application of PPM to solve two-dimensional problems will be addressed in Chapter 3.

2.1 One-dimensional advection equation in integral form

2.1.1 Notation

Before introducing the FV-SL schemes, let us establish some notation by introducing the concepts of a Δx -grid, a Δt -temporal grid, and the $(\Delta x, \Delta t, \lambda)$ -discretization, as well as the concept of grid function/winds.

Definition 2.1 (Δx -grid). *For a given interval $\Omega = [a, b]$ and a positive real number Δx such that $\Delta x = (b - a)/N$ for some positive integer N , we say that $\Omega_{\Delta x} = \{X_i\}_{i=1}^N$ is a Δx -grid for $[a, b]$ if*

$$X_i = [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}], \quad a = x_{\frac{1}{2}} < x_{\frac{3}{2}} < \dots < x_{N-\frac{1}{2}} < x_{N+\frac{1}{2}} = b,$$

and $\Delta x = x_{i+\frac{1}{2}} - x_{i-\frac{1}{2}}$. Each X_i is referred to as a control volume or cell, and $x_{i-\frac{1}{2}}$ and $x_{i+\frac{1}{2}}$ are the edges of the control volume X_i . The cell centroid is defined by

$$x_i = \frac{1}{2}(x_{i+\frac{1}{2}} + x_{i-\frac{1}{2}}), \quad \forall i = 1, \dots, N,$$

and Δx is the cell length.

Remark 2.1. We may extend the definition of cells X_i for values of i outside the range $1, \dots, N$ as $X_i = [a + (i - 1)\Delta x, a + i\Delta x]$. These cells are known as ghost cells (Figure 2.1.)

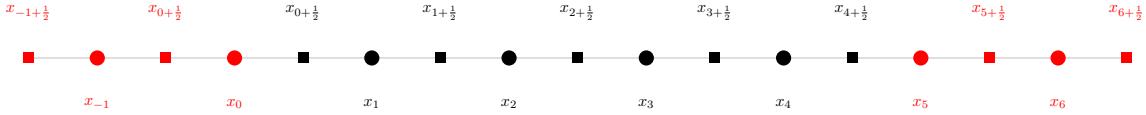


Figure 2.1: Illustration of a Δx -grid with $N = 4$ cells in its interior (in black) and 4 ghost cells (in red). The edges are denoted by squares and the cell centroids are denoted using circles.

Definition 2.2 (Δt -temporal grid). For a given interval $[0, T]$ and a positive real number Δt such that $\Delta t = T/N_T$ for some positive integer N_T , we say that $T_{\Delta t} = \{T_n\}_{n=0}^{N_T}$ a Δt -temporal grid for $[0, T]$ if

$$T_n = [t^n, t^{n+1}], \quad t^n = n\Delta t, \quad \Delta t = \frac{T}{N_T}, \quad \forall n = 0, \dots, N_T.$$

Definition 2.3 ($(\Delta x, \Delta t, \lambda)$ -discretization). Given $\Omega \times [0, T]$, $\Omega = [a, b]$ and positive real numbers Δx and Δt , we say that $(\Omega_{\Delta x}, T_{\Delta t})$ is a $(\Delta x, \Delta t, \lambda)$ -discretization of $\Omega \times [0, T]$ if $\Omega_{\Delta x}$ is a Δx -grid for Ω , $T_{\Delta t}$ is a Δt -temporal grid for $[0, T]$, and $\frac{\Delta t}{\Delta x} = \lambda$.

Remark 2.2. Whenever we refer to a Δx -grid, a Δt -temporal grid, or a $(\Delta x, \Delta t, \lambda)$ -discretization, X_i , N , t^n , and N_T are assumed to be implicitly defined.

Next, we introduce the definitions of grid functions at cell centroids and edges.

Definition 2.4 (Δx -grid function). For a Δx -grid, we say that Q is a Δx -grid function if $Q = (Q_1, \dots, Q_N) \in \mathbb{C}^N$.

Definition 2.5 (Δx -grid wind). For a Δx -grid, we say that u is a Δx -grid wind if $u = (u_{\frac{1}{2}}, \dots, u_{N+\frac{1}{2}}) \in \mathbb{C}^{N+1}$.

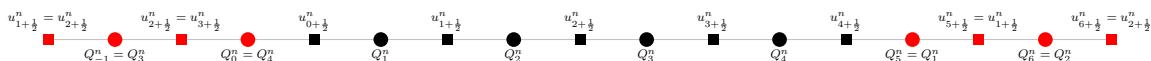


Figure 2.2: Illustration of Δx -grid function Q and u and its ghost cell values (in red) assuming periodicity.

The definition of a Δx -grid wind is based on the Arakawa grids (Arakawa & Lamb, 1977). Given $Q \in \mathbb{C}^N$, we define the p -norm as

$$\|Q\|_{p, \Delta x} = \begin{cases} \left(\sum_{i=1}^N |Q_i|^p \right)^{\frac{1}{p}} & \text{if } 1 \leq p < \infty, \\ \max_{i=1, \dots, N} |Q_i| & \text{otherwise.} \end{cases} \quad (2.1)$$

Using a similar notation as in Engwirda and Kelley (2016), we define the stencil and a grid function evaluated on a stencil as follows.

Definition 2.6 (stencil). For a Δx -grid, and each $i = 0, \dots, N$, we define a stencil as a set of the form $S_{i+\frac{1}{2}} = \{i - r + 1, \dots, i - 1, i, i + 1, \dots, i + s\}$.

Definition 2.7 (Grid function restricted to a stencil). *For a Δx -grid, a stencil $S_{i+\frac{1}{2}}$, and a Δx -grid function Q , we define $Q(S_{i+\frac{1}{2}}) = (Q_k)_{k \in S_{i+\frac{1}{2}}}$.*

These definitions provide the necessary notation for describing grid functions and their evaluations on stencils.

When computing $Q(S_{i+\frac{1}{2}})$, it may be necessary to access values of Q_i for i outside the range $1, \dots, N$. In such cases, we can consider Q_i as a ghost cell value. Since we assume periodic boundary conditions in this chapter, this issue is resolved by assuming periodicity in the grid function Q . For example, for a stencil that requires three ghost cell values at the left and at the right of the domain, we can assume $Q_0 = Q_N$, $Q_{-1} = Q_{N-1}$, $Q_{-2} = Q_{N-2}$, and $Q_{N+1} = Q_1$, $Q_{N+2} = Q_2$, $Q_{N+3} = Q_3$. The same applies to Δx -grid winds. Furthermore, considering functions $q, u : [a, b] \times [0, T] \rightarrow \mathbb{R}$ and a $(\Delta x, \Delta t, \lambda)$ -discretization of $[a, b] \times [0, T]$, we introduce the grid functions $q^n \in \mathbb{C}^N$ and $u^n \in \mathbb{C}^{N+1}$. Here, $q_i^n = q(x_i, t^n)$ and $u_{i+\frac{1}{2}}^n = u(x_{i+\frac{1}{2}}, t^n)$ for $n = 0, \dots, N_T$. These grid functions represent the discrete values of q and u at the cell centroids and edges, respectively, for each time level t^n (Figure 2.2).

Additionally, we introduce the average value of q in the i -th control volume at time t , denoted as $Q_i(t)$, defined by:

$$Q_i(t) = \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} q(x, t) dx. \quad (2.2)$$

Moreover, we define the Δx -grid function of average values as $Q(t) = (Q_1(t), Q_2(t), \dots, Q_N(t))$. Here, $Q_i(t)$ represents the average value of q in the i -th control volume at time t .

For the consideration of periodic boundary conditions, we can define spaces of periodic functions over the interval $\Omega = [a, b]$ as follows:

$$\mathcal{S}_P(\Omega) = \{q : \mathbb{R} \times [0, +\infty[\rightarrow \mathbb{R} : q(x + b - a, t) = q(x, t), \quad \forall x \in \mathbb{R}, \quad t > 0\}.$$

Similarly, the space of k -times periodically differentiable functions $C_P^k(\Omega)$ can be defined as:

$$C_P^k(\Omega) = \mathcal{S}_P(\Omega) \cup C^k(\mathbb{R} \times [0, \infty[),$$

where $C^k(\mathbb{R} \times [0, +\infty[)$ denotes the space of functions that are k times continuously differentiable in both the spatial and temporal variables. In summary, $\mathcal{S}_P(\Omega)$ represents the space of periodic functions, and $C_P^k(\Omega)$ represents the space of k -times periodically differentiable functions over the interval $\Omega = [a, b]$ subject to periodic boundary conditions.

2.1.2 The 1D advection equation

In this section, we will derive the integral form of the 1D advection equation with periodic boundary conditions over the interval $\Omega = [a, b]$. What is going to be presented here follows LeVeque (1990, 2002) closely. The advection equation with periodic boundary

conditions in its differential form is given by:

$$\begin{cases} \partial_t q(x, t) + \partial_x(uq)(x, t) = 0, & \forall (x, t) \in \mathbb{R} \times]0, +\infty[, \\ q(a, t) = q(b, t), & \forall t > 0, \\ q_0(x) = q(x, 0). \end{cases} \quad (2.3)$$

Here, $q \in C_p^1(\Omega)$ represents the advected quantity, and $u \in C_p^1(\Omega)$ represents the velocity. We will focus on Equation (2.3) over the domain $D = \Omega \times [0, T]$, where $T > 0$ is a finite time. A strong or classical solution to the advection equation is defined as a function q that is C^1 and satisfies Equation (2.3). In order to deduce the integral form of Equation (2.3), we consider $[x_1, x_2] \times [t_1, t_2] \subset D$. Integrating Equation (2.4) over $[x_1, x_2]$, we obtain:

$$\frac{d}{dt} \int_{x_1}^{x_2} q(x, t) dx = -((uq)(x_2, t) - (uq)(x_1, t)), \quad (2.4)$$

and integrating Equation (2.4) over $[t_1, t_2]$, we get

$$\int_{x_1}^{x_2} q(x, t_2) dx = \int_{x_1}^{x_2} q(x, t_1) dx - \left(\int_{t_1}^{t_2} (uq)(x_2, t) dt - \int_{t_1}^{t_2} (uq)(x_1, t) dt \right). \quad (2.5)$$

The presented problem, Problem 2.1, aims to find a solution, called weak solution, to the advection equation in its integral form, considering the given initial condition q_0 and velocity function u .

Problem 2.1. *Given an initial condition q_0 and a velocity function u we would like to find a weak solution q of the advection equation in the integral form:*

$$\int_{x_1}^{x_2} q(x, t_2) dx = \int_{x_1}^{x_2} q(x, t_1) dx + \int_{t_1}^{t_2} (uq)(x_1, t) dt - \int_{t_1}^{t_2} (uq)(x_2, t) dt,$$

$\forall [x_1, x_2] \times [t_1, t_2] \subset \Omega \times [0, T]$, and $q(x, 0) = q_0(x)$, $\forall x \in \Omega$, $q(a, t) = q(b, t)$, $\forall t \in [0, T]$.

We point out that, for Problem 2.1, the total mass in Ω at time t defined by:

$$M_{[a,b]}(t) = \int_a^b q(x, t) dx,$$

remains constant over time, i.e.,

$$M_{[a,b]}(t) = M_{[a,b]}(0), \quad \forall t \in [0, T].$$

This conservation of total mass property is highly desirable for numerical schemes aiming to approximate general conservation law solutions accurately.

Applying the steps from Equation (2.3) to Equation (2.5) in reverse order, one can verify that if q is a weak solution and q is C^1 , then it satisfies Equation (2.3). Therefore, Equation (2.3) and Problem (2.1) are equivalent when q is C^1 . However, Problem (2.1) can be formulated for functions that are not C^1 and have discontinuities. In fact, Problem (2.1) only requires that q and uq are locally integrable.

It is worth noting that Equation (2.5) holds for all x_1, x_2, t_1 , and t_2 such that $[x_1, x_2] \times [t_1, t_2] \subset D$. Therefore, let us consider a $(\Delta x, \Delta t, \lambda)$ -discretization of D and rewrite Equation (2.5) in terms of this discretization. By replacing t_1, t_2, x_1 , and x_2 with $t^n, t^{n+1}, x_{i-\frac{1}{2}}$, and $x_{i+\frac{1}{2}}$, respectively, in Equation (2.5), we obtain:

$$Q_i(t^{n+1}) = Q_i(t^n) - \frac{1}{\Delta x} \left(\int_{t^n}^{t^{n+1}} (uq)(x_{i+\frac{1}{2}}, t) dt - \int_{t^n}^{t^{n+1}} (uq)(x_{i-\frac{1}{2}}, t) dt \right), \quad \forall i = 1, \dots, N, \quad \forall n = 0, \dots, N_T - 1. \quad (2.6)$$

To achieve a more compact notation, we can use the centered difference notation:

$$\delta_x g(x_i, t) = g(x_{i+\frac{1}{2}}, t) - g(x_{i-\frac{1}{2}}, t), \quad (2.7)$$

for an arbitrary function g . Using this notation, Equation (2.6) can be rewritten as:

$$Q_i(t^{n+1}) = Q_i(t^n) - \frac{1}{\Delta x} \delta_x \left(\int_{t^n}^{t^{n+1}} (uq)(x_i, t) dt \right), \quad \forall i = 1, \dots, N, \quad \forall n = 0, \dots, N_T - 1. \quad (2.8)$$

Now we can define a discretized version of Problem 2.1 as Problem 2.2.

Problem 2.2. Let us consider the framework of Problem 2.1 and a $(\Delta x, \Delta t, \lambda)$ -discretization of $\Omega \times [0, T]$. Since we are operating within the framework of Problem 2.1, the following relationship holds:

$$Q_i(t^{n+1}) = Q_i(t^n) - \lambda \delta_x \left(\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} (uq)(x_i, t) dt \right), \quad \forall i = 1, \dots, N, \quad \forall n = 0, \dots, N_T - 1, \quad (2.9)$$

where $Q_i(t) = \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} q(x, t) dx$. Our objective now is to determine the values $Q_i(t^n)$, $\forall i = 1, \dots, N$, $\forall n = 0, \dots, N_T - 1$, given the initial values $Q_i(0)$, $\forall i = 1, \dots, N$. In other words, we aim to find the average values of q in each control volume X_i at the specified time instances.

It is important to note that no approximations have been made in problems (2.1) and (2.2). In Equation (2.9), we divided and multiplied by Δt to interpret $\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} (uq)(x_{i \pm \frac{1}{2}}, t) dt$ as a time-averaged flux. This interpretation is useful for deriving finite-volume schemes.

In Problem 2.2, we need to approximate the time-averaged flux at the cell edges $x_{i \pm \frac{1}{2}}$ to derive a finite-volume scheme. This flux, in principle, requires knowledge of q over the entire interval $[t^n, t^{n+1}]$. To overcome this, we can express the temporal integral as a spatial integral at time t^n . This approach avoids the need for information about q throughout the entire interval $[t^n, t^{n+1}]$. Furthermore, this spatial integral domain is closely related to the definition of the departure point.

To introduce the definition of departure point, for each $s \in [t^n, t^{n+1}]$, we consider the following Cauchy problem backward in time:

$$\begin{cases} \frac{\partial X}{\partial t}(t, s; x_{i+\frac{1}{2}}) = u(X(t, s; x_{i+\frac{1}{2}}), t), & t \in [t^n, s] \\ X(s, s; x_{i+\frac{1}{2}}) = x_{i+\frac{1}{2}}. \end{cases} \quad (2.10)$$

2.1 | ONE-DIMENSIONAL ADVECTION EQUATION IN INTEGRAL FORM

The point $X(t^n, s; x_{i+\frac{1}{2}})$ is called departure point at time t^n of the point $x_{i+\frac{1}{2}}$ at time s . In Figure 2.3 we illustrate the departure point idea.

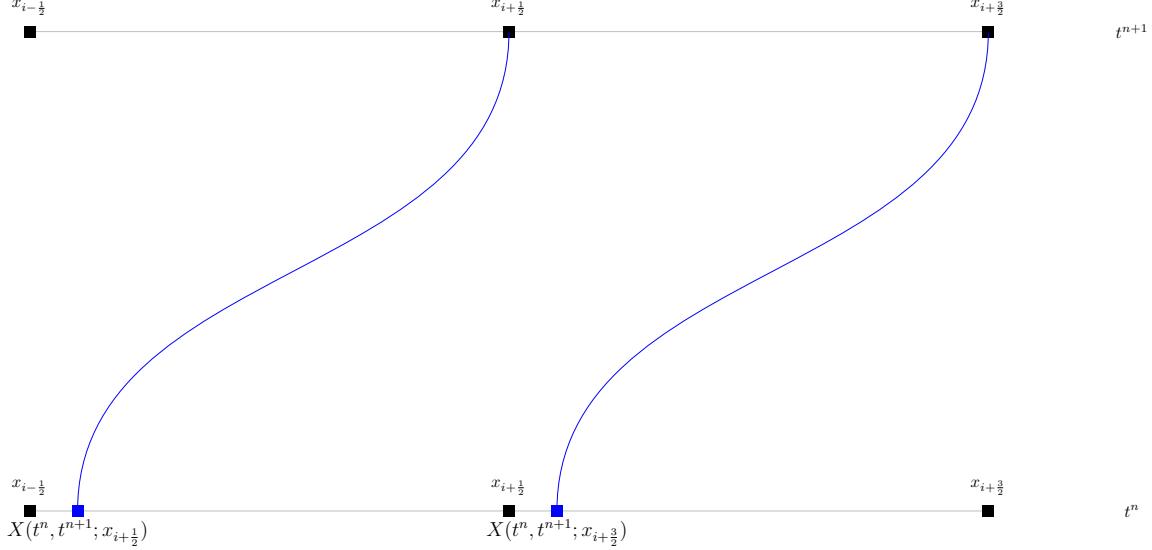


Figure 2.3: Illustration of the departure point of the cell edges from time t^{n+1} to t^n .

Integrating Equation (2.10) over the interval $[t, s]$, we get:

$$X(t, s; x_{i+\frac{1}{2}}) = x_{i+\frac{1}{2}} - \int_t^s u(X(\theta, s; x_{i+\frac{1}{2}}), \theta) d\theta. \quad (2.11)$$

In the following Proposition, we show how the time-averaged flux is related to a spatial integral over a interval depending on departure points.

Proposition 2.1. *Assume the framework of Problem 2.2. If q and u are C^1 functions, then:*

$$\int_{t^n}^{t^{n+1}} (uq)(x_{i+\frac{1}{2}}, s) ds = \int_{X(t^n, t^{n+1}; x_{i+\frac{1}{2}})}^{x_{i+\frac{1}{2}}} q(x, t^n) dx \quad (2.12)$$

Proof. Using the Leibniz rule for integration in Equation (2.11), it follows that:

$$\begin{aligned} \frac{\partial X}{\partial s}(t, s; x_{i+\frac{1}{2}}) &= - \left(u(x_{i+\frac{1}{2}}, s) + \int_t^s \frac{du}{ds}(X(\theta, s; x_{i+\frac{1}{2}}), \theta) d\theta \right) \\ &= -u(x_{i+\frac{1}{2}}, s) - \int_t^s \frac{\partial u}{\partial x}(X(\theta, s; x_{i+\frac{1}{2}}), \theta) \frac{\partial X}{\partial s}(\theta, s; x_{i+\frac{1}{2}}) d\theta. \end{aligned} \quad (2.13)$$

Taking the derivative with respect to t of Equation (2.13), we have:

$$\frac{\partial}{\partial t} \left(\frac{\partial X}{\partial s} \right) (t, s; x_{i+\frac{1}{2}}) = \frac{\partial u}{\partial x}(X(t, s; x_{i+\frac{1}{2}}), t) \frac{\partial X}{\partial s}(t, s; x_{i+\frac{1}{2}}). \quad (2.14)$$

Using standard ordinary differential equations techniques (ODE), we get that X that solves

Equations (2.13) and (2.14) is given by:

$$\frac{\partial X}{\partial s}(t, s; x_{i+\frac{1}{2}}) = -\exp \left(\int_t^s \frac{\partial u}{\partial x}(X(\theta, s; x_{i+\frac{1}{2}}), \theta) d\theta \right) u(x_{i+\frac{1}{2}}, s). \quad (2.15)$$

Computing q on the trajectory give by $X(t, s; x_{i+\frac{1}{2}})$ and taking its time derivative, we obtain:

$$\begin{aligned} \frac{dq}{dt}(X(t, s; x_{i+\frac{1}{2}}), t) &= \frac{\partial q}{\partial t}(X(t, s; x_{i+\frac{1}{2}}), t) + u(X(t, s; x_{i+\frac{1}{2}}), t) \frac{\partial q}{\partial x}(X(t, s; x_{i+\frac{1}{2}}), t) \\ &= -\frac{\partial u}{\partial x}(X(t, s; x_{i+\frac{1}{2}}), t) q(X(t, s; x_{i+\frac{1}{2}}), t), \end{aligned} \quad (2.16)$$

where we used that q satisfies the linear advection equation on its differential form and that $X(t, s; x_{i+\frac{1}{2}})$ solves Equation (2.10). Using again standard ODE techniques, we get that q that solves Equation (2.16) is given by:

$$q(X(t, s; x_{i+\frac{1}{2}}), t) = \exp \left(- \int_t^s \frac{\partial u}{\partial x}(X(\theta, s; x_{i+\frac{1}{2}}), \theta) d\theta \right) q(x_{i+\frac{1}{2}}, s). \quad (2.17)$$

Notice that if u does not depend on x , then q is constant along the trajectory $X(t, s; x_{i+\frac{1}{2}})$.

Let us consider the mapping $s \in [t^n, t^{n+1}] \rightarrow X(t^n, s, x_{i+\frac{1}{2}})$. Integrating q over all departure points at time t^n from $x_{i+\frac{1}{2}}$ at time s , we have

$$\int_{X(t^n, t^n; x_{i+\frac{1}{2}}) = x_{i+\frac{1}{2}}}^{X(t^n, t^{n+1}; x_{i+\frac{1}{2}})} q(x, t^n) dx = \int_{t^n}^{t^{n+1}} q(X(t^n, s; x_{i+\frac{1}{2}}), t^n) \frac{\partial X}{\partial s}(t^n, s; x_{i+\frac{1}{2}}) ds, \quad (2.18)$$

where we are just using the variable change integration formula. Then, it follows from Equations (2.15) and (2.17) with $t = t^n$ that:

$$\int_{x_{i+\frac{1}{2}}}^{X(t^n, t^{n+1}; x_{i+\frac{1}{2}})} q(x, t^n) dx = - \int_{t^n}^{t^{n+1}} (uq)(x_{i+\frac{1}{2}}, s) ds,$$

which is the desired formula. \square

With the aid of Proposition 2.1, we can rewrite Problem 2.2 in terms of the departure point, avoiding the need for knowledge about q over the entire interval $[t^n, t^{n+1}]$. This is described in Problem 2.3:

Problem 2.3. Assume the framework of Problem 2.1 and a $(\Delta x, \Delta t, \lambda)$ -discretization of $\Omega \times [0, T]$. Since we are in the framework of Problem 2.1, it follows that:

$$\begin{aligned} Q_i(t^{n+1}) &= Q_i(t^n) - \lambda \left(\frac{1}{\Delta t} \int_{X(t^n, t^{n+1}; x_{i+\frac{1}{2}})}^{x_{i+\frac{1}{2}}} q(x, t^n) dx - \frac{1}{\Delta t} \int_{X(t^n, t^{n+1}; x_{i-\frac{1}{2}})}^{x_{i-\frac{1}{2}}} q(x, t^n) dx \right), \\ &\forall i = 1, \dots, N, \quad \forall n = 0, \dots, N_T - 1, \end{aligned} \quad (2.19)$$

where $Q_i(t) = \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} q(x, t) dx$. Our problem now consists of finding the values $Q_i(t^n)$, $\forall i = 1, \dots, N$, $\forall n = 0, \dots, N_T - 1$, given the initial values $Q_i(0)$, $\forall i = 1, \dots, N$. In other words,

we would like to find the average values of q in each control volume X_i at the considered time instants.

At each time step t^n , we compute the values of $Q_i(t^{n+1})$ based on $Q_i(t^n)$ and the integrals of $q(x, t^n)$ over specific intervals. These intervals are defined by the departure points $X(t^n, t^{n+1}; x_{i+\frac{1}{2}})$ and $X(t^n, t^{n+1}; x_{i-\frac{1}{2}})$. To perform the computations, we need to determine the departure points from the edges of all control volumes and calculate the required integrals. This idea serves as the motivation for defining finite-volume Semi-Lagrangian schemes. These schemes involve estimating the departure points and reconstructing the function q at time t^n using its average values $Q_i(t^n)$, which enables us to compute the necessary integrals.

2.2 The finite-volume Semi-Lagrangian approach

Finally, we define the 1D FV-SL scheme problem as follows in Problem 2.3.

Problem 2.4 (1D FV-SL scheme). *Assume the framework defined in Problem 2.3. The finite-volume Semi-Lagrangian approach of Problem 2.3 consists of finding a scheme of the form:*

$$Q_i^{n+1} = Q_i^n - \lambda(F_{i+\frac{1}{2}}^n - F_{i-\frac{1}{2}}^n), \quad \forall i = 1, \dots, N, \quad \forall n = 0, \dots, N_T - 1, \quad (2.20)$$

where $Q^n \in \mathbb{R}^N$ is intended to be an approximation of $Q(t^n) \in \mathbb{R}^N$ in some sense. We define $Q_i^0 = Q_i(0)$ or $Q_i^0 = q_i^0$. The terms $F_{i\pm\frac{1}{2}}^n$ are known as numerical flux and are given by

$$F_{i\pm\frac{1}{2}}^n = \frac{1}{\Delta t} \int_{\tilde{x}_{i\pm\frac{1}{2}}^n}^{x_{i\pm\frac{1}{2}}} \tilde{q}(x; Q^n) dx, \quad (2.21)$$

where $\tilde{x}_{i\pm\frac{1}{2}}^n$ is an estimate of the departure point $X(t^n, t^{n+1}; x_{i\pm\frac{1}{2}})$, and \tilde{q} is a reconstruction function for q built with the values Q^n . Thus, $F_{i\pm\frac{1}{2}}^n$ approximates $\frac{1}{\Delta t} \int_{X(t^n, t^{n+1}; x_{i\pm\frac{1}{2}})}^{x_{i\pm\frac{1}{2}}} q(x, t^n) dx$.

For a 1D FV-SL the discrete total mass at the time-step n is given by

$$M^n = \Delta x \sum_{i=1}^N Q_i^n. \quad (2.22)$$

Therefore, the discrete total mass is constant for a 1D-FV scheme, which follows from a straightforward computation:

$$M^{n+1} = \Delta x \sum_{i=1}^N Q_i^{n+1} = M^n - \Delta t \sum_{i=1}^N (F_{i+\frac{1}{2}}^n - F_{i-\frac{1}{2}}^n) = M^n - \Delta t (F_{N+\frac{1}{2}}^n - F_{\frac{1}{2}}^n) = M^n,$$

where we are using that $F_{N+\frac{1}{2}}^n = F_{\frac{1}{2}}^n$, since we are assuming periodic boundary conditions.

We would like to highlight an important relationship between the average values of q and its values at the cell centroids. In Problem 2.4, we mentioned that the initial condition can be represented as q_i^0 instead of $Q_i(0)$. Moreover, when analyzing the convergence

of a FV-SL scheme, it is useful to compare Q_i^n with q_i^n since computing $Q_i(t^n)$ requires evaluating an analytical integral, which can be challenging in certain cases. In Proposition 2.2, we provide a simple proof that q_i^n approximates $Q_i(t^n)$ with second-order error when q is twice continuously differentiable.

Proposition 2.2. *If $q \in C_p^2(\Omega)$, then $Q_i(t^n) - q_i^n = C_1 \Delta x^2$, where $C_1 = \frac{1}{24} \frac{\partial^2 q}{\partial x^2}(\eta, t^n)$, $\eta \in X_i$.*

Proof. Just apply Theorem A.3 for the function $q(x, t^n)$. \square

The Problem of the convergence of 1D FV-SL schemes is addressed in Section A.5. Now we are going to address the problem of the departure point estimation and the reconstruction problem.

2.3 Departure point computation

2.3.1 RK1 scheme

Equation (2.11) enables us to compute or estimate the departure point. For instance, if u is constant, the departure point at time t^n for the point $x_{i+\frac{1}{2}}$ at time t^{n+1} is given by:

$$X(t^n, t^{n+1}; x_{i+\frac{1}{2}}) = x_{i+\frac{1}{2}} - u \Delta t. \quad (2.23)$$

In general, the estimated departure point, denoted by $\tilde{x}_{i+\frac{1}{2}}^n$, takes the form:

$$\tilde{x}_{i+\frac{1}{2}}^n = x_{i+\frac{1}{2}} - \tilde{u}_{i+\frac{1}{2}}^n \Delta t, \quad (2.24)$$

where $\tilde{u}_{i+\frac{1}{2}}^n$ represents the time-averaged wind and approximates:

$$\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} u(X(\theta, t^{n+1}; x_{i+\frac{1}{2}}), \theta), d\theta. \quad (2.25)$$

The departure point $\tilde{x}_{i+\frac{1}{2}}^n$ is said to be p -order accurate if there exists a constant C that does not depend on Δt , such that:

$$X(t^n, t^{n+1}; x_{i+\frac{1}{2}}) - \tilde{x}_{i+\frac{1}{2}}^n = C \Delta t^p. \quad (2.26)$$

One possible way of estimating the time-averaged wind is by using:

$$\tilde{u}_{i+\frac{1}{2}}^n = u_{i+\frac{1}{2}}^n, \quad (2.27)$$

as in the original PPM scheme by Colella and Woodward (1984). This scheme will be referred to as **RK1**. Our objective now is to determine the constant C and the value of p in Equation (2.26) in the following proposition. It is useful to introduce the concept of a material derivative beforehand:

$$\frac{Dh}{Dt} = \frac{\partial h}{\partial t} + u \frac{\partial h}{\partial x},$$

where h is a function belonging to \mathcal{C}^1 .

Proposition 2.3. *If $u \in \mathcal{C}^1$ and the time-averaged wind is computed using Equation (2.27), then the departure point from Equation (2.24) satisfies:*

$$X(t^n, t^{n+1}; x_{i+\frac{1}{2}}) - \tilde{x}_{i+\frac{1}{2}}^n = C\Delta t^2, \quad (2.28)$$

for a constant C that depends on u .

Proof. Using Corollary A.1 for the function $f(t) = u(X(t, t^{n+1}; x_{i+\frac{1}{2}}), t)$ in Equation (2.11), we obtain:

$$X(t^n, t^{n+1}; x_{i+\frac{1}{2}}) = x_{i+\frac{1}{2}} - u(X(t^n, t^{n+1}; x_{i+\frac{1}{2}}), t^n)\Delta t - \frac{1}{2} \frac{Du}{Dt}(X(\theta_1, t^{n+1}; x_{i+\frac{1}{2}}), \theta_1)\Delta t^2, \quad (2.29)$$

for $\theta_1 \in [t^n, t^{n+1}]$. Using Taylor's expansion of $u(X(t, t^{n+1}; x_{i+\frac{1}{2}}), t^n)$, we have:

$$u(X(t^n, t^{n+1}; x_{i+\frac{1}{2}}), t^n) = u_{i+\frac{1}{2}}^n - \left(u \frac{\partial u}{\partial x} \right)(X(\theta_2, t^{n+1}; x_{i+\frac{1}{2}}), t^n)\Delta t, \quad (2.30)$$

for $\theta_2 \in [t^n, t^{n+1}]$. Substituting Equation (2.30) into Equation (2.29), we obtain the desired constant C given by:

$$C = C(\theta_1, \theta_2) = -\frac{1}{2} \frac{Du}{Dt}(X(\theta_1, t^{n+1}; x_{i+\frac{1}{2}}), \theta_1) - \left(u \frac{\partial u}{\partial x} \right)(X(\theta_2, t^{n+1}; x_{i+\frac{1}{2}}), t^n). \quad (2.31)$$

□

2.3.2 RK2 scheme

Before presenting a higher-order estimate for the departure point, let us recall the definition of the CFL number.

Definition 2.8. *For Problem 2.4, the CFL number at an edge $x_{i+\frac{1}{2}}$ and at a time level t^n is defined by*

$$c_{i+\frac{1}{2}}^n = \frac{\Delta t}{\Delta x} u_{i+\frac{1}{2}}^n. \quad (2.32)$$

The CFL number is the maximum of the values $c_{i+\frac{1}{2}}^n$. The problem of estimating the departure point is very common in Semi-Lagrangian schemes, which are quite popular in atmospheric modeling. For a review of departure point calculation methods, we refer to Tumolo (2011, Chapter 3) and the references therein. There are different approaches to compute the departure point, such as integrating the ODE from Equation 2.1 using different time integrators (D. Durran, 2011) backward in time. The Runge-Kutta methods are a possible choice to compute the departure point (*cf. e.g.* Guo et al. (2014), Lu et al. (2022)). In this work, we shall consider a second-order Runge-Kutta method to compute the departure point, which we express in terms of $\tilde{u}_{i+\frac{1}{2}}^n$ using the following equations

(D. R. Durran, 2010):

$$\begin{aligned}\tilde{x}_{i+\frac{1}{2}}^{n+\frac{1}{2}} &= x_{i+\frac{1}{2}} - u_{i+\frac{1}{2}}^n \frac{\Delta t}{2} = x_{i+\frac{1}{2}} - c_{i+\frac{1}{2}}^n \frac{\Delta x}{2}, \\ \tilde{u}_{i+\frac{1}{2}}^n &= u\left(\tilde{x}_{i+\frac{1}{2}}^{n+\frac{1}{2}}, t^n + \frac{\Delta t}{2}\right).\end{aligned}\quad (2.33)$$

Notice that this scheme requires values of u at points that are not grid points, both in time and space. This problem is addressed firstly using a second-order extrapolation in time:

$$u_{i+\frac{1}{2}}^{n+\frac{1}{2}} = \frac{3}{2}u_{i+\frac{1}{2}}^n - \frac{1}{2}u_{i+\frac{1}{2}}^{n-1}. \quad (2.34)$$

Then we use linear interpolation in space:

$$\tilde{u}_{i+\frac{1}{2}}^n = \begin{cases} (1 - \alpha_{i+\frac{1}{2}}^n)u_{i+\frac{1}{2}-k}^{n+\frac{1}{2}} + \alpha_{i+\frac{1}{2}}^n u_{i-\frac{1}{2}-k}^{n+\frac{1}{2}} & \text{if } u_{i+\frac{1}{2}}^n \geq 0, \\ \alpha_{i+\frac{1}{2}}^n u_{i+\frac{3}{2}-k}^{n+\frac{1}{2}} + (1 - \alpha_{i+\frac{1}{2}}^n)u_{i+\frac{1}{2}-k}^{n+\frac{1}{2}} & \text{if } u_{i+\frac{1}{2}}^n < 0, \end{cases} \quad (2.35)$$

where $c_{i+\frac{1}{2}}^n = \alpha_{i+\frac{1}{2}}^n + k$, $k = \lfloor \frac{c_{i+\frac{1}{2}}^n}{2} \rfloor$, $\alpha_{i+\frac{1}{2}}^n \in [0, 1[$, and $\lfloor \cdot \rfloor$ is the floor function. This scheme leads to a third-order error in the departure point estimate (see e.g. D. R. Durran (2010, Section 7.1.2)). This scheme shall be referred to as **RK2**. Notice that for this scheme, we need ghost values for the velocity, depending on how large the CFL number is. Finally, the impact of the departure point approximation on the time-averaged flux is investigated in Section A.5.3.

2.4 Reconstruction

In this section, we will review the Piecewise-Parabolic Method (PPM). The analysis of its accuracy will be presented in Section A.4. PPM was originally proposed by Colella and Woodward (1984) for gas dynamic simulations, and its applicability to atmospheric simulations has been demonstrated by Carpenter et al. (1990). This method is based on utilizing parabolas to reconstruct the function using its average values, ensuring both mass conservation and monotonicity. PPM is an extension of the Piecewise-Linear Method introduced by Van Leer (1977), and it is implemented in the FV3 model using the dimension splitting method developed by Lin and Rood (1996).

Let's consider a function q defined in $\Omega = [a, b]$ and a Δx -grid covering Ω . We assume that we are given the average values $Q_i = \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} q(x) dx$ for each control volume X_i , where $i = 1, \dots, N$. In this context, it is convenient to define the Δx -grid function $Q = (Q_1, \dots, Q_N) \in \mathbb{R}^N$. To facilitate the discussion, we introduce the indicator function $\chi_i(x)$ for each control volume X_i , defined as:

$$\chi_i(x) = \begin{cases} 1 & \text{if } x \in X_i, \\ 0 & \text{otherwise.} \end{cases}$$

Drawing inspiration from Stoer and Bulirsch (2002, Chapter 1), we consider a family of

functions $\Phi(\xi; \mu)$ defined for $\xi \in [0, 1]$, depending on a parameter $\mu = (\mu_0, \mu_1, \dots, \mu_d) \in \mathbb{R}^{d+1}$. The reconstruction problem involves finding a piecewise function:

$$\tilde{q}(x; Q) = \sum_{i=1}^N \chi_i(x) q_i(x; Q), \quad (2.36)$$

where $q_i(x; Q) = \Phi\left(\frac{x-x_{i-\frac{1}{2}}}{\Delta x}; \alpha_i\right)$ and $\alpha_i = (\alpha_{i0}, \alpha_{i1}, \dots, \alpha_{id}) \in \mathbb{R}^{d+1}$. It is required that:

$$\frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} q P d(x; Q), dx = \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} q_i(x; Q), dx = \int_0^1 \Phi(\xi; \alpha_i), d\xi = Q_i,$$

which means that $q_i(x; Q)$ preserves the mass within each control volume X_i .

Notice that, given $q_i(x; Q) = \Phi\left(\frac{x-x_{i-\frac{1}{2}}}{\Delta x}; \alpha_i\right)$, it is reasonable to expect that $\Phi(0; \alpha_i)$ approximates $q_i(x_{i-\frac{1}{2}})$ and $\Phi(1; \alpha_i)$ approximates $q_i(x_{i+\frac{1}{2}})$. Additionally, if both q and Φ are sufficiently differentiable, $\Phi^{(l)}(0; \alpha_i)$ should approximate $(\Delta x)^l q^{(l)}(x_{i-\frac{1}{2}})$ and $\Phi^{(l)}(1; \alpha_i)$ should approximate $(\Delta x)^l q^{(l)}(x_{i+\frac{1}{2}})$, provided these derivatives exist.

One approach to estimating these values at the edges $x_{i+\frac{1}{2}}$ using the average values Q is by employing a reconstruction method based on primitive functions (LeVeque, 2002, Chapter 17). It is worth noting that if we define:

$$Q(x) = \int_a^x q(\xi), d\xi, \quad (2.37)$$

we have $Q^{(l)}(x) = q^{(l-1)}(x)$. Specifically, $Q^{(l)}(x_{i+\frac{1}{2}}) = q^{(l-1)}(x_{i+\frac{1}{2}})$ and $Q^{(l)}(x_{i+\frac{1}{2}}) = \Delta x \sum_{k=1}^i Q_k$, for all $i = 0, \dots, N$. Therefore, we can employ finite-difference schemes to estimate $q^{(l-1)}(x_{i+\frac{1}{2}})$ using the Δx -grid function Q , given that it is assumed to be known.

Let us assume that the l -th derivative of Q at $x_{i+\frac{1}{2}}$ is approximated using a stencil $S_{i+\frac{1}{2}}^{(l)}$ and weights $\beta_{k,i}^{(l)}$, where $k \in S_{i+\frac{1}{2}}^{(l)}$. When d is odd, we can seek a parameter $\alpha_i \in \mathbb{R}^{d+1}$ that ensures mass conservation and approximates q and its derivatives at the edges by solving the following system:

$$\begin{cases} \int_0^1 \Phi(\xi; \alpha_i), d\xi &= Q_i, \\ \Phi^{(l)}(0; \alpha_i) &= (\Delta x)^l \sum_{k \in S_{i-\frac{1}{2}}^{(l)}} \beta_{k,i}^{(l)} Q_k, \end{cases} \quad \text{for } l = 0, \dots, d-1. \quad (2.38)$$

If d is even, similarly we look for a parameter $\alpha_i \in \mathbb{R}^{d+1}$ that solves:

$$\begin{cases} \int_0^1 \Phi(\xi; \alpha_i), d\xi &= Q_i, \\ \Phi^{(l)}(0; \alpha_i) &= (\Delta x)^l \sum_{k \in S_{i-\frac{1}{2}}^{(l)}} \beta_{k,i}^{(l)} Q_k, \quad \text{for } l = 0, \dots, \frac{d}{2}-1, \\ \Phi^{(l)}(1; \alpha_i) &= (\Delta x)^l \sum_{k \in S_{i+\frac{1}{2}}^{(l)}} \beta_{k,i}^{(l)} Q_k, \quad \text{for } l = 0, \dots, \frac{d}{2}-1. \end{cases} \quad (2.39)$$

The reconstruction problem becomes linear when $\Phi(\xi; \mu)$ can be expressed as:

$$\Phi(\xi; \mu) = \sum_{k=0}^d \mu_k \Phi_k(\xi),$$

where Φ_k are functions defined on $[0, 1]$. In this case, Equation (2.38) and Equation (2.39) form $(d+1) \times (d+1)$ linear systems. It is common to assume that the Φ_k 's are linearly independent. Therefore, we have described a method that allows us to reconstruct a function from its average values, preserving its mass in each control volume, and approximating q at the edges. This method works for functions Φ_k as long as they are sufficiently differentiable. For example, choosing $d = 0$ and $\Phi_0(\xi) = 1$ gives us piecewise constant functions, as used in Godunov (1959). If we choose $d = 1$, $\Phi_0(\xi) = 1$, and $\Phi_1(\xi) = \xi$, we obtain a piecewise linear reconstruction, similar to Van Leer (1977). For polynomial reconstruction schemes, we refer to Engwirda and Kelley (2016) and the references therein.

2.4.1 The Piecewise-Parabolic Method

Hereafter, we are going the focus on the piecewise parabolic method from Colella and Woodward (1984) that uses $d = 2$, $\Phi_0(\xi) = 1$, $\Phi_1(\xi) = \xi$, $\Phi_2(\xi) = (1 - \xi)\xi$. In this case, we denote q_{Pd} by q_{PP} . In order to follow the notation from Colella and Woodward (1984), we write $\alpha_{0i} = q_{L,i}$, $\alpha_{1i} = \Delta q_i$ and $\alpha_{2i} = q_{6,i}$. Therefore, each q_i may be expressed as:

$$q_i(x; Q) = q_{L,i} + \Delta q_i z_i(x) + q_{6,i} z_i(x)(1 - z_i(x)), \quad \text{where } z_i(x) = \frac{x - x_{i-\frac{1}{2}}}{\Delta x}, \quad x \in X_i, \quad (2.40)$$

where the values $q_{L,i}$, Δq_i and $q_{6,i}$ will be specified latter. Note that each z_i is just a normalization function that maps X_i onto $[0, 1]$. It is easy to see that $\lim_{x \rightarrow x_{i-\frac{1}{2}}^+} q_i(x; Q) = q_{L,i}$. If we define $q_{R,i} = \lim_{x \rightarrow x_{i+\frac{1}{2}}^-} q_i(x; Q)$, then we have:

$$\Delta q_i = q_{R,i} - q_{L,i}. \quad (2.41)$$

The average value of q_i is given by:

$$\frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} q_i(x; Q) dx = \frac{(q_{L,i} + q_{R,i})}{2} + \frac{q_{6,i}}{6}. \quad (2.42)$$

Under the hypothesis of mass conservation, we have:

$$q_{6,i} = 6 \left(Q_i - \frac{(q_{L,i} + q_{R,i})}{2} \right). \quad (2.43)$$

Therefore, we have found the parameters Δq_i and $q_{6,i}$ as functions of the parameters $q_{L,i}$ and $q_{R,i}$, such that the parabola q_i from (2.36) guarantees mass conservation. To completely determine the parabola q_i , we need to set the values $q_{L,i}$ and $q_{R,i}$, which, as we have seen, represent the limits of q_i when x tends to the left and right boundaries of X_i , respectively. Hence, it is natural to seek for $q_{L,i}$ as an approximation of $q(x_{i-\frac{1}{2}})$ and $q_{R,i}$ as an approximation of $q(x_{i+\frac{1}{2}})$. As we mentioned before in after introducing Equation (2.37), this is achieved using finite-differences. An explicit expression for the approximation of $q(x_{i-\frac{1}{2}})$,

denoted by $q_{i+\frac{1}{2}}$, is given by (Colella & Woodward, 1984):

$$q_{i+\frac{1}{2}} = \frac{1}{2} \left(Q_{i+1} + Q_i \right) - \frac{1}{6} \left(\delta Q_{i+1} - \delta Q_i \right), \quad (2.44)$$

where δQ_i is the average slope in the i -th control-volume:

$$\delta Q_i = \frac{1}{2} \left(Q_{i+1} - Q_{i-1} \right). \quad (2.45)$$

We notice that Formula (2.45) may be rewritten more explicitly as:

$$q_{i+\frac{1}{2}} = \frac{7}{12} \left(Q_{i+1} + Q_i \right) - \frac{1}{12} \left(Q_{i+2} + Q_{i-1} \right). \quad (2.46)$$

The Formula (2.46) is fourth-order accurate if q is at least C^4 (Colella & Woodward, 1984). Indeed, we prove this later in Proposition A.1. An explicit expression for the values of $q_{R,i}$ and $q_{L,i}$ are given by:

$$q_{R,i} = q_{i+\frac{1}{2}} = \frac{7}{12} \left(Q_{i+1} + Q_i \right) - \frac{1}{12} \left(Q_{i+2} + Q_{i-1} \right), \quad (2.47)$$

$$q_{L,i} = q_{i-\frac{1}{2}} = \frac{7}{12} \left(Q_i + Q_{i-1} \right) - \frac{1}{12} \left(Q_{i+1} + Q_{i-2} \right). \quad (2.48)$$

We point out that a fifth-order accurate for the values of $q_{R,i}$ and $q_{L,i}$ is also possible, as it was developed by Putman and Lin (2007) based on the work Suresh and Huynh (1997). The fifth-order reconstruction formula reads:

$$q_{R,i} = \frac{1}{60} \left(2Q_{i-2} - 13Q_{i-1} + 47Q_i + 27Q_{i+1} - 3Q_{i+2} \right), \quad (2.49)$$

$$q_{L,i} = \frac{1}{60} \left(-3Q_{i-2} + 27Q_{i-1} + 47Q_i - 13Q_{i+1} + 2Q_{i+2} \right). \quad (2.50)$$

However, we notice that this reconstruction scheme allows discontinuity of the Piecewise-Parabolic function at the control volume edges since the stencil it is not symmetric. The PPM scheme, utilizing Equations (2.47) and (2.48), will be referred to as **PPM-0**. On the other hand, the PPM scheme, employing Equations (2.49) and (2.50), will be denoted as **PPM-PL07**.

2.4.2 Monotonization

This section is dedicated to presenting possible ways of ensuring the creation of new extrema values in the PPM reconstruction. We are going to present the original monotonic scheme from Colella and Woodward (1984) and an alternative scheme from Lin (2004), which was an attempt to reduce the diffusion of the original scheme Colella and Woodward (1984) and is currently employed in the FV3 dynamical core (L. Harris et al., 2021).

Limiter from Colella and Woodward (1984) - PPM-CW84

To avoid numerical oscillations in the parabolas, especially when discontinuities are present, Colella and Woodward (1984) ensures that the reconstructed value at cell edges (namely, $q_{i+\frac{1}{2}}$) does not stay outside of the range of its neighbors average values (Q_i and Q_{i+1}). This can be achieved by replacing the term δQ_i in Equation (2.44) by the values $\delta_m Q_i$ given by:

$$\delta_m Q_i = \begin{cases} \max(|\delta Q_i|, 2|Q_{i+1} - Q_i|, 2|Q_i - Q_{i-1}|) \cdot \text{sgn}(\delta Q_i) & \text{if } (Q_{i+1} - Q_i)(Q_i - Q_{i-1}) > 0, \\ 0 & \text{otherwise,} \end{cases} \quad (2.51)$$

where sgn denotes the sign function. To ensure, monotonicity we also must ensure that the parabola has values between $q_{R,i}$ and $q_{L,i}$. This step will introduce a discontinuity on the edges of the PPM approximation. If Q_i is the local maximum/minimum, then we make the parabola constant. This is expressed as:

$$q_{L,i} \leftarrow Q_i, \quad q_{R,i} \leftarrow Q_i, \quad \text{if } (Q_{R,i} - Q_i)(Q_i - Q_{L,i}) \geq 0 \quad (2.52)$$

This step eliminates the introduction of new extremes when we already have an extremum. The other case where we need to modify the values $q_{L,i}$ and $q_{R,i}$ is when the extrema of the parabola falls in $[x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}]$. It is easy to see from Equation (A.33) that, the extrema of the parabola falling in $[x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}]$ is equivalent to $|\Delta q_i| \leq |q_{6,i}|$. In this case, the values are updated as follows:

$$\begin{cases} q_{L,i} \leftarrow 3Q_i - 2q_{R,i} & \text{if } \Delta q_i \cdot q_{6,i} > (\Delta q_i)^2, \\ q_{R,i} \leftarrow 3Q_i - 2q_{L,i} & \text{if } -(\Delta q_i)^2 > \Delta q_i \cdot q_{6,i} \end{cases} \quad (2.53)$$

In this step, we are changing the value at the edge where the extreme is closer and ensuring again that no new extreme is created. This scheme is referred to as **PPM-CW84**.

Limiter from Lin (2004) - PPM-L04

Similarly to Colella and Woodward (1984), Lin (2004) reduces numerical oscillations in the parabolas by replacing the term δQ_i in Equation (2.44) with the values $\delta_m Q_i$ given by:

$$\delta_m Q_i = \max(|\delta Q_i|, 2\delta Q_{\min,i}, 2\delta Q_{\max,i}) \cdot \text{sgn}(\delta Q_i), \quad (2.54)$$

where $\delta Q_{\min,i} = Q_i - \min(Q_{i+1}, Q_i, Q_{i-1})$ and $\delta Q_{\max,i} = \max(Q_{i+1}, Q_i, Q_{i-1}) - Q_i$. The monotonicity is achieved by the following scheme:

$$q_{L,i} \leftarrow Q_i - \max(|\delta_m Q_i|, |q_{L,i} - Q_i|) \cdot \text{sgn}(\delta_m Q_i), \quad (2.55)$$

$$q_{R,i} \leftarrow Q_i - \max(|\delta_m Q_i|, |q_{R,i} - Q_i|) \cdot \text{sgn}(\delta_m Q_i). \quad (2.56)$$

This scheme may be further improved to reduce the diffusion even more, as described by Lin (2004), but we are not going to assess this approach here. This scheme is referred to as **PPM-L04**.

2.5 Flux

Let's consider the framework outlined in Problem 2.4. Assuming that $Q^n \in \mathbb{R}^N$ is known, our objective is to compute the values Q^{n+1} . To accomplish this, we utilize a scheme similar to the one presented in Problem 2.4, taking into account the presence of a reconstruction function $\tilde{q}(x; Q^n)$ as discussed in Section 2.4, and an initial estimation $\tilde{x}_{i+\frac{1}{2}}^n = x_{i+\frac{1}{2}} - \tilde{u}_{i+\frac{1}{2}}^n \Delta t$ for a time-averaged wind $\tilde{u}_{i+\frac{1}{2}}^n$ as explained in Section 2.3. The numerical flux function is then expressed as:

$$F_{i+\frac{1}{2}}^n = \frac{1}{\Delta t} \int_{x_{i+\frac{1}{2}} - \tilde{u}_{i+\frac{1}{2}}^n \Delta t}^{x_{i+\frac{1}{2}}} \tilde{q}(x; Q^n) dx. \quad (2.57)$$

Notice that if we define the averaged CFL number,

$$\tilde{c}_{i+\frac{1}{2}}^n = \tilde{u}_{i+\frac{1}{2}}^n \frac{\Delta t}{\Delta x},$$

where $\tilde{c}_{i+\frac{1}{2}}^n = k + \alpha_{i+\frac{1}{2}}^n$, $k = \lfloor \tilde{c}_{i+\frac{1}{2}}^n \rfloor$, $\alpha_{i+\frac{1}{2}}^n \in [0, 1[$, we can express the numerical flux as (Y. Chen et al., 2017; Lin & Rood, 1996):

$$F_{i+\frac{1}{2}}^n = \frac{1}{\Delta t} \begin{cases} \Delta x \sum_{l=0}^{k-1} Q_{i-l} + \int_{x_{i-k+\frac{1}{2}} - \alpha_{i+\frac{1}{2}}^n \Delta x}^{x_{i-k+\frac{1}{2}}} \tilde{q}(x; Q^n) dx, & \text{if } \tilde{u}_{i+\frac{1}{2}}^n \geq 0, \\ \Delta x \sum_{l=0}^{k-1} Q_{i-l} - \int_{x_{i-k+\frac{1}{2}} - \alpha_{i+\frac{1}{2}}^n \Delta x}^{x_{i-k+\frac{1}{2}}} \tilde{q}(x; Q^n) dx, & \text{if } \tilde{u}_{i+\frac{1}{2}}^n < 0. \end{cases} \quad (2.58)$$

where we used that \tilde{q} preserves the local mass.

We will provide explicit expressions for the integrals in Equation (2.58) when using the PPM method. For each control volume edge, denoted by $i = 0, \dots, N$, and $y > 0$, we define the following averages of the Piecewise-Parabolic approximation, as defined in Equation (2.36) for Q^n (Colella & Woodward, 1984):

$$F_{L,i+\frac{1}{2}}(y) = \frac{1}{y} \int_{x_{i+\frac{1}{2}} - y}^{x_{i+\frac{1}{2}}} \tilde{q}(x; Q^n) dx, \quad (2.59)$$

and

$$F_{R,i+\frac{1}{2}}(y) = \frac{1}{y} \int_{x_{i+\frac{1}{2}}}^{x_{i+\frac{1}{2}} + y} \tilde{q}(x; Q^n) dx. \quad (2.60)$$

If $y \leq \Delta x$, then both of the above integral domains are constrained to a single control volume. Thus, it follows from a straightforward computation using Equation (2.40) that:

$$F_{L,i+\frac{1}{2}}(y) = \frac{1}{y} \int_{x_{i+\frac{1}{2}} - y}^{x_{i+\frac{1}{2}}} q_i(x; Q^n) dx = q_{R,i} + \frac{(q_{6,i} - \Delta q_i)}{2\Delta x} y - \frac{q_{6,i}}{3\Delta x^2} y^2, \quad (2.61)$$

and

$$F_{R,i+\frac{1}{2}}(y) = \frac{1}{y} \int_{x_{i+\frac{1}{2}}}^{x_{i+\frac{1}{2}} + y} q_{i+1}(x; Q^n) dx = q_{L,i+1} + \frac{(q_{6,i+1} + \Delta q_{i+1})}{2\Delta x} y - \frac{q_{6,i+1}}{3\Delta x^2} y^2. \quad (2.62)$$

The numerical flux function for PPM is then defined by:

$$F_{i+\frac{1}{2}}^n = \frac{1}{\Delta t} \begin{cases} \Delta x \sum_{l=0}^{k-1} Q_{i-l} + \Delta x \alpha_{i+\frac{1}{2}}^n F_{L,i+\frac{1}{2}}(\alpha_{i+\frac{1}{2}}^n \Delta x) & \text{if } \tilde{u}_{i+\frac{1}{2}}^n \geq 0, \\ \Delta x \sum_{l=0}^{k-1} Q_{i-l} + \Delta x \alpha_{i+\frac{1}{2}}^n F_{R,i+\frac{1}{2}}(-\alpha_{i+\frac{1}{2}}^n \Delta x) & \text{if } \tilde{u}_{i+\frac{1}{2}}^n < 0, \end{cases} \quad (2.63)$$

where $\tilde{u}_{i+\frac{1}{2}}^n$ is the velocity used in the departure point estimation. If the monotonic scheme from Lin (2004) is employed, then $F_{i+\frac{1}{2}}$ uses the stencil $S_{i+\frac{1}{2}} = \{i-3, i-2, i-1, i, i+1, i+2, i+3\}$. Otherwise the stencil used by $F_{i+\frac{1}{2}}$ is given by $S_{i+\frac{1}{2}} = \{i-2, i-1, i, i+1, i+2\}$ for all the other schemes that we presented. The accuracy of this flux formulation is described in Section A.5.3.

2.6 Numerical experiments

This section is dedicated to presenting the numerical results of the PPM and its variations discussed here. We will consider the following reconstruction schemes: PPM-0, PPM-PL07, PPM-CW84, and PPM-L04, which were presented in Section 2.4, as well as the departure point schemes RK1 and RK2 described in Section 2.3. The code used in this section can be found in Appendix C.

For all the simulations presented here, we will consider the spatial domain $[0, 1]$ and the time interval $[0, 5]$. The relative change at time step n in the mass is computed as:

$$\frac{|M^n - M^0|}{|M_0|},$$

where M^n is given by Equation (2.22). For all the simulations, the mass is preserved with machine precision. Furthermore, we compute the initial average values $Q_i(0)$ using the initial values of q_i^0 at the control volume centroids for all simulations, which is second-order accurate by Proposition 2.2. In the error calculation, only when q_0 is given by Equation (2.66), we replace $Q_i(t^n)$ by its centroid value $q_i(t^n)$, which again gives a second-order approximation by Proposition 2.2.

As a first numerical experiment, we consider a discontinuous initial condition given by:

$$q_0(x) = \begin{cases} 1 & \text{if } x \in [0.4, 0.6], \\ 0 & \text{otherwise.} \end{cases} \quad (2.64)$$

for the linear advection equation with constant velocity, which we adopt as $u = 0.2$. It is easy to check that the exact solution of Problem 2.1 is given by $q_0(x - ut)$. We will use a CFL number equal to 0.8. The spatial domain will be $[0, 1]$, and the time integration interval will be $[0, 5]$. Since we are assuming periodic boundary conditions, the period is equal to 5. Therefore, the simulations presented here will advect an initial profile for one time period. The departure schemes RK1 and RK2 compute the departure point exactly in this case, so we will only use the RK1 scheme. In Figure 2.4, we present the obtained results. It is evident that the monotonic schemes exhibit a significant advantage. These schemes effectively prevent the strong oscillations observed in the non-monotonic schemes, as well as the generation of new extrema, which aligns with our expectations.

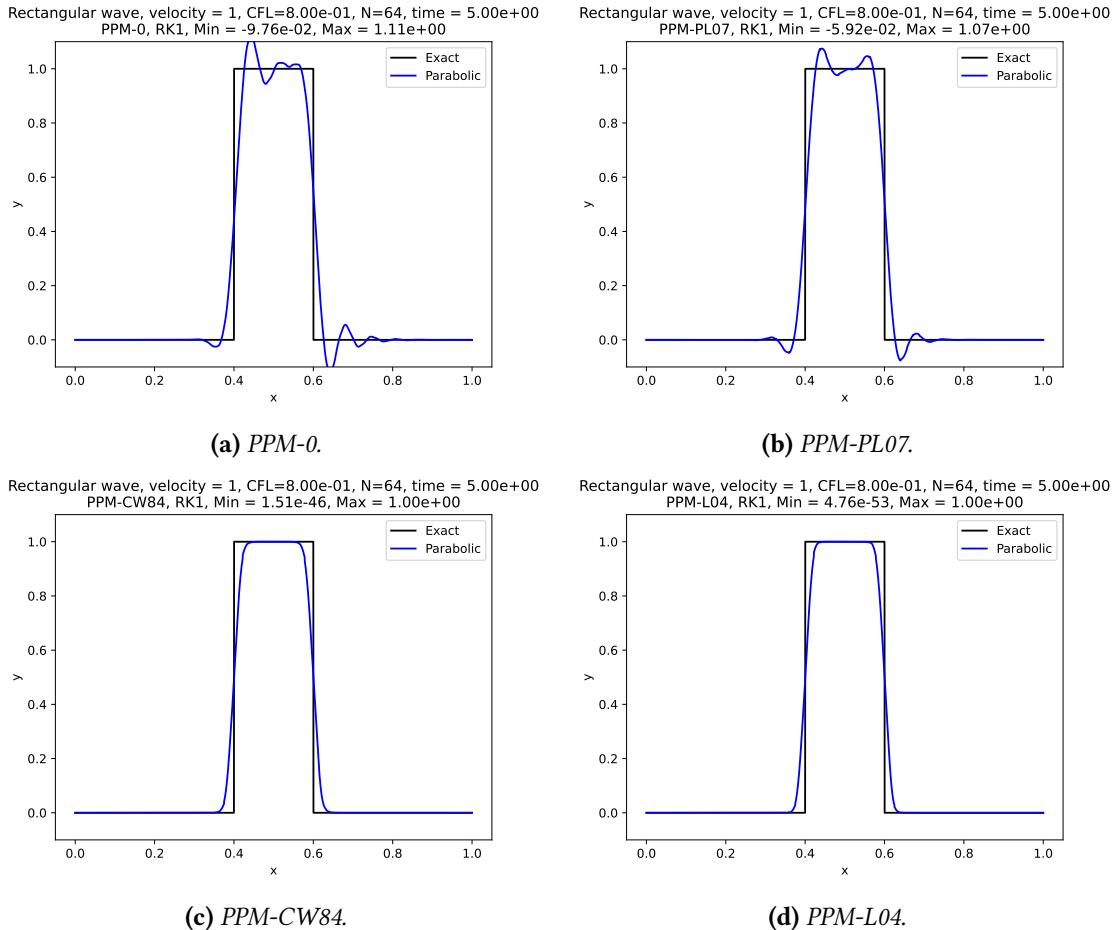


Figure 2.4: Linear advection experiment using a constant velocity equal to 0.2, a CFL number equal to 0.8, $N = 64$ cells, and the initial condition is given by Equation (2.64). These figures show the advected profile after 5 time units (one time period). Reconstruction schemes employed: PPM-0 (a), PPM-PL07 (b), PPM-CW84 (c) and PPM-L04 (d).

As a second experiment, we shall investigate the how the PPM schemes behave when the velocity is variable. We are going to consider the velocity

$$u(x, t) = u_0 \cos\left(\frac{\pi t}{T}\right) \sin^2\left(\pi\left(x - \frac{t}{T}\right)\right) + u_1. \quad (2.65)$$

We adopt the parameters $u_0 = u_1 = 0.2$, and $T = 5$. Following the approach in Trefethen (2000), we initialize the periodic Gaussian profile defined as:

$$q(x) = \exp(-10 \cos^2(\pi x)), \quad x \in [0, 1]. \quad (2.66)$$

The velocity function given by Equation (2.65) is based on the deformational flow test case in Nair and Lauritzen (2010). As the velocity is variable, we utilize the departure point schemes RK1 and RK2. In this case, the solution exhibits a period of 5 time units, meaning that the profile deforms and returns to its initial shape and position after 5 time units, allowing us to compute the error. Indeed, in Figure 2.5, we show how the solution behaves using a high-resolution ($N = 1024$), the PPM-L04 scheme and the RK1 departure point

scheme.

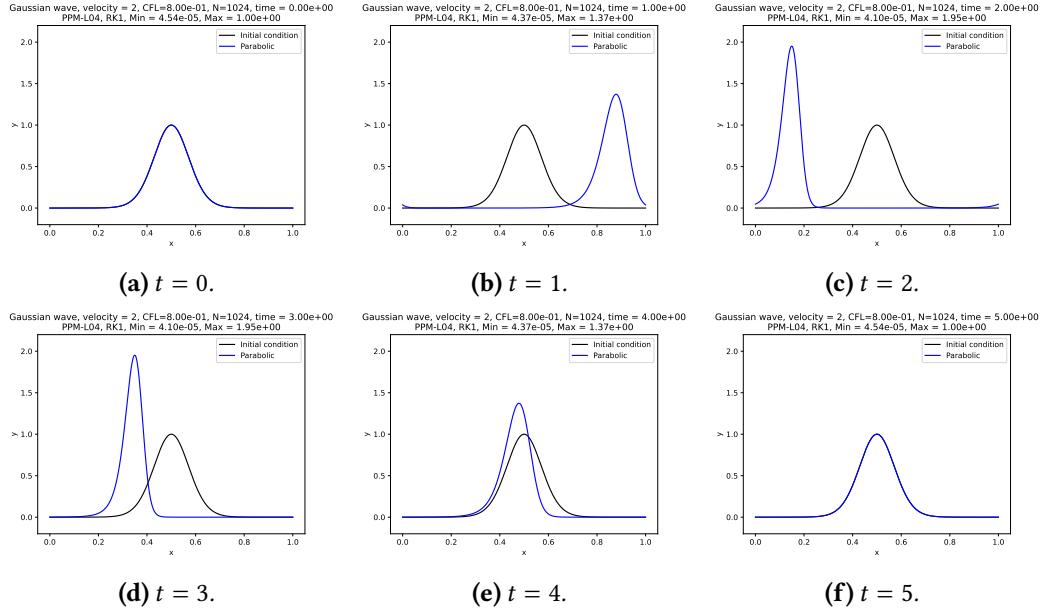


Figure 2.5: Linear advection experiment using the velocity from Equation (2.66), a CFL number equal to 0.8, $N = 1024$ cells, and the initial condition is given by Equation (2.66) (2.5a). These figures show the advected profile at 1 (2.5b), 2 (2.5c), 3 (2.5d), 4 (2.5e), and 5 (2.5f) time units. We are using the PPM-L04 scheme with the RK1 departure point scheme. .

To investigate the error convergence, we employ $\Delta x^{(k)}$ -grids with $\Delta x^{(k)} = 1/2^k$ for $k = 4, \dots, 10$. To measure the accuracy, we consider the relative error in the maximum norm as follows:

$$E_k = \max_{n=0, \dots, N_T} \frac{\|Q^n - Q(t^n)\|_{\infty, \Delta x}}{\|Q(t^n)\|_{\infty, \Delta x}}.$$

The convergence rate is defined by

$$CR_k = \frac{\ln \left(\frac{E_k}{E_{k-1}} \right)}{\ln 2}, \quad \text{for } k = 5, \dots, 10.$$

The difference between the RK1 and RK2 schemes becomes clear when observing the relative error in Figure 2.6 and the convergence rate in Figure 2.7. The RK1 scheme results in a first-order error in the departure point, which dominates the total error for all PPM schemes. This observation is in agreement with the discussion in Section 2.3. On the other hand, when employing the RK2 scheme, we can achieve a third-order accuracy for the PPM-0 and PPM-PL07 schemes, surpassing the expected second-order accuracy of the departure point scheme. This experiment illustrates the impact of the departure point calculation error on the overall error. Furthermore, regardless of the departure point scheme, the PPM-PL07 scheme exhibits a smaller error. Among the monotonic schemes, the PPM-L04 has a slightly smaller error compared to PPM-CW84.

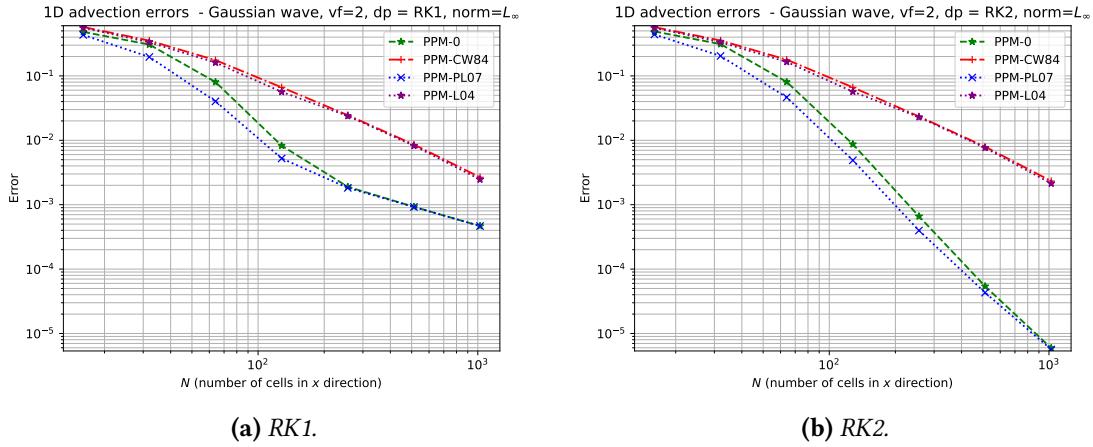


Figure 2.6: Relative error for different PPM schemes using the RK1 (left) and RK2 (right) departure point scheme for the initial condition given by Equation (2.66) and the variable velocity given by Equation (2.65).

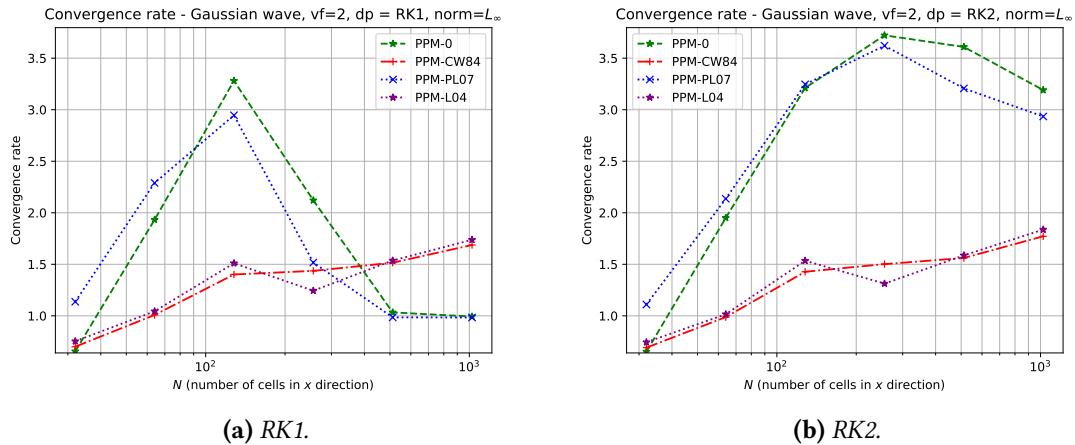


Figure 2.7: Convergence rate for different PPM schemes using the RK1 (left) and RK2 (right) departure point scheme for the initial condition given by Equation (2.66) and the variable velocity given by Equation (2.65).

2.7 Concluding remarks

In this chapter, we provided a general overview of 1D FV-SL schemes for the advection equation. We discussed the three essential tasks involved in these schemes. The first task is the reconstruction of a function from its average values. We employed the PPM method introduced by Colella and Woodward (1984) and its variants. We were able to achieve third-order accuracy in the reconstruction process, even without imposing monotonicity constraints. The second task involves computing the departure point of the control volume edges. For this purpose, we utilized the first-order departure point calculation method from Colella and Woodward (1984) known as RK1. Additionally, we explored a second-order approach by employing a two-stages Runge-Kutta scheme (RK2) to integrate the departure point ordinary differential equation (ODE). Lastly, the third task entails computing the

flux, which involves integrating the reconstructed function over a domain determined by the departure point.

From the numerical experiments, we observed that the PPM-PL07 (Putman & Lin, 2007), which uses fifth-order reconstruction at the edges, leads to third-order accuracy but is more accurate than the PPM-0 (Colella & Woodward, 1984) scheme, which uses fourth-order reconstruction at the edges. Regarding the monotonic schemes, we observed that both schemes were able to avoid overshoots, with PPM-L04 (Lin, 2004) being more accurate than the scheme PPM-CW84 proposed by Colella and Woodward (1984).

The difference between the departure point schemes became apparent when we performed a test with variable velocity. The simulation using the RK1 scheme resulted in a final first-order error, despite the scheme having third-order accuracy in space. However, the RK2 scheme preserved third-order accuracy despite being only second-order accurate. We expect that, in general, combining PPM with the RK2 scheme should result in at least second-order accuracy.

Clearly, the RK2 scheme is more computationally expensive since it requires time extrapolation and linear interpolation of the velocity field. One possible way to reduce its cost would be to use larger CFL numbers allowed by the FV-SL schemes, as discussed in Section 2.5.

Chapter 3

Two-dimensional finite-volume methods

In Chapter 2, we tackled the problem of solving the one-dimensional linear advection equation using the finite-volume method based on PPM. In this Chapter, we are interested in solving the two-dimensional linear advection equation using the finite-volume method. This step plays a key role in this work, since as we shall see in Chapter 5, the solution of the linear advection equation on the cubed-sphere boils down to solving one two-dimensional linear advection equations at each cube face with interpolation between the adjacent panels.

A natural way to derive a finite-volume method for the two-dimensional linear advection equation would be to extend the PPM for two dimensions. Indeed, a piecewise bi-parabolic extension of PPM was proposed by Rančić (1992) using a semi-lagrangian temporal discretization. However, the major problem of this method is its quite expensive computational cost. A popular alternative, and usually computationally cheaper, is to use dimension-splitting methods. These schemes replace two-dimensional problem with a sequence of one-dimensional problems. For instance, we can solve the two-dimensional linear advection equation by solving a sequence one-dimensional linear advection equations using the PPM from Chapter 2. Further, we could, in principle, use any numerical method that solves the one-dimensional linear advection. A comparison between two-dimension and dimension splitting semi-lagrangian schemes on the plane has been investigated by Y. Chen et al. (2017) using the PPM as the one-dimensional solver and distorted two-dimensional grids. Their major conclusion is that the dimension splitting schemes are more sensitive to grid distortions but it is computationally cheaper and more accurate than their two-dimensional methods, specially considering large CFL numbers.

The main aim of this chapter is to give a detailed description of the dimension splitting method proposed by Lin and Rood (1996), which is currently employed in the FV3 dynamical core, applied to the two-dimensional linear advection equation using the one-dimensional finite-volume schemes from Chapter 2. Similarly to Chapter 2, we start this Chapter start with a review of the two-dimensional advection equation in the integral form in Section 3.1, and in Section 3.2 we set the framework of general two-dimensional finite-volumes schemes. Section 3.3 presents the dimension splitting method and numerical

experiments are shown in Section 3.4.

3.1 Two-dimensional advection equation in integral form

Let us consider a C^1 velocity field given by $\mathbf{u} : \mathbb{R}^2 \rightarrow \mathbb{R}^2$, $\mathbf{u} = (u, v)$, where u is the velocity in x -direction and v is the velocity in x and y direction. The two-dimensional advection equation in the differential form in a domain $\Omega = [a, b] \times [c, d] \subset \mathbb{R}^2$ associated to the velocity field \mathbf{u} is given by:

$$\frac{\partial q}{\partial t}(x, y, t) + \nabla \cdot (q\mathbf{u})(x, y, t) = 0, \quad \forall (x, y, t) \in \Omega^\circ \times]0, +\infty[,\text{ }^1 \quad (3.1)$$

where $\nabla \cdot (q\mathbf{u})$ is the spatial divergence

$$\nabla \cdot (q\mathbf{u})(x, y, t) = \frac{\partial(uq)}{\partial x}(x, y, t) + \frac{\partial(vq)}{\partial y}(x, y, t). \quad (3.2)$$

We recall that we say the u is **non-divergent** if $\nabla \cdot \mathbf{u} = 0$. A classical or strong solution to the two-dimensional advection equation is a C^1 function q satisfying Equation (3.1). As we did in Section 2.1, our goal is to deduce an integral form of Equation (3.1). Thus, let us consider $[x_1, x_2] \times [y_1, y_2] \subset \Omega^\circ$ and $[t_1, t_2] \subset [0, +\infty[$. Integrating Equation (3.1) over $[x_1, x_2] \times [y_1, y_2]$ yields:

$$\begin{aligned} \frac{d}{dt} \left(\int_{x_1}^{x_2} \int_{y_1}^{y_2} q(x, y, t) dx dy \right) &= - \int_{y_1}^{y_2} \left((uq)(x_2, y, t) - (uq)(x_1, y, t) \right) dy \\ &\quad - \int_{x_1}^{x_2} \left((vq)(x, y_2, t) - (vq)(x, y_1, t) \right) dx. \end{aligned} \quad (3.3)$$

Integrating Equation (3.3) over the time interval $[t_1, t_2]$, we have:

$$\begin{aligned} \int_{x_1}^{x_2} \int_{y_1}^{y_2} q(x, y, t_{n+1}) dx dy &= \int_{x_1}^{x_2} \int_{y_1}^{y_2} q(x, y, t_n) dx dy \\ &\quad - \int_{t_1}^{t_2} \int_{y_1}^{y_2} \left((uq)(x_2, y, t) - (uq)(x_1, y, t) \right) dy dt \\ &\quad - \int_{t_1}^{t_2} \int_{x_1}^{x_2} \left((vq)(x, y_2, t) - (vq)(x, y_1, t) \right) dx dt. \end{aligned} \quad (3.4)$$

Equation (3.4) is the integral form of Equation (3.1). We say that q is a weak solution to the advection equation (3.1) if q satisfies the integral form (3.4), $\forall [x_1, x_2] \times [y_1, y_2] \subset \Omega^\circ$ and $\forall [t_1, t_2] \subset [0, +\infty[$. Similarly to Section 2.1, these problems are equivalent when q is a C^1 function. We consider an initial condition q_0 , $q(x, y, 0) = q_0(x, y)$, $\forall (x, y) \in \Omega$. Boundary conditions will be assumed bi-periodic. Therefore, we are again dealing with a Cauchy problem.

¹ Ω° denotes the interior of Ω . Namely, $\Omega^\circ =]a, b[\times]c, d[$.

To move in the direction of a discrete version of Equation (3.4), let us discretize the domain $D = \Omega \times [0, T]$ following the notations of Section 2.1. Given a positive integer N_T , we define the time step $\Delta t = \frac{T}{N_T}$, $t_n = n\Delta t$, for $n = 0, 1, \dots, N_T$. The spatial discretization is constructed through an uniformly spaced partition of Ω given by:

$$[a, b] = \bigcup_{i=1}^N X_i, \text{ where } X_i = [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}] \text{ and } a = x_{\frac{1}{2}} < x_{\frac{3}{2}} < \dots < x_{N-\frac{1}{2}} < x_{N+\frac{1}{2}} = b, \quad (3.5)$$

$$[c, d] = \bigcup_{j=1}^M Y_j, \text{ where } Y_j = [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}] \text{ and } c = y_{\frac{1}{2}} < y_{\frac{3}{2}} < \dots < y_{M-\frac{1}{2}} < y_{M+\frac{1}{2}} = d, \quad (3.6)$$

$$\Omega = \bigcup_{i=1}^N \bigcup_{j=1}^M \Omega_{ij}, \text{ where } \Omega_{ij} = X_i \times Y_j. \quad (3.7)$$

The regions Ω_{ij} are known as control volumes or cells. Similarly to Chapter 2 we employ the notations $\Delta x = x_{i+\frac{1}{2}} - x_{i-\frac{1}{2}}$, $\Delta y = y_{j+\frac{1}{2}} - y_{j-\frac{1}{2}}$ and $x_i = \frac{1}{2}(x_{i+\frac{1}{2}} + x_{i-\frac{1}{2}})$, $y_j = \frac{1}{2}(y_{j+\frac{1}{2}} + y_{j-\frac{1}{2}})$, $\forall i = 1, \dots, N$, $\forall j = 1, \dots, M$, to define the control volume lengths and centroids, respectively. Finally, we denote by $Q_{ij}(t)$ as the average values of state variable at time t in the control volume Ω_{ij} , that is:

$$Q_{ij}(t) = \frac{1}{\Delta x \Delta y} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} q(x, y, t) dx. \quad (3.8)$$

Substituting t_1, t_2, x_1, x_2, y_1 and y_2 by $t_n, t_{n+1}, x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}, y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}$, respectively, in Equation (3.4), we obtain:

$$\begin{aligned} Q_{ij}(t_{n+1}) &= Q_{ij}(t_n) - \frac{\Delta t}{\Delta x \Delta y} \delta_x \left(\frac{1}{\Delta t} \int_{t_1}^{t_2} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} (uq)(x_i, y, t) dy dt \right) \\ &\quad - \frac{\Delta t}{\Delta x \Delta y} \delta_y \left(\frac{1}{\Delta t} \int_{t_1}^{t_2} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} (vq)(x, y_j, t) dx dt \right), \end{aligned} \quad (3.9)$$

where we are using the centered finite-difference notation:

$$\delta_x h(x_i, y, t) = h(x_{i+\frac{1}{2}}, y, t) - h(x_{i-\frac{1}{2}}, y, t), \quad (3.10)$$

$$\delta_y h(x, y_j, t) = h(x, y_{j+\frac{1}{2}}, t) - h(x, y_{j-\frac{1}{2}}, t), \quad (3.11)$$

for any function h . The Equation (3.9) is useful to motivate two-dimensional finite-volume schemes, as we shall see in the next section.

3.2 The finite-volume approach

This Section is basically an extension to two dimensions of the concepts presented in Section ???. We introduce the following spaces of bi-periodic functions:

$$\mathcal{F}(\mathbb{T}^2) = \{q : \mathbb{R}^2 \rightarrow \mathbb{R}; \quad q(x + b - a, y) = q(x, y), \quad q(x, y + d - c) = q(x, y), \quad \forall (x, y) \in \mathbb{R}^2\},$$

$$\mathcal{F}(\mathbb{T}_T^2) = \{q : \mathbb{R}^2 \times [0, T] \rightarrow \mathbb{R}; \quad q(\cdot, \cdot, t) \in \mathcal{F}(\mathbb{T}^2), \quad \forall t \in [0, T]\},$$

$$\mathcal{C}^k(\mathbb{T}_T^2) = \{q \in \mathcal{C}^k(\mathbb{R}^2 \times [0, T]) : q \in \mathcal{F}(\mathbb{T}_T^2)\}.$$

where we are using the notation $\mathbb{T}_T^2 = \mathbb{T}^2 \times [0, T]$ and $\mathbb{T}^2 = [a, b] \times [c, d]$ denotes the torus of lengths $b - a$ and $d - c$. We are using the notation \mathbb{T}^2 since we may think of bi-periodic functions as functions defined on the torus of lengths $b - a$ and $d - c$. Whenever we use the notation \mathbb{T}^2 , a, b, c and d will be implicitly defined. We also introduce the following the locally integrable periodic functions:

$$\begin{aligned} L_{\text{loc}}^p(\Omega) &= \{q : \Omega \rightarrow \mathbb{R}; \quad \int_K |q(x, y)|^p dx dy < +\infty, \quad \text{for all compact sets } K \subset \Omega\}, \\ L_{\text{loc}}^p(\mathbb{T}^2) &= \{q \in \mathcal{F}(\mathbb{T}^2) : \quad q \in L_{\text{loc}}^p(\mathbb{R}^2)\}, \\ L_{\text{loc}}^{p,x,y}(\mathbb{T}_T^2) &= \{q \in \mathcal{F}(\mathbb{T}_T^2) : \forall t \in [0, T], \quad q(\cdot, \cdot, t) \in L_{\text{loc}}^p(\mathbb{R}^2)\}, \\ L_{\text{loc}}^{p,y,t}(\mathbb{T}_T^2) &= \{q \in \mathcal{F}(\mathbb{T}_T^2) : \forall x \in \mathbb{R}, \quad q(x, \cdot, \cdot) \in L_{\text{loc}}^p(\mathbb{R} \times [0, T])\}, \\ L_{\text{loc}}^{p,x,t}(\mathbb{T}_T^2) &= \{q \in \mathcal{F}(\mathbb{T}_T^2) : \forall y \in \mathbb{R}, \quad q(\cdot, y, \cdot) \in L_{\text{loc}}^p(\mathbb{R} \times [0, T])\}. \end{aligned}$$

3.2.1 Discretization of the problem

The problem of two-dimensional advection equation in the integral form presented Section 3.1 is written in a concise way in Problem 3.1.

Problem 3.1. *Given an initial condition $q_0 \in L_{\text{loc}}^1(\mathbb{T}^2) \cap L_{\text{loc}}^2(\mathbb{T}^2)$ and a velocity function $u \in L_{\text{loc}}^{2,y,t}(\mathbb{T}_T^2)$ in the x -direction, a velocity function $v \in L_{\text{loc}}^{2,x,t}(\mathbb{T}_T^2)$ in the y -direction, we would like to find a weak solution $q \in L_{\text{loc}}^{1,x,y}(\mathbb{T}_T^1) \cap L_{\text{loc}}^{2,x,t}(\mathbb{T}_T^2) \cap L_{\text{loc}}^{2,y,t}(\mathbb{T}_T^2)$ of the two-dimensional advection equation in the integral form:*

$$\begin{aligned} \int_{x_1}^{x_2} \int_{y_1}^{y_2} q(x, y, t) dx dy &= \int_{x_1}^{x_2} \int_{y_1}^{y_2} q(x, y, t) dx dy \\ &\quad - \int_{t_1}^{t_2} \int_{y_1}^{y_2} \left((uq)(x_2, y, t) - (uq)(x_1, y, t) \right) dy dt \\ &\quad - \int_{t_1}^{t_2} \int_{x_1}^{x_2} \left((vq)(x, y_2, t) - (vq)(x, y_1, t) \right) dx dt. \end{aligned}$$

$\forall [x_1, x_2] \times [y_1, y_2] \times [t_1, t_2] \subset [a, b] \times [c, d] \times [0, T]$, and $q(x, y, 0) = q_0(x, y)$, $\forall (x, y) \in [a, b] \times [c, d]$.

For Problem 3.1, the total mass in Ω is defined by:

$$M_\Omega(t) = \int_{\Omega} q(x, y, t) dx dy, \quad \forall t \in [0, T], \tag{3.12}$$

and is conserved within time:

$$M_\Omega(t) = M_\Omega(0), \quad \forall t \in [0, T]. \tag{3.13}$$

considering a discretization of the domain $[a, b] \times [0, T]$. Similar to Definitions 3.1 and 2.3, we introduce the concepts of $(\Delta x, \Delta y)$ -grid and $(\Delta x, \Delta y, \Delta t, \lambda_x, \lambda_y)$ discretization.

Definition 3.1 ($(\Delta x, \Delta y)$ -grid). *Given $[a, b] \times [c, d]$ and positive real numbers Δx and Δy such that $\Delta x = (b - a)/N$, $\Delta y = (d - c)/M$, for positive integers N and M , we say that a*

(N, M) -tuple $\mathcal{D} = (\Omega_{ij})_{i=1,\dots,N, j=1,\dots,M}$ is a $(\Delta x, \Delta y)$ -grid for $[a, b] \times [c, d]$ if $\Omega_{ij} = [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}] \times [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}]$, $a = x_{\frac{1}{2}} < x_{\frac{3}{2}} < \dots < x_{N-\frac{1}{2}} < x_{N+\frac{1}{2}} = b$, $c = y_{\frac{1}{2}} < y_{\frac{3}{2}} < \dots < y_{M-\frac{1}{2}} < y_{M+\frac{1}{2}} = d$, $\Delta x = x_{i+\frac{1}{2}} - x_{i-\frac{1}{2}}$, $\Delta y = y_{j+\frac{1}{2}} - y_{j-\frac{1}{2}}$. Each Ω_{ij} is called control volume or cell.

Remark 3.1. We may define the cells Ω_{ij} for i outside of the range $1, \dots, N$ or j outside of the range $1, \dots, M$ by $\Omega_{ij} = [a + (i-1)\Delta x, a + i\Delta x] \times [c + (j-1)\Delta y, c + j\Delta y]$. These cells are called ghost cells.

Definition 3.2 ($(\Delta x, \Delta y, \Delta t, \lambda_x, \lambda_y)$ -discretization). Given $[a, b] \times [c, d] \times [0, T]$ and positive real numbers $\Delta x, \Delta y$ and Δt , we say that $(\mathcal{D}, \mathcal{T})$ is a $(\Delta x, \Delta y, \Delta t, \lambda_x, \lambda_y)$ -discretization of $[a, b] \times [c, d] \times [0, T]$ if Ω is a $(\Delta x, \Delta y)$ -uniform grid for $[a, b] \times [c, d]$ and \mathcal{T} is a Δt -temporal grid for $[0, T]$, $\frac{\Delta t}{\Delta x} = \lambda_x$ and $\frac{\Delta t}{\Delta y} = \lambda_y$.

Remark 3.2. Whenever we mention a $(\Delta x, \Delta y)$ -grid, or a $(\Delta x, \Delta y, \Delta t, \lambda_x, \lambda_y)$ -discretization, then Ω_{ij} , N and M are implicitly defined.

Section 3.1 introduced a version of Problem 3.1 considering a discretization of the domain $[a, b] \times [c, d] \times [0, T]$. This version is also summarized in Problem 3.2.

Problem 3.2. Assume the framework of Problem 3.1 and that $(\mathcal{D}, \mathcal{T})$ is a $(\Delta x, \Delta y, \Delta t, \lambda_x, \lambda_y)$ -discretization of $[a, b] \times [c, d] \times [0, T]$. Since we are in the framework of Problem 3.1, it follows that:

$$\begin{aligned} Q_{ij}(t_{n+1}) &= Q_{ij}(t_n) - \lambda_x \delta_x \left(\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} (uq)(x_i, y, t) dy dt \right) \\ &\quad - \lambda_y \delta_y \left(\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} (vq)(x, y_j, t) dx dt \right), \end{aligned}$$

where $Q_{ij}(t) = \frac{1}{\Delta x \Delta y} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} q(x, y, t) dx dy$.

Our problem now consists of finding the values $Q_{ij}(t_n)$, $\forall i = 1, \dots, N$, $\forall j = 1, \dots, M$, $\forall n = 1, \dots, N_T$, given the initial values $Q_{ij}(0)$, $\forall i = 1, \dots, N$, $\forall j = 1, \dots, M$. In other words, we would like to find the average values of q in each control volume Ω_{ij} at the considered time instants.

Next, we introduce the definitions of grid functions at cell centroids and C-grid functions.

Definition 3.3 ($(\Delta x, \Delta y)$ -grid function). For a $(\Delta x, \Delta y)$ -grid \mathcal{D} , we say that $Q \in \mathbb{R}^{N \times M}$ is a $(\Delta x, \Delta y)$ -grid function, where we assume that Q is ordered by the lines in the x direction of the grid, i.e.,

$$Q = (Q_{11}, Q_{21}, \dots, Q_{M1}, Q_{12}, Q_{22}, \dots, Q_{M2}, \dots, Q_{1N}, Q_{2N}, \dots, Q_{NM}).$$

We denote the space of $(\Delta x, \Delta y)$ -grid functions by $\mathbb{R}^{\Delta x \times \Delta y}$.

Definition 3.4 ($(\Delta x, \Delta y)$ -C grid wind). For a $(\Delta x, \Delta y)$ -grid \mathcal{D} , we say that (u, v) is a $(\Delta x, \Delta y)$ -C grid wind if $u \in \mathbb{R}^{(N+1) \times M}$, $v \in \mathbb{R}^{N \times (M+1)}$, where we assume that u is ordered by the

lines in the x direction of the grid, i.e.,

$$u = (u_{\frac{1}{2},1}, u_{\frac{3}{2},1}, \dots, u_{N+\frac{1}{2},1}, u_{\frac{1}{2},2}, u_{\frac{3}{2},2}, \dots, u_{N+\frac{1}{2},2}, \dots, u_{\frac{1}{2},M}, u_{\frac{3}{2},M}, \dots, u_{N+\frac{1}{2},M}),$$

v is ordered by the lines in the y direction of the grid, i.e.,

$$v = (v_{1,\frac{1}{2}}, v_{1,\frac{3}{2}}, \dots, v_{1,M+\frac{1}{2}}, v_{2,\frac{1}{2}}, v_{2,\frac{3}{2}}, \dots, v_{2,M+\frac{1}{2}}, \dots, v_{N,\frac{1}{2}}, v_{N,\frac{3}{2}}, \dots, v_{N,M+\frac{1}{2}}).$$

Remark 3.3. When computing stencils, we may need values of Q_{ij} such that the indexes i and j are out of the range $1, \dots, N$, $1, \dots, M$. These values are called ghost cell values. Since we are under the assumption of periodic boundary conditions, this problem is overcome by assuming periodicity on the grid function Q . The same applies for $(\Delta x, \Delta y)$ -C grid functions.

Finally, we define the two-dimensional (2D) finite-volume (FV) scheme problem as follows in Problem 3.3.

Remark 3.4. For Problem 2.2, we define the $(\Delta x, \Delta y)$ -grid functions q^n and $Q(t^n)$, where $q_{ij}^n = q(x_i, y_j, t^n)$, $Q(t^n)_{ij} = Q_{ij}(t^n)$, for $n = 0, \dots, N_T$. We also define the $(\Delta x, \Delta y)$ -C grid wind (u^n, v^n) where $u_{i+\frac{1}{2},j}^n = u(x_{i+\frac{1}{2}}, y_j, t^n)$ and $v_{i,j+\frac{1}{2}}^n = v(x_i, y_{j+\frac{1}{2}}, t^n)$.

Problem 3.3 (2D-FV scheme). Assume the framework defined in Problem 3.2. The finite-volume approach of Problem 3.1 consists of a finding a scheme of the form:

$$\begin{aligned} Q_{ij}^{n+1} &= Q_{ij}^n - \lambda_x \delta_i F_{i,j}^n - \lambda_y \delta_j G_{i,j}^n, \\ \forall i &= 1, \dots, N, \quad \forall j = 1, \dots, M, \quad \forall n = 0, \dots, N_T - 1, \end{aligned} \tag{3.14}$$

where $\delta_i F_{ij}^n = F_{i+\frac{1}{2},j}^n - F_{i-\frac{1}{2},j}^n$, $\delta_j G_{ij}^n = G_{i,j+\frac{1}{2}}^n - G_{i,j-\frac{1}{2}}^n$ and Q^n is intended to be an approximation of $Q(t_n)$ in some sense. We define $Q_{ij}^0 = Q_{ij}(0)$ or $Q_{ij}^0 = q_{ij}^0$.

The term $F_{i+\frac{1}{2},j}^n = \mathbb{F}(Q^n, \tilde{u}^n, \tilde{v}^n; i, j)$ is known as numerical flux in the x direction, where \mathbb{F} is the numerical flux function, and it approximates $\frac{1}{\Delta t} \int_{t_n}^{t_{n+1}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} (uq)(x_{i+\frac{1}{2}}, y, t) dy dt$, $\forall i = 0, 1, \dots, N$, and $G_{i,j+\frac{1}{2}}^n = \mathbb{G}(Q^n, \tilde{u}^n, \tilde{v}^n; i, j)$ is known as numerical flux in the y direction, where \mathbb{G} is the numerical flux function, and it approximates $\frac{1}{\Delta t} \int_{t_n}^{t_{n+1}} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} (vq)(x, y_{j+\frac{1}{2}}, t) dx dt$, $\forall j = 0, 1, \dots, M$, or, in other words, they estimate the time-averaged fluxes at the control volume Ω_{ij} boundaries. The values $\tilde{u}_{i+\frac{1}{2},j}^n$ and $\tilde{v}_{i,j+\frac{1}{2}}^n$ are related to the time-averaged velocities and depend on values of $u_{i+\frac{1}{2},j}^n$ and $v_{i,j+\frac{1}{2}}^n$.

Remark 3.5. A scheme of the form from Equation (3.14) is referred to as a 2D-FV scheme and it is also known as a conservative scheme.

Remark 3.6. For Problem 3.3, we define the CFL number in the x and y direction by $\max\{|u_{i+\frac{1}{2},j}^n|\}$ and $\max\{|v_{i,j+\frac{1}{2}}^n|\}$, respectively. The CFL number is maximum between these numbers and we say that the CFL conditions is satisfied if the CFL number is less than one.

Definition 3.5 (Discrete divergence). For Problem 3.3, we define the discrete divergence as

a grid function $\mathbb{D}^n = \mathbb{D}^n(Q, \tilde{u}^n, \tilde{v}^n)$ given by

$$\mathbb{D}_{ij}^n = \left(\frac{\mathbb{F}(Q, \tilde{u}^n, \tilde{v}^n; i, j) - \mathbb{F}(Q, \tilde{u}^n, \tilde{v}^n; i-1, j)}{\Delta x} \right) + \left(\frac{\mathbb{G}(Q, \tilde{u}^n, \tilde{v}^n; i, j) - \mathbb{G}(Q, \tilde{u}^n, \tilde{v}^n; i, j-1)}{\Delta y} \right). \quad (3.15)$$

For a 2D-FV the discrete total mass at the time-step n is given by

$$M^n = \Delta x \Delta y \sum_{i=1}^N \sum_{j=1}^M Q_{ij}^n.$$

Therefore, the discrete total mass is constant for a 2D-FV scheme, which follows from a straightforward computation:

$$\begin{aligned} M^{n+1} &= \Delta x \sum_{i=1}^N \sum_{j=1}^M Q_{ij}^{n+1} = M^n - \Delta t \sum_{i=1}^N \sum_{j=1}^M (F_{i+\frac{1}{2}, j}^n - F_{i-\frac{1}{2}, j}^n) - \Delta t \sum_{i=1}^N \sum_{j=1}^M (G_{i, j+\frac{1}{2}}^n - G_{i, j-\frac{1}{2}}^n) \\ &= M^n - \Delta t \sum_{j=1}^M (F_{N+\frac{1}{2}, j}^n - F_{\frac{1}{2}, j}^n) - \Delta t \sum_{i=1}^N (G_{i, M+\frac{1}{2}}^n - G_{i, \frac{1}{2}}^n) = M^n, \end{aligned}$$

where we are using that $F_{N+\frac{1}{2}, j}^n = F_{\frac{1}{2}, j}^n$, $G_{i, M+\frac{1}{2}}^n = G_{i, \frac{1}{2}}^n$ since we are assuming bi-periodic boundary conditions.

3.2.2 Convergence, consistency and stability

As we mentioned in Problem 3.3, the initial condition may be assumed as q_{ij}^0 or $Q_{ij}(0)$. For two-dimensional simulations, we are going to assume q_{ij}^0 as initial data to avoid the computation of integrals. Furthermore, the errors will be calculated using the values q_{ij}^n instead of $Q_{ij}(t_n)$. Similarly to Proposition 2.2, we have that the centroid value approximates the average value with second order, as Proposition 3.1 shows.

Proposition 3.1. *If $q \in C^2$, then $|Q_{ij}(t^n) - q_{ij}^n| \leq C_1 \Delta x^2 + C_2 \Delta x \Delta y + C_3 \Delta y^2$, where C_1 , C_2 and C_3 are constants.*

Proof. Just apply Theorem A.4 for the function $q(x, y, t^n)$. □

The notions of convergence, consistency and stability for a 2D-FV schemes are straightforward from these notions for 1D-FV schemes (see Subsections A.5.1 and A.5.2). Indeed, in the context of Problem 3.3, we define the operators $\mathcal{H}_{\Delta x, \Delta y, n} : \mathbb{R}^{\Delta x \times \Delta y} \rightarrow \mathbb{R}^{\Delta x \times \Delta y}$ whose (i, j) entry is given by:

$$\begin{aligned} [\mathcal{H}_{\Delta x, \Delta y, n}(Q)]_{ij} &= Q_{ij} - \lambda_x \left(\mathbb{F}(Q, \tilde{u}^n, \tilde{v}^n; i, j) - \mathbb{F}(Q, \tilde{u}^n, \tilde{v}^n; i-1, j) \right) \\ &\quad - \lambda_y \left(\mathbb{G}(Q, \tilde{u}^n, \tilde{v}^n; i, j) - \mathbb{G}(Q, \tilde{u}^n, \tilde{v}^n; i, j-1) \right) \end{aligned}$$

for $i = 1, \dots, N$, $j = 1, \dots, M$, $n = 0, \dots, N_T - 1$. The 2D-FV is then expressed as

$$Q^{n+1} = \mathcal{H}_{\Delta x, \Delta y, n}(Q^n).$$

The local error truncation $\tau^n \in \mathbb{R}^{\Delta x \times \Delta y}$ is given by

$$Q(t^{n+1}) = \mathcal{H}_{\Delta x, \Delta y, n}(Q(t^n)) + \Delta t \tau^n.$$

The error equation is given by

$$E^{n+1} = \mathcal{H}_{\Delta x, \Delta y, n}(Q(t^n)) - \mathcal{H}_{\Delta x, \Delta y, n}(Q^n) + \Delta t \tau^n. \quad (3.16)$$

Given $r = (r_{ij})_{i=1, \dots, N, j=1, \dots, M} \in \mathbb{R}^{\Delta x \times \Delta y}$, we define the p -norm by

$$\|r\|_{p, \Delta x \times \Delta y} = \begin{cases} \left(\sum_{i=1}^N \sum_{j=1}^M |r_{ij}|^p \right)^{\frac{1}{p}} & \text{if } 1 \leq p < \infty, \\ \max_{i=1, \dots, N, j=1, \dots, M} |r_{ij}| & \text{otherwise.} \end{cases} \quad (3.17)$$

The stability in the p -norm is defined as in the 1D case.

Definition 3.6. A 2D-FV scheme is stable in the p -norm if

$$\|\mathcal{H}_{\Delta x, \Delta y, n}(Q) - \mathcal{H}_{\Delta x, \Delta y, n}(P)\|_{p, \Delta x \times \Delta y} \leq (1 + \alpha \Delta t) \|Q - P\|_{p, \Delta x \times \Delta y}, \quad (3.18)$$

for all $Q, P \in \mathbb{R}^{\Delta x \times \Delta y}$ and α is a constant that does not depend neither on Δx , Δy , Δt nor on n .

If a 2D-FV scheme is stable in the p -norm, similarly to Equation (A.55) we have:

$$\|E^{n+1}\|_{p, \Delta x \times \Delta y} \leq e^{\alpha T} (\|E^0\|_{p, \Delta x \times \Delta y} + T \max_{n=1, \dots, N_T} \|\tau^n\|_{p, \Delta x \times \Delta y}).$$

Again, we point out that from Proposition 3.1, we have that the initial error E^0 shall be second-order accurate. Consistency is defined as in Definition A.1 and convergence is defined as in Definition A.2.

The Von Neumann analysis can be applied when $\mathcal{H}_{\Delta x, \Delta y, n}$ is linear, since we are considering periodic boundary conditions. The idea is the same as in the one-dimensional case, we just apply the operator $\mathcal{H}_{\Delta x, \Delta y, n}$ on the Fourier modes to obtain the amplification factor. We introduce the nodes $\theta_i = i \frac{2\pi}{N}$, $i = 1, \dots, N$, $\Delta\theta = \frac{2\pi}{N}$, $\theta_i = (\theta_1, \theta_2, \dots, \theta_N)$, $\phi_j = j \frac{2\pi}{M}$, $j = 1, \dots, M$, $\Delta\phi = \frac{2\pi}{M}$, $\phi = (\phi_1, \phi_2, \dots, \phi_M)$. For $k_1 = 1, \dots, N$, $k_2 = 1, \dots, M$, the two-dimensional Fourier mode $\mathbf{k} = (k_1, k_2)$ from $\mathbb{C}^{N \times M}$ has its (i, j) entry given by $[e^{ik\theta}]_{ij} = e^{ik_1\theta_i} e^{ik_2\phi_j}$.

Notice that if $q, u, v \in C^3$, we can rewrite Equation (A.49) as:

$$\begin{aligned} \tau_{ij}^n = & \left[\frac{1}{\Delta x \Delta y \Delta t} \int_{t^n}^{t^{n+1}} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \nabla \cdot (\mathbf{u} q)(x, y, t) dy dx dt - \left(\frac{\mathbb{F}(Q, \tilde{u}^n, \tilde{v}^n; i, j) - \mathbb{F}(Q, \tilde{u}^n, \tilde{v}^n; i-1, j)}{\Delta x} \right. \right. \\ & \left. \left. - \left(\frac{\mathbb{G}(Q, \tilde{u}^n, \tilde{v}^n; i, j) - \mathbb{G}(Q, \tilde{u}^n, \tilde{v}^n; i, j-1)}{\Delta y} \right) \right) \right]. \end{aligned}$$

Using the midpoint rule for integration (Theorem A.4), the mean value theorem for integrals (Theorem A.2) and recalling the discrete divergence (Definition 3.5), we have:

$$\tau_{ij}^n = \frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} \nabla \cdot (\mathbf{u}q)(x_i, y_j, t) dt - \mathbb{D}_{ij}^n + O(\Delta x^2) + O(\Delta y^2). \quad (3.19)$$

Therefore, in order to investigate the consistency, we may compare how well the discrete divergence approximates the divergence.

3.3 Dimension splitting

Before introducing the dimension splitting scheme from Lin and Rood (1996), it is useful to investigate a little bit of general operator splitting schemes, since the dimension splitting technique is a particular case of operator splitting methods. For a time interval $[0, T]$, we consider a Δt -temporal grid. We consider the abstract Cauchy problems

$$\begin{cases} \frac{dq}{dt}(t) = Aq(t), & t \in [t^n, t^{n+1}], \\ q(t^n) = q_n, \end{cases}$$

for $n = 0, \dots, N_T - 1$, where $q(t) \in \mathcal{B}$ for some Banach space \mathcal{B} , and $A : \mathcal{B} \rightarrow \mathcal{B}$ is a linear operator following the framework of Richtmyer and Morton (1968, Chapter 3). We are interested in finding $q(t^{n+1})$ given q_n . Assuming that $A = A_1 + A_2$ for two linear operators $A_1, A_2 : \mathcal{B} \rightarrow \mathcal{B}$, we consider the following abstract Cauchy sub-problems:

$$\begin{cases} \frac{dq^1}{dt}(t) = A_1 q(t), & t \in [t^n, t^{n+1}], \\ q^1(t^n) = q_n, \end{cases}$$

and

$$\begin{cases} \frac{dq^{21}}{dt}(t) = A_2 q(t), & t \in [t^n, t^{n+1}], \\ q^{21}(t^n) = q^1(t^{n+1}). \end{cases}$$

Then we can approximate $q(t_0 + \Delta t)$ by $q^{21}(t^n + \Delta t)$ with error $O(\Delta t)$, if A_1 and A_2 do not commute; otherwise, this method is exact. This approach is known as Lie-Trotter splitting. Observe the Lie-Trotter splitting may be performed in a reverse order solving the sub-problems

$$\begin{cases} \frac{dq^2}{dt}(t) = A_2 q(t), & t \in [t^n, t^{n+1}], \\ q^2(t^n) = q_n, \end{cases}$$

and

$$\begin{cases} \frac{dq^{21}}{dt}(t) = A_1 q(t), & t \in [t^n, t^{n+1}], \\ q^{12}(t^n) = q^1(t^{n+1}), \end{cases}$$

and again we estimate $q(t^{n+1})$ by $q^{12}(t^{n+1})$ with error $O(\Delta t)$. As observed by Strang (1968), we can consider

$$q^*(t^{n+1}) = \frac{q^{21}(t^{n+1}) + q^{12}(t^{n+1})}{2}, \quad (3.20)$$

which is a second-order ($O(\Delta t^2)$) symmetric scheme to approximate $q(t^{n+1})$. We shall refer to this scheme as average Lie-Trotter splitting. This process of averaging two Lie-Trotter splitting is a particular case of methods known in the literature as weighted sequential splitting methods. Furthermore, this process of averaging schemes may be extended to achieve higher-order schemes (c.f., e.g. Jia and Li (2011)). For an accuracy analysis of weighted sequential splitting methods we refer to Csomós et al. (2005).

We point out that one of the most widely used in the literature second-order splitting schemes is the Strang splitting (Strang, 1968). This scheme requires the solution of 3 sub-problems per time-step, with one of them at time $t_n + \frac{\Delta t}{2}$, while the average Lie-Trotter splitting requires the solution of 4 sub-problems per time-step. Hence, the Strang splitting is computationally cheaper. However, as we shall see in this Chapter, the average Lie-Trotter splitting applied for the linear advection equation allows a modification that eliminates a splitting error that appears when we consider a constant scalar field and non-divergent velocity as observed by Lin and Rood (1996).

To move towards to the scheme from Lin and Rood (1996), let us consider Problem 3.1 in its differential form:

$$\begin{cases} \frac{\partial q}{\partial t}(x, y, t) + \frac{\partial(uq)}{\partial x}(x, y, t) + \frac{\partial(vq)}{\partial y}(x, y, t) = 0, & \forall(x, y, t) \in \mathbb{T}_T^2 \\ q(x, y, 0) = q_0(x, y), \end{cases}$$

where $q, u, v \in C^1(\mathbb{T}_T^2)$. We are going to consider the one-dimensional advection equation in the x -direction

$$\frac{\partial q^x}{\partial t}(x, y, t) + \frac{\partial(uq^x)}{\partial x}(x, y, t) = 0,$$

for each $y = y_j$, and the one-dimensional advection equation in the y -direction

$$\frac{\partial q^y}{\partial t}(x, y, t) + \frac{\partial(vq^y)}{\partial y}(x, y, t) = 0,$$

for each $x = x_i$. We shall assume that these problems are solved using a 1D-FV scheme as in Problem 2.3 with numerical flux functions \mathcal{F} and \mathcal{G} , respectively. We introduce the auxiliary functions

$$\mathbf{F}(Q_{ij}^n) = -\lambda_x (\mathcal{F}(Q_{ij}^n(S_{i+\frac{1}{2}}); \tilde{u}_{i+\frac{1}{2},j}^n) - \mathcal{F}(Q_{ij}^n(S_{i-\frac{1}{2}}); \tilde{u}_{i-\frac{1}{2},j}^n)),$$

and

$$\mathbf{G}(Q_{ij}^n) = -\lambda_y (\mathcal{G}(Q_{ij}^n(S_{j+\frac{1}{2}}); \tilde{v}_{i,j+\frac{1}{2}}^n) - \mathcal{G}(Q_{ij}^n(S_{j-\frac{1}{2}}); \tilde{v}_{i,j-\frac{1}{2}}^n)),$$

which are the numerical flux update of the 1D-FV schemes in the x and y direction,

respectively, that is

$$\begin{aligned}\mathcal{F}(Q_{ij}^n(\mathcal{S}_{i+\frac{1}{2}}); \tilde{u}_{i+\frac{1}{2},j}^n) &= \frac{1}{\Delta t} \int_{t_n}^{t^{n+1}} (uq)(x_{i+\frac{1}{2}}, y_j, t) dt + O(\Delta t^P), \\ \mathcal{G}(Q_{ij}^n(\mathcal{S}_{j+\frac{1}{2}}); \tilde{v}_{i,j+\frac{1}{2}}^n) &= \frac{1}{\Delta t} \int_{t_n}^{t^{n+1}} (vq)(x_i, y_{j+\frac{1}{2}}, t) dt + O(\Delta t^P),\end{aligned}$$

where P is the order of accuracy of the flux and \tilde{u}^n and \tilde{v}^n are the time-averaged velocities. The Lie-Trotter splitting is obtained by solving the advection in the x direction

$$Q_{ij}^{x,n+1} = Q_{ij}^n + \mathbf{F}(Q_{ij}^n),$$

for $j = 1, \dots, M$, and then we advect in the y direction with initial data $Q_{ij}^{x,n+1}$

$$Q_{ij}^{yx,n+1} = Q_{ij}^{x,n+1} + \mathbf{G}(Q_{ij}^{x,n+1}),$$

for $i = 1, \dots, N$. To get the average Lie-Trotter splitting we repeat the process in the reverse order by solving the advection equation in the y direction

$$Q_{ij}^{y,n+1} = Q_{ij}^n + \mathbf{G}(Q_{ij}^n),$$

for $i = 1, \dots, N$, and then we advect in the x -direction with initial data $Q_{ij}^{y,n+1}$

$$Q_{ij}^{xy,n+1} = Q_{ij}^{y,n+1} + \mathbf{F}(Q_{ij}^{y,n+1}),$$

for $j = 1, \dots, M$, and thus we have the average Lie-Trotter solution:

$$Q_{ij}^{n+1} = \frac{(Q^{xy,n+1} + Q^{yx,n+1})}{2} = Q_{ij}^n + \frac{1}{2}\mathbf{F}(Q_{ij}^n) + \frac{1}{2}\mathbf{G}(Q_{ij}^n) + \frac{1}{2}\mathbf{F}\left(Q_{ij}^n + \frac{1}{2}\mathbf{G}(Q_{ij}^n)\right) + \frac{1}{2}\mathbf{G}\left(Q_{ij}^n + \frac{1}{2}\mathbf{F}(Q_{ij}^n)\right),$$

assuming that the numerical flux functions are linear in the input Q , we may rewrite a computationally cheaper version of the average Lie-Trotter splitting as (Lin & Rood, 1996):

$$Q_{ij}^{n+1} = \frac{(Q^{xy,n+1} + Q^{yx,n+1})}{2} = Q_{ij}^n + \mathbf{F}\left(Q_{ij}^n + \frac{1}{2}\mathbf{G}(Q_{ij}^n)\right) + \mathbf{G}\left(Q_{ij}^n + \frac{1}{2}\mathbf{F}(Q_{ij}^n)\right). \quad (3.21)$$

The numerical flux functions defined in Chapter 2 are indeed linear the input Q if there are monotonic constrain, as we mention in Chapter 2, but we are going to consider this scheme even when there are monotonic constraints since it requires fewer operations. For a while let us assume that the numerical flux functions are exactly the time-averaged fluxes, which they must approximate, as pointed out in Problem 2.3. Under this assumption, we have:

$$\begin{aligned}\mathbf{F}(Q_{ij}^n) &= -\lambda_x \delta_x \left(\frac{1}{\Delta t} \int_{t_n}^{t^{n+1}} (uq)(x_i, y_j, t) dt \right), \\ \mathbf{G}(Q_{ij}^n) &= -\lambda_y \delta_y \left(\frac{1}{\Delta t} \int_{t_n}^{t^{n+1}} (vq)(x_i, y_j, t) dt \right).\end{aligned}$$

Further, if we assume that $q = \bar{q}$ is constant and $\nabla \cdot \mathbf{u} = 0$ then the solution remains constant and then, assuming also that \mathbf{u} does not depend on t , then \mathbf{F} and \mathbf{G} are given by

$$\begin{aligned}\mathbf{F}(Q_{ij}^n) &= -\bar{q}\lambda_x\delta_x u(x_i, y_j), \\ \mathbf{G}(Q_{ij}^n) &= -\bar{q}\lambda_y\delta_y v(x_i, y_j).\end{aligned}$$

However, if we compute the updated solution using Equation (3.21), we have that the error is given by

$$Q_{ij}^{n+1} - \bar{q} = -\Delta t \left(\frac{\delta_x u(x_i, y_j)}{\Delta x} + \frac{\delta_y v(x_i, y_j)}{\Delta y} \right) - \Delta t^2 \bar{q} \left(\frac{\delta_y v \delta_x u(x_i, y_j) + \delta_x u \delta_y v(x_i, y_j)}{2\Delta x \Delta y} \right) \quad (3.22)$$

$$= \Delta t(O(\Delta x^2) + O(\Delta y^2)) - \Delta t^2 \bar{q} \left(\frac{\delta_y v \delta_x u(x_i, y_j) + \delta_x u \delta_y v(x_i, y_j)}{2\Delta x \Delta y} \right). \quad (3.23)$$

Thus, the terms in the equation above multiplied by Δt^2 are related to a splitting error, even if we consider the exact fluxes. Aiming to eliminate the error from, Lin and Rood (1996) proposes to consider a modification of the average Lie-Trotter splitting as

$$Q_{ij}^{n+1} = Q_{ij}^n + \mathbf{F} \left(Q_{ij}^n + \frac{1}{2} \mathbf{g}(Q_{ij}^n) \right) + \mathbf{G} \left(Q_{ij}^n + \frac{1}{2} \mathbf{f}(Q_{ij}^n) \right), \quad (3.24)$$

where \mathbf{f} and \mathbf{g} are called inner advective operators and approximate $-\Delta t u \frac{\partial q}{\partial x}$ and $-\Delta t v \frac{\partial q}{\partial y}$.

In this work, we shall consider the following inner advective operator proposed by Lin (2004) (hereafter, **L04**) and the one proposed by Putman and Lin (2007) (hereafter, **PL07**). The PL07 scheme is currently used in the FV3 dynamical core. We also shall consider the average Lie-Trotter splitting (hereafter, **AVLT**). All the expressions of each inner advective operator mentioned are shown in Table 3.1. It is easy to see that both operators L04 and PL07 eliminate the term multiplied by Δt^2 that appeared in Equation (3.22) when we apply these operators for a constant grid function Q^n and a non-divergent velocity field in Equation (3.24). Therefore, these inner advective operators eliminate the splitting error for a constant field and a non-divergent velocity field.

Scheme	$\mathbf{f}(Q_{ij}^n)$	$\mathbf{g}(Q_{ij}^n)$
AVLT	$\mathbf{F}(Q_{ij}^n)$	$\mathbf{G}(Q_{ij}^n)$
L04	$\mathbf{F}(Q_{ij}^n) + Q_{ij}^n \frac{\Delta t}{\Delta x} (\tilde{u}_{i+\frac{1}{2},j}^n - \tilde{u}_{i-\frac{1}{2},j}^n)$	$\mathbf{G}(Q_{ij}^n) + Q_{ij}^n \frac{\Delta t}{\Delta y} (\tilde{v}_{i,j+\frac{1}{2}}^n - \tilde{v}_{i,j-\frac{1}{2}}^n)$
PL07	$\frac{1}{2} \left(-Q_{ij}^n + \frac{Q_{ij}^n + \mathbf{F}(Q_{ij}^n)}{1 - \frac{\Delta t}{\Delta x} (\tilde{u}_{i+\frac{1}{2},j}^n - \tilde{u}_{i-\frac{1}{2},j}^n)} \right)$	$\frac{1}{2} \left(-Q_{ij}^n + \frac{Q_{ij}^n + \mathbf{G}(Q_{ij}^n)}{1 - \frac{\Delta t}{\Delta y} (\tilde{v}_{i,j+\frac{1}{2}}^n - \tilde{v}_{i,j-\frac{1}{2}}^n)} \right)$

Table 3.1: Expression of the inner advective operators considered in this work. AVLT stands for the average Lie-Trotter scheme, while L04 and PL07 stands for the inner advective operators from Lin (2004) and from Putman and Lin (2007), respectively.

Recalling the definition of discrete divergence (Definition 3.5) we have:

$$\mathbb{D}_{ij}^n = -\frac{1}{\Delta t} \left[\mathbf{F} \left(Q_{ij}^n + \frac{1}{2} \mathbf{g}(Q_{ij}^n) \right) + \mathbf{G} \left(Q_{ij}^n + \frac{1}{2} \mathbf{f}(Q_{ij}^n) \right) \right], \quad (3.25)$$

and as pointed out in Section 3.2.2, we may use the discrete divergence to check the scheme consistency.

3.4 Numerical experiments

To assess the dimension splitting schemes introduce previously, we are going to consider the linear advection equation on $[0, 1] \times [0, 1]$ with bi-biperiodic boundary conditions. For the 1D schemes, we consider the reconstruction schemes PPM-0, PPM-PL07, PPM-CW84, PPM-L04 and the departure point schemes as in Section 2.6. For all simulations, the CFL number equal to 0.8 and the time integration interval is $[0, 5]$, which represents the period of the exact solution for all tests. We employ $(\Delta x^{(k)}, \Delta y^{(k)})$ -grids with $\Delta x^{(k)} = \Delta y^{(k)} = 1/2^k$ for $k = 4, \dots, 10$. We introduce the relative error in the maximum norm:

$$E_k = \max_{n=0, \dots, N_T} \frac{\|Q^n - q(t^n)\|_{\infty, \Delta x \times \Delta y}}{\|q(t^n)\|_{\infty, \Delta x \times \Delta y}},$$

and the convergence rate is defined as in Section ??, as well the total mass variation, which in all experiments presented here is preserved with machine precision. Notice that in the error computation we are going to use the centroid values instead of the exact average values to avoid computations of analytical integrals. As we pointed in Proposition 3.1, this approximation introduces a second-order error.

3.4.1 Linear advection equation with constant velocity simulations

For a constant velocity we shall adopt the $\mathbf{u} = (0.2, -0.2)$. We are going to consider as initial condition a Gaussian profile given by:

$$q_0(x, y) = \exp(-10 \cos^2(2\pi x)) \exp(-10 \cos^2(2\pi y)), \quad (x, y) \in [0, 1] \times [0, 1], \quad (3.26)$$

and a rectangular profile given by:

$$q_0(x, y) = \begin{cases} 1 & \text{if } (x, y) \in [0.4, 0.6] \times [0.4, 0.6], \\ 0 & \text{otherwise.} \end{cases} \quad (3.27)$$

The exact solution of Problem 3.1 in this case is $q_0(x - ut)$ for both q_0 . Since the velocity field is constant, all the splitting schemes introduce in Section 3.3 are the same, hence we are just going to consider the AVLT splitting. Besides that, it is easy to see that the Lie-Trotter splitting is exact in this case (cf. eg. LeVeque, 1990, p. 202-203). For 1D schemes, we use the RK1 to compute the departure point, since this scheme is exact if the velocity is constant.

The conclusions for the constant velocity tests are very similar to the 1D tests from Section ???. In fact, the error and convergence rate (Figure ??) for the Gaussian profile is very similar to the one-dimensional case (Figure 2.6). This behavior is due to the fact of no splitting error is introduced when the velocity is constant. From Figure 3.1, we can see that the AVLT splitting preserves the monotonicity when we use the monotonic 1D schemes PPM-CW84 and PPM-L04. For the non-monotonic schemes, we observe as in Figure ?? that PPM-PL07 produces less numerical dispersion than PPM-0.

3.4 | NUMERICAL EXPERIMENTS

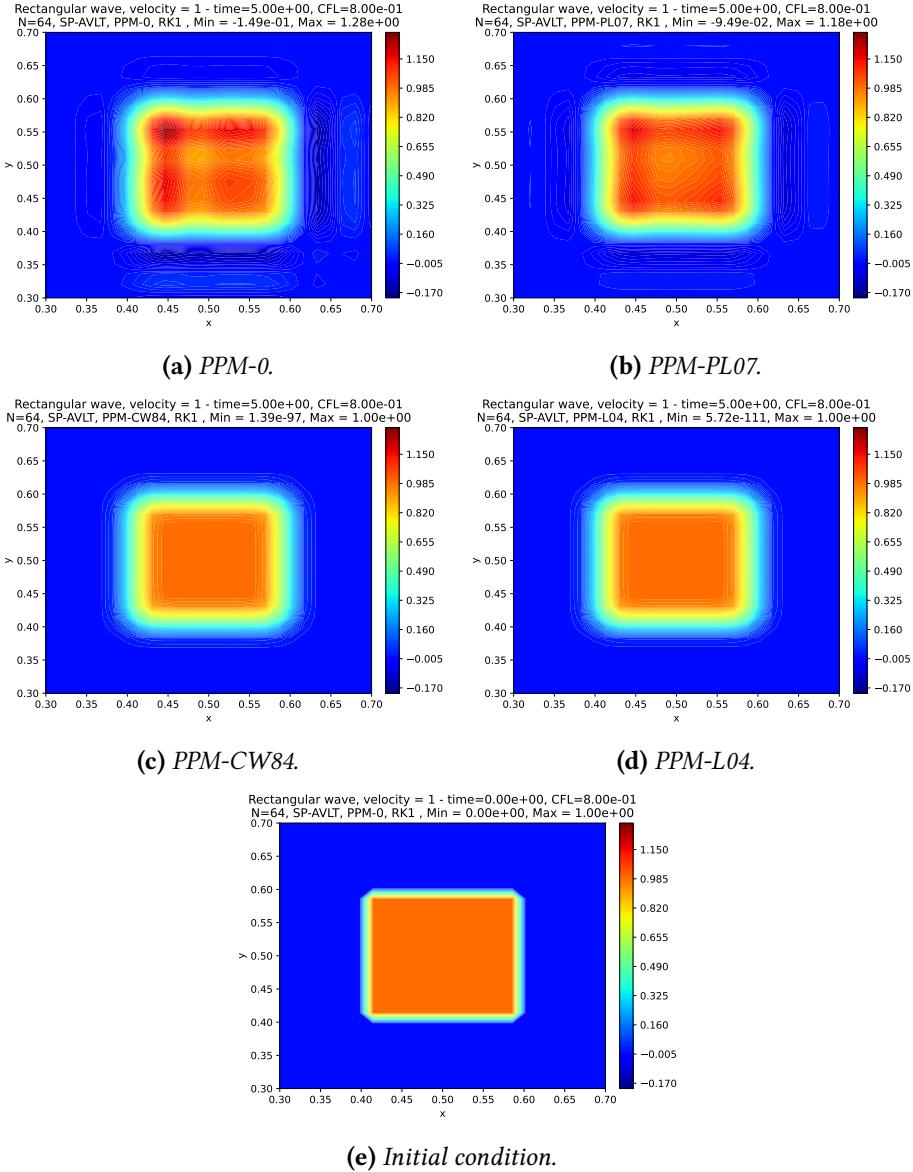


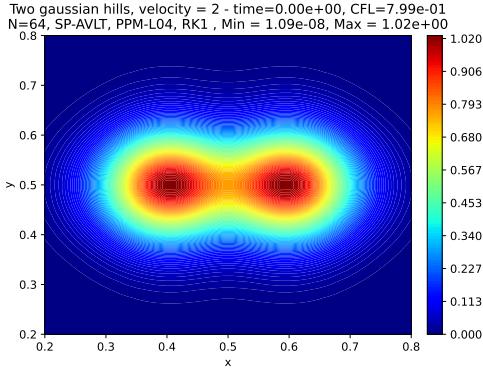
Figure 3.1: Linear advection experiment using a constant velocity equal to $\mathbf{u} = (0.2, -0.2)$, a CFL number equal to 0.8, $N = M = 64$, and the initial condition is given by Equation (3.27). We use the schemes PPM-0, PPM-PL07, PPM-CW84 and PPM-L04 with AVLT splitting. These figures show the advected profile after 5 time units (one time period). The initial condition is shown in (e).

3.4.2 Linear advection equation with variable velocity simulations

For variable velocity testing, we consider two Gaussian hills given by:

$$q_0(x, y) = \exp(-10 \cos^2(2\pi(x - 0.1))) \exp(-10 \cos^2(2\pi y)) + \exp(-10 \cos^2(2\pi(x + 0.1))) \exp(-10 \cos^2(2\pi y)), \quad (3.28)$$

defined in $[0, 1] \times [0, 1]$, whose graph is shown in Figure 3.2a.



(a) Initial condition from Equation (3.28).

We consider the velocity proposed by Nair and Lauritzen (2010):

$$\begin{cases} u(x, y, t) &= \sin(\pi x)^2 \sin(2\pi y) \cos\left(\frac{\pi t}{T}\right), \\ v(x, y, t) &= -\sin(\pi y)^2 \sin(2\pi x) \cos\left(\frac{\pi t}{T}\right), \end{cases} \quad (3.29)$$

where $T = 5$. We also consider the Cartesian version of the deformational flow test case on the sphere from Nair and Lauritzen (2010) proposed by Y. Chen et al. (2017). The velocity is given by:

$$\begin{cases} u(x, y, t) &= c \frac{\pi}{L_y} \sin^2(\alpha_1)(2 \cos(\alpha_2) \sin(\alpha_2))(\cos(\alpha_3)) - \frac{L_x}{T}, \\ v(x, y, t) &= \frac{-c}{\pi} \frac{2\pi}{L_x}(2 \sin(\alpha_1) \cos(\alpha_1) \cos^2(\alpha_2)) \cos(\alpha_3), \end{cases} \quad (3.30)$$

where $L_x = 2\pi$, $L_y = \pi$, $T = 5$, $c = \frac{10}{T}(\frac{L_x}{2\pi})^2$, $\alpha_1 = 2\pi\left(\frac{X}{L_x} - \frac{t}{T}\right)$, $\alpha_2 = \frac{\pi Y}{L_y}$, $\alpha_3 = \frac{\pi t}{T}$, $X = -\pi + 2\pi x$, $Y = -\frac{\pi}{2} + \pi y$. Y. Chen et al. (2017) uses periodic boundary conditions in the x -direction and zero-gradient in the y -direction. We shall employ biperiodic boundary conditions to simplify the problem. Both velocity fields are divergence-free and they deform the initial condition after 5 time units the scalar field returns to its initial position and shape, therefore we can compute the error.

In Figure ?? we depict the results obtained using the two Gaussian hills and the velocity field from Equation (3.30) using the PPM-L04 with AVLT, L04, and PL07 splitting and RK1 as the departure point scheme. We can observe how the scalar field has deformed and returns to its initial position. However, we can notice that the PL07 and L04 splitting are more diffusive than the AVLT splitting when we look at the solution after 5 time units in Figure ???. Employing the RK1 scheme and also considering the PPM-PL07 scheme, from Figure 3.4 and Figure 3.5 we observe that the AVLT has the larger error considering all schemes have first-order of accuracy (Figure ??). However, when we use the RK2 scheme, the most accurate schemes are obtained with AVLT with PPM-L07, which reaches third-order (Figure ??), despite AVLT being second-order, followed by AVLT with PPM-L04 (Figure ??). The other schemes seem only to not improve but get a larger error using RK2. Therefore, we conclude that the L04 and PL07 splitting introduce a first-order error. Furthermore, these schemes do not seem to differ from each other.

In Figure 3.3 we depict the results using the same setup and schemes as before but using

3.4 | NUMERICAL EXPERIMENTS

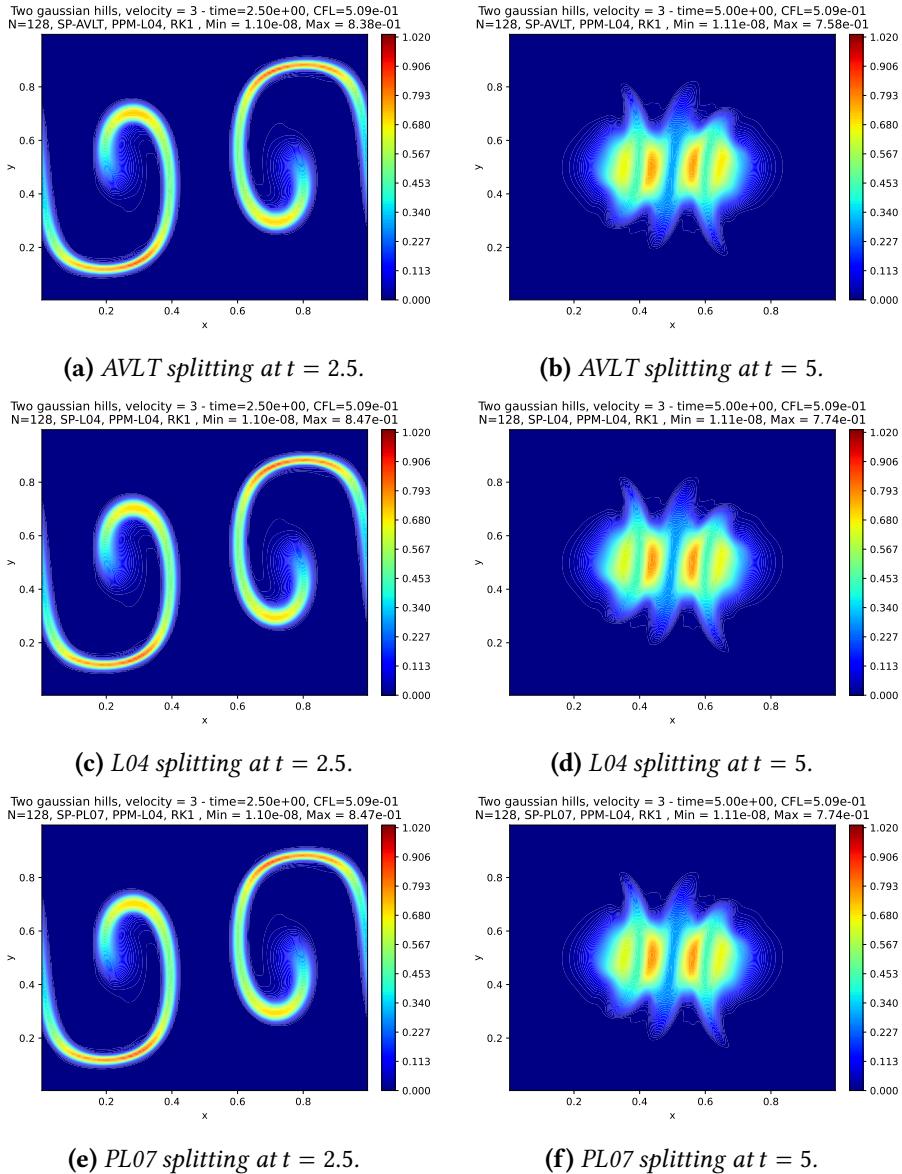


Figure 3.3: Linear advection experiment using the velocity from Equation (3.29), a CFL number equal to 0.8, $N = M = 128$, and the initial condition is given by Equation (3.28). We use the scheme PPM-L04 with AVLT (a and b), L04 (c and d) and PL07 (e and f) splitting and RK1 departure point scheme. These figures show the advected profile after 2.5 (left) and 5 (right) time units (one time period).

the velocity from Equation (3.30). As in the previous test, we see the velocity deformatring the Gaussian hills. Again the PL07 and L04 schemes produce almost the same results, however, in contrast with previous test, they are less diffusive than the AVLT. From Figure 3.4a, notice that, when using the RK1 schemes, the PL07 and L04 schemes are slightly more accurate than AVLT for the 1D PPM-PL07 schemes. All these schemes reaches a converge-order superior to two (Figure 3.5a). When we use the 1D scheme PPM-L04, all the splitting method are very similar, regardless the departure point scheme. However, when use the RK2 schemes, the AVLT is the most accurate and reaches third-order (Figure 3.5b).

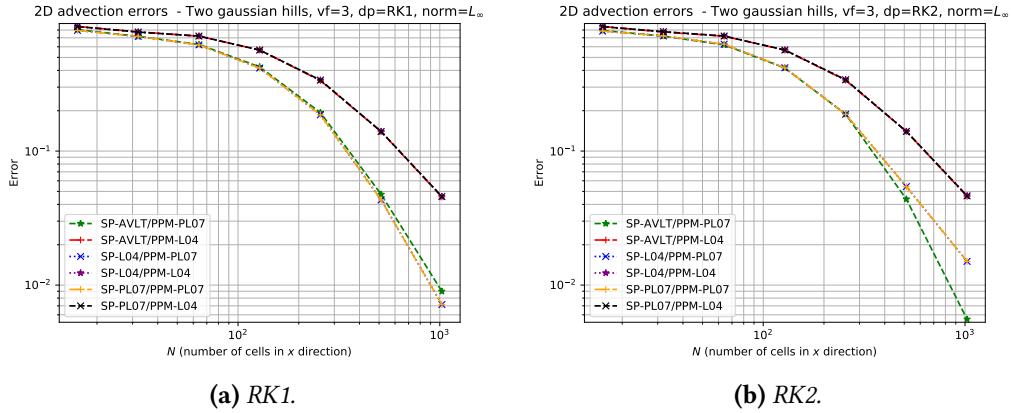


Figure 3.4: Convergence of the error for the schemes PPM-PL07 and PPM-L04 with AVLT/L04/PL07 splitting applied to the linear advection problem using a velocity from Equation (3.30), a CFL number equal to 0.8, a final time of integration equal to 5 time units and the initial condition given by Equation (3.26). The departure points are computed using the RK1 (left) and RK2 (right) schemes.

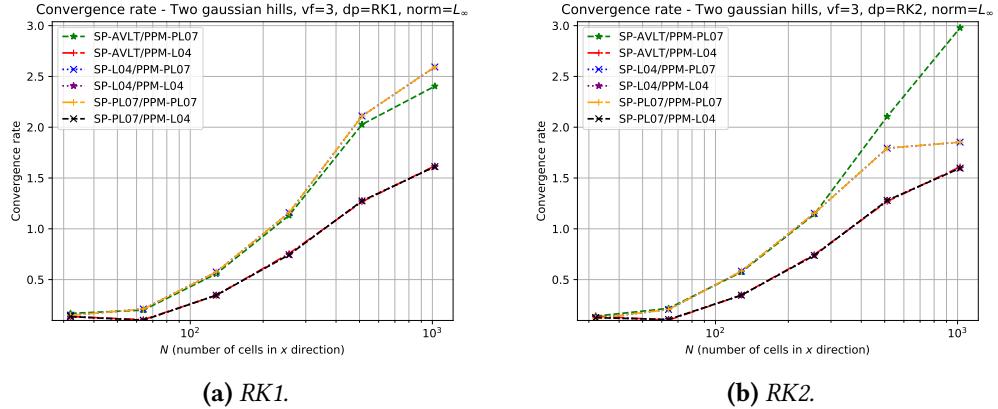


Figure 3.5: Similar to Figure 3.4 but considering the convergence rate.

3.5 Concluding remarks

In this chapter, we introduced the dimension-splitting method which replaces the solution of the 2D advection equation by the solution of multiple 1D advection equations. The average of two Lie-Trotter splitting, which is second-order accurate, was modified to ensure the preservation of a constant scalar field with a divergence-free velocity, which the classical averaging Lie-Trotter splitting lacks, following the methodology used in FV3.

From the constant velocity simulations, we conclude that all the splitting schemes are equal and they do not introduce any splitting error. In fact, the splittings are exact. We could observe that all conclusions for the 1D simulations were valid in the 2D case, with mass conservation and monotonicity being respected in the 2D case whenever we employ the monotonic limiters in the 1D subproblems.

For variable velocity simulations, we investigated two flow deformation test cases.

We observed that all splitting schemes preserved the monotonicity. The schemes PL07 and L04 yields very similar results and they introduced a first-order error, which was observed when we employed the RK2 scheme to compute the departure point. In this case, the AVLT reached third-order of accuracy, which is better than expected since this scheme is only second-order. Therefore, we can conclude that a more accurate departure point calculation benefits more the AVLT splitting. However, when using a first-order departure point computation, the splitting schemes PL07 and L04 produced slightly smaller errors.

Chapter 4

Cubed-sphere grids

Consistent Treatment of Boundaries: Boundary conditions play a crucial role in maintaining mass conservation. The treatment of boundaries should ensure that there is no spurious mass exchange between the model domain and the exterior. Various approaches, such as using ghost cells or extrapolation methods, can be employed to enforce mass conservation at the boundaries.

The cubed-sphere grid was originally proposed by Sadourny (1972) and was reinvestigated by Ronchi et al. (1996) and Rančić et al. (1996). As is usual to Planotonic grids, we start with a Platonic solid, in this case, a cube, circumscribed in a sphere and project its faces on the sphere. The original cubed-sphere proposed by Sadourny (1972), called equidistant cubed-sphere, leads to non-uniform grid; a solution to this problem was proposed with the introduction of angular coordinates, leading to a quasi-uniform grid called equiangular cubed-sphere. The cubed sphere then consists of six panels, each one having a local Cartesian coordinate system, which make it easier to extend methods from the plane to the sphere. Indeed, Putman and Lin (2007) extends the dimension splitting from Lin and Rood (1996), presented in Chapter 3, to the cubed-sphere.

There are essentially two major challenges when working on the cubed-sphere: 1) the non-orthogonal grid system; 2) the discontinuity of the coordinate system at the cube edges. Challenge 1) is more related to the appearance of metric terms in the equations, which adds more computational cost requiring, for instance, several conversions between contravariant/covariant components of a velocity field. The second challenge, perhaps the most problematic, is related to the computation of stencils along the cube edges, where the coordinate system is discontinuous. A possible way to compute the stencil at the edges is to extend the local coordinate of each panel to its neighbor panels, adding ghost cells in the so-called halo region. In this case, the equiangular cubed sphere has ghost cell values, lying on the same geodesics containing the data from its neighbor panels, which allows us to use one-dimensional high-order Lagrange interpolation (for a review of this method, see Zerroukat and Allen (2022)). This approach was already investing since the work of Ronchi et al. (1996) and it is widely used in the literature (X. Chen, 2021; Croisille, 2013; Katta et al., 2015a, 2015b). Putman and Lin (2007), on the other hand, uses extrapolation at grid values near the cube edges. Another approach that avoids the need for interpolation/extrapolation near the edges is the conformal cubed-sphere developed by Rančić et al. (1996). This grid

leads to an orthogonal and continuous coordinate system near the edges, but it generates grid singularities near the cube corners, similar to the pole problem. An improved and more uniform conformal grid, called Uniform Jacobian cubed sphere, was later proposed Rančić et al. (2017). Each approach is likely to generate grid imprinting and one of our goals this work is to investigate how much grid-imprinting each method generates.

This chapter aims to review and investigated the geometrical properties of the cubed-sphere. Besides that, we also aim to investigate the process of interpolating/extrapolating near the cube edges. We start with a basic review of the cubed-sphere mappings in Section 4.1, while Section 4.2 investigates the interpolation/extrapolation near the cube edges with some numerical experiments.

4.1 Cubed-sphere mappings

4.1.1 Equidistant cubed-sphere

We consider a sphere radius $R > 0$ and $a = \frac{R}{\sqrt{3}}$ representing the half-length of the cube, and the family of maps $\Psi_p : [-a, a] \times [-a, a] \rightarrow \mathbb{S}_R^2$, $p = 1, \dots, 6$, where:

$$\Psi_1(x, y) = \frac{R}{\sqrt{a^2 + x^2 + y^2}}(a, x, y), \quad (4.1)$$

$$\Psi_2(x, y) = \frac{R}{\sqrt{a^2 + x^2 + y^2}}(-x, a, y), \quad (4.2)$$

$$\Psi_3(x, y) = \frac{R}{\sqrt{a^2 + x^2 + y^2}}(-a, -x, y), \quad (4.3)$$

$$\Psi_4(x, y) = \frac{R}{\sqrt{a^2 + x^2 + y^2}}(x, -a, y), \quad (4.4)$$

$$\Psi_5(x, y) = \frac{R}{\sqrt{a^2 + x^2 + y^2}}(-y, x, a), \quad (4.5)$$

$$\Psi_6(x, y) = \frac{R}{\sqrt{a^2 + x^2 + y^2}}(y, x, -a). \quad (4.6)$$

The family of maps $\{\Psi_p, p = 1, \dots, 6\}$ allow us to cover the sphere. Here p denotes a panel, and they are defined as Figure 4.1 shows. The derivative of the maps Ψ_p are given by:

$$D\Psi_1(x, y) = \frac{R}{(a^2 + x^2 + y^2)^{3/2}} \begin{bmatrix} -ax & -ay \\ a^2 + y^2 & -xy \\ -xy & a^2 + x^2 \end{bmatrix}, \quad (4.7)$$

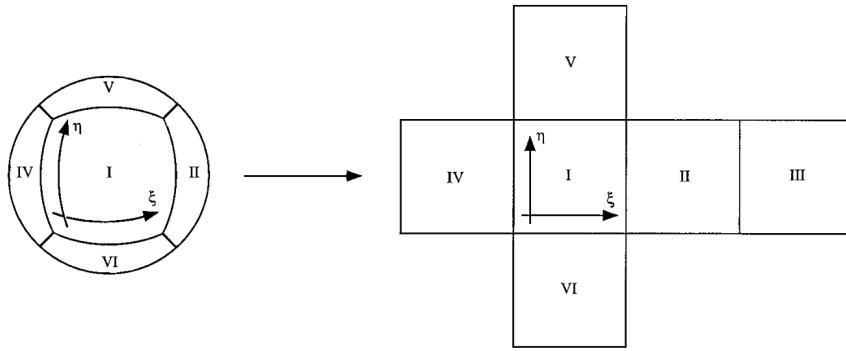


Figure 4.1: Cubed-sphere panels definition. Figure taken from Ronchi et al. (1996).

$$D\Psi_2(x, y) = \frac{R}{(a^2 + x^2 + y^2)^{3/2}} \begin{bmatrix} -(a^2 + y^2) & xy \\ -ax & -ay \\ -xy & a^2 + x^2 \end{bmatrix}, \quad (4.8)$$

$$D\Psi_3(x, y) = \frac{R}{(a^2 + x^2 + y^2)^{3/2}} \begin{bmatrix} ax & ay \\ -(a^2 + y^2) & xy \\ -xy & a^2 + x^2 \end{bmatrix}, \quad (4.9)$$

$$D\Psi_4(x, y) = \frac{R}{(a^2 + x^2 + y^2)^{3/2}} \begin{bmatrix} a^2 + y^2 & -xy \\ ax & ay \\ -xy & a^2 + x^2 \end{bmatrix}, \quad (4.10)$$

$$D\Psi_5(x, y) = \frac{R}{(a^2 + x^2 + y^2)^{3/2}} \begin{bmatrix} xy & -(a^2 + x^2) \\ a^2 + y^2 & -xy \\ -ax & -ay \end{bmatrix}, \quad (4.11)$$

$$D\Psi_6(x, y) = \frac{R}{(a^2 + x^2 + y^2)^{3/2}} \begin{bmatrix} -xy & a^2 + x^2 \\ a^2 + y^2 & -xy \\ ax & ay \end{bmatrix}. \quad (4.12)$$

With the aid of the derivative, we may define a basis of tangent vectors $\{\mathbf{g}_1, \mathbf{g}_2\}$ on each point on the sphere by:

$$\mathbf{g}_1(x, y; p) = D\Psi_p(x, y) \begin{bmatrix} 1 \\ 0 \end{bmatrix}, \mathbf{g}_2(x, y; p) = D\Psi_p(x, y) \begin{bmatrix} 0 \\ 1 \end{bmatrix}. \quad (4.13)$$

In other words, we have $\{\mathbf{g}_1(x, y; p), \mathbf{g}_2(x, y; p)\} \subset T_{\Psi_p(x, y)} \mathbb{S}_R^2$, $\forall (x, y) \in [-a, a] \times [-a, a]$ (see Appendix B for the definition of tangent space). Notice that

$$[D\Psi_p(x, y)]^T D\Psi_p(x, y) = \frac{R^2}{(a^2 + x^2 + y^2)^2} \begin{bmatrix} a^2 + x^2 & -xy \\ -xy & a^2 + y^2 \end{bmatrix}, \quad (4.14)$$

does not depend on p . Hence, it makes sense to define the matrix $G_\Psi(x, y) = [D\Psi_p(x, y)]^T D\Psi_p(x, y)$ which is known as metric tensor. It is easy to see that:

$$G_\Psi(x, y) = \begin{bmatrix} \langle \mathbf{g}_1(x, y; p), \mathbf{g}_1(x, y; p) \rangle & \langle \mathbf{g}_1(x, y; p), \mathbf{g}_2(x, y; p) \rangle \\ \langle \mathbf{g}_1(x, y; p), \mathbf{g}_2(x, y; p) \rangle & \langle \mathbf{g}_2(x, y; p), \mathbf{g}_2(x, y; p) \rangle \end{bmatrix}, \quad (4.15)$$

and that $G_\Psi(x, y)$ is positive-definite, $\forall(x, y) \in [-a, a] \times [-a, a]$. The Jacobian of the metric tensor $G_\Psi(x, y)$ is then given by:

$$\sqrt{|\det G_\Psi(x, y)|} = \frac{R^2}{(a^2 + x^2 + y^2)^{3/2}} a. \quad (4.16)$$

4.1.2 Equiangular cubed-sphere

Another cubed-sphere mapping is the equiangular mapping (Ronchi et al., 1996), which leads to a more uniform grid. This mapping is a composition of equidistant mapping with angular coordinates. We consider again $a = \frac{R}{\sqrt{3}}$ and we define the family of maps $\Phi_p : [-\frac{\pi}{4}, \frac{\pi}{4}] \times [-\frac{\pi}{4}, \frac{\pi}{4}] \rightarrow \mathbb{S}_R^2$, $p = 1, \dots, 6$, given by $\Phi_p(x, y) = \Psi_p(a \tan x, a \tan y)$. The coordinates $(a \tan x, a \tan y)$ are called angular coordinates. By the chain rule:

$$D\Phi_p(x, y) = a D\Psi_p(a \tan x, a \tan y) \begin{bmatrix} \frac{1}{\cos^2 x} & 0 \\ 0 & \frac{1}{\cos^2 y} \end{bmatrix}, \quad (4.17)$$

and therefore we can define the following tangent vectors

$$\mathbf{r}_1(x, y; p) = D\Phi_p(x, y) \begin{bmatrix} 1 \\ 0 \end{bmatrix} = \frac{a}{\cos^2 x} \mathbf{g}_1(\tan x, \tan y; p), \quad (4.18)$$

$$\mathbf{r}_2(x, y; p) = D\Phi_p(x, y) \begin{bmatrix} 0 \\ 1 \end{bmatrix} = \frac{a}{\cos^2 y} \mathbf{g}_2(\tan x, \tan y; p), \quad (4.19)$$

that is, $\{\mathbf{r}_1(x, y; p), \mathbf{r}_2(x, y; p)\} \subset T_{\Phi_p(x, y)} \mathbb{S}_R^2$, $\forall(x, y) \in [-\frac{\pi}{4}, \frac{\pi}{4}] \times [-\frac{\pi}{4}, \frac{\pi}{4}]$. Again, it makes sense to define the matrix

$$G_\Phi(x, y) = [D\Phi_p(x, y)]^T D\Phi_p(x, y) \quad (4.20)$$

$$= a^2 [D\Psi_p(a \tan x, a \tan y)]^T \begin{bmatrix} \frac{1}{\cos^4 x} & 0 \\ 0 & \frac{1}{\cos^4 y} \end{bmatrix} D\Psi_p(a \tan x, a \tan y), \quad (4.21)$$

that does not depend on p and is the metric tensor. It is easy to see that:

$$G_\Phi(x, y) = \begin{bmatrix} \langle \mathbf{r}_1(x, y; p), \mathbf{r}_1(x, y; p) \rangle & \langle \mathbf{r}_1(x, y; p), \mathbf{r}_2(x, y; p) \rangle \\ \langle \mathbf{r}_1(x, y; p), \mathbf{r}_2(x, y; p) \rangle & \langle \mathbf{r}_2(x, y; p), \mathbf{r}_2(x, y; p) \rangle \end{bmatrix}, \quad (4.22)$$

and that $G_\Phi(x, y)$ is positive-definite, $\forall(x, y) \in [-\frac{\pi}{4}, \frac{\pi}{4}] \times [-\frac{\pi}{4}, \frac{\pi}{4}]$. The Jacobian of the metric tensor $G_\Phi(x, y)$ is then given by:

$$\begin{aligned} \sqrt{|\det G_\Phi(x, y)|} &= \frac{a}{\cos^2 x \cos^2 y} \frac{R^2}{(a^2 + a^2 \tan^2 x + a^2 \tan^2 y)^{3/2}} a \\ &= \frac{R^2}{\cos^2 x \cos^2 y} \frac{1}{(1 + \tan^2 x + \tan^2 y)^{3/2}}. \end{aligned} \quad (4.23)$$

Hereafter, we shall denote $g(x, y) = |\det G_\Phi(x, y)|$.

4.1.3 Examples

To represent the cubed-sphere grid, we consider the notation of Section 3.2. We shall assume that we have a $(\Delta x, \Delta y)$ -grid of $[-a, a]^2$, with $\Delta x = \Delta y$, $N = M$, where a depends on the mapping considered. Similarly to Chapter 3, we may extend the grid to ghost cells. Hence, for each $p = 1, \dots, 6$, we apply the mapping Ψ_p or Φ_p on each local coordinate system grid to generate the cubed-sphere with $6N^2$ cells. In Figure 4.2, we depict an example of the grids generated using the equidistant and equiangular mappings for $N = 20$.

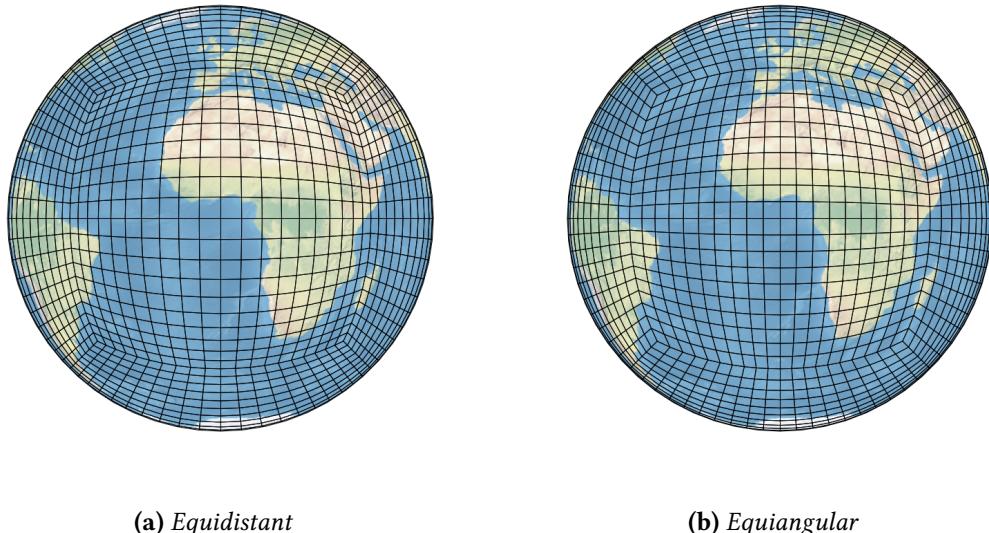


Figure 4.2: Equidistant (a) and equiangular (b) cubed-spheres generated with $N = 20$.

From Figure 4.2, we notice that the equiangular cubed-sphere is much more uniform than the equidistant cubed-sphere. As pointed out by Rančić et al. (1996), the ratio between the maximum and minimum cell area on the equiangular cubed-sphere is approximately 1.3, while the same ratio is approximately 5.2 on the equidistant cubed-sphere.

4.2 Edges treatment

4.2.1 Ghost cells interpolation

Hereafter, we are going to use the notation $(x, y; p)$ for $p = 1, \dots, 6$, to represent a point on the sphere obtained through the equiangular cubed-sphere mapping. Let us assume that we have a function $q : \mathbb{S}_R^2 \rightarrow \mathbb{R}$ given at the cell centroids and let us denote these values by $q_{ijp} = q(x_i, y_j; p)$, for $i, j = 1, \dots, N$, $p = 1, \dots, 6$. We wish to estimate these values outside of the range $1, \dots, N$, that is, we wish to estimate them at ghost cell positions.

As we pointed out before, this can be done by noticing that the ghost cells on the local Cartesian systems are mapped onto the geodesic of a adjacent panel, which allows us to use Lagrange interpolation to obtain the ghost cell values. To illustrate this process in Panel 1, in Figure 4.3 we depict the values of q_{ijp} using green circles for Panel 1 and black

circles for the other panels, for $N = 8$. Assuming a halo size equal to 3, we also show the target value at ghost cell positions using yellow and magenta circles. Observe that the dashed yellow lines in Figure 4.3 illustrate how the ghost cell points lie in geodesics containing grid position from adjacent panels. Except for the magenta circles, all the ghost cell values may be obtained using 1D Lagrange interpolation using the surrounding black circles on the geodesic. This can be performed for all panels. After that, the magenta circles may be interpolated using also the values obtained in the first step of interpolation (see cyan circles in Figure 4.3), preserving the order of accuracy of the interpolation, supposing this one is fixed.

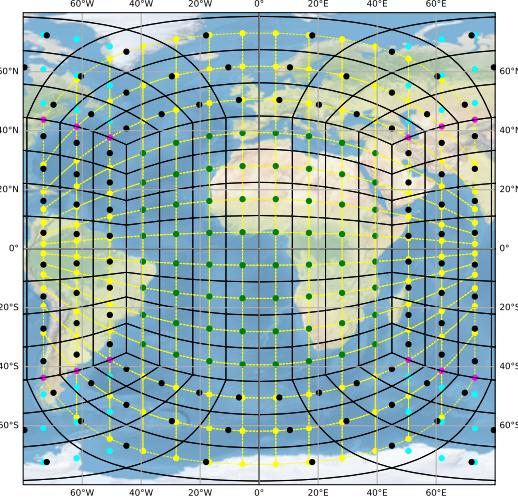


Figure 4.3: Equiangular cubed-sphere panel 1 with $N = 8$: centroid at panel 1 (green circles) and others panel centroids (black circles), ghost cell points at panel 1 (yellow and magenta circles) and others panels (cyan circles).

We are going to show a numerical example of this interpolation process using a halo region of size 4 and assuming the radius of the sphere is equal to one. We shall consider the following trigonometric function, which is the divergence of a velocity field, as in Peixoto and Barros (2013) in our tests:

$$q(\lambda, \phi) = \frac{1}{\cos(\phi)} \left(-2 \cos^3(\phi) \sin(\lambda) \cos(\lambda) + 16 \sin^2(\lambda) \cos(\lambda) \cos^3(\phi) \sin(\phi) \right), \quad (4.24)$$

whose graph is depicted in Figure 4.4. We are going to consider the relative error in the maximum norm and the convergence rate at the ghost cell positions defined analogously as in Section 3.4. We also consider values os N given by 2^k , for $k = 4, \dots, 10$, in order to compute the error and convergence rate. In Figures 4.5 we show the error and convergence rate for the trigonometric function (Equation (4.24)), respectively, considering polynomials of degrees from 0 up to 4. As both graphs show, we were able to achieve the expected order of convergence. Next, we shall use the ghost cell interpolation when computing the stencils from PPM edge reconstructions presented in Chapter 2 in each direction of a panel

and also compare the results obtained with extrapolation.

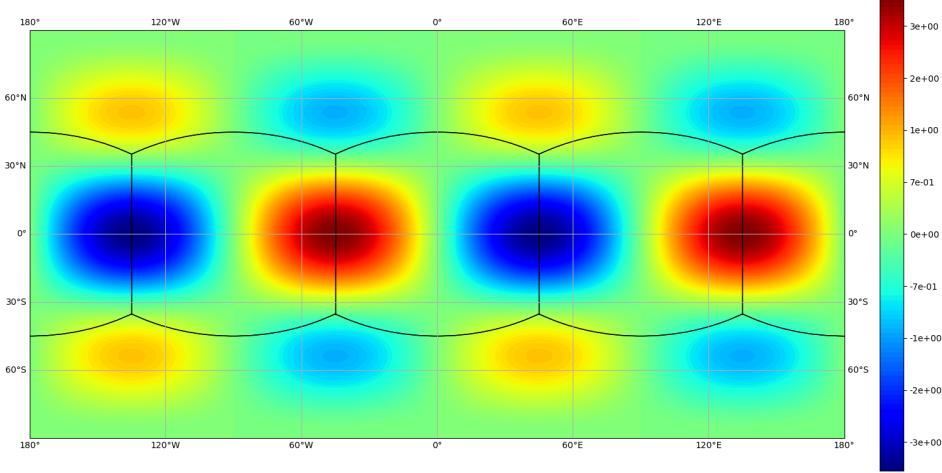


Figure 4.4: Trigonometric function defined by Equation (4.24).

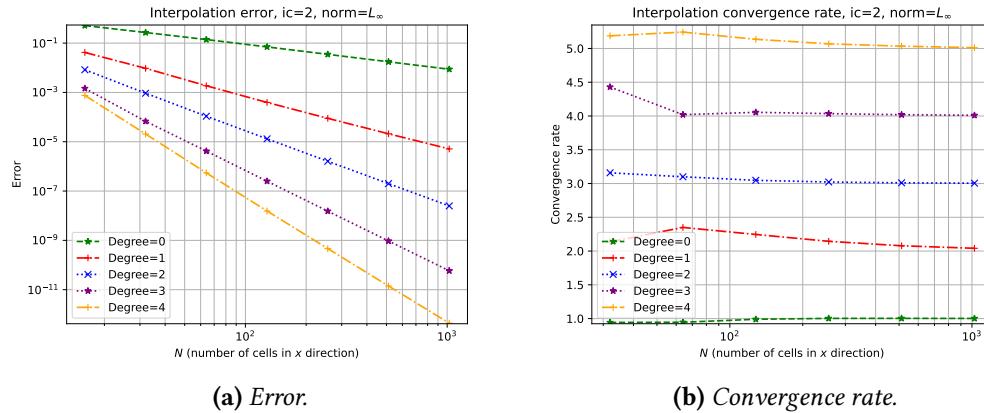


Figure 4.5: Relative error convergence (a) and convergence rate (b) for the ghost cell interpolation process for different polynomial degrees, using the trigonometric function given by Equation (4.24).

4.2.2 Edges reconstruction

Let denote a control volume of the cubed-sphere by Ω_{ijp} , that is:

$$\Omega_{ijp} = \Phi_p([x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}] \times [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}]), \quad 1 \leq i, j \leq N, \quad 1 \leq p \leq 6.$$

We define the average values of a function q with the aid of the metric tensor $g(x, y)$ (Equation 4.23):

$$Q_{ijp} = \frac{1}{|\Omega_{ijp}|} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} q(x, y; p) \sqrt{g(x, y)} dx dy,$$

where $|\Omega_{ijp}|$ is the control volume area given by:

$$|\Omega_{ijp}| = \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \sqrt{g(x, y)} dx dy.$$

Similar to Proposition 3.1, we may approximate the average value using the centroid value, that is

$$Q_{ijp} - q_{ijp} = O(\Delta x^2),$$

where $q_{ijp} = q(x_i, y_j; p)$, recalling that on the cubed-sphere we always assume $\Delta x = \Delta y$. In this work, we shall always approximate the average values since our schemes are expected to be at most second-order, this approximation does not deteriorate the convergence order.

Let us consider the following problem: given the values q_{ijp} we wish to find approximations of the function q at the control volume edge midpoints denoted by $q_{ijp}^{L,x} \approx q_{i-\frac{1}{2},j,p}$, $q_{ijp}^{R,x} \approx q_{i+\frac{1}{2},j,p}$, $q_{ijp}^{L,y} \approx q_{i,j-\frac{1}{2},p}$, $q_{ijp}^{R,y} \approx q_{i,j+\frac{1}{2},p}$, where we also using the notations $q_{i+\frac{1}{2},j,p} \approx q(x_{i+\frac{1}{2}}, y_j; p)$, $q_{i,j+\frac{1}{2},p} \approx q(x_i, y_{j+\frac{1}{2}}; p)$. These points are illustrated in Figure 4.6 for panel 1.

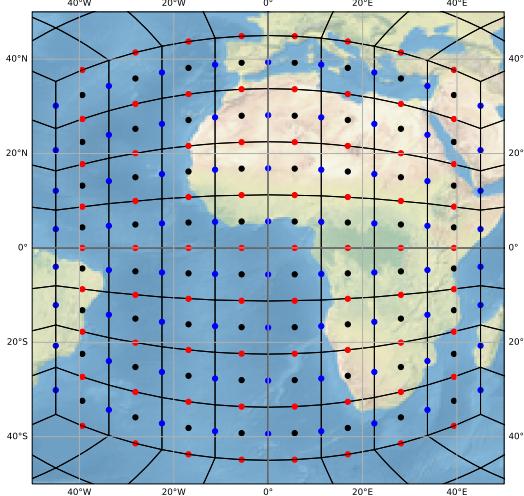


Figure 4.6: The reconstruction problem on the cubed-sphere panel 1: we are given the centroid values of a function (black circles) and we wish to estimate these values at the edges midpoints in the x direction (blue circles) and y direction (red circles). This figure uses an equiangular cubed-sphere with $N = 8$.

We can estimate the desired values using the one-dimensional reconstruction schemes from Sections 2.4 and 2.4.2 by performing the PPM reconstruction in the x and y directions independently. Notice that all the schemes from Sections 2.4 and 2.4.2 are expected to be second-order accurate due to centroid point approximation. The major difference here is when we compute the stencil near the cube edges. Unlike the previous chapters where we assumed periodic boundary conditions, the boundary conditions are related to the

adjacent panels. One way to overcome this problem is just to add ghost cell layers and use the process described in Section 4.2.1, hence all the stencils may be computed. This approach shall be referred to as **ET-R96** (ET stands for edge treatment) since this approach of extending the grid lines and using Lagrange interpolation was originally proposed in Ronchi et al. (1996).

Another way to fill the ghost cell values used for instance in Sadourny (1972), is just to ignore the discontinuity of the coordinate system and use the values of the cells in the adjacent panels as the ghost cell values. This scheme is labeled **ET-S72**. An approach that avoids the use of ghost cells has been developed by Putman and Lin (2007) using extrapolation at the cells surrounding to the cube edge. We are going to describe this scheme that we shall name **ET-PL07**. This scheme uses the following extrapolations:

$$\begin{aligned} q_{1,j,p}^{L,x} &= \frac{1}{2} \left(3Q_{1,j,p} - Q_{2,j,p} \right), \\ q_{N,j,p}^{R,x} &= \frac{1}{2} \left(3Q_{N,j,p} - Q_{N-1,j,p} \right), \\ q_{i,1,p}^{L,y} &= \frac{1}{2} \left(3Q_{i,1,p} - Q_{i,2,p} \right), \\ q_{i,N,p}^{R,y} &= \frac{1}{2} \left(3Q_{i,N,p} - Q_{i,N-1,p} \right), \end{aligned}$$

at the points that are located on the cube edges. The other edge values are estimated as:

$$\begin{aligned} q_{1,j,p}^{R,x} &= \frac{1}{14} \left(3Q_{1,j,p} + 11Q_{2,j,p} - 2(Q_{3,j,p} - Q_{1,j,p}) \right), \\ q_{2,j,p}^{L,x} &= q_{1,j,p}^{R,x}, \\ q_{N,j,p}^{L,x} &= \frac{1}{14} \left(3Q_{N,j,p} + 11Q_{N-1,j,p} - 2(Q_{N-2,j,p} - Q_{N,j,p}) \right), \\ q_{N-1,j,p}^{R,x} &= q_{N,j,p}^{L,x}, \end{aligned}$$

in the x direction and in the y direction we use the formulas

$$\begin{aligned} q_{i,1,p}^{R,y} &= \frac{1}{14} \left(3Q_{i,1,p} + 11Q_{i,2,p} - 2(Q_{i,3,p} - Q_{i,1,p}) \right), \\ q_{i,2,p}^{L,y} &= q_{i,1,p}^{R,y}, \\ q_{i,N,p}^{L,y} &= \frac{1}{14} \left(3Q_{i,N,p} + 11Q_{i,N-1,p} - 2(Q_{i,N-2,p} - Q_{i,N,p}) \right), \\ q_{i,N-1,p}^{R,y} &= q_{i,N,p}^{L,y}. \end{aligned}$$

We are going to use the trigonometric function (Equation (4.24)) as before on the unit sphere to compare the schemes ET-R96, ET-S72 and ET-PL07. The scheme ET-R96 uses cubic polynomials. We introduce the relative errors:

$$\begin{aligned} e_{i-\frac{1}{2},j,p} &= (|q_{i-\frac{1}{2},j,p} - q_{ijp}^{L,x}|)/|q_{i-\frac{1}{2},j,p}|, \\ e_{i+\frac{1}{2},j,p} &= (|q_{i+\frac{1}{2},j,p} - q_{ijp}^{R,x}|)/|q_{i+\frac{1}{2},j,p}|, \\ e_{i,j-\frac{1}{2},p} &= (|q_{i,j-\frac{1}{2},p} - q_{ijp}^{L,y}|)/|q_{i,j-\frac{1}{2},p}|, \\ e_{i,j+\frac{1}{2},p} &= (|q_{i,j+\frac{1}{2},p} - q_{ijp}^{R,y}|)/|q_{i,j+\frac{1}{2},p}|, \\ e_{ijp} &= \max\{e_{i-\frac{1}{2},j,p}, e_{i+\frac{1}{2},j,p}, e_{i,j-\frac{1}{2},p}, e_{i,j+\frac{1}{2},p}\}, \\ E &= \max\{e_{ijp}\}. \end{aligned}$$

We are going to compute E for different values of N as in the numerical experiments of Section 4.2.1.

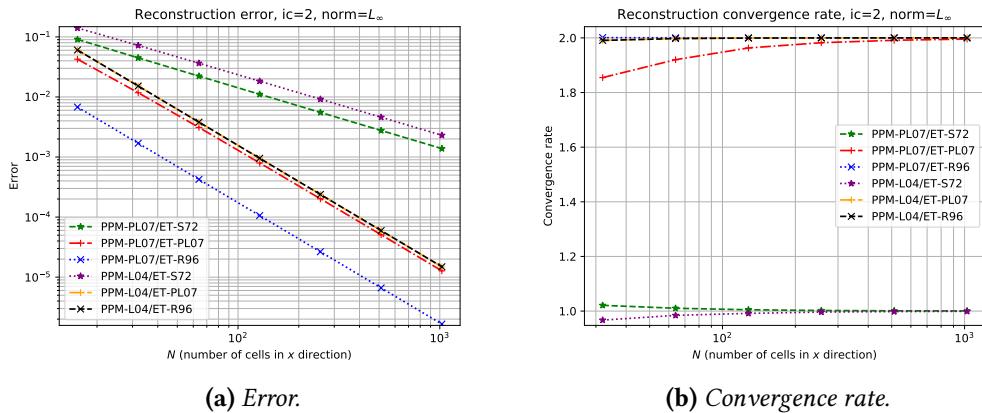


Figure 4.7: Relative error convergence (a) and convergence rate (b) for the reconstruction problem for different edge treatments (ET), using the trigonometric function given by Equation (4.24).

In Figure 4.7 we show the errors and the convergence rate using the PPM-PL07 and PPM-L04 reconstruction schemes. We can observe that all the schemes converge to zero with second-order. The difference between the edge treatments ET-S72, ET-PL07 and ET-R96 schemes may be observed in Figure 4.8 and Figure 4.9, for the PPM-PL07 and PPM-L04 reconstruction schemes, respectively. We notice that the cube edges appear on the error graph when we use the ET-S72 and ET-PL07 schemes. This is an example of grid imprinting. Although all schemes are second-order, the scheme ET-R96 do not seem to produce grid imprinting in the reconstruction problem.

4.2 | EDGES TREATMENT

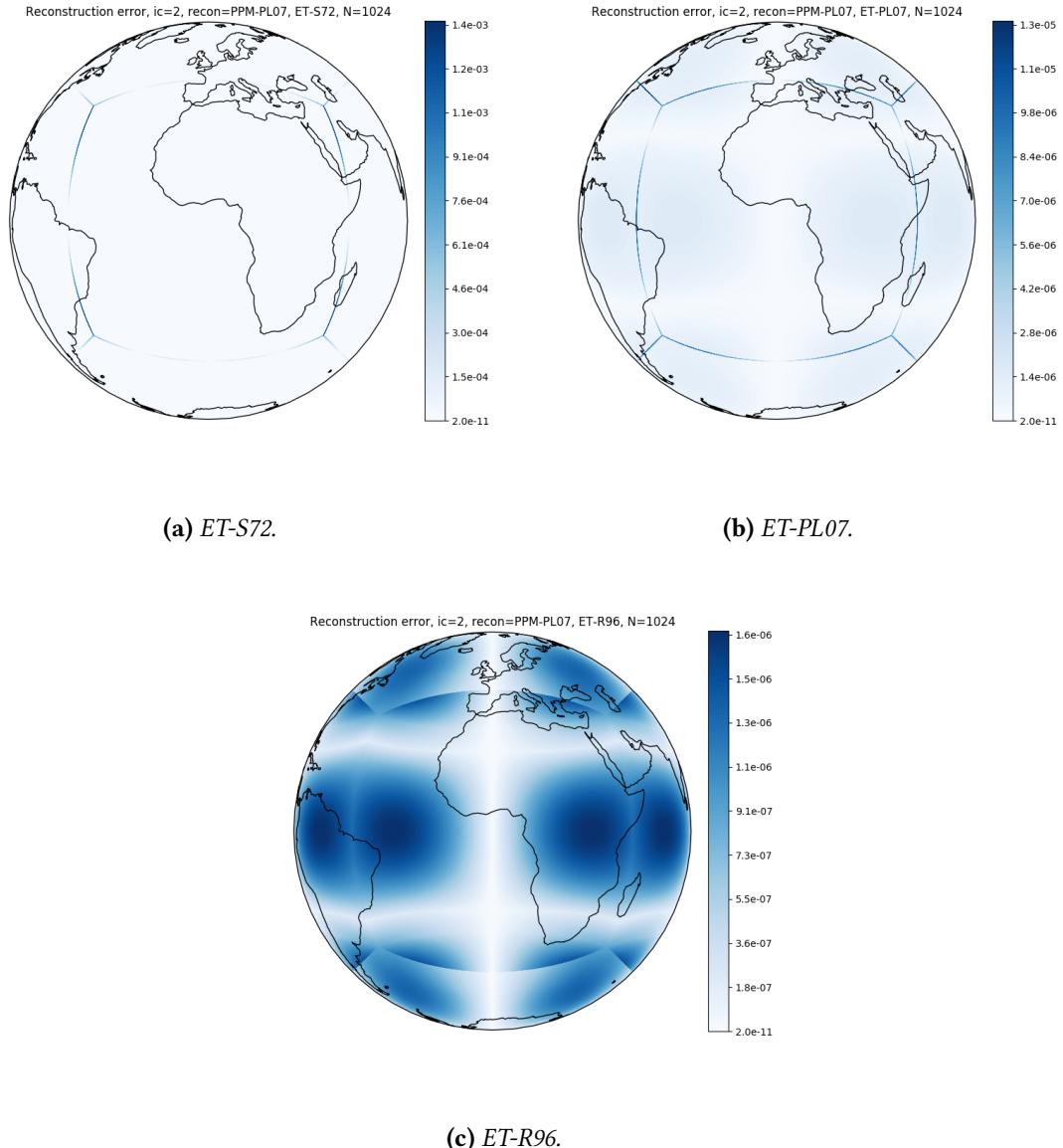


Figure 4.8: Error for the edge midpoint values considering the trigonometric function (Equation (4.24)) with edges treatment schemes ET-S72 (a), ET-PL07 (b) and ET-R96 for the cubed-sphere with $N = 1024$ using the reconstruction PPM-PL07.

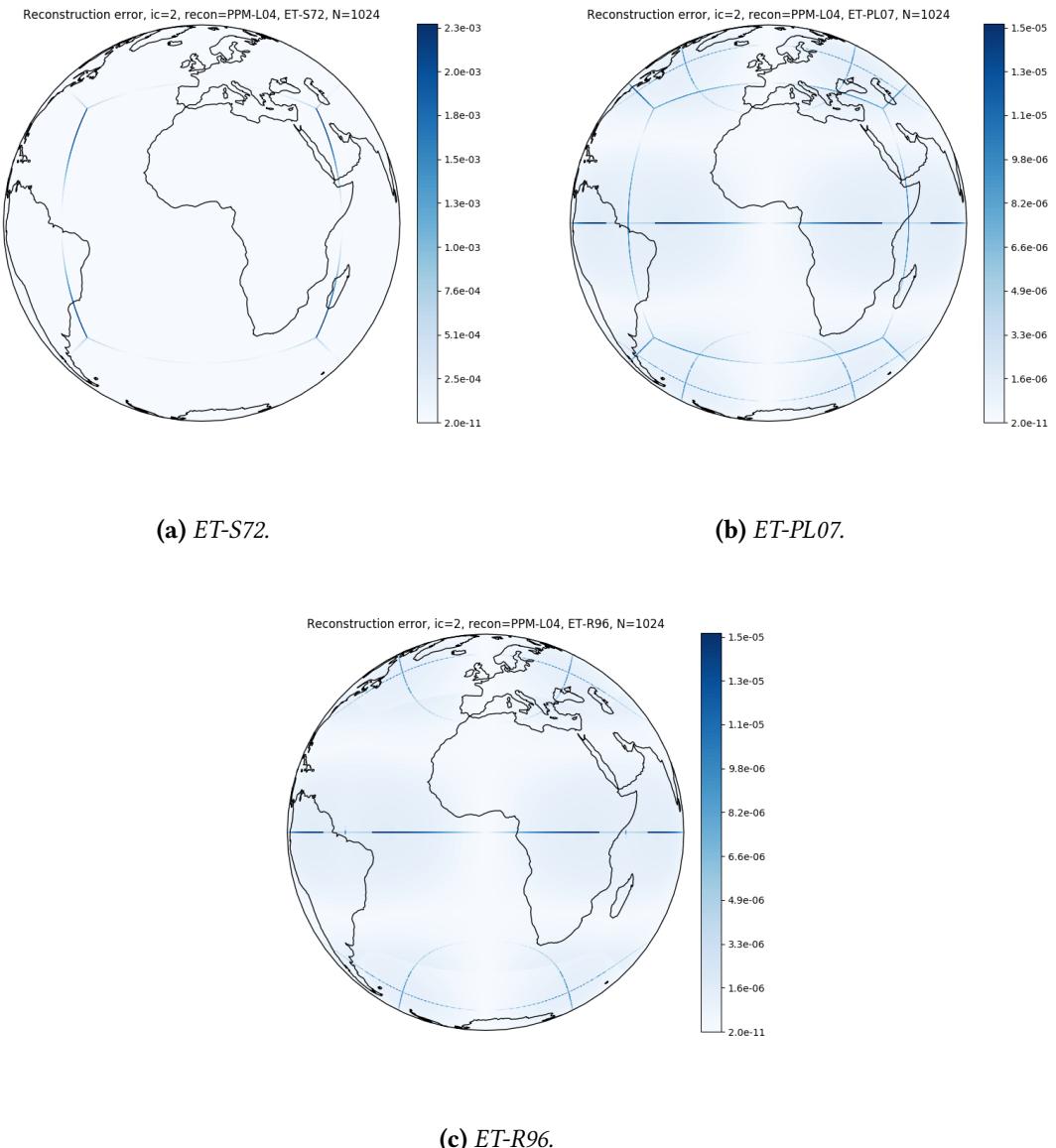


Figure 4.9: As Figure in 4.8 but using the PPM-L04 scheme.

Chapter 5

Cubed-sphere finite-volume methods

5.1 Advection finite-volume scheme

In this Chapter, we show how we can use the dimension splitting method presented in Chapter 3 to solve the advection equation on the cubed-sphere with base on Putman and Lin (2007).

We denote by $\Psi_p : [-a, a] \times [-a, a] \rightarrow \mathbb{S}_R^2$, $p = 1, \dots, 6$, as a cubed-sphere mapping introduce in Chapter 4. We introduce the notations:

- $(x, y; p)$ represents a point on the cubed-sphere using a cubed-sphere mapping;
 - $[-a, a]^2 = \bigcup_{i,j=1}^N [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}] \times [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}]$;
 - $\Delta x = x_{i+\frac{1}{2}} - x_{i-\frac{1}{2}}$, $\Delta y = y_{j+\frac{1}{2}} - y_{j-\frac{1}{2}}$;
 - $\Omega_{ijp} = \Psi_p([x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}] \times [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}])$ are the cubed-sphere control-volumes;
 - $\mathbf{g}_1(x, y; p) = D\Psi_p(x, y) \begin{bmatrix} 1 \\ 0 \end{bmatrix}$ and $\mathbf{g}_2(x, y; p) = D\Psi_p(x, y) \begin{bmatrix} 0 \\ 1 \end{bmatrix}$ are the tangent vectors;
 - $g_\Psi(x, y) = \begin{bmatrix} \langle \mathbf{g}_1(x, y; p), \mathbf{g}_1(x, y; p) \rangle & \langle \mathbf{g}_1(x, y; p), \mathbf{g}_2(x, y; p) \rangle \\ \langle \mathbf{g}_1(x, y; p), \mathbf{g}_2(x, y; p) \rangle & \langle \mathbf{g}_2(x, y; p), \mathbf{g}_2(x, y; p) \rangle \end{bmatrix}$ is the metric tensor;
 - $\sqrt{\det g_\Psi(x, y)}$ is the metric tensor Jacobian;
 - $|\Omega_{ijp}| = \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \sqrt{\det g_\Psi(x, y)} dx dy$
- are the control-volume areas
- $Q_{ijp}(t) = \frac{1}{|\Omega_{ijp}|} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} q(x, y, t; p) \sqrt{\det g_\Psi(x, y)} dx dy$

are the averages of q on the control-volumes;

- $u_{i+\frac{1}{2},j,p}^n = u(x_{i+\frac{1}{2}}, y_j, t_n; p);$
- $v_{i,j+\frac{1}{2},p}^n = v(x_i, y_{j+\frac{1}{2}}, t_n; p).$

Given a tangent velocity field \mathbf{u} on the sphere, we denote its contravariant components by \tilde{u} and \tilde{v} . For a give a detailed discussion on contravariant representations in Appendix B. The advection equation on panel the p of the cubed-sphere is given by:

$$\frac{\partial}{\partial t} q + \frac{1}{\sqrt{\det g_\Psi}} \left(\frac{\partial}{\partial x} (\tilde{u} \sqrt{\det g_\Psi} q) + \frac{\partial}{\partial y} (\tilde{v} \sqrt{\det g_\Psi} q) \right) = 0,$$

$\forall (x, y, t) \in [-a, a]^2 \times [0, T]$, $q = q(x, y, t; p)$. Its integral form is given by:

$$\begin{aligned} Q_{ijp}(t_{n+1}) &= Q_{ijp}(t_n) - \frac{\Delta t}{|\Omega_{ijp}|} \delta_x \left(\frac{1}{\Delta t} \int_{t_1}^{t_2} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} (\tilde{u} \sqrt{\det g_\Psi} q)(x_i, y, t; p) dy dt \right) \\ &\quad - \frac{\Delta t}{|\Omega_{ijp}|} \delta_y \left(\frac{1}{\Delta t} \int_{t_1}^{t_2} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} (\tilde{v} \sqrt{\det g_\Psi} q)(x, y_j, t; p) dx dt \right), \end{aligned}$$

Hence, we can use the dimension splitting presented in Chapter 3 to the variable $\sqrt{\det g_\Psi} q$. However, when computing the stencils near to the cube edges, we need to approximate the values of q in the ghost cells in order to compute the stencils.

Appendix A

Numerical Analysis

A.1 Finite-difference estimates

This Section aims to prove all finite-difference error estimations used throughout this appendix. All the proves are very simple and consist of applying Taylor's expansions, as it is usual when computing the accuracy order of many numerical schemes.

Lemma A.1. *Let $F \in C^5(\mathbb{R})$, $x_0 \in \mathbb{R}$ and $h > 0$. Then, the following identity holds:*

$$F'(x_0) = \frac{4}{3} \left(\frac{F(x_0 + h) - F(x_0 - h)}{2h} \right) - \frac{1}{3} \left(\frac{F(x_0 + 2h) - F(x_0 - 2h)}{4h} \right) + C_1 h^4, \quad (\text{A.1})$$

where C_1 is a constant that depends only on F and h .

Proof. Given $\delta \in]0, 2h]$, then $x_0 + \delta \in]x_0, x_0 + 2h]$ and $x_0 - \delta \in]x_0 - 2h, x_0]$. Then, we get using the Taylor expansion of F :

$$\begin{aligned} F(x_0 + \delta) &= F(x_0) + F'(x_0)\delta + F^{(2)}(x_0)\frac{\delta^2}{2} + F^{(3)}(x_0)\frac{\delta^3}{3!} + F^{(4)}(x_0)\frac{\delta^4}{4!} + F^{(5)}(\theta_\delta)\frac{\delta^5}{5!}, \quad \theta_\delta \in [x_0, x_0 + \delta], \\ F(x_0 - \delta) &= F(x_0) - F'(x_0)\delta + F^{(2)}(x_0)\frac{\delta^2}{2} - F^{(3)}(x_0)\frac{\delta^3}{3!} + F^{(4)}(x_0)\frac{\delta^4}{4!} - F^{(5)}(\theta_{-\delta})\frac{\delta^5}{5!}, \quad \theta_{-\delta} \in [x_0 - \delta, x_0]. \end{aligned}$$

Thus:

$$\frac{F(x_0 + \delta) - F(x_0 - \delta)}{2\delta} = F'(x_0) + F^{(3)}(x_0)\frac{\delta^2}{3!} + \left(F^{(5)}(\theta_\delta) + F^{(5)}(\theta_{-\delta}) \right) \frac{\delta^4}{2 \cdot 5!}, \quad (\text{A.2})$$

Applying Equation (A.2) for $\delta = h$ and $\delta = 2h$, we get, respectively:

$$\frac{F(x_0 + h) - F(x_0 - h)}{2h} = F'(x_0) + F^{(3)}(x_0)\frac{h^2}{3!} + \left(F^{(5)}(\theta_h) + F^{(5)}(\theta_{-h}) \right) \frac{h^4}{2 \cdot 5!}, \quad \theta_h \in [x_0, x_0 + h], \quad \theta_{-h} \in [x_0 - h, x_0], \quad (\text{A.3})$$

and

$$\frac{F(x_0 + 2h) - F(x_0 - 2h)}{4h} = F'(x_0) + F^{(3)}(x_0) \frac{4h^2}{3!} + \left(F^{(5)}(\theta_{2h}) + F^{(5)}(\theta_{-2h}) \right) \frac{16h^4}{2 \cdot 5!}, \quad (\text{A.4})$$

$$\theta_{2h} \in [x_0, x_0 + 2h], \quad \theta_{-2h} \in [x_0 - 2h, x_0].$$

Using Equations (A.3) and (A.4), we obtain:

$$\frac{4}{3} \left(\frac{F(x_0 + h) - F(x_0 - h)}{2h} \right) = \frac{4}{3} F'(x_0) + F^{(3)}(x_0) \frac{4h^2}{3 \cdot 3!} + \left(F^{(5)}(\theta_h) + F^{(5)}(\theta_{-h}) \right) \frac{h^4}{2 \cdot 5!}, \quad (\text{A.5})$$

$$\frac{1}{3} \left(\frac{F(x_0 + 2h) - F(x_0 - 2h)}{4h} \right) = \frac{1}{3} F'(x_0) + F^{(3)}(x_0) \frac{4h^2}{3 \cdot 3!} + \left(F^{(5)}(\theta_{2h}) + F^{(5)}(\theta_{-2h}) \right) \frac{16h^4}{3 \cdot 2 \cdot 5!} \quad (\text{A.6})$$

Subtracting Equation (A.6) from Equation (A.5) we get the desired Equation (A.1) with

$$C_1 = \frac{1}{720} \left(3F^{(5)}(\theta_h) + 3F^{(5)}(\theta_{-h}) - 16F^{(5)}(\theta_{2h}) - 16F^{(5)}(\theta_{-2h}) \right), \quad (\text{A.7})$$

where $\theta_h \in [x_0, x_0 + h]$, $\theta_{-h} \in [x_0 - h, x_0]$, $\theta_{2h} \in [x_0, x_0 + 2h]$, $\theta_{-2h} \in [x_0 - 2h, x_0]$. Using the intermediate value theorem, we can express C_1 in a more compact way as

$$C_1 = \frac{1}{720} \left(6F^{(5)}(\eta_1) - 32F^{(5)}(\eta_2) \right), \quad (\text{A.8})$$

where $\eta_1, \eta_2 \in [x_0 - 2h, x_0 + 2h]$, which concludes the proof. \square

Lemma A.2. *Let $F \in C^4(\mathbb{R})$, $x_0 \in \mathbb{R}$ and $h > 0$. Then, the following identity holds:*

$$F''(x_0) = \frac{-2F(x_0 - 2h) + 15F(x_0 - h) - 28F(x_0) + 20F(x_0 + h) - 6F(x_0 + 2h) + F(x_0 + 3h)}{6h^2} + C_2 h^2, \quad (\text{A.9})$$

where C_2 is a constant that depends only on F and h .

Proof. From the Taylor's expansion, we have:

$$\begin{aligned} F(x_0 - 2h) &= F(x_0) - 2F'(x_0)h + 2F^{(2)}(x_0)h^2 - \frac{8}{6}F^{(3)}(x_0)h^3 + \frac{16}{24}F^{(4)}(\theta_{-2h})h^4, \\ F(x_0 - h) &= F(x_0) - F'(x_0)h + \frac{1}{2}F^{(2)}(x_0)h^2 - \frac{1}{6}F^{(3)}(x_0)h^3 + \frac{1}{24}F^{(4)}(\theta_{-h})h^4, \\ F(x_0 + h) &= F(x_0) + F'(x_0)h + \frac{1}{2}F^{(2)}(x_0)h^2 + \frac{1}{6}F^{(3)}(x_0)h^3 + \frac{1}{24}F^{(4)}(\theta_h)h^4, \\ F(x_0 + 2h) &= F(x_0) + 2F'(x_0)h + 2F^{(2)}(x_0)h^2 + \frac{8}{6}F^{(3)}(x_0)h^3 + \frac{16}{24}F^{(4)}(\theta_{2h})h^4, \\ F(x_0 + 3h) &= F(x_0) + 3F'(x_0)h + \frac{9}{2}F^{(2)}(x_0)h^2 + \frac{27}{6}F^{(3)}(x_0)h^3 + \frac{81}{24}F^{(4)}(\theta_{3h})h^4, \end{aligned}$$

where $\theta_{-2h} \in [x_0 - 2h, x_0 - h]$, $\theta_{-h} \in [x_0 - h, x_0]$, $\theta_h \in [x_0, x_0 + h]$, $\theta_{2h} \in [x_0 + h, x_0 + 2h]$, $\theta_{3h} \in [x_0 + 2h, x_0 + 3h]$. Multiplying these equations by their respective coefficients given in Equation (A.9), one get:

$$\begin{aligned} -2F(x_0 - 2h) &= -2F(x_0) + 4F'(x_0)h - 4F^{(2)}(x_0)h^2 + \frac{16}{6}F^{(3)}(x_0)h^3 - \frac{32}{24}F^{(4)}(\theta_{-2h})h^4, \\ 15F(x_0 - h) &= 15F(x_0) - 15F'(x_0)h + \frac{15}{2}F^{(2)}(x_0)h^2 - \frac{15}{6}F^{(3)}(x_0)h^3 + \frac{15}{24}F^{(4)}(\theta_{-h})h^4, \\ -28F(x_0) &= -28F(x_0), \\ 20F(x_0 + h) &= 20F(x_0) + 20F'(x_0)h + 10F^{(2)}(x_0)h^2 + \frac{20}{6}F^{(3)}(x_0)h^3 + \frac{20}{24}F^{(4)}(\theta_h)h^4, \\ -6F(x_0 + 2h) &= -6F(x_0) - 12F'(x_0)h - 12F^{(2)}(x_0)h^2 - 8F^{(3)}(x_0)h^3 - \frac{96}{24}F^{(4)}(\theta_{2h})h^4, \\ F(x_0 + 3h) &= F(x_0) + 3F'(x_0)h + \frac{9}{2}F^{(2)}(x_0)h^2 + \frac{27}{6}F^{(3)}(x_0)h^3 + \frac{81}{24}F^{(4)}(\theta_{3h})h^4. \end{aligned}$$

Summing all these equations, we get the desired Formula (A.9) with C_2 given by:

$$C_2 = \frac{1}{24} \left(32F^{(4)}(\theta_{-2h}) - 15F^{(4)}(\theta_{-h}) - 20F^{(4)}(\theta_h) + 96F^{(4)}(\theta_{2h}) - 81F^{(4)}(\theta_{3h}) \right). \quad (\text{A.10})$$

Using the intermediate value theorem, we can express C_2 in a more compact way as

$$C_2 = \frac{1}{24} \left(128F^{(5)}(\eta_1) - 116F^{(5)}(\eta_2) \right), \quad (\text{A.11})$$

where $\eta_1, \eta_2 \in [x_0 - 2h, x_0 + 3h]$, which concludes the proof. \square

Lemma A.3. Let $F \in C^4(\mathbb{R})$, $x_0 \in \mathbb{R}$ and $h > 0$. Then, the following identity holds:

$$F^{(3)}(x_0) = \frac{F(x_0 - 2h) - 7F(x_0 - h) + 16F(x_0) - 16F(x_0 + h) + 7F(x_0 + 2h) - F(x_0 + 3h)}{2h^3} + C_3 h, \quad (\text{A.12})$$

where C_3 is a constant that depends only on F and h .

Proof. From the Taylor's expansion, we have:

$$\begin{aligned} F(x_0 - 2h) &= F(x_0) - 2F'(x_0)h + 2F^{(2)}(x_0)h^2 - \frac{8}{6}F^{(3)}(x_0)h^3 + \frac{16}{24}F^{(4)}(\theta_{-2h})h^4, \\ F(x_0 - h) &= F(x_0) - F'(x_0)h + \frac{1}{2}F^{(2)}(x_0)h^2 - \frac{1}{6}F^{(3)}(x_0)h^3 + \frac{1}{24}F^{(4)}(\theta_{-h})h^4, \\ F(x_0 + h) &= F(x_0) + F'(x_0)h + \frac{1}{2}F^{(2)}(x_0)h^2 + \frac{1}{6}F^{(3)}(x_0)h^3 + \frac{1}{24}F^{(4)}(\theta_h)h^4, \\ F(x_0 + 2h) &= F(x_0) + 2F'(x_0)h + 2F^{(2)}(x_0)h^2 + \frac{8}{6}F^{(3)}(x_0)h^3 + \frac{16}{24}F^{(4)}(\theta_{2h})h^4, \\ F(x_0 + 3h) &= F(x_0) + 3F'(x_0)h + \frac{9}{2}F^{(2)}(x_0)h^2 + \frac{27}{6}F^{(3)}(x_0)h^3 + \frac{81}{24}F^{(4)}(\theta_{3h})h^4, \end{aligned}$$

where $\theta_{-2h} \in [x_0 - 2h, x_0 - h]$, $\theta_{-h} \in [x_0 - h, x_0]$, $\theta_h \in [x_0, x_0 + h]$, $\theta_{2h} \in [x_0 + h, x_0 + 2h]$, $\theta_{3h} \in [x_0 + 2h, x_0 + 3h]$. Multiplying these equations by their respective coefficients given in Equation (A.12), one get:

$$\begin{aligned} F(x_0 - 2h) &= F(x_0) - 2F'(x_0)h + \frac{4}{2}F^{(2)}(x_0)h^2 - \frac{8}{6}F^{(3)}(x_0)h^3 + \frac{16}{24}F^{(4)}(\theta_{-2h})h^4, \\ -7F(x_0 - h) &= -7F(x_0) + 7F'(x_0)h - \frac{7}{2}F^{(2)}(x_0)h^2 + \frac{7}{6}F^{(3)}(x_0)h^3 - \frac{7}{24}F^{(4)}(\theta_{-h})h^4, \\ 16F(x_0) &= 16F(x_0), \\ -16F(x_0 + h) &= -16F(x_0) - 16F'(x_0)h - \frac{16}{2}F^{(2)}(x_0)h^2 - \frac{16}{6}F^{(3)}(x_0)h^3 - \frac{16}{24}F^{(4)}(\theta_h)h^4, \\ 7F(x_0 + 2h) &= 7F(x_0) + 14F'(x_0)h + \frac{28}{2}F^{(2)}(x_0)h^2 + \frac{56}{6}F^{(3)}(x_0)h^3 + \frac{112}{24}F^{(4)}(\theta_{2h})h^4, \\ -F(x_0 + 3h) &= -F(x_0) - 3F'(x_0)h - \frac{9}{2}F^{(2)}(x_0)h^2 - \frac{27}{6}F^{(3)}(x_0)h^3 - \frac{81}{24}F^{(4)}(\theta_{3h})h^4. \end{aligned}$$

Summing all these equations, we have:

$$F(x_0 - 2h) - 7F(x_0 - h) + 16F(x_0) - 16F(x_0 + h) + 7F(x_0 + 2h) - F(x_0 + 3h) = 2F^{(3)}(x_0)h^3 - 2C_3 h^4,$$

we get the desired Formula (A.12) with C_3 given by:

$$C_3 = \frac{1}{48} \left(-16F^{(4)}(\theta_{-2h}) + 7F^{(4)}(\theta_{-h}) + 16F^{(4)}(\theta_h) - 112F^{(4)}(\theta_{2h}) + 81F^{(4)}(\theta_{3h}) \right). \quad (\text{A.13})$$

Using the intermediate value theorem, we can express C_3 in a more compact way as

$$C_3 = \frac{1}{48} \left(104F^{(5)}(\eta_1) - 128F^{(5)}(\eta_2) \right), \quad (\text{A.14})$$

where $\eta_1, \eta_2 \in [x_0 - 2h, x_0 + 3h]$, which concludes the proof. \square

A.2 Lagrange interpolation

Given real numbers, called nodes, $x_0 < x_1 < \dots < x_m$, we define the k -th Lagrange polynomial by

$$L_k(x) = \prod_{j=0, j \neq k}^m \frac{x - x_j}{x_k - x_j}.$$

They satisfy $L_k(x_j) = \delta_{kj}$, where δ_{kj} is the Kronecker delta. Given a function f defined at the nodes x_j , its interpolating polynomial of degree m is given by:

$$P_m(x) = \sum_{k=0}^m f(x_k) L_k(x).$$

Indeed, this polynomial interpolates f since $P_m(x_j) = f(x_j)$. It is well known that P_m always exists and is unique. Besides that, we have the following error formula for Lagrange interpolation.

Theorem A.1. *Let $f \in C^{m+1}(\mathbb{R})$. Then, there is ξ in the smallest interval containing x_0, \dots, x_m, x such that:*

$$f(x) - P_m(x) = \omega(x) \frac{f^{(m+1)}(\xi)}{(m+1)!}, \quad (\text{A.15})$$

where $\omega(x) = (x - x_0)(x - x_1) \dots (x - x_m)$.

Proof. See Stoer and Bulirsch (2002, Theorem 2.1.4.1. on p. 49). \square

A.3 Numerical integration

The following mean value theorem for integrals is a very useful tool when working with numerical integration errors.

Theorem A.2 (Mean value theorem for integrals). *If $f \in C([a, b])$, and g is a integrable function in $[a, b]$ whose sign does not change in $[a, b]$, then there exists $c \in]a, b[$ such that*

$$\int_a^b f(x)g(x) dx = f(c) \int_a^b g(x) dx.$$

Proof. See Courant and John (1999, p. 143). \square

A.3.1 Midpoint rule

When considering finite-volume schemes, it is useful to compare the average value on a control volume of a function with its value at the control volume centroid. In the following theorems, for the one and two dimensional cases, respectively, we show that the value of a function at the centroid of a control volume given a second-order approximation to its average value on the control volume.

Theorem A.3. *If $f \in C^2([x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}])$, then*

$$\frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} f(x) dx - f(x_i) = C_1 \Delta x^2, \quad (\text{A.16})$$

where C_1 is a constant that depends only on f , and $x_i = \frac{x_{i+\frac{1}{2}} + x_{i-\frac{1}{2}}}{2}$, $\Delta x = x_{i+\frac{1}{2}} - x_{i-\frac{1}{2}}$.

Proof. From Taylor's expansion, it follows that, for $x \in [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}]$, we have:

$$f(x) = f(x_i) + f'(x_i)(x - x_i) + f''(\xi) \frac{(x - x_i)^2}{2}, \quad (\text{A.17})$$

for some ξ between x and x_i . Therefore:

$$\begin{aligned} \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} f(x) dx - f(x_i) &= \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \left(f'(x_i)(x - x_i) + f''(\xi) \frac{(x - x_i)^2}{2} \right) dx \\ &= \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} f''(\xi) \frac{(x - x_i)^2}{2} dx. \end{aligned}$$

Using the mean value theorem for integrals (see Theorem A.2), we have:

$$\frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} f(x) dx - f(x_i) = f''(\eta) \frac{1}{\Delta x} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \frac{(x - x_i)^2}{2} dx = f''(\eta) \frac{\Delta x^2}{24}$$

for some $\eta \in [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}]$, from which the proposition follows with

$$C_1 = \frac{1}{24} f''(\eta). \quad (\text{A.18})$$

□

Theorem A.4. *If $f \in C^2([x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}] \times [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}])$, then*

$$\left| \frac{1}{\Delta x \Delta y} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} f(x, y) dx dy - f(x_i, y_j) \right| \leq C_1 \Delta x^2 + C_2 \Delta x \Delta y + C_3 \Delta y^2, \quad (\text{A.19})$$

where C_1, C_2 and C_3 are constants that depend only on f , and $x_i = \frac{x_{i+\frac{1}{2}} + x_{i-\frac{1}{2}}}{2}$, $y_i = \frac{y_{j+\frac{1}{2}} + y_{j-\frac{1}{2}}}{2}$, $\Delta x = x_{i+\frac{1}{2}} - x_{i-\frac{1}{2}}$, $\Delta y = y_{j+\frac{1}{2}} - y_{j-\frac{1}{2}}$.

Proof. From Taylor's expansion, it follows that, for $x \in [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}]$, $y \in [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}]$, we have:

$$\begin{aligned} f(x, y) &= f(x_i, y_j) + \frac{\partial f}{\partial x}(x_i, y_j)(x - x_i) + \frac{\partial f}{\partial y}(x_i, y_j)(y - y_j) \\ &\quad + \frac{1}{2} \left(\frac{\partial^2 f}{\partial x^2}(\xi, \theta)(x - x_i)^2 + 2 \frac{\partial^2 f}{\partial x \partial y}(\xi, \theta)(x - x_i)(y - y_j) \frac{\partial^2 f}{\partial y^2}(\xi, \theta)(y - y_j)^2 \right) \end{aligned}$$

for some ξ between x and x_i , and θ between y and y_j . Therefore:

$$\begin{aligned} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} f(x, y) dy dx - \Delta x \Delta y f(x_i, y_j) &= \frac{1}{2} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \frac{\partial^2 f}{\partial x^2}(\xi, \theta)(x - x_i)^2 dy dx + \\ \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \frac{\partial^2 f}{\partial x \partial y}(\xi, \theta)(x - x_i)(y - y_j) dy dx + \frac{1}{2} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \frac{\partial^2 f}{\partial y^2}(\xi, \theta)(y - y_j)^2 dy dx \end{aligned}$$

Using the mean value theorem for integrals (see Theorem A.2), we have:

$$\begin{aligned} \frac{1}{\Delta x \Delta y} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} f(x, y) dx dy - f(x_i, y_j) &= \frac{\partial^2 f}{\partial x^2}(\eta_1, \lambda_1) \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \frac{(x - x_i)^2}{2 \Delta x \Delta y} dx dy \\ &\quad + \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \frac{\partial^2 f}{\partial x \partial y}(\xi, \theta) \frac{(x - x_i)(y - y_j)}{\Delta x \Delta y} dx dy + \frac{\partial^2 f}{\partial x^2}(\eta_2, \lambda_2) \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \frac{(y - y_j)^2}{2 \Delta x \Delta y} dx dy \\ &= \frac{\partial^2 f}{\partial x^2}(\eta_1, \lambda_1) \frac{\Delta x^2}{24} + \frac{\partial^2 f}{\partial y^2}(\eta_2, \lambda_2) \frac{\Delta y^2}{24} + \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \int_{y_{j-\frac{1}{2}}}^{y_{j+\frac{1}{2}}} \frac{\partial^2 f}{\partial x \partial y}(\xi, \theta) \frac{(x - x_i)(y - y_j)}{\Delta x \Delta y} dx dy \end{aligned}$$

for $\eta_1, \eta_2 \in [x_{i-\frac{1}{2}}, x_{i+\frac{1}{2}}]$, $\lambda_1, \lambda_2 \in [y_{j-\frac{1}{2}}, y_{j+\frac{1}{2}}]$, from which the proposition follows. \square

A.3.2 Multi-step schemes

Let us consider the following problem: given a function $f \in C^{m+1}([0, T])$, a discretization of $[0, T]$ given by $t^n = n\Delta t$, $\Delta t = \frac{T}{N_T}$, for some $N_T \in \mathbb{N}$, we wish to estimate $\int_{t^n}^{t^{n+1}} f(t) dt$ using the values $f(t_{n-k})$, for $k = 0, \dots, m$. This kind of problem arises, for instance, when we are interested in computing departure points as in Equation 2.11. We can estimate the desired integral by computing the interpolating polynomial of $f(t_{n-k})$, for $k = 0, \dots, m$ and then integrating this polynomial. This approach is exactly what is used in multi-step Adams-Basforth methods. On the next theorem, we give an expression the error of this approach.

Theorem A.5. If $f \in C^{m+1}([0, T])$, $t^n = n\Delta t$, $n = 0, \dots, N_T$, $\Delta t = \frac{T}{N_T}$ for some $N_T \in \mathbb{N}$, then:

$$\int_{t^n}^{t^{n+1}} f(t) dt = \Delta t \sum_{k=0}^m \left(\int_0^1 L_k(s) ds \right) f(t_{n-k}) + \frac{(\Delta t)^{k+1}}{(m+1)!} f^{(m+1)}(\eta) \int_0^1 \omega(s) ds, \quad (\text{A.20})$$

where $\omega(s) = s(s+1) \cdots (s+m)$, $\eta \in [t^{n-m}, t^n]$.

Proof. We introduce auxiliary functions $\theta(s) = s\Delta t + t_n$, $s \in [-m, 1]$ and $g(s) = f(\theta(s))$. It

is clear that $f(t_{n-k}) = g(-k)$, for $k = -1, 0, \dots, m$. Hence, we can write:

$$\int_{t^n}^{t^{n+1}} f(t) dt = \Delta t \int_0^1 f(\theta(s)) ds = \Delta t \int_0^1 g(s) ds. \quad (\text{A.21})$$

Defining the nodes $s_k = -k$ for $k = 0, \dots, m$, it follows from Theorem A.1 that the interpolating polynomial P_m of $g(s_k)$ satisfies:

$$g(s) - P_m(s) = \omega(s) \frac{g^{(m+1)}(\xi)}{(m+1)!}, \quad (\text{A.22})$$

where $\xi \in [-m, 1]$. Substituting Equation (A.22) in Equation (A.21), we obtain

$$\int_{t^n}^{t^{n+1}} f(t) dt = \Delta t \sum_{k=0}^m \left(\int_0^1 L_k(s) ds \right) g(-k) + \frac{\Delta t}{(m+1)!} \int_0^1 g^{(m+1)}(\xi) \omega(s) ds. \quad (\text{A.23})$$

Since $\omega(s)$ does not change its sign in $[0, 1]$ it follows from Theorem A.2 that:

$$\int_{t^n}^{t^{n+1}} f(t) dt = \Delta t \sum_{k=0}^m \left(\int_0^1 L_k(s) ds \right) g(-k) + \frac{\Delta t}{(m+1)!} g^{(m+1)}(\bar{\xi}) \int_0^1 \omega(s) ds, \quad (\text{A.24})$$

for some $\bar{\xi} \in [-m, 1]$. Notice that by the chain rule we get $g^{(m+1)}(s) = (\Delta t)^k f^{(m+1)}(\theta(s))$, therefore Equation (A.24) in terms of f reads:

$$\int_{t^n}^{t^{n+1}} f(t) dt = \Delta t \sum_{k=0}^m \left(\int_0^1 L_k(s) ds \right) f(t_{n-k}) + \frac{(\Delta t)^{k+1}}{(m+1)!} f^{(m+1)}(\eta) \int_0^1 \omega(s) ds, \quad (\text{A.25})$$

where $\eta \in [t^{n-m}, t^n]$, which is the desired identity. \square

In the following corollaries, we give the explicit formulas for Equation (A.25) for $m = 0, m = 1, m = 2$. This is achieved by computing the terms $\int_0^1 L_k(s) ds$ and $\int_0^1 \omega(s) ds$, which are trivial to be computed.

Corollary A.1. *If $f \in C^1([0, T])$, $t^n = n\Delta t$, $n = 0, \dots, N_T$, $\Delta t = \frac{T}{N_T}$ for some $N_T \in \mathbb{N}$, then:*

$$\int_{t^n}^{t^{n+1}} f(t) dt = \Delta t f(t_n) + \frac{\Delta t^2}{2} f'(\bar{t}), \quad (\text{A.26})$$

for some $\bar{t} \in [t^n, t^{n+1}]$.

Corollary A.2. *If $f \in C^2([0, T])$, $t^n = n\Delta t$, $n = 0, \dots, N_T$, $\Delta t = \frac{T}{N_T}$ for some $N_T \in \mathbb{N}$, then:*

$$\int_{t^n}^{t^{n+1}} f(t) dt = \frac{\Delta t}{2} (3f(t_n) - f(t_{n-1})) + \frac{5\Delta t^3}{12} f^{(2)}(\bar{t}), \quad (\text{A.27})$$

for some $\bar{t} \in [t^{n-1}, t^{n+1}]$.

Corollary A.3. If $f \in C^3([0, T])$, $t^n = n\Delta t$, $n = 0, \dots, N_T$, $\Delta t = \frac{T}{N_T}$ for some $N_T \in \mathbb{N}$, then:

$$\int_{t^n}^{t^{n+1}} f(t) dt = \frac{\Delta t}{12}(23f(t_n) - 16f(t_{n-1}) + 5f(t_{n-2})) + \frac{3\Delta t^4}{8}f^{(3)}(\bar{t}), \quad (\text{A.28})$$

for some $\bar{t} \in [t^{n-2}, t^{n+1}]$.

When using these schemes for and ODE written in its integral form, $m = 0$ gives the classical Euler method; for $m = 1$ we get the second-order Adams-Bashforth scheme and for $m = 2$ we have the third-order Adams-Bashforth scheme.

A.4 PPM reconstruction accuracy analysis

In this Section, we are going to investigate the accuracy of the PPM reconstruction process. As we pointed out in Section 2.4.1, the approximation of q at the control volumes edges given by Equation (2.46) is fourth-order accurate when $q \in C^4(\mathbb{R})$. This is proved as a Corollary of the following Proposition A.1.

Proposition A.1. Let $q \in C^4(\mathbb{R})$, $\bar{x} \in \mathbb{R}$ and $h > 0$. Then, the following identity holds:

$$q(\bar{x}) = \frac{7}{12} \left(\frac{1}{h} \int_{\bar{x}}^{\bar{x}+h} q(x) dx + \frac{1}{h} \int_{\bar{x}-h}^{\bar{x}} q(x) dx \right) - \frac{1}{12} \left(\frac{1}{h} \int_{\bar{x}+h}^{\bar{x}+2h} q(x) dx + \frac{1}{h} \int_{\bar{x}-2h}^{\bar{x}-h} q(x) dx \right) + C_1 h^4, \quad (\text{A.29})$$

where C_1 is a constant that depends on q and h .

Proof. We define $Q(x) = \int_a^x q(\xi) d\xi$ for fixed $a \in \mathbb{R}$ as in Equation (2.37). It follows that:

$$\begin{aligned} \int_{\bar{x}}^{\bar{x}+h} q(\xi) d\xi + \int_{\bar{x}-h}^{\bar{x}} q(\xi) d\xi &= Q(\bar{x} + h) - Q(\bar{x} - h), \\ \int_{\bar{x}+h}^{\bar{x}+2h} q(\xi) d\xi + \int_{\bar{x}-2h}^{\bar{x}-h} q(\xi) d\xi &= Q(\bar{x} + 2h) - Q(\bar{x} - 2h) - (Q(\bar{x} + h) - Q(\bar{x} - h)). \end{aligned}$$

Using these identities, Equation (A.29) may be rewritten as:

$$q(\bar{x}) = \frac{4}{3} \left(\frac{Q(\bar{x} + h) - Q(\bar{x} - h)}{2h} \right) - \frac{1}{3} \left(\frac{Q(\bar{x} + 2h) - Q(\bar{x} - 2h)}{4h} \right) + C_1 h^4, \quad (\text{A.30})$$

which consists of finite-difference approximations. Thus, Equation (A.29) follows from Lemma A.1 with:

$$C_1 = C_1(\mu_1, \mu_2) = \frac{1}{720} \left(6q^{(4)}(\mu_1) - 32q^{(4)}(\mu_2) \right), \quad (\text{A.31})$$

where $\mu_1, \mu_2 \in [\bar{x} - 2h, \bar{x} + 2h]$, which concludes the proof. \square

Corollary A.4. It follows from Proposition A.1 with $\bar{x} = x_{i+\frac{1}{2}}$ and $h = \Delta x$ that $q_{i+\frac{1}{2}}$ given by

Equation (2.46) satisfies:

$$q(x_{i+\frac{1}{2}}) - q_{i+\frac{1}{2}} = C_1 \Delta x^4, \quad (\text{A.32})$$

with C_1 given by Equation (A.31), whenever $q \in C^4(\mathbb{R})$.

Remark A.1. Similarly, one can show that the formulas are given by Equation (2.49) and Equation (2.49) are fifth-order accurate.

The parabolic function from (2.40) given with coefficients specified before approximates q with order 3 when $q \in C^4(\mathbb{R})$. In order to check this, for $x \in X_i$ we rewrite Equation (2.40) as:

$$q_i(x; Q) = q_{L,i} + \frac{(\Delta q_i + q_{6,i})}{\Delta x} (x - x_{i-\frac{1}{2}}) - \frac{q_{6,i}}{\Delta x^2} (x - x_{i-\frac{1}{2}})^2 \quad (\text{A.33})$$

and we write q using its Taylor expansion assuming $q \in C^4(\mathbb{R})$:

$$q(x) = q(x_{i-\frac{1}{2}}) + q'(x_{i-\frac{1}{2}})(x - x_{i-\frac{1}{2}}) + \frac{q''(x_{i-\frac{1}{2}})}{2}(x - x_{i-\frac{1}{2}})^2 + \frac{q^{(3)}(\theta_i)}{6}(x - x_{i-\frac{1}{2}})^3, \quad (\text{A.34})$$

where $\theta_i \in X_i$. Comparing Equation (A.33) with Equation (A.34), it is reasonable to seek to some bound to the expressions:

$$q'(x_{i-\frac{1}{2}}) - \frac{(\Delta q_i + q_{6,i})}{\Delta x}, \quad (\text{A.35})$$

and:

$$\frac{q''(x_{i-\frac{1}{2}})}{2} - \left(-\frac{q_{6,i}}{\Delta x^2} \right). \quad (\text{A.36})$$

We have seen that term $q_{L,i}$ gives a fourth-order approximation to $q(x_{i-\frac{1}{2}})$. The Corollary A.5 shall prove that the term (A.35) has a bound proportional to Δx^2 , and the Corollary A.6 shall prove that the term (A.36) is bounded by a constant times Δx .

Before proving the desired bounds, it is useful to rewrite some terms explicitly as functions of the values of the Δx -grid function Q . Combining Equation (2.43) with Equations (2.47) and (2.48), we may write $q_{6,i}$ as:

$$q_{6,i} = \frac{1}{4} \left(Q_{i-2} - 6Q_{i-1} + 10Q_i - 6Q_{i+1} + Q_{i+2} \right). \quad (\text{A.37})$$

Recalling the definition of Δq_i from Equation (2.41), and applying Equations (2.47) and (2.48), we may express Δq_i as:

$$\Delta q_i = \frac{1}{12} \left(Q_{i-2} - 8Q_{i-1} + 8Q_{i+1} - Q_{i+2} \right). \quad (\text{A.38})$$

Finally, we combine Equations (A.37) and (A.38) and write their sum as:

$$\frac{(\Delta q_i + q_{6,i})}{\Delta x} = \frac{2Q_{i-2} - 13Q_{i-1} + 15Q_i - 5Q_{i+1} + Q_{i+2}}{6\Delta x}. \quad (\text{A.39})$$

The next Proposition A.2 proves that Equation (A.39) approximates $q'(x_{i-\frac{1}{2}})$ with order

2.

Proposition A.2. Let $q \in C^3(\mathbb{R})$, $\bar{x} \in \mathbb{R}$ and $h > 0$. Then, the following identity holds:

$$q'(\bar{x}) = \frac{1}{6h} \left(\frac{2}{h} \int_{\bar{x}-2h}^{\bar{x}-h} q(x) dx - \frac{13}{h} \int_{\bar{x}-h}^{\bar{x}} q(x) dx + \frac{15}{h} \int_{\bar{x}}^{\bar{x}+h} q(x) dx - \frac{5}{h} \int_{\bar{x}+h}^{\bar{x}+2h} q(x) dx + \frac{1}{h} \int_{\bar{x}+2h}^{\bar{x}+3h} q(x) dx \right) + C_2 h^2, \quad (\text{A.40})$$

where C_2 is a constant that depends on q and h .

Proof. We consider again $Q(x) = \int_a^x q(\xi) d\xi$ for $a \in \mathbb{R}$ fixed as in Equation (2.37). Like in Proposition A.2, we have:

$$\begin{aligned} & \frac{1}{6h} \left(\frac{2}{h} \int_{\bar{x}-2h}^{\bar{x}-h} q(x) dx - \frac{13}{h} \int_{\bar{x}-h}^{\bar{x}} q(x) dx + \frac{15}{h} \int_{\bar{x}}^{\bar{x}+h} q(x) dx - \frac{5}{h} \int_{\bar{x}+h}^{\bar{x}+2h} q(x) dx + \frac{1}{h} \int_{\bar{x}+2h}^{\bar{x}+3h} q(x) dx \right) \\ &= \frac{1}{6h} \left(\frac{2}{h} (Q(\bar{x} - h) - Q(\bar{x} - 2h)) - \frac{13}{h} (Q(\bar{x}) - Q(\bar{x} - h)) + \frac{15}{h} (Q(\bar{x} + h) - Q(\bar{x})) \right. \\ &\quad \left. - \frac{5}{h} (Q(\bar{x} + 2h) - Q(\bar{x} + h)) + \frac{1}{h} (Q(\bar{x} + 3h) - Q(\bar{x} + 2h)) \right) \\ &= \frac{1}{6h^2} \left(-2Q(\bar{x} - 2h) + 15Q(\bar{x} - h) - 28Q(\bar{x}) + 20Q(\bar{x} + h) - 6Q(\bar{x} + 2h) + Q(\bar{x} + 3h) \right), \end{aligned}$$

which consists of the finite-difference scheme from Lemma A.2. Therefore, Equation (A.40) follows from Lemma A.2 with:

$$C_2 = C_2(\mu_1, \mu_2) = \frac{1}{24} \left(128q^{(3)}(\mu_1) - 116q^{(3)}(\mu_2) \right), \quad (\text{A.41})$$

where $\mu_1, \mu_2 \in [x_0 - 2h, x_0 + 3h]$, which concludes the proof. \square

Corollary A.5. It follows from Proposition A.2 with $\bar{x} = x_{i-\frac{1}{2}}$ and $h = \Delta x$ that Δq_i given by Equation (A.38) and $q_{6,i}$ given by Equation (A.37) satisfy:

$$q'(x_{i-\frac{1}{2}}) - \frac{(\Delta q_i + q_{6,i})}{\Delta x} = C_2 \Delta x^2, \quad (\text{A.42})$$

with C_2 given by Equation (A.41), whenever $q \in C^3(\mathbb{R})$.

Now, we analyse the following expression:

$$-\frac{2q_{6,i}}{\Delta x^2} = -\frac{1}{2\Delta x^2} \left(Q_{i-2} - 6Q_{i-1} + 10Q_i - 6Q_{i+1} + Q_{i+2} \right). \quad (\text{A.43})$$

deduced from Equation (A.37) and we prove in Proposition A.3 that Equation (A.43) approximates $q''(x_{i-\frac{1}{2}})$ with order 1.

Proposition A.3. Let $q \in C^3(\mathbb{R})$, $\bar{x} \in \mathbb{R}$ and $h > 0$. Then, the following identity holds:

$$\begin{aligned} q''(\bar{x}) = \frac{1}{2h^2} \left(-\frac{1}{h} \int_{\bar{x}-2h}^{\bar{x}-h} q(x) dx + \frac{6}{h} \int_{\bar{x}-h}^{\bar{x}} q(x) dx - \frac{10}{h} \int_{\bar{x}}^{\bar{x}+h} q(x) dx \right. \\ \left. + \frac{6}{h} \int_{\bar{x}+h}^{\bar{x}+2h} q(x) dx - \frac{1}{h} \int_{\bar{x}+2h}^{\bar{x}+3h} q(x) dx \right) + C_3 h, \end{aligned} \quad (\text{A.44})$$

where C_3 is a constant that depends on q and h .

Proof. Similarly to Proposition A.2 using the same function Q , we have:

$$\begin{aligned} \frac{1}{2h^2} \left(-\frac{1}{h} \int_{\bar{x}-2h}^{\bar{x}-h} q(x) dx + \frac{6}{h} \int_{\bar{x}-h}^{\bar{x}} q(x) dx - \frac{10}{h} \int_{\bar{x}}^{\bar{x}+h} q(x) dx + \frac{6}{h} \int_{\bar{x}+h}^{\bar{x}+2h} q(x) dx - \frac{1}{h} \int_{\bar{x}+2h}^{\bar{x}+3h} q(x) dx \right) \\ = \frac{1}{2h^2} \left(-\frac{1}{h} (Q(\bar{x} - h) - Q(\bar{x} - 2h)) + \frac{6}{h} (Q(\bar{x}) - Q(\bar{x} - h)) - \frac{10}{h} (Q(\bar{x} + h) - Q(\bar{x})) \right. \\ \left. + \frac{6}{h} (Q(\bar{x} + 2h) - Q(\bar{x} + h)) - \frac{1}{h} (Q(\bar{x} + 3h) - Q(\bar{x} + 2h)) \right) \\ = \frac{1}{2h^3} \left(Q(\bar{x} - 2h) - 7Q(\bar{x} - h) + 16Q(\bar{x}) - 16Q(\bar{x} + h) + 7Q(\bar{x} + 2h) - Q(\bar{x} + 3h) \right), \end{aligned}$$

which consists of the finite-difference scheme from Lemma A.3. Therefore, Equation (A.44) follows from Lemma A.3 with:

$$C_3 = C_3(\mu_1, \mu_2) = \frac{1}{48} \left(104q^{(3)}(\mu_1) - 128q^{(3)}(\mu_2) \right), \quad (\text{A.45})$$

where $\mu_1, \mu_2 \in [x_0 - 2h, x_0 + 3h]$, which concludes the proof. \square

Corollary A.6. It follows from Proposition A.3 with $\bar{x} = x_{i-\frac{1}{2}}$ and $h = \Delta x$ that $q_{6,i}$ given by Equation (2.46) satisfies:

$$q''(x_{i-\frac{1}{2}}) - \left(-\frac{2q_{6,i}}{\Delta x^2} \right) = C_3 \Delta x, \quad (\text{A.46})$$

with C_3 given by Equation (A.45), whenever $q \in C^3(\mathbb{R})$.

With the aid of Corollaries A.4, A.5, and A.6, we are able to prove that the PPM reconstruction approximates q with order 3. Indeed, we prove this on the follow up Proposition A.4.

Proposition A.4. Let $q \in C^4([a, b])$. Then, the Piecewise-Parabolic function given by Equation (2.40) with the parameters $q_{R,i}$ and $q_{L,i}$ obeying Equations (2.47) and (2.48) gives a third-order approximation to q on the control volume X_i . Namely, there exist constants M_1 and M_2 such that

$$|q(x) - q_i(x; Q)| \leq M_1 \Delta x^4 + M_2 \Delta x^3, \quad \forall x \in X_i.$$

Proof. For $x \in X_i$, from Equations (A.34) and (A.33), we have:

$$\begin{aligned} q(x) - q_i(x; Q) &= (q'(x_{i-\frac{1}{2}}) - q_{L,i}) + \left(q'(x_{i-\frac{1}{2}}) - \frac{(\Delta q_i + q_{6,i})}{\Delta x} \right) (x - x_{i-\frac{1}{2}}) \\ &\quad + \left(\frac{q''(x_{i-\frac{1}{2}})}{2} + \frac{q_{6,i}}{\Delta x^2} \right) (x - x_{i-\frac{1}{2}})^2 + \frac{q^{(3)}(\theta_i)}{6} (x - x_{i-\frac{1}{2}})^3. \end{aligned}$$

Using this fact with Corollaries A.4, A.5, and A.6, we have:

$$q(x) - q_i(x; Q) = C_1 \Delta x^4 + C_2 \Delta x^2 (x - x_{i-\frac{1}{2}}) + \frac{C_3}{2} \Delta x (x - x_{i-\frac{1}{2}})^2 + C_4 (x - x_{i-\frac{1}{2}})^3,$$

where C_1, C_2 and C_3 are given by Equations (A.31), (A.41) and (A.45), respectively, and

$$C_4 = C_4(\theta_i) = \frac{q^{(3)}(\theta_i)}{6}. \quad (\text{A.47})$$

For $x \in X_i$, we have $|x - x_{i-\frac{1}{2}}| \leq \Delta x$, thus:

$$|q(x) - q_i(x; Q)| \leq M_1 \Delta x^4 + M_2 \Delta x^3,$$

where

$$\begin{aligned} M_1 &= \frac{38}{720} \sup_{\xi \in [a,b]} |q^{(4)}(\xi)|, \\ M_2 &= \left(\frac{244}{24} + \frac{232}{96} + \frac{1}{6} \right) \sup_{\xi \in [a,b]} |q^{(3)}(\xi)| = \frac{143}{12} \sup_{\xi \in [a,b]} |q^{(3)}(\xi)|, \end{aligned}$$

which concludes the proof. \square

Remark A.2. Replacing the formulas for $q_{R,i}$ and $q_{L,i}$ given by Equations (2.47) and (2.48) by the formulas given by Equations (2.49) and (2.50), does not change the order of convergence of the parabolic approximation.

A.5 Convergence of 1D FV-SL schemes

A.5.1 Consistency and convergence

Hereafter, we are going to use the notations introduced in Section 2.1.1. To move towards the convergence of 1D-FV schemes, for Problem 2.4 we introduce the local truncation error (LTE hereafter) τ_i^n following LeVeque (2002):

$$Q_i(t^{n+1}) = Q_i(t^n) - \lambda \left(\mathcal{F}(Q(t^n)(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n) - \mathcal{F}(Q(t^n)(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n) \right) + \Delta t \tau_i^n. \quad (\text{A.48})$$

Notice the LTE is obtained by replacing the exact solution in Equation (2.20). Since $Q_i(t^n)$ is the exact solution of Equation (2.9), the LTE may be rewritten as

$$\tau_i^n = \frac{1}{\Delta x} \left[\left(\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} (uq)(x_{i+\frac{1}{2}}, t) dt - \mathcal{F}(Q(t^n))(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n \right) + \left(\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} (uq)(x_{i-\frac{1}{2}}, t) dt - \mathcal{F}(Q(t^n))(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n \right) \right]. \quad (\text{A.49})$$

The LTE gives a measure of how well the 1D-FV scheme approximates the integral form of the considered conservation law. Another interpretation of the LTE is that the LTE gives the error obtained after applying the scheme for a single time-step using the exact solution. The 1D-FV scheme is said to be consistent if the LTE converges to zero.

We define $\tau^n = (\tau_1^n, \dots, \tau_N^n)$, which represent the LTES at the time-step n . Now we can define consistency.

Definition A.1 (Consistency). *Let us consider the framework of Problem 2.4. A 1D-FV scheme is said to be consistency in the p -norm if for any sequence of $(\Delta x^{(k)}, \Delta t^{(k)}, \lambda)$ -discretizations, $k \in \mathbb{N}$, with $\lim_{k \rightarrow \infty} \Delta x^{(k)} = \lim_{k \rightarrow \infty} \Delta t^{(k)} = 0$, we have:*

$$\lim_{k \rightarrow \infty} \left[\max_{1 \leq n \leq N_T^{(k)}} \|\tau^n\|_{p, \Delta x^{(k)}} \right] = 0,$$

and it is said to be consistent with order P in the p -norm if

$$\max_{1 \leq n \leq N_T^{(k)}} \|\tau^n\|_{p, \Delta x^{(k)}} = O(\Delta x^P).$$

From Equation (A.49), it follows that we basically need to ensure that the numerical flux function \mathcal{F} converges to the time-averaged flux at edges when $\Delta x \rightarrow 0$ in order to guarantee consistency. In Section 2.5 we shall address how the numerical flux from PPM approximates the time-averaged flux at edges.

At last, we define the point-wise error at time-step n by:

$$E_i^n = Q_i(t^n) - Q_i^n, \quad i = 1, \dots, N,$$

and we define the vector of errors by $E^n = (E_1^n, \dots, E_N^n)$.

Definition A.2 (Convergence). *Let us consider the framework of Problem 2.4. A 1D-FV scheme is said to be convergent in the p -norm if for any sequence of $(\Delta x^{(k)}, \Delta t^{(k)}, \lambda)$ -discretizations, $k \in \mathbb{N}$, with $\lim_{k \rightarrow \infty} \Delta x^{(k)} = \lim_{k \rightarrow \infty} \Delta t^{(k)} = 0$, we have:*

$$\lim_{k \rightarrow \infty} \left[\max_{1 \leq n \leq N_T^{(k)}} \|E^n\|_{p, \Delta x^{(k)}} \right] = 0,$$

and it is said to converge with order P in the p -norm if

$$\max_{1 \leq n \leq N_T^{(k)}} \|E^n\|_{p, \Delta x^{(k)}} = O(\Delta x^P).$$

Subtracting Equation (2.20) from Equation (A.48) we get the following equation for the error:

$$\begin{aligned} E_i^{n+1} &= E_i^n - \lambda \left[\left(\mathcal{F}(Q(t^n)(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n) - \mathcal{F}(Q^n(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n) \right) \right. \\ &\quad \left. - \left(\mathcal{F}(Q(t^n)(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n) - \mathcal{F}(Q^n(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n) \right) \right] + \tau_i^n \Delta t. \end{aligned} \quad (\text{A.50})$$

Notice that if $q, u \in C^3$, we can rewrite Equation A.49 as:

$$\tau_i^n = \left[\frac{1}{\Delta x \Delta t} \int_{t^n}^{t^{n+1}} \int_{x_{i-\frac{1}{2}}}^{x_{i+\frac{1}{2}}} \frac{\partial(uq)}{\partial x}(x, t) dx dt - \left(\frac{\mathcal{F}(Q(t^n)(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n)) - \mathcal{F}(Q(t^n)(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n)}{\Delta x} \right) \right].$$

Using the midpoint rule for integration (Theorem A.3) and the mean value theorem for integrals (Theorem A.2), we have:

$$\begin{aligned} \tau_i^n &= \left[\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} \left(\frac{\partial(uq)}{\partial x}(x_i, t) + \frac{\Delta x^2}{24} \frac{\partial^2(uq)}{\partial x^2}(\xi, t) \right) dt - \left(\frac{\mathcal{F}(Q(t^n)(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n)) - \mathcal{F}(Q(t^n)(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n)}{\Delta x} \right) \right] \\ &= \left[\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} \frac{\partial(uq)}{\partial x}(x_i, t) dt - \left(\frac{\mathcal{F}(Q(t^n)(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n)) - \mathcal{F}(Q(t^n)(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n)}{\Delta x} \right) \right] + \frac{\Delta x^2}{24} \frac{\partial^3(uq)}{\partial x^3}(\xi, \bar{t}), \end{aligned} \quad (\text{A.51})$$

for $\xi \in X_i$ and $\bar{t} \in [t^n, t^{n+1}]$. Therefore, if $q, u \in C^3$ the scheme is consistent, if and only if, $\frac{1}{\Delta t} \int_{t^n}^{t^{n+1}} \frac{\partial(uq)}{\partial x}(x_i, t) dt$ is approximated by $\frac{\mathcal{F}(Q(t^n)(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n)) - \mathcal{F}(Q(t^n)(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n)}{\Delta x}$. This shall be very useful when we consider two-dimensional schemes, where we are going to use the discrete operators to estimate the divergence of velocity fields.

A.5.2 Stability

In order to define the concept of stability, it is useful to introduce an operator representation of 1D-FV schemes. In the context of Problem 2.4, we define the operators $\mathcal{H}_{\Delta x, n} : \mathbb{R}^{\Delta x} \rightarrow \mathbb{R}^{\Delta x}$ whose i -th entry is given by:

$$[\mathcal{H}_{\Delta x, n}(Q)]_i = Q_i - \lambda \left(\mathcal{F}(Q(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n) - \mathcal{F}(Q(S_{i-\frac{1}{2}}), \tilde{u}_{i-\frac{1}{2}}^n) \right), \quad (\text{A.52})$$

for $i = 1, \dots, N$, $n = 0, \dots, N_T - 1$. Notice that the dependence on n is due to the velocity that may be allowed to vary with time. As it is usual, we are assuming periodicity in the entries of Q when we apply the operator $\mathcal{H}_{\Delta x, n}$. Thus, Equation (2.20) may be rewritten in a vector form by

$$Q^{n+1} = \mathcal{H}_{\Delta x, n}(Q^n),$$

and Equation (A.48) in a vector form reads

$$Q(t^{n+1}) = \mathcal{H}_{\Delta x, n}(Q(t^n)) + \Delta t \tau^n,$$

and the error equation (A.50) is given by

$$E^{n+1} = \mathcal{H}_{\Delta x, n}(Q(t^n)) - \mathcal{H}_{\Delta x, n}(Q^n) + \Delta t \tau^n. \quad (\text{A.53})$$

The stability theory focus on uniformly bounding the norm of $\mathcal{H}_{\Delta x, n}(Q(t^n)) - \mathcal{H}_{\Delta x, n}(Q^n)$ (LeVeque, 2002). We define stability as follows.

Definition A.3 (Stability). *In the context of Problem 2.4, a 1D-FV scheme is stable in the p -norm if for any $(\Delta x, \Delta t, \lambda)$ -discretization of $[a, b] \times [0, T]$ we have:*

$$\|\mathcal{H}_{\Delta x, n}(Q) - \mathcal{H}_{\Delta x, n}(P)\|_{p, \Delta x} \leq (1 + \alpha \Delta t) \|Q - P\|_{p, \Delta x}, \quad (\text{A.54})$$

for all $Q, P \in \mathbb{R}^{\Delta x}$ and α is a constant that does not depend neither on Δx nor on Δt .

Assuming that the scheme is stable in the p -norm, then it follows from Equation (A.53) that:

$$\begin{aligned} \|E^{n+1}\|_{p, \Delta x} &\leq \|\mathcal{H}_{\Delta x, n}(Q(t^n)) - \mathcal{H}_{\Delta x, n}(Q^n)\|_{p, \Delta x} + \Delta t \max_{n=1, \dots, N_T} \|\tau^n\|_{p, \Delta x} \\ &\leq (1 + \alpha \Delta t) \|E^n\|_{p, \Delta x} + \Delta t \max_{n=1, \dots, N_T} \|\tau^n\|_{p, \Delta x} \\ &\leq (1 + \alpha \Delta t)^n \|E^0\|_{p, \Delta x} + \Delta t \max_{n=1, \dots, N_T} \|\tau^n\|_{p, \Delta x} \sum_{k=0}^{n-1} (1 + \alpha \Delta t)^k \\ &\leq e^{\alpha T} (\|E^0\|_{p, \Delta x} + T \max_{n=1, \dots, N_T} \|\tau^n\|_{p, \Delta x}), \end{aligned} \quad (\text{A.55})$$

where we used $n \Delta t \leq T$, $T = N \Delta t$ and the inequality $e^t > 1 + t$. When computing the initial average values using the value at the cell centroid, the initial error E^0 converges to zero provided q is twice continuously differentiable by Proposition 2.2. Therefore, it follows that if the scheme is stable and consistent then it is convergent. Furthermore, if it is stable and consistent with order P , then the convergence order is at least equal to $\min\{P, 2\}$. In the case where both the conservation law and $\mathcal{H}_{\Delta x, n}$ are linear, this result is a particular case of the Lax-Ritchmyer stability and the convergence is guaranteed by the Lax equivalence theorem (LeVeque, 2002). In this Chapter, we are interested only in the linear advection equation. However, as we shall see in Section 2.5, the operator $\mathcal{H}_{\Delta x, n}$ may become non-linear when monotonicity constraints are activated.

Notice that, if $\mathcal{H}_{\Delta x, n}$ is linear, then stability is equivalent to require that

$$\|\mathcal{H}_{\Delta x, n}\|_{p, \Delta x} \leq 1 + \alpha \Delta t,$$

where

$$\|\mathcal{H}_{\Delta x, n}\|_{p, \Delta x} = \sup_{Q \in \mathbb{R}^{\Delta x}} \frac{\|\mathcal{H}_{\Delta x, n}(Q)\|_{p, \Delta x}}{\|Q\|_{p, \Delta x}},$$

is the operator p -norm.

For linear operators, we may use the discrete Fourier transform (Trefethen, 2000) to estimate the 2-norm of $\mathcal{H}_{\Delta x, n}$. This approach is known as Von Neumann stability analysis. We define the nodes $\theta_i = i \frac{2\pi}{N}$, $i = 1, \dots, N$, $\Delta\theta = \frac{2\pi}{N}$, $\theta = (\theta_1, \theta_2, \dots, \theta_N)$. The imaginary unit

is denoted by ι . The Fourier modes are given by:

$$e^{ik\theta} = (e^{ik\theta_1}, e^{ik\theta_2}, \dots, e^{ik\theta_N}) \in \mathbb{C}^N,$$

for $k = 1, \dots, N$. Each k is referred to wavenumber and θ_k is called dimensionless wavenumber. The Fourier modes form an orthogonal basis of \mathbb{C}^N with respect to the inner product

$$\langle Q, P \rangle = \frac{1}{N} \sum_{i=1}^N Q_i \bar{P}_i,$$

for $P, Q \in \mathbb{C}$ and \bar{z} denotes the complex conjugate of z . Given $Q \in \mathbb{R}^{\Delta x}$, we may express it in terms of the Fourier modes

$$Q = \sum_{k=1}^N a_k \exp(ik\theta),$$

where $a_k \in \mathbb{C}$. The 2-norm of Q is then given by:

$$\|Q\|_{2,\Delta x} = \sqrt{N \sum_{k=1}^N |a_k|^2}.$$

The idea of Von Neumann stability analysis is to apply the operator $\mathcal{H}_{\Delta x,n}$ on each Fourier mode and analyze how it modifies its amplitude. For ease of analysis, we assume that the velocity is constant, which implies that the operator $\mathcal{H}_{\Delta x,n}$ has constant coefficients and does not depend on n . For the general case, where the velocity is not constant, the stability can be ensured using the frozen coefficients method (Strikwerda, 2004, p. 59). This method boils down to performing multiple times the stability analysis with a constant velocity being equal to each one of the possible values of the velocity on the grid. If the scheme is stable for all the possible constant velocities, then stability is ensured. Since the operator is supposed to be linear with constant coefficients and we are assuming periodic boundaries conditions, we may write:

$$\mathcal{H}_{\Delta x,n}(e^{ik\theta}) = \rho(k) e^{ik\theta},$$

where the term $\rho(k)$ is called amplification factor and it is an eigenvalue of $\mathcal{H}_{\Delta x,n}$. The norm of $\mathcal{H}_{\Delta x,n}(Q)$ is bounded by:

$$\|\mathcal{H}_{\Delta x,n}(Q)\|_{2,\Delta x}^2 = N \sum_{k=1}^N |a_k|^2 |\rho(k)|^2 \leq \max_{k=1,\dots,N} |\rho(k)|^2 \|Q\|_{2,\Delta x}^2.$$

Therefore:

$$\|\mathcal{H}_{\Delta x,n}\|_{2,\Delta x} \leq \max_{k=1,\dots,N} |\rho(k)|.$$

If we show that $\max_{k=1,\dots,N} |\rho(k)| \leq 1 + \alpha \Delta t$, with α independent of Δt , N and n , then we ensure the stability of $\mathcal{H}_{\Delta x,n}$. Generally speaking, the numerical flux can be written as a linear function

$$\mathcal{F}(Q(S_{i+\frac{1}{2}}), \tilde{u}_{i+\frac{1}{2}}^n) = \sum_{l \in S_{i+\frac{1}{2}}} \alpha_{l,i} Q_l,$$

when no monotonicity constraint is imposed, where the coefficients $\alpha_{l,i}$ depend on \tilde{u}^n , Δt and Δx . We can then express ρ in terms of $\alpha_{l,i}$. Indeed, when we apply the operator $\mathcal{H}_{\Delta x,n}$ in a Fourier mode, we get:

$$\begin{aligned} [\mathcal{H}_{\Delta x,n}(e^{ik\theta})]_i &= e^{ik\theta_i} - \lambda \left(\sum_{l \in S_{i+\frac{1}{2}}} \alpha_{l,i} e^{ik\theta_l} - \sum_{l \in S_{i-\frac{1}{2}}} \alpha_{l,i-1} e^{ik\theta_{l-1}} \right) \\ &= e^{ik\theta_i} \left(1 - \lambda \left(\sum_{l \in S_{i+\frac{1}{2}}} \alpha_{l,i} e^{ik\theta_{l-i}} - \sum_{l \in S_{i-\frac{1}{2}}} \alpha_{l,i-1} e^{ik\theta_{l-1-i}} \right) \right). \end{aligned}$$

Hence, the amplification factor has the form

$$\rho(k) = 1 - \lambda \left(\sum_{l \in S_{i+\frac{1}{2}}} \alpha_{l,i} e^{ik\theta_{l-i}} - \sum_{l \in S_{i-\frac{1}{2}}} \alpha_{l,i-1} e^{ik\theta_{l-1-i}} \right). \quad (\text{A.56})$$

In Section 2.5 we shall analyse $|\rho(k)|$ in terms of the PPM coefficients.

A.5.3 Flux accuracy analysis

With the stencil coefficients, we can compute the amplification factor (Equation (A.56)) for the PPM and the hybrid PPM schemes, both without monotonization. We assume a constant velocity equal to one and $N = 100$ (number of control volumes). In Figure A.1 we show the amplification factor for both PPM and hybrid PPM schemes considering different CFL numbers. We can observe that both schemes damp most of the Fourier modes for larger k , regardless of the CFL number. Besides that, the hybrid scheme is more effective when reducing the Fourier modes amplitude. We point out that both schemes are exact when the CFL number is equal to 1. From this analysis, we can conclude that the PPM and hybrid PPM schemes satisfy the Von Neumann stability criteria when the CFL restriction is respected.

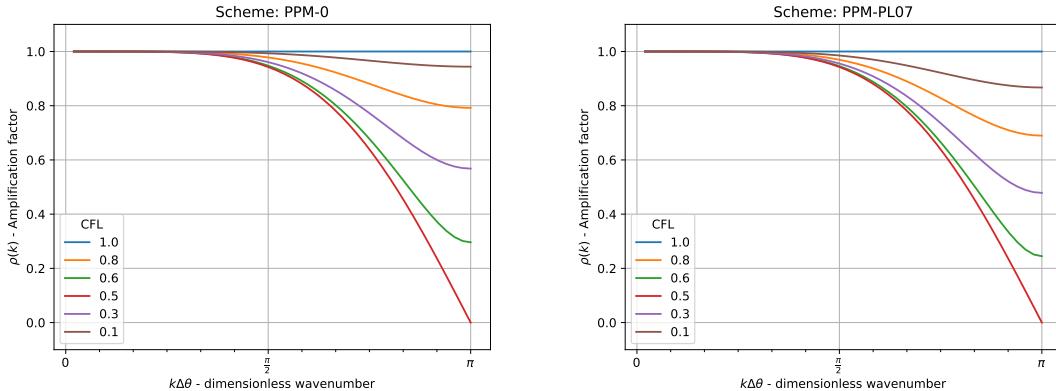


Figure A.1: Amplification factor for the PPM (left) and hybrid PPM (right) schemes for different CFL numbers.

Appendix B

Spherical coordinates and geometry

Given $R > 0$, we denote the sphere of radius R centered at the origin of \mathbb{R}^3 :

$$\mathbb{S}_R^2 = \{(X, Y, Z) \in \mathbb{R}^3 : X^2 + Y^2 + Z^2 = R^2\}.$$

The tangent space at $P \in \mathbb{S}_R^2$ by $T_P \mathbb{S}^2$. It is easy to see that:

$$T_P \mathbb{S}_R^2 = \{Q \in \mathbb{R}^3 : \langle P, Q \rangle = 0\},$$

where $\langle \cdot, \cdot \rangle$ denotes the standard inner product of \mathbb{R}^3 . The tangent bundle is denoted by:

$$T \mathbb{S}_R^2 = \bigcup_{P \in \mathbb{S}_R^2} T_P \mathbb{S}_R^2.$$

We are going to consider three ways to represent an element of \mathbb{S}_R^2 : using (X, Y, Z) coordinates, or using (λ, ϕ) latitude-longitude coordinates, or, at last, using the cubed-sphere coordinates (x, y, p) , where (x, y) are the cube face coordinates and $p \in \{1, 2, \dots, 6\}$ stands for a cube panel, as presented in Chapter 4.

B.1 Conversions between latitude-longitude and contravariant coordinates

We consider the latitude-longitude mapping $\Psi_{ll} : [0, 2\pi] \times [-\frac{\pi}{2}, \frac{\pi}{2}] \rightarrow \mathbb{S}_R^2$, given by:

$$X(\lambda, \phi) = R \cos \phi \cos \lambda, \tag{B.1}$$

$$Y(\lambda, \phi) = R \cos \phi \sin \lambda, \tag{B.2}$$

$$Z(\lambda, \phi) = R \sin \phi, \tag{B.3}$$

The derivative or Jacobian matrix of the mapping Ψ_{ll} is given by:

$$D\Psi_{ll}(\lambda, \phi) = R \begin{bmatrix} -\cos \phi \sin \lambda & -\sin \phi \cos \lambda \\ \cos \phi \cos \lambda & \sin \phi \sin \lambda \\ 0 & \cos \phi \end{bmatrix} \quad (\text{B.4})$$

Using this matrix columns, we can define the tangent vectors:

$$\mathbf{g}_\lambda(\lambda, \phi) = D\Psi_{ll}(\lambda, \phi) \begin{bmatrix} 1 \\ 0 \end{bmatrix}, \quad \mathbf{g}_\phi(\lambda, \phi) = D\Psi_{ll}(\lambda, \phi) \begin{bmatrix} 0 \\ 1 \end{bmatrix}, \quad (\text{B.5})$$

We normalize the vectors \mathbf{g}_λ and \mathbf{g}_ϕ and we obtain unit tangent vectors on the sphere at $\Phi_{ll}(\lambda, \phi)$:

$$\mathbf{e}_\lambda(\lambda, \phi) = \begin{bmatrix} -\sin \lambda \\ \cos \lambda \\ 0 \end{bmatrix}, \quad \mathbf{e}_\phi(\lambda, \phi) = \begin{bmatrix} -\sin \phi \cos \lambda \\ -\sin \phi \sin \lambda \\ \cos \phi \end{bmatrix}, \quad (\text{B.6})$$

Let us consider a tangent vector field $\mathbf{u} : \mathbb{S}_R^2 \rightarrow T\mathbb{S}_R^2$ on the sphere, i.e., $\mathbf{u}(P) \in T_P\mathbb{S}_R^2$, $\forall P \in \mathbb{S}_R^2$. We may express this vector fields in latitude-longitude coordinates as:

$$\mathbf{u}(\lambda, \phi) = u_\lambda(\lambda, \phi)\mathbf{e}_\lambda(\lambda, \phi) + v_\phi(\lambda, \phi)\mathbf{e}_\phi(\lambda, \phi). \quad (\text{B.7})$$

Or, we may also represent this vector field using the basis obtained by cubed-sphere coordinates:

$$\mathbf{u}(x, y; p) = \tilde{u}(x, y; p)\mathbf{e}_x(x, y; p) + \tilde{v}(x, y; p)\mathbf{e}_y(x, y; p). \quad (\text{B.8})$$

This representation is known as contravariant representation. In order to relate the latitude-longitude representation with the contravariant representation, we notice that:

$$\mathbf{e}_x(x, y; p) = \langle \mathbf{e}_x, \mathbf{e}_\lambda \rangle \mathbf{e}_\lambda(\lambda, \phi) + \langle \mathbf{e}_x, \mathbf{e}_\phi \rangle \mathbf{e}_\phi(\lambda, \phi), \quad (\text{B.9})$$

$$\mathbf{e}_y(x, y; p) = \langle \mathbf{e}_y, \mathbf{e}_\lambda \rangle \mathbf{e}_\lambda(\lambda, \phi) + \langle \mathbf{e}_y, \mathbf{e}_\phi \rangle \mathbf{e}_\phi(\lambda, \phi), \quad (\text{B.10})$$

which holds since the vectors $\mathbf{e}_\lambda(\lambda, \phi)$ and $\mathbf{e}_\phi(\lambda, \phi)$ are orthogonal. Replacing Equations (B.9) and (B.10) in Equation (B.8), we obtain the values (u_λ, v_ϕ) in terms of the contravariant components (\tilde{u}, \tilde{v}) as the following matrix equation:

$$\begin{bmatrix} u_\lambda(\lambda, \phi) \\ v_\phi(\lambda, \phi) \end{bmatrix} = \begin{bmatrix} \langle \mathbf{e}_x, \mathbf{e}_\lambda \rangle & \langle \mathbf{e}_y, \mathbf{e}_\lambda \rangle \\ \langle \mathbf{e}_x, \mathbf{e}_\phi \rangle & \langle \mathbf{e}_y, \mathbf{e}_\phi \rangle \end{bmatrix} \begin{bmatrix} \tilde{u}(x, y; p) \\ \tilde{v}(x, y; p) \end{bmatrix}. \quad (\text{B.11})$$

Conversely, we may express the contravariant components in terms of latitude-longitude components by inverting Equation (B.11):

$$\begin{bmatrix} \tilde{u}(x, y; p) \\ \tilde{v}(x, y; p) \end{bmatrix} = \frac{1}{\langle \mathbf{e}_x, \mathbf{e}_\lambda \rangle \langle \mathbf{e}_y, \mathbf{e}_\phi \rangle - \langle \mathbf{e}_y, \mathbf{e}_\lambda \rangle \langle \mathbf{e}_x, \mathbf{e}_\phi \rangle} \begin{bmatrix} \langle \mathbf{e}_y, \mathbf{e}_\phi \rangle & -\langle \mathbf{e}_y, \mathbf{e}_\lambda \rangle \\ -\langle \mathbf{e}_x, \mathbf{e}_\phi \rangle & \langle \mathbf{e}_x, \mathbf{e}_\lambda \rangle \end{bmatrix} \begin{bmatrix} u_\lambda(\lambda, \phi) \\ v_\phi(\lambda, \phi) \end{bmatrix}. \quad (\text{B.12})$$

B.2 Covariant/contravariant conversion

Given Equation Let us consider again a tangent vector field $\mathbf{u} : \mathbb{S}_R^2 \rightarrow T\mathbb{S}_R^2$ on the sphere, the contravariant representation of \mathbf{u} is given by Equation (B.8). The covariant components (u, v) are given by:

$$u(x, y; p) = \langle \mathbf{u}(x, y; p), e_x(x, y; p) \rangle, \quad (\text{B.13})$$

$$v(x, y; p) = \langle \mathbf{u}(x, y; p), e_y(x, y; p) \rangle. \quad (\text{B.14})$$

Replacing Equation (B.8) in Equations (B.13) and (B.14) we obtain the relation covariant components in terms of the contravariant terms:

$$\begin{bmatrix} u(x, y; p) \\ v(x, y; p) \end{bmatrix} = \begin{bmatrix} 1 & \langle \mathbf{e}_x, \mathbf{e}_y \rangle \\ \langle \mathbf{e}_x, \mathbf{e}_y \rangle & 1 \end{bmatrix} \begin{bmatrix} \tilde{u}(x, y; p) \\ \tilde{v}(x, y; p) \end{bmatrix}. \quad (\text{B.15})$$

Denoting the angle between \mathbf{e}_x and \mathbf{e}_y by α , we have $\langle \mathbf{e}_x, \mathbf{e}_y \rangle = \cos \alpha$. Thus, we may express the contravariant components in terms of the covariant terms inverting Equation (B.15):

$$\begin{bmatrix} \tilde{u}(x, y; p) \\ \tilde{v}(x, y; p) \end{bmatrix} = \frac{1}{\sin^2 \alpha} \begin{bmatrix} 1 & -\cos \alpha \\ -\cos \alpha & 1 \end{bmatrix} \begin{bmatrix} u(x, y; p) \\ v(x, y; p) \end{bmatrix}. \quad (\text{B.16})$$

Notice that combining Equations (B.15) and (B.16) with Equations (B.11) and (B.12) one may get relations between the latitude-longitude components and the covariant components.

Appendix C

Code availability

The codes needed for this work have been built openly at GitHub. The PPM implementation for the one-dimensional advection equation used in Chapter 2 is available at <https://github.com/luanfs/py-ppm>. The dimension-splitting implementation for the advection equation on the plane used in Chapter 3 is available at <https://github.com/luanfs/py-dimension-splitting>. At last, all the grid tools for the cubed sphere used Chapters 4 and 5, including the finite volume model on this grid, is available in a Python version at <https://github.com/luanfs/py-cubed-sphere> and in a Fortran 90 version at <https://github.com/luanfs/cubed-sphere>.

References

- Arakawa, A., & Lamb, V. R. (1977). Computational design of the basic dynamical processes of the ucla general circulation model. In *General circulation models of the atmosphere* (pp. 173–265). Elsevier. <https://doi.org/https://doi.org/10.1016/B978-0-12-460817-7.50009-4>. (Cit. on pp. 4, 9)
- Barros, S., Dent, D., Isaksen, L., Robinson, G., Mozdzynski, G., & Wollenweber, F. (1995). The ifs model: A parallel production weather code. *Parallel Computing*, 21(10), 1621–1638. [https://doi.org/https://doi.org/10.1016/0167-8191\(96\)80002-0](https://doi.org/https://doi.org/10.1016/0167-8191(96)80002-0) (cit. on p. 3)
- Benacchio, T., & Wood, N. (2016). Semi-implicit semi-lagrangian modelling of the atmosphere: A met office perspective. *Communications in Applied and Industrial Mathematics*, 7(3), 4–25. <https://doi.org/doi:10.1515/caim-2016-0020> (cit. on p. 1)
- Carpenter, R. L., Droegemeier, K. K., Woodward, P. R., & Hane, C. E. (1990). Application of the piecewise parabolic method (ppm) to meteorological modeling. *Monthly Weather Review*, 118(3), 586–612. [https://doi.org/10.1175/1520-0493\(1990\)118<0586:AOTPPM>2.0.CO;2](https://doi.org/10.1175/1520-0493(1990)118<0586:AOTPPM>2.0.CO;2) (cit. on pp. 4, 8, 18)
- Chen, X. (2021). The lmars based shallow-water dynamical core on generic gnmonic cubed-sphere geometry [e2020MS002280 2020MS002280]. *Journal of Advances in Modeling Earth Systems*, 13(1), e2020MS002280. <https://doi.org/https://doi.org/10.1029/2020MS002280> (cit. on p. 49)
- Chen, Y., Weller, H., Pring, S., & Shaw, J. (2017). Comparison of dimensionally split and multi-dimensional atmospheric transport schemes for long time steps. *Quarterly Journal of the Royal Meteorological Society*, 143(708), 2764–2779. <https://doi.org/https://doi.org/10.1002/qj.3125> (cit. on pp. 23, 29, 44)
- Colella, P., & Woodward, P. R. (1984). The piecewise parabolic method (ppm) for gas-dynamical simulations. *Journal of Computational Physics*, 54(1), 174–201. [https://doi.org/https://doi.org/10.1016/0021-9991\(84\)90143-8](https://doi.org/https://doi.org/10.1016/0021-9991(84)90143-8) (cit. on pp. 4, 7, 16, 18, 20–23, 27, 28)
- Cooley, J. W., & Tukey, J. W. (1965). An algorithm for the machine calculation of complex fourier series. *Mathematics of Computation*, 19(90), 297–301. <http://www.jstor.org/stable/2003354> (cit. on p. 2)
- Courant, R., & John, F. (1999). In *Introduction to calculus and analysis i*. Springer Berlin, Heidelberg. <https://doi.org/https://doi.org/10.1007/978-3-642-58604-0>. (Cit. on p. 67)
- Croisille, J.-P. (2013). Hermitian compact interpolation on the cubed-sphere grid. *Journal of Scientific Computing*, 57. <https://doi.org/10.1007/s10915-013-9702-3> (cit. on p. 49)

- Csomós, P., Faragó, I., & Havasi, Á. (2005). Weighted sequential splittings and their analysis [Numerical Methods and Computational Mechanics]. *Computers and Mathematics with Applications*, 50(7), 1017–1031. <https://doi.org/https://doi.org/10.1016/j.camwa.2005.08.004> (cit. on p. 38)
- Dennis, J., Edwards, J., Evans, K., Guba, O., Lauritzen, P., Mirin, A., St-Cyr, A., Taylor, M., & Worley, P. (2012). Cam-se: A scalable spectral element dynamical core for the community atmosphere model. *Internat. J. High Perf. Comput. Appl.*, 26, 74–89. <https://doi.org/10.1177/1094342011428142> (cit. on p. 4)
- Durran, D. (2011). Time discretization: Some basic approaches. In *Numerical techniques for global atmospheric models* (pp. 75–104). Springer Berlin Heidelberg. https://doi.org/10.1007/978-3-642-11640-7_5. (Cit. on p. 17)
- Durran, D. R. (2010). Semi-lagrangian methods. In *Numerical methods for fluid dynamics: With applications to geophysics* (pp. 357–391). Springer New York. https://doi.org/10.1007/978-1-4419-6412-0_7. (Cit. on p. 18)
- Eliassen, E., Machenhauer, B., & Rasmussen, E. (1970). On a numerical method for integration of the hydrodynamical equations with a spectral representation of the horizontal fields. <https://doi.org/10.13140/RG.2.2.13894.88645> (cit. on p. 2)
- Engwirda, D., & Kelley, M. (2016). A weno-type slope-limiter for a family of piecewise polynomial methods. <https://doi.org/10.48550/ARXIV.1606.08188>. (Cit. on pp. 8, 9, 20)
- Figueroa, S., Bonatti, J., Kubota, P., Grell, G., Morrison, H., R. M. Barros, S., Fernandez, J., Ramirez-Gutierrez, E., Siqueira, L., Luzia, G., Silva, J., Silva, J., Pendharkar, J., Capistrano, V., Alvim, D., Enore, D., Diniz, F., Satyamurty, P., Cavalcanti, I., & Panetta, J. (2016). The brazilian global atmospheric model (bam): Performance for tropical rainfall forecasting and sensitivity to convective scheme and horizontal resolution. *Weather Forecast.*, 31(5), 1547–1572. <https://doi.org/10.1175/WAF-D-16-0062.1> (cit. on p. 3)
- Giraldo, F. X., Kelly, J. F., & Constantinescu, E. M. (2013). Implicit-explicit formulations of a three-dimensional nonhydrostatic unified model of the atmosphere (numa). *SIAM Journal on Scientific Computing*, 35(5), B1162–B1194. <https://doi.org/10.1137/120876034> (cit. on p. 4)
- Godunov, S. (1959). A difference method for numerical calculation of discontinuous solutions of the equations of hydrodynamics. *Mat. Sb.*, 47(89):3, 271–306 (cit. on pp. 7, 20).
- Guo, W., Nair, R. D., & Qiu, J.-M. (2014). A conservative semi-lagrangian discontinuous galerkin scheme on the cubed sphere. *Monthly Weather Review*, 142(1), 457–475. <https://doi.org/10.1175/MWR-D-13-00048.1> (cit. on p. 17)
- Harris, L., Chen, X., Putman, W., Zhou, L., & Chen, J.-H. (2021). A scientific description of the gfdl finite-volume cubed-sphere dynamical core. *Series : NOAA technical memorandum OAR GFDL ; 2021-001*. <https://doi.org/https://doi.org/10.25923/6nhs-5897> (cit. on pp. 8, 21)
- Harris, L. M., & Lin, S.-J. (2013). A two-way nested global-regional dynamical core on the cubed-sphere grid. *Monthly Weather Review*, 141(1), 283–306. <https://doi.org/10.1175/MWR-D-11-00201.1> (cit. on pp. 4, 5)

REFERENCES

- Jia, H., & Li, K. (2011). A third accurate operator splitting method. *Mathematical and Computer Modelling*, 53(1), 387–396. <https://doi.org/https://doi.org/10.1016/j.mcm.2010.09.005> (cit. on p. 38)
- Katta, K. K., Nair, R. D., & Kumar, V. (2015a). High-order finite volume shallow water model on the cubed-sphere: 1d reconstruction scheme. *Applied Mathematics and Computation*, 266, 316–327. <https://doi.org/https://doi.org/10.1016/j.amc.2015.04.053> (cit. on p. 49)
- Katta, K. K., Nair, R. D., & Kumar, V. (2015b). High-order finite-volume transport on the cubed sphere: Comparison between 1d and 2d reconstruction schemes. *Monthly Weather Review*, 143(7), 2937–2954. <https://doi.org/https://doi.org/10.1175/MWR-D-13-00176.1> (cit. on p. 49)
- Kent, J., Melvin, T., & Wimmer, G. A. (2022). A mixed finite element discretisation of the shallow water equations. *Geoscientific Model Development Discussions*, 2022, 1–17. <https://doi.org/10.5194/gmd-2022-225> (cit. on p. 4)
- Krishnamurti, T., Hardiker, V., Bedi, H., & Ramaswamy, L. (2006). *An introduction to global spectral modeling* (Vol. 35). <https://doi.org/10.1007/0-387-32962-5>. (Cit. on p. 2)
- Lauritzen, P. H., Ullrich, P. A., & Nair, R. D. (2011). Atmospheric transport schemes: Desirable properties and a semi-lagrangian view on finite-volume discretizations. In P. Lauritzen, C. Jablonowski, M. Taylor, & R. Nair (Eds.), *Numerical techniques for global atmospheric models* (pp. 185–250). Springer Berlin Heidelberg. https://doi.org/10.1007/978-3-642-11640-7_8. (Cit. on p. 8)
- Leonard, B. P., Lock, A. P., & MacVean, M. K. (1996). Conservative explicit unrestricted-time-step multidimensional constancy-preserving advection schemes. *Monthly Weather Review*, 124(11), 2588–2606. [https://doi.org/https://doi.org/10.1175/1520-0493\(1996\)124<2588:CEUTSM>2.0.CO;2](https://doi.org/https://doi.org/10.1175/1520-0493(1996)124<2588:CEUTSM>2.0.CO;2) (cit. on p. 7)
- LeVeque, R. J. (1985). A large time step generalization of godunov's method for systems of conservation laws. *SIAM Journal on Numerical Analysis*, 22(6), 1051–1073. <https://doi.org/10.1137/0722063> (cit. on p. 7)
- LeVeque, R. J. (1990). *Numerical methods for conservation laws*. Birkhäuser Basel. <https://doi.org/10.1007/978-3-0348-5116-9>. (Cit. on pp. 10, 42)
- LeVeque, R. J. (2002). *Finite volume methods for hyperbolic problems*. Cambridge University Press. <https://doi.org/10.1017/CBO9780511791253>. (Cit. on pp. 10, 19, 75, 78)
- Lin, S.-J. (2004). A “vertically lagrangian” finite-volume dynamical core for global models. *Monthly Weather Review*, 132(10), 2293–2307. [https://doi.org/10.1175/1520-0493\(2004\)132<2293:AVLFDC>2.0.CO;2](https://doi.org/10.1175/1520-0493(2004)132<2293:AVLFDC>2.0.CO;2) (cit. on pp. 4, 8, 21, 22, 24, 28, 40)
- Lin, S.-J., Chao, W. C., Sud, Y. C., & Walker, G. K. (1994). A class of the van leer-type transport schemes and its application to the moisture transport in a general circulation model. *Monthly Weather Review*, 122(7), 1575–1593. [https://doi.org/10.1175/1520-0493\(1994\)122<1575:ACOTVL>2.0.CO;2](https://doi.org/10.1175/1520-0493(1994)122<1575:ACOTVL>2.0.CO;2) (cit. on p. 4)
- Lin, S.-J., & Rood, R. B. (1996). Multidimensional flux-form semi-lagrangian transport schemes. *Monthly Weather Review*, 124(9), 2046–2070. [https://doi.org/10.1175/1520-0493\(1996\)124<2046:MFFSLT>2.0.CO;2](https://doi.org/10.1175/1520-0493(1996)124<2046:MFFSLT>2.0.CO;2) (cit. on pp. 4, 7, 18, 23, 29, 37–40, 49)
- Lin, S.-J., & Rood, R. B. (1997). An explicit flux-form semi-lagrangian shallow-water model on the sphere. *Quarterly Journal of the Royal Meteorological Society*, 123(544), 2477–2498. <https://doi.org/https://doi.org/10.1002/qj.49712354416> (cit. on p. 4)

- Lu, F., Zhang, F., Wang, T., Tian, G., & Wu, F. (2022). High-order semi-lagrangian schemes for the transport equation on icosahedron spherical grids. *Atmosphere*, 13(11). <https://doi.org/10.3390/atmos13111807> (cit. on p. 17)
- Müller, A., Deconinck, W., Kühnlein, C., Mengaldo, G., Lange, M., Wedi, N., Bauer, P., Smolarkiewicz, P. K., Diamantakis, M., Lock, S.-J., Hamrud, M., Saarinen, S., Mozdzynski, G., Thiemert, D., Clinton, M., Bénard, P., Voitus, F., Colavolpe, C., Marguinaud, P., ... New, N. (2019). The escape project: Energy-efficient scalable algorithms for weather prediction at exascale. *Geoscientific Model Development*, 12(10), 4425–4441. <https://doi.org/10.5194/gmd-12-4425-2019> (cit. on p. 3)
- Nair, R. D., & Lauritzen, P. H. (2010). A class of deformational flow test cases for linear transport problems on the sphere. *Journal of Computational Physics*, 229(23), 8868–8887. <https://doi.org/https://doi.org/10.1016/j.jcp.2010.08.014> (cit. on pp. 25, 44)
- Orszag, S. A. (1970). Transform method for the calculation of vector-coupled sums: Application to the spectral form of the vorticity equation. *Journal of Atmospheric Sciences*, 27(6), 890–895. [https://doi.org/10.1175/1520-0469\(1970\)027<0890:TMFTCO>2.0.CO;2](https://doi.org/10.1175/1520-0469(1970)027<0890:TMFTCO>2.0.CO;2) (cit. on p. 2)
- Peixoto, P. (2016). Accuracy analysis of mimetic finite volume operators on geodesic grids and a consistent alternative. *J. Comput. Phys.*, 310, 127–160. <https://doi.org/10.1016/j.jcp.2015.12.058> (cit. on p. 5)
- Peixoto, P., & Barros, S. R. M. (2013). Analysis of grid imprinting on geodesic spherical icosahedral grids. *J. Comput. Phys.*, 237, 61–78. <https://doi.org/10.1016/j.jcp.2012.11.041> (cit. on pp. 5, 54)
- Putman, W. M. (2007). *Development of the finite-volume dynamical core on the cubed-sphere* (Doctoral dissertation). Florida State University. Florida, US. http://purl.flvc.org/fsu/fd/FSU_migr_etd-0511. (Cit. on p. 4)
- Putman, W. M., & Lin, S.-J. (2007). Finite-volume transport on various cubed-sphere grids. *Journal of Computational Physics*, 227(1), 55–78. <https://doi.org/https://doi.org/10.1016/j.jcp.2007.07.022> (cit. on pp. 4, 5, 8, 21, 28, 40, 49, 57, 61)
- Rančić, M., Purser, R. J., & Mesinger, F. (1996). A global shallow-water model using an expanded spherical cube: Gnomonic versus conformal coordinates. *Quarterly Journal of the Royal Meteorological Society*, 122(532), 959–982. <https://doi.org/https://doi.org/10.1002/qj.49712253209> (cit. on pp. 49, 53)
- Rančić, M. (1992). Semi-lagrangian piecewise biparabolic scheme for two-dimensional horizontal advection of a passive scalar. *Monthly Weather Review*, 120(7), 1394–1406. [https://doi.org/10.1175/1520-0493\(1992\)120<1394:SLPBSF>2.0.CO;2](https://doi.org/10.1175/1520-0493(1992)120<1394:SLPBSF>2.0.CO;2) (cit. on p. 29)
- Rančić, M., Purser, R. J., Jović, D., Vasic, R., & Black, T. (2017). A nonhydrostatic multiscale model on the uniform jacobian cubed sphere. *Monthly Weather Review*, 145(3), 1083–1105. <https://doi.org/10.1175/MWR-D-16-0178.1> (cit. on pp. 4, 50)
- Randall, D. A., Bitz, C. M., Danabasoglu, G., Denning, A. S., Gent, P. R., Gettelman, A., Griffies, S. M., Lynch, P., Morrison, H., Pincus, R., & Thuburn, J. (2018). 100 years of earth system model development. *Meteorological Monographs*, 59, 12.1–12.66. <https://doi.org/10.1175/AMSMONOGRAPH-D-18-0018.1> (cit. on pp. 1, 3)
- Richtmyer, R. D., & Morton, K. W. (1968). Difference methods for initial-value problems. *SIAM Review*, 10(3), 381–383. <https://doi.org/10.1137/1010073> (cit. on p. 37)

REFERENCES

- Ringler, T., Thuburn, J., Klemp, J., & Skamarock, W. (2010). A unified approach to energy conservation and potential vorticity dynamics on arbitrarily structured C-grids. *J. Comput. Phys.*, 229, 3065–3090. <https://doi.org/10.1016/j.jcp.2009.12.007> (cit. on p. 5)
- Ronchi, C., Iacono, R., & Paolucci, P. (1996). The “cubed sphere”: A new method for the solution of partial differential equations in spherical geometry. *Journal of Computational Physics*, 124(1), 93–114. <https://doi.org/https://doi.org/10.1006/jcph.1996.0047> (cit. on pp. 4, 49, 51, 52, 57)
- Sadourny, R. (1972). Conservative finite-difference approximations of the primitive equations on quasi-uniform spherical grids. *Monthly Weather Review*, 100(2), 136–144. [https://doi.org/10.1175/1520-0493\(1972\)100<0136:CFAOTP>2.3.CO;2](https://doi.org/10.1175/1520-0493(1972)100<0136:CFAOTP>2.3.CO;2) (cit. on pp. 4, 49, 57)
- Samenow, J. (2019). *National weather service launches upgraded, improved global forecast model*. Retrieved July 29, 2022, from <https://www.washingtonpost.com/weather/2019/06/12/national-weather-service-launches-upgraded-improved-global-forecast-model/>. (Cit. on p. 4)
- Santos, L. F., & Peixoto, P. S. (2021). Topography-based local spherical voronoi grid refinement on classical and moist shallow-water finite-volume models. *Geoscientific Model Development*, 14(11), 6919–6944. <https://doi.org/10.5194/gmd-14-6919-2021> (cit. on p. 5)
- Skamarock, W., Klemp, J., Duda, M., Fowler, L., Park, S.-H., & Ringler, T. (2012). A multiscale nonhydrostatic atmospheric model using centroidal Voronoi tessellations and C-grid staggering. *Mon. Weather Rev.*, 140(09), 3090–3105. <https://doi.org/10.1175/MWR-D-11-00215.1> (cit. on p. 5)
- Staniforth, A., & Thuburn, J. (2012). Horizontal grids for global weather and climate prediction models: A review. *Q. J. Roy. Meteor. Soc.*, 138, 1–26. <https://doi.org/10.1002/qj.958> (cit. on p. 3)
- Stoer, J., & Bulirsch, R. (2002). In *Introduction to numerical analysis*. Springer New York, NY. <https://doi.org/https://doi.org/10.1007/978-0-387-21738-3>. (Cit. on pp. 18, 67)
- Strang, G. (1968). On the construction and comparison of difference schemes. *SIAM Journal on Numerical Analysis*, 5(3), 506–517. <https://doi.org/10.1137/0705041> (cit. on p. 38)
- Strikwerda, J. C. (2004). *Finite difference schemes and partial differential equations, second edition*. Society for Industrial; Applied Mathematics. <https://doi.org/10.1137/1.9780898717938>. (Cit. on p. 79)
- Suresh, A., & Huynh, H. (1997). Accurate monotonicity-preserving schemes with runge–kutta time stepping. *Journal of Computational Physics*, 136(1), 83–99. <https://doi.org/https://doi.org/10.1006/jcph.1997.5745> (cit. on p. 21)
- Thuburn, J. (2011). Conservation in dynamical cores: What, how and why? In *Numerical techniques for global atmospheric models* (pp. 345–355). Springer Berlin Heidelberg. https://doi.org/10.1007/978-3-642-11640-7_11. (Cit. on p. 3)
- Thuburn, J., Ringler, T., Skamarock, W., & Klemp, J. (2009). Numerical representation of geostrophic modes on arbitrarily structured C-grids. *J. Comput. Phys.*, 228, 8321–8335. <https://doi.org/10.1016/j.jcp.2009.08.006> (cit. on p. 5)
- Trefethen, L. N. (2000). *Spectral methods in matlab*. Society for Industrial; Applied Mathematics. <https://doi.org/10.1137/1.9780898719598>. (Cit. on pp. 25, 78)

- Tumolo, G. (2011). *A semi-implicit, semi-lagrangian, p-adaptative discontinuous galerkin method for the rotating shallow-water equations: Analysis and numerical experiments* (Doctoral dissertation). University of Trieste. <https://core.ac.uk/download/pdf/41173373.pdf>. (Cit. on p. 17)
- Ullrich, P. A., Jablonowski, C., Kent, J., Lauritzen, P. H., Nair, R., Reed, K. A., Zarzycki, C. M., Hall, D. M., Dazllich, D., Heikes, R., Konor, C., Randall, D., Dubos, T., Meurdesoif, Y., Chen, X., Harris, L., Kühnlein, C., Lee, V., Qaddouri, A., ... Viner, K. (2017). Dcmip2016: A review of non-hydrostatic dynamical core design and intercomparison of participating models. *Geoscientific Model Development*, 10(12), 4477–4509. <https://doi.org/10.5194/gmd-10-4477-2017> (cit. on p. 3)
- Van Leer, B. (1977). Towards the ultimate conservative difference scheme. iv. a new approach to numerical convection. *Journal of Computational Physics*, 23(3), 276–299. [https://doi.org/https://doi.org/10.1016/0021-9991\(77\)90095-X](https://doi.org/https://doi.org/10.1016/0021-9991(77)90095-X) (cit. on pp. 4, 7, 18, 20)
- Weller, H. (2012). Controlling the computational modes of the arbitrarily structured c grid, *Mon. Weather. Rev.*, 140(10), 3220–3234. <https://doi.org/doi.org/10.1175/MWR-D-11-00221.1> (cit. on p. 5)
- Whitaker, J. (2015). *Hiwpp non-hydrostatic dynamical core tests: Results from idealized test cases*. Retrieved November 5, 2022, from https://www.weather.gov/media/sti/nggps/HIWPP_idealized_tests-v8%20revised%2005212015.pdf. (Cit. on p. 4)
- White, L., & Adcroft, A. (2008). A high-order finite volume remapping scheme for nonuniform grids: The piecewise quartic method (pqm). *Journal of Computational Physics*, 227(15), 7394–7422. <https://doi.org/https://doi.org/10.1016/j.jcp.2008.04.026> (cit. on p. 8)
- Williamson, D., Drake, J., Hack, J., Jakob, R., & Swarztrauber, P. (1992). A standard test set for numerical approximations to the shallow water equations in spherical geometry. *J. Comput. Phys.*, 102, 211–224. [https://doi.org/10.1016/S0021-9991\(05\)80016-6](https://doi.org/10.1016/S0021-9991(05)80016-6) (cit. on p. 5)
- Williamson, D. L. (2007). The evolution of dynamical cores for global atmospheric models. *Journal of the Meteorological Society of Japan. Ser. II*, 85B, 241–269. <https://doi.org/10.2151/jmsj.85B.241> (cit. on pp. 1, 2)
- Wood, N., Staniforth, A., White, A., Allen, T., Diamantakis, M., Gross, M., Melvin, T., Smith, C., Vosper, S., Zerroukat, M., & Thuburn, J. (2014). An inherently mass-conserving semi-implicit semi-lagrangian discretization of the deep-atmosphere global non-hydrostatic equations. *Quarterly Journal of the Royal Meteorological Society*, 140(682), 1505–1520. <https://doi.org/https://doi.org/10.1002/qj.2235> (cit. on p. 1)
- Woodward, P. R. (1986). Piecewise-parabolic methods for astrophysical fluid dynamics. In K.-H. A. Winkler & M. L. Norman (Eds.), *Astrophysical radiation hydrodynamics* (pp. 245–326). Springer Netherlands. https://doi.org/10.1007/978-94-009-4754-2_8 (Cit. on p. 8)
- Zerroukat, M., & Allen, T. (2022). On the corners of the cubed-sphere grid. *Quarterly Journal of the Royal Meteorological Society*, 148(743), 778–783. <https://doi.org/https://doi.org/10.1002/qj.4230> (cit. on p. 49)

REFERENCES

- Zheng, Y., & Marguinaud, P. (2018). Simulation of the performance and scalability of message passing interface (mpi) communications of atmospheric models running on exascale supercomputers. *Geoscientific Model Development*, 11(8), 3409–3426. <https://doi.org/10.5194/gmd-11-3409-2018> (cit. on p. 3)