## NONLINEAR DYNAMICS AND SYSTEMS THEORY

An International Journal of Research and Surveys
Volume 19

$$
\text { Number } 1
$$

Existence of Solutions for a Biological Model Using Topological Degree Theory
C.H.D. Alliera

Solution of 2D Fractional Order Integral Equations by Bernstein Polynomials Operational Matrices $\qquad$
M. Asgari, R. Ezzati and H. Jafari

Singular Analysis of Reduced ODEs of Rotating Stratified Boussinesq Equations Through the Mirror Transformations.
B.S. Desale and K.D. Patil

Real Time Analysis of Signed Marked Random Measures with Applications to Finance and Insurance . $\qquad$
Jewgeni H. Dshalalow and Kizza M. Nandyose

Decentralized Stabilization for a Class of Nonlinear Interconnected Systems Using SDRE Optimal Control Approach
A. Feydi, S. Elloumi and N. Benhadj Braiek

Analysis and Adaptive Control Synchronization of a Novel 3-D Chaotic System.
nnachi
F. Hannachi

Mathematical Model of $C d$ for Circular Cylinder Using Two Passive Controls at $\mathrm{Re}=5000$
C. Imron C.J. Kumalasari, B. Widodo and T.Y. Yuwono

## Nonlinear Dynamics and Systems Theory

## An International Journal of Research and Surveys

```
EDITOR-IN-CHIEF A.A.MARTYNYUK
S.P.Timoshenko Institute of Mechanics
National Academy of Sciences of Ukraine, Kiev, Ukraine
REGIONAL EDITORS
P.BORNE, Lille, France M.FABRIZIO, Bologna, Italy Europe
M.BOHNER, Rolla, USA HAO WANG, Edmonton, Canada USA and Canada
T.A.BURTON, Port Angeles, USA C.CRUZ-HERNANDEZ, Ensenada, Mexico
USA and Latin America
M.ALQURAN, Irbid, Jordan
Jordan and Middle East
K.L.TEO, Perth, Australia
Australia and New Zealand
```


## Nonlinear Dynamics and Systems Theory

An International Journal of Research and Surveys

EDITOR-IN-CHIEF A.A.MARTYNYUK
The S.P.Timoshenko Institute of Mechanics, National Academy of Sciences of Ukraine, Nesterov Str. 3, 03680 MSP, Kiev-57, UKRAINE / e-mail: journalndst@gmail.com

MANAGING EDITOR
Department of Mathematics, University of Ioannina
Department of Mathematics, University of Ioannina
45110 Ioannina, HELLAS (GREECE) / e-mail: ipstav@cc.uoi.gr
ADVISORY EDITOR A.G.MAZKO
Institute of Mathematics of NAS of Ukraine, Kiev (Ukraine)
e-mail: mazko@imath.kiev.ua

## REGIONAL EDITOR

M.ALQURAN (Jordan), e-mail: marwan04@just.edu.j BORNE (France), e-mail: Pierre.Borne@ec-ille A.BURTON (USA), e-mail taburton@olypen
C. CRUZ-HERNANDEZ (Mexico), e-mail: ccruz@cicese.mx M.FABRIZIO (Italy), e-mail: mauro.fabrizio@unibo.it HAO WANG (Canada), e-mail: hao8@ualberta.ca K.L.TEO (Australia), e-mail: K.L.Teo@curtin.edu.au

| EDITORIAL BOARD |  |
| :---: | :---: |
| Aleksandrov, A.Yu. (Russia) | Khusainov, D.Ya. (Ukraine) |
| Artstein, Z. (Israel) | Kloeden, P. (Germany) |
| Awrejcewicz, J. (Poland) | Kokologiannaki, C. (Greece) |
| Benrejeb, M. (Tunisia) | Krishnan, E.V. (Oman) |
| Braiek, N.B. (Tunisia) | Limarchenko, O.S. (Ukraine) |
| Chen Ye-Hwa (USA) | Nguang Sing Kiong (New Zealand) |
| Corduneanu, C. (USA) | Okninski, A. (Poland) |
| D'Anna, A. (Italy) | Peng Shi (Australia) |
| De Angelis, M. (Italy) | Peterson, A. (USA) |
| Denton, Z. (USA) | Radziszewski, B. (Poland) |
| Vasundhara Devi, J. (India) | Shi Yan (Japan) |
| Djemai, M. (France) | Siljak, D.D. (USA) |
| Dshalalow, J.H. (USA) | Sivasundaram, S. (USA) |
| Eke, F.O. (USA) | Sree Hari Rao, V. (India) |
| Georgiou, G. (Cyprus) | Stavrakakis, N.M. (Greece) |
| Honglei Xu (Australia) | Vatsala, A. (USA) |
| Izobov, N.A. (Belarussia) | Zuyev, A.L. (Germany) |
| Jafari, H. (South African Repu |  |
| ADVISORY COMPUTER SCIENCE EDITORS <br> A.N.CHERNIENKO and L.N.CHERNETSKAYA, Kiev, Ukraine |  |
| ADVISORY LINGUISTIC EDITOR S.N.RASSHYVALOVA, Kiev, Ukraine |  |
|  |  |

## INSTRUCTIONS FOR CONTRIBUTORS

(1) General. Nonlinear Dynamics and Systems Theory (ND\&ST) is an international journal devoted to publishing peer-refereed, high quality, original papers, brief notes and review articles focusing on nonlinear dynamics and systems theory and their practical applications in engineering, physical and life sciences. Submission of a manuscript is a representation that the submis is ben appred by all or the authors and by the institution where the work was carried out. It also represents that the manuscript has not been previously published, has not been copyrighted, is not being submitted for publication elsewhere, and that the authors have agreed that the copyright in the article shall be assigned exclusively to InforMath Publishing Group by signing a transfer of copyright form. Before submission, the authors should visit the website

## http://www.e-ndst.kiev.ua

for information on the preparation of accepted manuscripts. Please download the archive Sample_NDST.zip Samplefilename.tex).
(2) Manuscript and Correspondence. Manuscripts should be in English and must meet common standards of usage and grammar. To submit a paper, send by e-mail a file in PDF format directly to

Nesterov str 3 . 03057 , e-mail: journalndst@gmail.com
or to one of the Regional Editors or to a member of the Editorial Board. Final version of the manuscript must typeset using LaTex program which is prepared in accordance with the style file of the Journal. Manuscript texts should contain the title of the article, name(s) of the author(s) and complete affiliations. Each article requires an abstract not exceeding 150 words. Formulas and citations should not be included in the abstract. AMS subject classifications and key words must be included in all accepted papers. Each article requires a running head (abbreviated form of the title) of no more than 30 characters. The sizes for regular papers, survey articles, brief notes, letters to editors and book reviews are: (i) $10-14$ pages for regular papers, (ii) up to 24 pages for survey articles, and (iii) 2-3 pages for brief notes, letters to the ditor and book review
(3) Tables, Graphs and Illustrations. Each figure must be of a quality suitable for direct should be drawn professionally a caption. Drawings should include all relevant details and hard copy of the artwork, it is necessary to attach the electronic file of the artwork (preferably hard copy of the at in PCX format).
(4) References. Each entry must be cited in the text by author(s) and number or by number All references should be listed in their alphabetic order. Use please the following style:
Journal: [1] H. Poincare, Title of the article. Title of the Journal volume
(issue) (year) pages. [Language]
Book: [2] A.M. Lyapunov, Title of the Book. Name of the Publishers, Town, year.
Proceeding: [3] R. Bellman, Title of the article. In: Title of the Book. (Eds.). Name of the Publishers, Town, year, pages. [Language]
(5) Proofs and Sample Copy. Proofs sent to authors should be returned to the Editorial Office with corrections within three days after receipt. The corresponding author will receive a sample copy of the issue of the Journal for which his/her paper is published.
(6) Editorial Policy. Every submission will undergo a stringent peer review process. An editor will be assigned to handle the review process of the paper. He/she will secure at least two reviewers' reports. The decision on acceptance, rejection or acceptance subject to revision will be made based on these reviewers' reports and the editor's own reading of the paper.

# NONLINEAR DYNAMICS AND SYSTEMS THEORY 

An International Journal of Research and Surveys
Published by InforMath Publishing Group since 2001
Volume 19 Number 1 ..... 2019
CONTENTS
Existence of Solutions for a Biological Model Using Topological Degree Theory ..... 1
C.H.D. Alliera
Solution of 2D Fractional Order Integral Equations by Bernstein Polynomials Operational Matrices ..... 10
M. Asgari, R. Ezzati and H. Jafari
Singular Analysis of Reduced ODEs of Rotating Stratified Boussinesq Equations Through the Mirror Transformations ..... 21
B.S. Desale and K.D. Patil
Real Time Analysis of Signed Marked Random Measures with Applications to Finance and Insurance ..... 36
Jewgeni H. Dshalalow and Kizza M. Nandyose
Decentralized Stabilization for a Class of Nonlinear Interconnected Systems Using SDRE Optimal Control Approach ..... 55
A. Feydi, S. Elloumi and N. Benhadj Braiek
Analysis and Adaptive Control Synchronization of a Novel 3-D Chaotic System ..... 68
F. Hannachi
Mathematical Model of $C_{d}$ for Circular Cylinder Using Two Passive Controls at $\mathrm{Re}=5000$ ..... 79C. Imron C.J. Kumalasari, B. Widodo and T.Y. Yuwono
A Variety of New Solitary-Solutions for the Two-mode Modified Korteweg-de Vries Equation ..... 88A. Jaradat, M.S.M. Noorani, M. Alquran and H.M. Jaradat
Existence of Renormalized Solutions for Some Strongly Parabolic Problems in Musielak-Orlicz-Sobolev Spaces ..... 97A. Talha, A. Benkirane and M.S.B. Elemine Vall

# NONLINEAR DYNAMICS AND SYSTEMS THEORY 

# An International Journal of Research and Surveys 

Impact Factor from SCOPUS for 2017: SNIP - 0.707, SJR - 0.316

Nonlinear Dynamics and Systems Theory (ISSN 1562-8353 (Print), ISSN 18137385 (Online)) is an international journal published under the auspices of the S.P. Timoshenko Institute of Mechanics of National Academy of Sciences of Ukraine and Curtin University of Technology (Perth, Australia). It aims to publish high quality original scientific papers and surveys in areas of nonlinear dynamics and systems theory and their real world applications.

## AIMS AND SCOPE

Nonlinear Dynamics and Systems Theory is a multidisciplinary journal. It publishes papers focusing on proofs of important theorems as well as papers presenting new ideas and new theory, conjectures, numerical algorithms and physical experiments in areas related to nonlinear dynamics and systems theory. Papers that deal with theoretical aspects of nonlinear dynamics and/or systems theory should contain significant mathematical results with an indication of their possible applications. Papers that emphasize applications should contain new mathematical models of real world phenomena and/or description of engineering problems. They should include rigorous analysis of data used and results obtained. Papers that integrate and interrelate ideas and methods of nonlinear dynamics and systems theory will be particularly welcomed. This journal and the individual contributions published therein are protected under the copyright by International InforMath Publishing Group.

## PUBLICATION AND SUBSCRIPTION INFORMATION

Nonlinear Dynamics and Systems Theory will have 4 issues in 2019, printed in hard copy (ISSN 1562-8353) and available online (ISSN 1813-7385), by InforMath Publishing Group, Nesterov str., 3, Institute of Mechanics, Kiev, MSP 680, Ukraine, 03057. Subscription prices are available upon request from the Publisher, EBSCO Information Services (mailto:journals@ebsco.com), or website of the Journal: http: //e-ndst.kiev.ua. Subscriptions are accepted on a calendar year basis. Issues are sent by airmail to all countries of the world. Claims for missing issues should be made within six months of the date of dispatch.

## ABSTRACTING AND INDEXING SERVICES

Papers published in this journal are indexed or abstracted in: Mathematical Reviews / MathSciNet, Zentralblatt MATH / Mathematics Abstracts, PASCAL database (INISTCNRS) and SCOPUS.

# Existence of Solutions for a Biological Model Using Topological Degree Theory 

C. H. D. Alliera*<br>Departamento de Matemáticas, Facultad de Ciencias Exactas y Naturales, Universidad de Buenos Aires.

Received: January 11, 2018; Revised: December 21, 2018


#### Abstract

Topological degree theory is a useful tool for studying systems of differential equations. In this work, a biological model is considered. Specifically, we prove the existence of positive $T$-periodic solutions of a system of delay differential equations for a model with feedback arising on circadian oscillations in the drosophila period gene protein.


Keywords: differential equations with delay; periodic solutions; models with feedback; topological degree; drosophila; circadian cycle.

Mathematics Subject Classification (2010): 34K13, 92B05.

## 1 Introduction

The study of cellular control has been developed in many papers on mathematical analysis to determine the existence of stable oscillations in mRNA regulatory processes, see [5] and to understand circadian cycles and, in particular, of the cellular machinery that produces them, see [7].

In all cases, search for conditions on the parameters of the proposed systems has been carried out with the purpose of determining conditions for the existence of stable cycles and the cycles when the system solution may be even chaotic.

Let us consider a model proposed by Goldbeter [3], who showed the variation on PER: the period of messenger of Ribo-Nucleic Acid (mRNA) in Drosophila (often called "fruit flies") related to circadian rhythms. Our model does not consider temperature variation as shown in [6]. Here, a nonautonomous version of the model is considered with the aim of proving the existence of periodic solutions by means of a powerful topological tool: the Leray-Schauder degree (see [1] and [2]). In the original model, the existence of a positive steady state can be shown, under appropriate conditions, by the use of the Brouwer degree. As we shall see, when the parameters are replaced by periodic functions, essentially the same conditions yield the existence of positive periodic solutions.

[^0]

Figure 1: Model for the circadian variation in PER.

## 2 The Model

The following simplified model was proposed in [3]. Some more complex alternative models have been studied with light interaction and timeless (TIM) proteins (see [4]).

### 2.1 General features

1. This negative feedback will be described by an equation of Hill type in which $n$ denotes the degree of cooperativity, and $K(t)$ is the threshold repression function.
2. To simplify the model, we consider that $P_{N}$ behaves directly as a repressor.
3. The constants $K_{s}, K_{i}$ and $V_{j}$ denote the maximum rate and Michaelis constant of the kinase(s) and the phosphatase(s) involved in the reversible phosphorylation of $P_{0}$ into $P_{1}$, and of $P_{1}$ into $P_{2}$ are not negative.
4. Maximum accumulation rate of cytosol is denoted by $V_{s}$.
5. Cytosol is degraded enzymically, in a Michaelian manner, at a maximum rate $V_{m}$.
6. Functions of this system are:
(a) Cytosolic concentration is denoted by $M$.
(b) We consider only three states of the protein: unphosphorylated $\left(P_{0}\right)$, monophosphorylated $\left(P_{1}\right)$ and bisphosphorylated $\left(P_{2}\right)$.
(c) Fully phosphorylated form of PER $\left(P_{2}\right)$ is degraded in a Michaelian manner, at a maximum rate $V_{d}$, and also transported into the nucleus, at a rate characterized by the apparent first-order rate constant $k_{1}$.
7. The rate of synthesis of PER, proportional to $M$, is characterized by an apparent first-order rate constant $K_{s}$.
8. Transport of the nuclear, bisphosphorylated form of PER $\left(P_{N}\right)$ into the cytosol is characterized by the apparent first-order rate constant $k_{2}$.
9. The model could be readily extended to include a larger number of phosphorylated residues.

With this in mind, our non-autonomous version of Goldbeter's system reads:

$$
\begin{align*}
\frac{d M}{d t} & =\frac{V_{S}(t) K_{1}(t)^{n}}{K_{1}^{n}(t)+P_{N}(t)^{n}}-\frac{V_{m}(t) M(t)}{K_{m_{1}}(t)+M(t)}, \\
\frac{d P_{0}}{d t} & =K_{s}(t) M(t)+\frac{V_{2}(t) P_{1}(t)}{K_{2}(t)+P_{1}(t)}-\frac{V_{1}(t) P_{0}(t)}{K_{1}(t)+P_{0}(t)}, \\
\frac{d P_{1}}{d t} & =\frac{V_{1}(t) P_{0}(t)}{K_{1}(t)+P_{0}(t)}+\frac{V_{4}(t) P_{2}(t)}{K_{4}(t)+P_{2}(t)}-P_{1}(t)\left(\frac{V_{2}(t)}{K_{2}(t)+P_{1}(t)}+\frac{V_{3}(t)}{K_{3}(t)+P_{1}(t)}\right), \\
\frac{d P_{2}}{d t} & =\frac{V_{3}(t) P_{1}(t)}{K_{3}(t)+P_{1}(t)}+k_{2}(t) P_{N}(t)-P_{2}(t)\left(k_{1}(t)+\frac{V_{4}(t)}{K_{4}(t)+P_{2}(t)}+\frac{V_{d}(t)}{K_{d}(t)+P_{2}(t)}\right), \\
\frac{d P_{N}}{d t} & =k_{1}(t) P_{2}(t)-k_{2}(t) P_{N}(t), \tag{1}
\end{align*}
$$

where $K_{i}, i=1,2,3,4, d, m_{1}, s, k_{1}, k_{2}$ and $V_{j}, j=1,2,3,4, S, m, d$ are strictly positive, continuous $T$-periodic functions. We shall prove that, under accurate assumptions to be specified below, the system admits at least one positive $T$-periodic solution.

## 3 Existence of Positive Periodic Solutions

In order to apply the topological degree method to problem (1), let us consider the space of continuous $T$-periodic vector functions

$$
C_{T}:=\left\{u \in C\left(\mathbb{R}, \mathbb{R}^{5}\right): u(t)=u(t+T) \text { for all } t\right\}
$$

equipped with the standard uniform norm, and the positive cone

$$
\mathcal{K}:=\left\{u \in C_{T}: u_{j} \geq 0, j=1, \ldots, 5\right\} .
$$

Thus, the original problem can be written as $L u=N u$, where $L: C^{1} \cap C_{T} \rightarrow C$ is given by $L u:=u^{\prime}$ and the nonlinear operator $N: \mathcal{K} \rightarrow C_{T}$ is defined as the right-hand side of system (1). For convenience, the average of a function $u$ shall be denoted by $\bar{u}$, namely $\bar{u}:=\frac{1}{T} \int_{0}^{T} u(t) d t$. Also, identifying $\mathbb{R}^{5}$ with the subset of constant functions of $C_{T}$, we may define the function $\phi:[0,+\infty)^{5} \rightarrow \mathbb{R}^{5}$ given by $\phi(x):=\overline{N x}$.

For the reader's convenience, let us summarize the basic properties of the LeraySchauder degree which, roughly speaking, can be regarded as an algebraic count of the zeros of a mapping $F: \bar{\Omega} \rightarrow E$, where $E$ is a Banach space and $\Omega \subset E$ is open and bounded. In more precise terms, assume that $F=I-K$, where $K$ is compact and $F \neq 0$ on $\partial \Omega$. The degree $\operatorname{deg}_{L S}(F, \Omega, 0)$ is defined as the Brouwer degree $\operatorname{deg}_{B}$ of its
restriction $\left.F\right|_{V}: \Omega \cap V \rightarrow V$, where $V$ is an accurate finite-dimensional subspace of $E$. In particular, if the range of $K$ is finite dimensional, then one may take $V$ as the subspace spanned by $\operatorname{Im}(K)$. If $\operatorname{deg}_{L S}(F, \Omega, 0)$ is different from 0 , then $F$ vanishes in $\Omega$; moreover, the degree is invariant over a continuous homotopy $F_{\lambda}:=I-K_{\lambda}$ with $K_{\lambda}$ being compact and $F_{\lambda} \neq 0$ over $\partial \Omega$. Finally, we recall that if $\Delta: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is a diffeomorphism and $0 \in \Delta(A)$ for some open bounded $A \subset \mathbb{R}^{n}$, then $\operatorname{deg}_{B}(\Delta, A, 0)$ is just the sign of the Jacobian determinant of $\Delta$ at the (unique) pre-image of 0 . The following continuation theorem is a direct consequence of the standard topological degree methods (see e.g. [1]).

Theorem 3.1 Assume there exists $\Omega \subset \mathcal{K}^{\circ}$ being open and bounded such that:
a) The problem $L u=\lambda N u$ has no solutions on $\partial \Omega$ for $0<\lambda<1$.
b) $\phi(u) \neq 0$ for all $u \in \partial \Omega \cap \mathbb{R}^{5}$.
c) $\operatorname{deg}_{B}\left(\phi, \Omega \cap \mathbb{R}^{5}, 0\right) \neq 0$.

Then (1) has at least one solution in $\bar{\Omega}$.

### 3.1 A priori bounds

Firstly, we shall find appropriate bounds for the solution of the problem $L u=\lambda N u$ with $\lambda \in(0,1)$. For convenience, let us fix the following notation for the minima and maxima of all the functions involved in the model, namely

$$
0<v_{i} \leq V_{i}(t) \leq \mathcal{V}_{i}, 0<\kappa_{j} \leq K_{j}(t) \leq \mathcal{K}_{j}, 0<\hat{k}_{l} \leq k_{l}(t) \leq \mathbb{k}_{l}, \forall i, j, l
$$

Now assume that $u \in \mathcal{K}^{\circ}$ satisfies $L u=\lambda N u$ for some $0<\lambda<1$. Let us firstly consider a value $t^{*}$ at which $M$ achieves an absolute maximum, then $M^{\prime}\left(t^{*}\right)=0$ and hence

$$
\frac{V_{S}\left(t^{*}\right) K_{1}\left(t^{*}\right)^{n}}{K_{1}^{n}\left(t^{*}\right)+P_{N}\left(t^{*}\right)^{n}}=\frac{V_{m}\left(t^{*}\right) M\left(t^{*}\right)}{K_{m_{1}}\left(t^{*}\right)+M\left(t^{*}\right)} \geq \frac{v_{m} M\left(t^{*}\right)}{\mathcal{K}_{m_{1}}+M\left(t^{*}\right)}:=b_{M}\left(M\left(t^{*}\right)\right),
$$

where the increasing function

$$
b_{M}(x):=\frac{v_{m} x}{\mathcal{K}_{m_{1}}+x}
$$

has inverse such that

$$
b_{M}^{-1}(y):=\frac{\mathcal{K}_{m_{1}} y}{v_{m}-y}
$$

If

## Hypothesis 3.1

$$
v_{m}>\mathcal{V}_{S}
$$

then

$$
M\left(t^{*}\right)=b_{M}^{-1}\left(\frac{V_{S}\left(t^{*}\right) K_{1}\left(t^{*}\right)^{n}}{K_{1}^{n}\left(t^{*}\right)+P_{N}\left(t^{*}\right)^{n}}\right)<b_{M}^{-1}\left(V_{S}\left(t^{*}\right)\right) \leq \frac{\mathcal{V}_{S} \mathcal{K}_{m_{1}}}{v_{M}-\mathcal{V}_{S}}:=\mathcal{M}
$$

Next, suppose that $P_{0}$ achieves its absolute maximum at some point, denoted again $t^{*}$, then

$$
K_{s}\left(t^{*}\right) M\left(t^{*}\right)+\frac{V_{2}\left(t^{*}\right) P_{1}\left(t^{*}\right)}{K_{2}\left(t^{*}\right)+P_{1}\left(t^{*}\right)}=\frac{V_{1}\left(t^{*}\right) P_{0}\left(t^{*}\right)}{K_{1}\left(t^{*}\right)+P_{0}\left(t^{*}\right)} \geq \frac{v_{1} P_{0}\left(t^{*}\right)}{\mathcal{K}_{1}+P_{0}\left(t^{*}\right)}:=b_{0}\left(P_{0}\left(t^{*}\right)\right) .
$$

Again, we define an increasing and invertible function:

$$
b_{0}(x):=\frac{v_{1} x}{\mathcal{K}_{1}+x} \rightarrow b_{0}^{-1}(y):=\frac{\mathcal{K}_{1} y}{v_{1}-y} .
$$

Thus, under the condition

## Hypothesis 3.2

$$
\mathcal{K}_{S} \mathcal{M}+\mathcal{V}_{2}<v_{1}
$$

we deduce that

$$
P_{0}\left(t^{*}\right)=b_{0}^{-1}\left(K_{s}\left(t^{*}\right) M\left(t^{*}\right)+\frac{V_{2}\left(t^{*}\right) P_{1}\left(t^{*}\right)}{K_{2}\left(t^{*}\right)+P_{1}\left(t^{*}\right)}\right)<\frac{\mathcal{K}_{S} \mathcal{M}+\mathcal{V}_{2}}{v_{1}-\left(\mathcal{K}_{S} \mathcal{M}+\mathcal{V}_{2}\right)} \mathcal{K}_{1}:=\mathcal{P}_{0}
$$

Next, an upper bound $\mathcal{P}_{1}$ for $P_{1}$ is readily obtained in the following way. Let us denote again by $t^{*}$ a value at which $P_{1}$ achieves its absolute maximum, then

$$
\frac{V_{1}\left(t^{*}\right) P_{0}\left(t^{*}\right)}{K_{1}\left(t^{*}\right)+P_{0}\left(t^{*}\right)}+\frac{V_{4}\left(t^{*}\right) P_{2}\left(t^{*}\right)}{K_{4}\left(t^{*}\right)+P_{2}\left(t^{*}\right)}=P_{1}\left(t^{*}\right)\left(\frac{V_{2}\left(t^{*}\right)}{K_{2}\left(t^{*}\right)+P_{1}\left(t^{*}\right)}+\frac{V_{3}\left(t^{*}\right)}{K_{3}\left(t^{*}\right)+P_{1}\left(t^{*}\right)}\right) .
$$

When $P_{1}\left(t^{*}\right) \gg 0$, the right-hand side gets close to $V_{2}\left(t^{*}\right)+V_{3}\left(t^{*}\right)$, while the left-hand side is always less than or equal to $\frac{\mathcal{V}_{1} \mathcal{P}_{0}}{\kappa_{1}+\mathcal{P}_{0}}+\mathcal{V}_{4}$ because $\frac{P_{2}}{K_{4}\left(t^{*}\right)+P_{4}} \leq 1$ and $\frac{x}{\kappa_{1}+x}$ increase when $x=\mathcal{P}_{0}$.

Thus, the existence of $\mathcal{P}_{1}$ is guaranteed by the condition

## Hypothesis 3.3

$$
\frac{\mathcal{V}_{1} \mathcal{P}_{0}}{\kappa_{1}+\mathcal{P}_{0}}+\mathcal{V}_{4}<\min _{t \in \mathbb{R}}\left\{V_{2}(t)+V_{3}(t)\right\}
$$

The remaining upper bounds are obtained as follows. In the first place, define a new variable $Q:=P_{N}+P_{2}$ which satisfies the equation:

$$
\frac{d Q}{d t}=\frac{V_{3}(t) P_{1}(t)}{K_{3}(t)+P_{1}(t)}-P_{2}(t)\left(\frac{V_{4}(t)}{K_{4}(t)+P_{2}(t)}+\frac{V_{d}(t)}{K_{d}(t)+P_{2}(t)}\right) .
$$

If $Q$ achieves its absolute maximum at $t^{*}$, then

$$
\frac{\mathcal{V}_{3} \mathcal{P}_{1}}{\kappa_{3}+\mathcal{P}_{1}} \geq \frac{V_{3}\left(t^{*}\right) P_{1}\left(t^{*}\right)}{K_{3}\left(t^{*}\right)+P_{1}\left(t^{*}\right)}-P_{2}\left(t^{*}\right)\left(\frac{V_{4}\left(t^{*}\right)}{K_{4}\left(t^{*}\right)+P_{2}\left(t^{*}\right)}+\frac{V_{d}\left(t^{*}\right)}{K_{d}\left(t^{*}\right)+P_{2}\left(t^{*}\right)}\right) .
$$

As before, if the condition

## Hypothesis 3.4

$$
\frac{\mathcal{V}_{3} \mathcal{P}_{1}}{\kappa_{3}+\mathcal{P}_{1}}<\min _{t \in \mathbb{R}}\left(V_{4}(t)+V_{d}(t)\right)
$$

is assumed, then $P_{2}\left(t^{*}\right) \leq \tilde{P}$ for some $\tilde{P}$. Moreover, from the fourth equation of the system we deduce the existence of a constant $C$ such that $\frac{d P_{2}}{d t} \geq-C P_{2}(t)$. Hence we obtain, for all $t$, that $P_{2}(t) \leq e^{C T} \tilde{P}:=\mathcal{P}_{2}$. Then $Q^{\prime}(t)$ is bounded. Besides, there exist $\hat{t}$ critical point of $\mathcal{P}_{N}$, in consequence

$$
k_{1}(\hat{t}) P_{2}(\hat{t})=k_{2}(\hat{t}) P_{N}(\hat{t})
$$

then $P_{N}(\hat{t})$ verifies:

$$
P_{N}(\hat{t}) \leq \frac{k_{1}^{*}}{k_{2 *}} \mathcal{P}_{2}
$$

thus

$$
Q(\hat{t})=P_{N}(\hat{t})+P_{2}(\hat{t}) \leq \mathfrak{Q}_{0}:=\left(\frac{k_{1}^{*}}{k_{2 *}}+1\right) \mathcal{P}_{2}
$$

In this way, knowing that $Q^{\prime} \leq \mathfrak{Q}_{1}$, by integrating up to a certain $t$ in the interval $\mathcal{J}:=[\hat{t}, \hat{t}+T]$ follows:

$$
Q(t)=Q(\hat{t})+\int_{\hat{t}}^{t} Q^{\prime}(t) \leq \mathfrak{Q}_{0}+\mathfrak{Q}_{1} \underbrace{(t-\hat{t})}_{\leq T}, t \in \mathcal{J}
$$

in this way, there is also a $\mathcal{P}_{N}$ of $P_{N}(t)$, then

$$
P_{N}(t) \leq Q(t) \leq \mathfrak{Q}_{0}+\mathfrak{Q}_{1} T:=\mathcal{P}_{N}
$$

After upper bounds are established, we proceed with the lower bounds as follows. Assume that $M$ achieves its absolute minimum at some $t_{*}$, then we use again the fact that $M^{\prime}\left(t_{*}\right)=0$ to obtain:

$$
\frac{V_{m}\left(t_{*}\right) M\left(t_{*}\right)}{K_{m_{1}}\left(t_{*}\right)+M\left(t_{*}\right)}=\frac{V_{S}\left(t_{*}\right) K_{1}\left(t_{*}\right)^{n}}{K_{1}^{n}\left(t_{*}\right)+P_{N}\left(t_{*}\right)^{n}} \geq \frac{v_{S} \kappa_{1}^{n}}{\kappa_{1}^{n}+\mathcal{P}_{N}^{n}} .
$$

As we did before:

$$
\frac{V_{m}\left(t_{*}\right) M\left(t_{*}\right)}{K_{m_{1}}\left(t_{*}\right)+M\left(t_{*}\right)} \geq \frac{v_{m} M\left(t_{*}\right)}{\kappa_{m_{1}}+M\left(t_{*}\right)}
$$

lets define the increasing and bijective function

$$
\hat{b}_{M}(x):=\frac{v_{m} x}{\kappa_{m_{1}}+x}, \hat{b}_{M}^{-1}(y):=\frac{\kappa_{m_{1}} y}{v_{m}-y}
$$

this inverse is increasing too, thus:

$$
M\left(t_{*}\right) \geq \hat{b}_{M}^{-1}\left(\frac{v_{S} \kappa_{1}^{n}}{\mathcal{K}_{1}^{n}+\mathcal{P}_{N}^{n}}\right):=\mathfrak{m}
$$

This shows that $M_{1}(t) \geq \mathfrak{m}$ for some positive constant $\mathfrak{m}$. In the same way, we find a lower bound $\mathfrak{p}_{0}$ for $P_{0}$ using the fact that

$$
\frac{V_{1}\left(t_{*}\right) P_{0}\left(t_{*}\right)}{K_{1}\left(t_{*}\right)+P_{0}\left(t_{*}\right)}=K_{s}\left(t_{*}\right) M\left(t_{*}\right)+\frac{V_{2}\left(t_{*}\right) P_{1}\left(t_{*}\right)}{K_{2}\left(t_{*}\right)+P_{1}\left(t_{*}\right)} \geq \kappa_{s} \mathfrak{m} .
$$

We know that

$$
\frac{V_{1}\left(t_{*}\right) P_{0}\left(t_{*}\right)}{K_{1}\left(t_{*}\right)+P_{0}\left(t_{*}\right)} \geq \frac{v_{1} P_{0}\left(t_{*}\right)}{\kappa_{1}+P_{0}\left(t_{*}\right)},
$$

this function is increasing and its inverse is also increasing:

$$
\hat{b}_{1}(x):=\frac{v_{1} x}{\kappa_{1}+x} \rightarrow \hat{b}_{1}^{-1}(y):=\frac{\kappa_{1} y}{v_{1}-y},
$$

therefore, it is defined $\mathfrak{p}_{0}:=\hat{b}_{1}^{-1}\left(\kappa_{s} \mathfrak{m}\right)$.

Next, suppose that $P_{1}$ achieves its absolute minimum at $t_{*}$, then

$$
P_{1}\left(t_{*}\right)\left(\frac{V_{2}\left(t_{*}\right)}{K_{2}\left(t_{*}\right)+P_{1}\left(t_{*}\right)}+\frac{V_{3}\left(t_{*}\right)}{K_{3}\left(t_{*}\right)+P_{1}\left(t_{*}\right)}\right)>\frac{V_{1}\left(t_{*}\right) P_{0}\left(t^{*}\right)}{K_{1}\left(t_{*}\right)+P_{0}\left(t_{*}\right)} \geq \frac{v_{1} \mathfrak{p}_{0}}{\mathcal{K}_{1}+\mathfrak{p}_{0}}>0
$$

which yields the existence of a positive lower bound $\mathfrak{p}_{1}:=\frac{v_{1} \mathfrak{p}_{0}}{\mathcal{K}_{1}+\mathfrak{p}_{0}}$. Finally, positive lower bounds for $P_{2}$ and $P_{N}$ are obtained by means of the function $Q=P_{2}+P_{N}$. Indeed, if $Q$ achieves its absolute minimum at some $t_{*}$, then

$$
P_{2}\left(t_{*}\right)\left(\frac{V_{4}\left(t_{*}\right)}{K_{4}\left(t_{*}\right)+P_{2}\left(t_{*}\right)}+\frac{V_{d}\left(t_{*}\right)}{K_{d}\left(t_{*}\right)+P_{2}\left(t_{*}\right)}\right) \geq \frac{v_{3} \mathfrak{p}_{1}}{\mathcal{K}_{3}+\mathfrak{p}_{1}}
$$

and we deduce that $P_{2}\left(t_{*}\right)$ cannot be arbitrarily small. As before, using the fact that $P_{2}^{\prime} \geq-C P_{2}$ it is seen that $P_{2}(t) \geq e^{-C T} P_{2}\left(t_{*}\right)$ and the conclusion follows. This, in turn, yields a lower bound $\mathfrak{p}_{N}>0$ for $P_{N}$.

## 4 Main Theorem

We are already in conditions of defining the open set $\Omega \subset \mathcal{K}^{\circ}$ as

$$
\begin{aligned}
& \Omega:=( \\
&\left(M, P_{0}, P_{1}, P_{2}, P_{N}\right) \in C_{T}: \mathfrak{m}<M(t)<\mathcal{M}, \mathfrak{p}_{0}<P_{0}(t)<\mathcal{P}_{0}, \\
&\left.\mathfrak{p}_{1}<P_{1}(t)<\mathcal{P}_{1}, \mathfrak{p}_{2}<P_{2}(t)<\mathcal{P}_{2}, \mathfrak{p}_{N}<P_{N}(t)<\mathcal{P}_{N}\right\} .
\end{aligned}
$$

Theorem 4.1 Assume that the previous conditions (3.1), (3.2), (3.3) and (3.4) hold. Then problem (1) has at least one positive $T$-periodic solution.

Proof. In the previous section, the first condition of the continuation theorem was verified. It remains to prove that $b$ ) and $c$ ) are fulfilled as well. With this aim, set $\mathcal{Q}:=\Omega \cap \mathbb{R}^{5}$ and recall that the function $\phi: \overline{\mathcal{Q}} \rightarrow \mathbb{R}^{5}$ is defined by $\phi(x)=\overline{N x}$. We claim that each coordinate $\phi_{j}$ has different signs at the corresponding opposite faces of $\mathcal{Q}$.

Indeed, compute for example $\phi_{1}\left(\mathcal{M}, P_{0}, P_{1}, P_{2}, P_{N}\right)$ and $\phi_{1}\left(\mathfrak{m}, P_{0}, P_{1}, P_{2}, P_{N}\right)$ for $\mathfrak{p}_{j} \leq$ $P_{j} \leq \mathcal{P}_{j}$ :

$$
\begin{aligned}
\phi_{1}\left(\mathcal{M}, P_{0}, P_{1}, P_{2}, P_{N}\right) & =\frac{1}{T} \int_{0}^{T}\left(\frac{V_{S}(t) K_{1}(t)^{n}}{K_{1}^{n}(t)+P_{N}}-\frac{V_{m}(t) \mathcal{M}}{K_{m_{1}}(t)+\mathcal{M}}\right) d t \\
& <\mathcal{V}_{S}-\frac{v_{m} \mathcal{M}}{\mathcal{K}_{m_{1}}+\mathcal{M}}=0 \\
\phi_{1}\left(\mathfrak{m}, P_{0}, P_{1}, P_{2}, P_{N}\right) & =\frac{1}{T} \int_{0}^{T}\left(\frac{V_{S}(t) K_{1}(t)^{n}}{K_{1}^{n}(t)+P_{N}}-\frac{V_{m}(t) \mathfrak{m}}{K_{m_{1}}(t)+\mathfrak{m}}\right) d t \\
> & \frac{v_{S} \kappa_{1}^{n}}{\mathcal{K}_{1}^{n}+\mathcal{P}_{N}^{n}}-\frac{\mathcal{V}_{m} \mathfrak{m}}{\kappa_{m_{1}}+\mathfrak{m}} \geq 0
\end{aligned}
$$

provided that $\mathfrak{m}$ is small enough. In the same way, making the lower bounds smaller if necessary, we deduce that

$$
\begin{aligned}
& \phi_{2}\left(M, \mathcal{P}_{0}, P_{1}, P_{2}, P_{N}\right)<0<\phi_{2}\left(M, \mathfrak{p}_{0}, P_{1}, P_{2}, P_{N}\right) \\
& \phi_{3}\left(M, P_{0}, \mathcal{P}_{1}, P_{2}, P_{N}\right)<0<\phi_{3}\left(M, p_{0}, \mathfrak{p}_{1}, P_{2}, P_{N}\right)
\end{aligned}
$$

$$
\begin{aligned}
& \phi_{4}\left(M, P_{0}, P_{1}, \mathcal{P}_{2}, P_{N}\right)<0<\phi_{4}\left(M, p_{0}, p_{1}, \mathfrak{p}_{2}, P_{N}\right), \\
& \phi_{5}\left(M, P_{0}, P_{1}, P_{2}, \mathcal{P}_{N}\right)<0<\phi_{5}\left(M, p_{0}, p_{1}, p_{2}, \mathfrak{p}_{N}\right) .
\end{aligned}
$$

Thus, condition b) of the continuation theorem is verified. Moreover, we may define a homotopy as follows. Consider the center of $\mathcal{Q}$ given by

$$
\wp:=\left(\frac{\mathcal{M}+\mathfrak{m}}{2}, \frac{\mathcal{P}_{0}+\mathfrak{p}_{0}}{2}, \frac{\mathcal{P}_{1}+\mathfrak{p}_{1}}{2}, \frac{\mathcal{P}_{2}+\mathfrak{p}_{2}}{2}, \frac{\mathcal{P}_{N}+\mathfrak{p}_{N}}{2}\right)
$$

and the function $\mathcal{H}: \overline{\mathcal{Q}} \times[0 ; 1] \rightarrow \mathbb{R}^{5}$ given by

$$
\mathcal{H}(x, \lambda)=(1-\lambda)(\wp-x)+\lambda \phi .
$$

We need to verify that $\mathcal{H}$ does not vanish at $\partial \mathcal{Q}$. To this end, suppose, for example, that $\mathcal{H}\left(\mathcal{M}, P_{0}, P_{1}, P_{2}, P_{N}\right)=0$ for some $\hat{\lambda} \in[0 ; 1]$, then

$$
0=\mathcal{H}_{1}(\mathcal{M}, \hat{\lambda})=(1-\hat{\lambda}) \underbrace{\left(\frac{\mathcal{M}+\mathfrak{m}}{2}-\mathcal{M}\right)}_{<0}+\hat{\lambda} \underbrace{\phi_{1}\left(\mathcal{M}, P_{0}, P_{1}, P_{2}, P_{N}\right)}_{<0}<0
$$

which is a contradiction. All the remaining cases follow in an analogous way. By the homotopy invariance of the Brouwer degree, it follows that

$$
\operatorname{deg}_{B}(\phi, \mathcal{Q}, 0)=\operatorname{deg}_{B}(\wp-I, \mathcal{Q}, 0)=(-1)^{5} \neq 0 .
$$

This proves the third condition of the continuation theorem and, therefore, the existence of a $T$-periodic solution is deduced.

## 5 Conclusion

Topological degree was used for proving existence of stable equilibrium in a generic model of circadian cycle. This theory allowed to demonstrate the existence of positive periodic solutions when parameters are replaced by fixed periodic functions. The relevance of finding periodic solutions in biological models relies mainly on the fact that periodic functions represent natural cycles, such as hormonal processes.

We show that topological degree can be successfully applied to find positive periodic orbits for some of these models in the non-autonomous case. It is worthy mentioning that, for diverse biological cycles, the behaviour is characterized by models with periodic parameters; thus, the present paper provides a useful mathematical tool to understand such models.

For future work, it would be interesting to consider a more general situation, in which the parameters are not periodic but almost-periodic functions, which attracted the attention of many researchers in the last decades. Here, the topological degree cannot be used anymore because of the lack of compactness of the associated operator; thus, a different approach is required, such as the use of fixed points in cones under monotonicity conditions that avoid the compactness assumption.

## Acknowledgment

This work was partially supported by project UBACyT 20020120100029BA. I am grateful to Dr. Pablo Amster for his helpful comments.

## References

[1] Amster, P. Topological Methods in the Study of Boundary Value Problems. New York: Springer, 2014.
[2] Amster, P. and Idels, L. Existence theorems for some abstract nonlinear non-autonomous systems with delays. Commun. Nonlinear Sci. Numer. Simulat. 19 (2014) 2974-2982.
[3] Goldbeter, A. A model for circadian oscillations in Drosophila Period Protein (PER). Biological Sciences 261 (1362) (1995) 319-324.
[4] Gonze, D., Leloup, J.C. and Goldbeter, A. Theoretical models for circadian rhythms in Neurospora and Drosophila. C.R. Acad. Sci. Paris, Sciences de la vie / Life Sciences 323 (2000) 57-67.
[5] Griffith, J. S. Mathematics of Cellular Control Processes I. Negative Feedback to One Gene J. Theoret. Biol. 20 (1968) 202-208.
[6] Majercak, J., Sidote, D., Hardin, P.E. and Edery, I. How a Circadian Clock Adapts to Seasonal Decreases in Temperature and Day Length Neuron 24 (1999) 219-230.
[7] Paetkau, V., Edwards, R. and Illner, R. A Model for generating circadian rithm by coupling ultradian oscillators. Theoretical Biology and Medical Modelling 3 (12) (2006) 1-10.

# Solution of 2D Fractional Order Integral Equations by Bernstein Polynomials Operational Matrices 

M. Asgari ${ }^{1}$, R. Ezzati ${ }^{1, *}$ and H. Jafari ${ }^{2,3}$<br>${ }^{1}$ Department of Mathematics, Karaj Branch, Islamic Azad University, Karaj, Iran<br>${ }^{2}$ Department of Mathematical Sciences, University of South Africa, UNISA003, South Africa<br>${ }^{3}$ Department of Mathematics, University of Mazandaran, Babolsar, Iran

Received: October 11, 2017; Revised: December 12, 2018


#### Abstract

In this paper, we construct a new two-dimensional Bernstein polynomials operational matrix for solving 2 -dimensional fractional order Volterra integral equations (2DFOVIE). By using this operational matrix, we reduce the original problem to a linear or nonlinear system of algebraic equations. We present some numerical examples to show the efficiency of the proposed method.


Keywords: two-dimensional fractional integral equations; two-dimensional Bernstein polynomials; block pulse operational matrix; operational matrix of integration.

Mathematics Subject Classification (2010): 26A33, 45G05.

## 1 Introduction

In the last few decades, various engineering and scientific problems involving fractional calculus were discussed. For example, electrochemical process [1, 2], earthquakes [3], economics [4], bioengineering [5], orthogonal splin collocation [6] and fractional optimal control problems $[7,8]$. There are several analytical and numerical methods for solving one-dimensional and two-dimensional differential and integral equations of fractional order such as the Adomian decomposition [9], Variational iteration method [10, 11], Transform method [12], Homotopy perturbation method [13], and the methods of Harr and Chebyshev wavelet $[14,15]$ and Bernstein polynomials $[16,17]$.

The Bernstein polynomials play a conspicuous role in several areas of mathematics. These polynomials have been commonly used in the solution of differential equations,

[^1]integral equations, fractional optimal control problems and approximation theory $[7,8$, 17-23]. In this work, we consider the following type of 2DVIEFO
\[

$$
\begin{equation*}
u(x, y)-I_{0}^{q} u^{p}(x, y)=g(x, y), \quad q=(\alpha, \beta) \in(0, \infty) \times(0, \infty) \tag{1}
\end{equation*}
$$

\]

where $g(x, y)$ is a known function and $I_{0}^{q} u(x, y)$ is the left-sided mixed Riemann-Liouville integral of order $q$ which is defined as [24]

$$
\begin{equation*}
\left(I_{0}^{q} u\right)(x, y)=\frac{1}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}(x-\xi)^{\alpha-1}(y-\tau)^{\beta-1} u(\xi, \tau) d \tau d \xi \tag{2}
\end{equation*}
$$

Note: For $\alpha>0$, the Riemann-Liouville integral $\left(I^{\alpha}\right)$ on the Lebesgue space $L^{1}[a, b]$ is defined as

$$
\begin{equation*}
\left(I_{0}^{\alpha} u\right)(t)=\left(I^{\alpha} u\right)(t)=\frac{1}{\Gamma(\alpha)} \int_{0}^{t}(t-\tau)^{\alpha-1} u(\tau) d \tau \tag{3}
\end{equation*}
$$

In particular, for (2), we have

1. $\left(I_{0}^{0} u\right)(x, y)=u(x, y)$,
2. $\left(I_{0}^{\sigma} u\right)(x, y)=\int_{0}^{x} \int_{0}^{y} u(\xi, \tau) d \tau d \xi, \quad(x, y) \in J, \sigma=(1,1)$,
3. $\left(I_{0}^{r} u\right)(x, 0)=\left(I_{0}^{r} u\right)(0, y)=0, \quad x \in[0, a], y \in[0, b]$,
4. $I_{0}^{r} x^{\lambda} y^{\omega}=\frac{\Gamma(1+\lambda) \Gamma(1+\omega)}{\Gamma(1+\lambda+\alpha) \Gamma(1+\omega+\beta)} x^{\lambda+\alpha} y^{\omega+\beta}, \quad(x, y) \in J, \quad \lambda, \omega \in(-1, \infty)$.

We are looking for $u \in L^{1}(J), J:=[0, a] \times[0, b]$. The existence and uniqueness of (1) is investigated in [25].

We want to obtain the numerical solution of (1) by using two-dimensional Bernstein polynomials and block pulse functions. The rest of this paper is organized as follows. First, we briefly review some general concepts concerning one-dimensional and two-dimensional Bernstein polynomials, block pulse functions and derive the Bernstein polynomials operational matrix of two-dimensional integration of fractional order. In Section 3, the method is applied to solve linear or nonlinear 2DVIEFO. Section 4 exhibits an error estimation for the presented method. Section 5 illustrates several numerical examples to show the convergence and accuracy of the proposed method.

## 2 Bernstein Polynomials and Block Pulse Functions

### 2.1 One dimensional Bernstein polynomials (1D-BPs)

The $n$th degree Bernstein polynomials $(B P s)$ on the interval $[0,1]$ are defined as

$$
\begin{equation*}
B_{i, n}(\tau)=\binom{n}{i} \tau^{i}(1-\tau)^{n-i}, \quad 0 \leq i \leq n \tag{4}
\end{equation*}
$$

The $B P s$ on $[0,1]$ have the following properties [7]:

1. $B_{i, n}(\tau) \geq 0, i=0,1, \ldots, n, \quad \tau \in[0,1]$,
2. $\sum_{i=0}^{n} B_{i, n}(t)=1$,
3. $B_{i, n}(\tau)=(1-\tau) B_{i, n-1}(\tau)+\tau B_{i-1, n-1}(\tau), \quad i=0,1, \ldots, n$,
4. $B_{i, n}(\tau)=\sum_{k=0}^{n-i}(-1)^{k}\binom{n}{i}\binom{n-i}{k} \tau^{i+k}, \quad i=0,1, \ldots, n$.

Theorem 2.1 [26] Suppose that $H=L^{2}[0,1]$ is a Hilbert space with the inner product and $X=\operatorname{Span}\left\{B_{0, n}(t), B_{1, n}(t), \ldots, B_{n, n}(t)\right\}$ is a closed subspace with finite dimensions, therefore $X$ is a complete subspace of $H$. So, if $u \in H$ is an arbitrary element, it has a unique best approximation out of $X$ such as $x_{0}$, that is

$$
\begin{equation*}
\exists x_{0} \in Y \quad \text { s.t. } \forall x \in X, \quad\left\|u-x_{0}\right\|_{2} \leq\|u-x\|_{2} \tag{5}
\end{equation*}
$$

where $\|u\|_{2}=\sqrt{<u, u>},\langle u, v\rangle=\int_{0}^{1} u(\tau) v(\tau) d \tau$.
Thus, there exist unique coefficients $c_{0}, c_{1}, \ldots, c_{n}$ such that

$$
\begin{equation*}
u(t) \simeq x_{0}=\sum_{i=0}^{n} c_{i} B_{i, n}(t)=c^{T} \varphi(t) \tag{6}
\end{equation*}
$$

where $c^{T}=\left[c_{0}, c_{1}, \ldots, c_{n}\right], \quad \varphi(\tau)=\left[B_{0, n}(\tau), B_{1, n}(\tau), \ldots, B_{n, n}(\tau)\right]^{T}$.
Lemma 2.1 If $\varphi_{n}(\tau)=\left[B_{0, n}(\tau), B_{1, n}(\tau), \ldots, B_{n, n}(\tau)\right]^{T}$ is a complete basis, then $\varphi_{n}(t)=A T_{n}(t)$, where $A$ is an $(n+1) \times(n+1)$ upper triangular matrix with

$$
a_{i+1, j+1}= \begin{cases}(-1)^{j-i}\binom{n}{i}\binom{n-i}{j-i}, & i \leqslant j  \tag{7}\\ 0, & i>j\end{cases}
$$

for $i, j=0,1, \ldots, n$ and $T_{n}(\tau)=\left[1, \tau, \tau^{2}, \ldots, \tau^{n}\right]^{T}$.

### 2.2 BPF and operational matrix

A set of BPF on $[0,1)$ is defined as follows:

$$
b_{i}(t)=\left\{\begin{array}{l}
1, \quad \frac{i}{m} \leq t<\frac{i+1}{m}, \quad i, j=0,1, \ldots, m-1,  \tag{8}\\
0, \quad \text { otherwise }
\end{array}\right.
$$

The above functions are orthogonal and disjoint, i.e.
$b_{i}(t) b_{j}(t)=\left\{\begin{array}{ll}b_{i}(t) & i=j, \\ 0 & i \neq j,\end{array}\right.$ and $\int_{0}^{1} b_{i}(t) b_{j}(t) d t=\frac{1}{m} \delta_{i j}$, where $\delta_{i j}$ is the Kronecker delta.

If $B_{m}(\tau)=\left[b_{0}(\tau), b_{1}(\tau), \ldots, b_{m-1}(\tau)\right]^{T}$, the block pulse operational matrix of the fractional order integration $F^{\alpha}$ is [27]

$$
I^{\alpha} B_{m}(\tau)=F^{\alpha} B_{m}(\tau)
$$

where

$$
F^{\alpha}=\frac{1}{m^{\alpha}} \frac{1}{\Gamma(\alpha+2)}\left[\begin{array}{cccccc}
1 & \xi_{1} & \xi_{2} & \xi_{3} & \ldots & \xi_{m-1}  \tag{9}\\
0 & 1 & \xi_{1} & \xi_{2} & \ldots & \xi_{m-2} \\
0 & 0 & 1 & \xi_{1} & \ldots & \xi_{m-3} \\
\vdots & \vdots & \ddots & \ddots & \ddots & \vdots \\
0 & 0 & \ldots & 0 & 1 & \xi_{1} \\
0 & 0 & \ldots & 0 & 0 & 1
\end{array}\right]
$$

with $\xi_{s}=(s+1)^{\alpha+1}-2 s^{\alpha+1}+(s-1)^{\alpha+1}$.

### 2.3 Operational matrix for fractional integral equation(1D)

If

$$
\varphi(\tau)=\varphi_{n}(\tau)=\left[B_{0, n}(\tau), B_{1, n}(\tau), \ldots, B_{n, n}(\tau)\right]^{T}
$$

then for the fractional integral equation (3), we have

$$
\begin{equation*}
I^{\alpha} \varphi_{n}(\tau)=P^{\alpha} \varphi_{n}(\tau) \tag{10}
\end{equation*}
$$

with $n=m-1$, the Bernstein polynomial might be expanded into an $m$-term BPF as

$$
\begin{equation*}
\varphi_{m}(\tau)=\phi_{m \times m} B_{m}(\tau), \tag{11}
\end{equation*}
$$

now

$$
\begin{equation*}
I^{\alpha} \varphi_{m}(\tau)=I^{\alpha} \phi_{m \times m} B_{m}(\tau)=\phi_{m \times m} I^{\alpha} B_{m}(\tau)=\phi_{m \times m} F^{\alpha} B_{m}(\tau) . \tag{12}
\end{equation*}
$$

From equations (11) and (12), we have

$$
\begin{equation*}
I^{\alpha} \varphi_{m}(\tau)=\phi_{m \times m} F^{\alpha} B_{m}(\tau)=\phi_{m \times m} F^{\alpha} \phi_{m \times m}^{-1} \varphi_{m}(\tau) \tag{13}
\end{equation*}
$$

Therefore,

$$
\begin{equation*}
P_{m \times m}^{\alpha}=\phi_{m \times m} F^{\alpha} \phi_{m \times m}^{-1} . \tag{14}
\end{equation*}
$$

$P^{\alpha}$ is called an operational matrix for fractional integration based on the Bernstein polynomials [28].

### 2.4 Two-dimensional Bernstein polynomials (2D-BPs)

The Bernstein polynomials of degree $m n$ on the interval $[0,1] \times[0,1]$ are defined by

$$
\begin{equation*}
B_{(i, m)(j, n)}(\mu, \nu)=\binom{m}{i}\binom{n}{j} \mu^{i}(1-\mu)^{m-i} \nu^{j}(1-\nu)^{n-j} \tag{15}
\end{equation*}
$$

for $i=0,1, \ldots, m, j=0,1, \ldots, n$.
Similar to the 1D case, we have [19]:

1. $B_{(i, m)(j, n)}(\mu, \nu) \geq 0$,
2. $B_{(i, m)(j, n)}(\mu, \nu)=B_{(i, m)}(\mu) B_{(j, n)}(\nu)$,
3. $B_{(i, m)(j, n)}(\mu, \nu)=\sum_{k=0}^{m-i} \sum_{t=0}^{n-j}(-1)^{r+t}\binom{m}{i}\binom{n}{j}\binom{m-i}{k}\binom{n-j}{t} \mu^{i+k} \nu^{j+t}$,
4. $Q=<B_{(i, m)(j, n)}(\mu, \nu), B_{(k, m)(t, n)}(\mu, \nu)>$
$=\int_{0}^{1} \int_{0}^{1} B_{(i, m)(j, n)}(\mu, \nu) B_{(k, m)(t, n)}(\mu, \nu) d \mu d \nu=\frac{\binom{m}{i}\binom{n}{j}\binom{m}{k}\binom{n}{t}}{(2 m+1)(2 n+1)\binom{2 m}{i+k}\binom{2 n}{j+t}}$,
for $i, k=0,1, \ldots, m, \quad j, t=0,1, \ldots, n$.
Now, if we define $(m+1) \times(n+1)$-vector

$$
\begin{gather*}
\varphi_{m n}(\mu, \nu)=\left[B_{(0, m)(0, n)}(\mu, \nu), \ldots, B_{(0, m)(n, n)}(\mu, \nu)\right. \\
\left.\ldots, B_{(m, m)(0, n)}(\mu, \nu), \ldots, B_{(m, m)(n, n)}(\mu, \nu)\right]^{T} \tag{16}
\end{gather*}
$$

where $(\mu, \nu) \in[0,1] \times[0,1]$, then $\varphi_{m n}(\mu, \nu)$ is a complete basis.

### 2.5 Function expansion with 2D-BPs

We expand $u(\mu, \nu) \in L^{2}([0,1] \times[0,1])$ by 2D-BPs as

$$
\begin{equation*}
u(\mu, \nu)=\sum_{i=0}^{\infty} \sum_{j=0}^{\infty} u_{i j} \varphi_{i j}(\mu, \nu) \simeq \sum_{i=0}^{m} \sum_{j=0}^{n} u_{i j} \varphi_{i j}(\mu, \nu)=U^{T} \varphi(\mu, \nu)=\varphi^{T}(\mu, \nu) U \tag{17}
\end{equation*}
$$

where $\varphi(\mu, \nu)$ and $U$ are $(m+1)(n+1)$ vectors. Components $u_{i j}$ of $U$ are obtained as

$$
\begin{equation*}
u_{i j}=<u(\mu, \nu), \varphi(\mu, \nu)>=\int_{0}^{1} \int_{0}^{1} u(\mu, \nu) B_{(i, m)(j, n)}(\mu, \nu) d \mu d \nu \tag{18}
\end{equation*}
$$

Similarly, let $k(\mu, \nu, s, t)$ be defined on $[0,1] \times[0,1] \times[0,1] \times[0,1]$. It can be expanded with respect to $2 \mathrm{D}-\mathrm{BPs}$ as

$$
\begin{equation*}
k(\mu, \nu, s, t) \simeq \varphi^{T}(\mu, \nu) K \psi(s, t) \tag{19}
\end{equation*}
$$

where $\varphi(\mu, \nu)$ and $\psi(s, t)$ are 2D-BPs vectors of dimension $\left(m_{1}+1\right)\left(n_{1}+1\right)$ and $\left(m_{2}+\right.$ 1) $\left(n_{2}+1\right)$, respectively, and $K$ is the $\left(m_{1}+1\right)\left(n_{1}+1\right) \times\left(m_{2}+1\right)\left(n_{2}+1\right)$ two-dimensional Bernstein polynomials coefficient matrix.

### 2.6 Operational matrix for fractional integral equation(2D)

Suppose $B_{(i, m)}(\mu)=A_{1} T_{m}(\mu)$ and $B_{(j, n)}(\nu)=A_{2} T_{n}(\nu)$. Then

$$
\varphi_{m n}(\mu, \nu)=M T_{m n}(\mu, \nu)
$$

where

$$
T_{m n}(\mu, \nu)=\left[1, \nu, \nu^{2}, \ldots, \nu^{n}, \mu, \mu \nu, \ldots, \mu \nu^{n}, \ldots, \mu^{m}, \mu^{m} \nu, \ldots, \mu^{m} \nu^{n}\right]^{T}
$$

and $M=A_{1} \otimes A_{2}$ and $\otimes$ denotes the Kronecker product.
Now, we present two-dimensional Bernstein polynomials operational matrices of fractional mode. Let $\varphi_{m n}(\mu, \nu)$ be defined as in (16). The fractional integration of the $\varphi_{m n}(\mu, \nu)$ can be approximately obtained as

$$
\begin{equation*}
\frac{1}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}(x-\xi)^{\alpha-1}(y-\tau)^{\beta-1} \varphi_{m n}(\xi, \tau) d \xi d \tau \simeq P^{r} \varphi_{m n}(\mu, \nu) \tag{20}
\end{equation*}
$$

where $P^{r}$ is a $(m+1)(n+1) \times(m+1)(n+1)$ matrix and is called an operational matrix. Let operational matrices $P^{\alpha}$ and $P^{\beta}$ satisfy (14), i.e.

$$
\begin{gather*}
I^{\alpha} \varphi_{m}(\mu)=P^{\alpha} \varphi_{m}(\mu)=\phi_{m \times m} F^{\alpha} \phi_{m \times m}^{-1} \varphi_{m}(\mu) \\
I^{\beta} \varphi_{n}(\nu)=P^{\beta} \varphi_{n}(\nu)=\phi_{n \times n} F^{\beta} \phi_{n \times n}^{-1} \varphi_{n}(\nu) . \tag{21}
\end{gather*}
$$

From the disjointness property of two-dimensional Bernstein polynomials, we get

$$
\begin{gathered}
\frac{1}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}(x-\xi)^{\alpha-1}(y-\tau)^{\beta-1} \varphi_{m n}(\xi, \tau) d \xi d \tau=\frac{1}{\Gamma(\alpha)} \int_{0}^{x}(x-\xi)^{\alpha-1} \varphi_{m}(\xi) d \xi \\
\times \frac{1}{\Gamma(\beta)} \int_{0}^{y}(y-\tau)^{\beta-1} \varphi_{n}(\tau) d \tau
\end{gathered}
$$

By using (21), we have

$$
\begin{equation*}
P^{r}=P^{\alpha} \otimes P^{\beta} \tag{22}
\end{equation*}
$$

### 2.7 Product operational matrix

In view of (1), we have $u^{p}(x, y)$. So, we need to evaluate the product of $\left.\varphi_{( } x, y\right)$ and $\varphi^{T}(x, y)$, which is called the product matrix.

Lemma 2.2 Suppose that $C_{(m+1)(n+1)}$ is an arbitrary vector. The operational matrix of product $\hat{C}_{(m+1)(n+1) \times(m+1)(n+1)}$ using BPs can be given as follows [29] :

$$
\begin{equation*}
\varphi(x, y) \varphi^{T}(x, y) C \simeq \hat{C}^{T} \varphi(x, y) \tag{23}
\end{equation*}
$$

Corollary 2.1 Suppose $u(x, y)=U^{T} \varphi(x, y)=\varphi^{T}(x, y) U$ and $\hat{U}$ is the operational matrix of product. Then

$$
\begin{equation*}
(u(x, y))^{k}=\varphi^{T}(x, y) \bar{U}_{k} \tag{24}
\end{equation*}
$$

where $k \in N$ and $\bar{U}_{k}=\hat{U}^{k-1} U$.
Proof. By using Lemma 2.2, for $k=2$, we get

$$
(u(x, y))^{2}=U^{T} \varphi(x, y) \varphi^{T}(x, y) U=\varphi^{T}(x, y) \hat{U} U=\varphi^{T}(x, y) \bar{U}_{2} .
$$

Also, if $k=3$,

$$
(u(x, y))^{3}=U^{T} \varphi(x, y) \varphi^{T}(x, y) \hat{U} U=\varphi^{T}(x, y) \hat{U}^{2} U=\varphi^{T}(x, y) \bar{U}_{3} .
$$

So, by induction we have

$$
(u(x, y))^{k}=U^{T} \varphi(x, y) \varphi^{T}(x, y) \hat{U}^{k-2} U=\varphi^{T}(x, y) \hat{U}^{k-1} U=\varphi^{T}(x, y) \bar{U}_{k}
$$

## 3 Solving 2DFOVIE

In this section, two-dimensional Bernstein polynomials are applied to solve equation(1). Using the procedures mentioned in Section 2, we approximate functions $(u(x, y))^{p}$, $k(x, y, s, t)$ and $f(x, y)$ as follows:

$$
\begin{align*}
(u(x, y))^{p} & =\varphi^{T}(x, y) \bar{U}_{p}=\bar{U}_{p}^{T} \varphi(x, y) \\
f(x, y) & =\varphi^{T}(x, y) F=F^{T} \varphi(x, y)  \tag{25}\\
k(x, y, s, t) & =\varphi^{T}(x, y) K \varphi(x, y)
\end{align*}
$$

where the $(m+1)(n+1) \times 1$ vectors $\bar{U}_{p}, F$ and $(m+1)(n+1) \times(m+1)(n+1)$ matrix $K$ are 2D-BPs coefficients of $(u(x, y))^{p}, f(x, y)$ and $k(x, y, s, t)$ respectively. Substituting equations(25) in equation(1), we have:
$\varphi^{T}(x, y) U-\frac{1}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}(x-s)^{\alpha-1}(y-t)^{\beta-1} \varphi^{T}(x, y) K \varphi(s, t) \varphi^{T}(s, t) \bar{U}_{p} d t d s=\varphi^{T}(x, y) F$.
By using (23), we get

$$
\varphi^{T}(x, y) U-\frac{\varphi^{T}(x, y) K \hat{\bar{U}}_{p}^{T}}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}(x-s)^{\alpha-1}(y-t)^{\beta-1} \varphi(s, t) d t d s=\varphi^{T}(x, y) F
$$

From equation(20) and the above equation, we obtain

$$
\varphi^{T}(x, y) U-\varphi^{T}(x, y) K \hat{\bar{U}}_{p}^{T} P^{r} \varphi(x, y)=\varphi^{T}(x, y) F,
$$

or

$$
\begin{equation*}
U-K \hat{\bar{U}}_{p}^{T} P^{r} \varphi(x, y)=F \tag{26}
\end{equation*}
$$

Now, we collocate equation(26) in $(m+1)(n+1)$ Newton-Cotes nodes as

$$
x_{i}=\frac{2 i-1}{2(m+1)}, \quad y_{j}=\frac{2 j-1}{2(n+1)}, \quad i=1,2, \ldots, m+1, \quad j=1,2, \ldots, n+1 .
$$

So, we have a linear $(p=1)$ or nonlinear $(p \geq 1)$ algebraic system

$$
\begin{equation*}
U-B \psi=F \tag{27}
\end{equation*}
$$

where $B=K \hat{\bar{U}}_{p}^{T} P^{r}$, and

$$
\psi=\left[\varphi\left(x_{1}, y_{1}\right), \varphi\left(x_{1}, y_{2}\right), \ldots, \varphi\left(x_{1}, y_{n+1}\right), \ldots \varphi\left(x_{m+1}, y_{1}\right), \ldots, \varphi\left(x_{m+1}, y_{n+1}\right)\right]^{T}
$$

## 4 Error analysis

Theorem 4.1 Suppose $u(x, y)$ is an exact solution of the equation (1) and $\hat{u}(x, y)$ shows its approximate solution by Bernstein polynomials, and

1. $\left|(x-\xi)^{\alpha-1}(y-\tau)^{\beta-1} k(x, y, \xi, \tau)\right|<C$,
2. $(u(x, y))^{p}$ is a Lipschitz continuous function, i.e.

$$
\left|(u(x, y))^{p}-(\hat{u}(x, y))^{p}\right| \leq L|u(x, y)-\hat{u}(x, y)|
$$

where $L$ is a Lipschitz constant
3. $m_{1}=m_{2}=m$.

Then $\hat{u}(x, y)$ converges to $u(x, y)$, if $0<\frac{L C}{\Gamma(\alpha) \Gamma(\beta)}<1$.
Proof.

$$
\|u(x, y)-\hat{u}(x, y)\|_{\infty}=\max _{0 \leq x, y \leq 1}|u(x, y)-\hat{u}(x, y)|
$$

$$
=\max _{0 \leq x, y \leq 1}\left|\frac{1}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}(x-\xi)^{\alpha-1}(y-\tau)^{\beta-1} k(x, y, \xi, \tau)\left((u(\xi, \tau))^{p}-(\hat{u}(\xi, \tau))^{p}\right) d \xi d \tau\right|
$$

$$
\leq \max _{0 \leq x, y \leq 1} \frac{1}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}\left|(x-\xi)^{\alpha-1}(y-\tau)^{\beta-1} k(x, y, \xi, \tau)\right|\left|(u(\xi, \tau))^{p}-(\hat{u}(\xi, \tau))^{p}\right| d \xi d \tau
$$

$$
\leq \max _{0 \leq x, y \leq 1} \frac{C L}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}|u(\xi, \tau)-\hat{u}(\xi, \tau)| d \xi d \tau
$$

$$
\leq \frac{C L x y}{\Gamma(\alpha) \Gamma(\beta)}\|u(\xi, \tau)-\hat{u}(\xi, \tau)\|_{\infty} \leq \frac{C L}{\Gamma(\alpha) \Gamma(\beta)}\|u(\xi, \tau)-\hat{u}(\xi, \tau)\|_{\infty}
$$

Therefore we get

$$
\begin{equation*}
\|u(x, y)-\hat{u}(x, y)\|_{\infty} \leq \frac{C L}{\Gamma(\alpha) \Gamma(\beta)}\|u(\xi, \tau)-\hat{u}(\xi, \tau)\|_{\infty} \tag{28}
\end{equation*}
$$

Equation (28) shows that if $0<\frac{L C}{\Gamma(\alpha) \Gamma(\beta)}<1$, then $\|u(\xi, \tau)-\hat{u}(\xi, \tau)\|_{\infty} \longrightarrow 0$.

## 5 Numerical Examples

To demonstrate the validity and applicability of this scheme, we use the present method for the following four examples. In view of (2), we rewrite (1) in the following form of 2DFOVIE:

$$
\begin{equation*}
u(x, y)-\frac{1}{\Gamma(\alpha) \Gamma(\beta)} \int_{0}^{x} \int_{0}^{y}(x-\xi)^{\alpha-1}(y-\tau)^{\beta-1} k(x, y, \xi, \tau) u^{p}(\xi, \tau) d \xi d \tau=g(x, y) \tag{29}
\end{equation*}
$$

Now, for different values of $\alpha, \beta, k(x, y, \xi, \tau), p$ and $g(x, y)$, we solve (29).
Example 5.1 Let $\alpha=\frac{5}{3}, \beta=\frac{7}{3}, k(x, y, \xi, \tau)=\xi \tau \sqrt{x y}, p=1$ and $g(x, y)=x^{3}\left(y^{2}-\right.$ $y)-\frac{x^{\frac{17}{3}} y^{\frac{13}{3}} \sqrt{x y}(9 y-16)}{5000}$ The exact solution is $u(x, y)=x^{3}\left(y^{2}-y\right)$. We applied the proposed method to solve this example for various values of $m$ and $n$. Also, we compare the numerical results with the exact solution. The results are tabulated in Table 1.

| $x=y$ | $m=n=1$ | $m=n=2$ | $m=n=3$ |
| :---: | :---: | :---: | :---: |
| 0.0 | $6.292 \times 10^{-6}$ | $3.091 \times 10^{-6}$ | $7.394 \times 10^{-6}$ |
| 0.1 | $7.702 \times 10^{-5}$ | $3.942 \times 10^{-4}$ | $4.401 \times 10^{-5}$ |
| 0.2 | $1.261 \times 10^{-3}$ | $2.814 \times 10^{-3}$ | $9.460 \times 10^{-5}$ |
| 0.3 | $5.645 \times 10^{-3}$ | $4.417 \times 10^{-3}$ | $1.492 \times 10^{-4}$ |
| 0.4 | $1.533 \times 10^{-2}$ | $3.212 \times 10^{-3}$ | $2.022 \times 10^{-4}$ |
| 0.5 | $3.121 \times 10^{-2}$ | $1.926 \times 10^{-4}$ | $2.515 \times 10^{-4}$ |
| 0.6 | $5.180 \times 10^{-2}$ | $3.579 \times 10^{-3}$ | $2.919 \times 10^{-4}$ |
| 0.7 | $7.198 \times 10^{-2}$ | $4.819 \times 10^{-3}$ | $3.020 \times 10^{-4}$ |
| 0.8 | $8.187 \times 10^{-2}$ | $2.975 \times 10^{-3}$ | $2.257 \times 10^{-4}$ |
| 0.9 | $6.556 \times 10^{-2}$ | $5.010 \times 10^{-4}$ | $5.289 \times 10^{-5}$ |

Table 1: The maximum absolute errors in Example 5.1.

Example 5.2 Let $\alpha=\beta=\frac{5}{2}, k(x, y, \xi, \tau)=\sqrt{x y \xi}, p=2$ and $f(x, y)=x \sqrt{y}-$ $\frac{1}{420} x^{\frac{11}{2}} y^{4}$ with the exact solution $u(x, y)=x \sqrt{y}$. The maximum absolute errors are shown in Table 2.

Example 5.3 Let $\alpha=\frac{5}{2}, \beta=\frac{7}{2}, k(x, y, \xi, \tau)=(y+\xi) e^{-2 \tau}, p=2$ and $f(x, y)=x e^{y}-\frac{1024 x^{\frac{9}{2}} y^{\frac{7}{2}}(6 x+11 y)}{1091475 \pi}$ with the exact solution $u(x, y)=x e^{y}$. The maximum absolute errors are shown in Table 3.

| $x=y$ | $m=n=1$ | $m=n=2$ | $m=n=3$ |
| :---: | :---: | :---: | :---: |
| 0.0 | $5.603 \times 10^{-5}$ | $2.592 \times 10^{-6}$ | $1.327 \times 10^{-5}$ |
| 0.1 | $3.064 \times 10^{-3}$ | $1.273 \times 10^{-3}$ | $1.611 \times 10^{-3}$ |
| 0.2 | $4.032 \times 10^{-3}$ | $4.748 \times 10^{-3}$ | $1.497 \times 10^{-3}$ |
| 0.3 | $1.220 \times 10^{-2}$ | $4.694 \times 10^{-3}$ | $1.253 \times 10^{-3}$ |
| 0.4 | $1.818 \times 10^{-2}$ | $1.276 \times 10^{-3}$ | $3.797 \times 10^{-3}$ |
| 0.5 | $2.009 \times 10^{-2}$ | $3.948 \times 10^{-3}$ | $4.052 \times 10^{-3}$ |
| 0.6 | $1.666 \times 10^{-2}$ | $8.789 \times 10^{-3}$ | $1.451 \times 10^{-3}$ |
| 0.7 | $6.937 \times 10^{-3}$ | $1.070 \times 10^{-2}$ | $2.806 \times 10^{-3}$ |
| 0.8 | $9.786 \times 10^{-3}$ | $6.921 \times 10^{-3}$ | $5.652 \times 10^{-3}$ |
| 0.9 | $3.410 \times 10^{-2}$ | $5.490 \times 10^{-3}$ | $2.132 \times 10^{-3}$ |

Table 2: The maximum absolute errors in Example 5.2.

| $x=y$ | $m=n=1$ | $m=n=2$ | $m=n=4$ |
| :---: | :---: | :---: | :---: |
| 0.0 | $9.890 \times 10^{-5}$ | $3.578 \times 10^{-4}$ | $7.921 \times 10^{-4}$ |
| 0.1 | $6.034 \times 10^{-3}$ | $6.324 \times 10^{-4}$ | $9.468 \times 10^{-4}$ |
| 0.2 | $1.666 \times 10^{-3}$ | $3.307 \times 10^{-4}$ | $1.104 \times 10^{-3}$ |
| 0.3 | $9.537 \times 10^{-3}$ | $1.114 \times 10^{-3}$ | $1.266 \times 10^{-3}$ |
| 0.4 | $2.339 \times 10^{-2}$ | $8.532 \times 10^{-4}$ | $1.424 \times 10^{-3}$ |
| 0.5 | $3.514 \times 10^{-2}$ | $6.995 \times 10^{-4}$ | $1.566 \times 10^{-3}$ |
| 0.6 | $3.935 \times 10^{-2}$ | $3.095 \times 10^{-3}$ | $1.673 \times 10^{-3}$ |
| 0.7 | $2.987 \times 10^{-2}$ | $5.105 \times 10^{-3}$ | $1.675 \times 10^{-3}$ |
| 0.8 | $3.150 \times 10^{-4}$ | $4.622 \times 10^{-3}$ | $1.370 \times 10^{-3}$ |
| 0.9 | $5.916 \times 10^{-2}$ | $1.445 \times 10^{-3}$ | $2.511 \times 10^{-4}$ |

Table 3: The maximum absolute errors in Example 5.3.

Example 5.4 As the last example, let $\alpha=\frac{3}{2}, \beta=\frac{5}{2}, k(x, y, \xi, \tau)=\sqrt{x y \tau}, p=2$ and $f(x, y)=\sqrt{y}\left(\frac{-1}{180} x^{3} y^{\frac{7}{2}}+\sqrt{\frac{x}{3}}\right)$ The exact solution of this example is $u(x, y)=\frac{\sqrt{3 x y}}{3}$. The maximum absolute errors are shown in Table 4. Also, the obtained numerical results are compared with the method of block pulse operational matrix (BPOM) proposed in $[23,30]$.

## 6 Conclusion

A new approach to obtain numerical solution of 2DFOVIE based on the operational matrices of Bernstein polynomials has been presented. With the help of the operational matrix of fractional integration $P^{r}$ and the collocation method, the given 2DFOVIE is reduced to a linear or nonlinear system of algebraic equations. Illustrative examples show that the proposed method can be a suitable method for solving these equations. All of computations are done by Mathematica 9 .

| $x=y$ | $m=n=1$ | $m=n=2$ | $m=n=3$ | $\frac{m_{1}=m_{2}=32}{B P O M}$ |
| :--- | :---: | :---: | :---: | :---: |
| 0.0 | $4.091 \times 10^{-2}$ | $1.701 \times 10^{-2}$ | $9.354 \times 10^{-3}$ | $9.386 \times 10^{-3}$ |
| 0.1 | $1.171 \times 10^{-2}$ | $4.572 \times 10^{-3}$ | $5.766 \times 10^{-3}$ | $1.561 \times 10^{-2}$ |
| 0.2 | $1.017 \times 10^{-2}$ | $1.183 \times 10^{-2}$ | $3.740 \times 10^{-3}$ | $8.812 \times 10^{-3}$ |
| 0.3 | $2.472 \times 10^{-2}$ | $9.513 \times 10^{-3}$ | $2.911 \times 10^{-3}$ | $1.630 \times 10^{-2}$ |
| 0.4 | $3.196 \times 10^{-2}$ | $1.934 \times 10^{-3}$ | $7.428 \times 10^{-3}$ | $8.239 \times 10^{-3}$ |
| 0.5 | $3.186 \times 10^{-2}$ | $7.003 \times 10^{-3}$ | $7.270 \times 10^{-3}$ | $1.410 \times 10^{-2}$ |
| 0.6 | $2.444 \times 10^{-2}$ | $1.382 \times 10^{-2}$ | $2.893 \times 10^{-3}$ | $7.665 \times 10^{-3}$ |
| 0.7 | $9.702 \times 10^{-3}$ | $1.545 \times 10^{-2}$ | $3.149 \times 10^{-3}$ | $1.430 \times 10^{-2}$ |
| 0.8 | $1.236 \times 10^{-2}$ | $9.258 \times 10^{-3}$ | $6.781 \times 10^{-3}$ | $7.091 \times 10^{-3}$ |
| 0.9 | $4.176 \times 10^{-2}$ | $6.980 \times 10^{-3}$ | $2.666 \times 10^{-3}$ | $1.260 \times 10^{-2}$ |

Table 4: The maximum absolute errors in Example 5.4.

## References

[1] Ichise, M., Nagayanagi, Y. and Kojima, T. An analog simulation of non-integer order transfer functions for analysis of electrode processes. J. Electronical. Chem. Interfacial Electrochem 33 (1971) 253-265.
[2] Sun, H. H., Onaral, B. and Tsao, Y. Application of positive reality principle to metal electrode linear polarization phenomena. IEEE Trans. Biomed. Eng. BME-31 (10) (1984) 664-674.
[3] He, J. H. Nonlinear oscillation with fractional derivative and its applications. In: International Conference on Vibrating Engineering 98, China, 1998, 288-291.
[4] Baillie, R. T. Long memory processes and fractional integration in econometrics. J. Econometrics 73 (1996) 50-59.
[5] Magin, R. L. Fractional calculus in bioengineering. Crit. Rev. Biomed. Eng. 32 (1) (2004) 91-104.
[6] Pandy, R. K., Singh, O. P. and Singh, V. K. Efficient algorithms to solve singular integral equations of Abel type. Computers and Mathematics with Applications 57 (4) (2009) 664676.
[7] Jafari, H. and Tajadodi, H. Fractional order optimal control problems via the operational matrices of Bernstein polynomials. U.P.B Sci. Bull., Series A 76 (3) (2014) 115-128.
[8] Alipour, M. and Rostamy, D. BPs operational matrices for solving time varying fractional optimal control problems. The Journal of Mathematics and Computer Science 6 (2013) 292-304.
[9] Kharrat, M. Closed-form solution of European option under fractional Heston model. Nonlinear Dynamics and Systems Theory 18 (2) (2018) 191-195.
[10] Mohammed, M.A. and Fadhel, F.S. Solution of two-dimensional fractional order Volterra integro-differential equations. Joyrnal of AL-Nahrain university 12 (4) (2009) 185-189.
[11] Das, S. Analytical solution of a fractional diffusion equation by variational iteration method. Comput. Math. Appl. 57 (2009) 483-487.
[12] Erturk, V. S., Momani, S. and Odibat, Z. Application of generalized differential transform method to multi-order fractional differential equations. Comm. Nonlinear Sci. Numer. Simmulat. 13 (2008) 1642-1654.
[13] Zhang, X., Liu, J., Zhao, J. and Tang, B. Homotopy perturbation method for twodimensional time-fractional wave equation. Appl. Math. Modelling 38 (23) (2014) 55455552.
[14] Saeedi, H., Mollahasani, N., Mohseni moghadam, M. and Chuev, G. N. An operational Harr wavelet method for solving fractional Volterra integral equations. Int. J. Appl. Math. Sci. 21 (3) (2011) 535-547.
[15] Zhu, L. and Fan, Q. Solving fractional nonlinear Fredholm integro-differential by the second kind Chebyshev wavelet. Comm. Nonlinear Sci. Numer. Simmulat. 17 (2012) 2333-2341.
[16] Doha, E.H., Bhrawy, A.H. and Saker, M.A. Integrals of Bernstein polynomials: an application for the solution of high even-order differential equations. Appl. Math. Lett. 24 (2011) 559-565.
[17] Maleknejad, K., Hashemizadeh, E. and Basirat, B. Computational method based on Bernstein operational matrices for nonlinear VolterraFredholmHammerstein integral equations. Commun. Nonlinear Sci. Numer. Simul. 17 (1) (2012) 52-61.
[18] Hosseini Shekarabi, F., Maleknejad, K. and Ezzati, R. Application of two-dimensional Bernstein polynomials for solving mixed VolterraFredholm integral equations. Afr. Mat. 26 (2015) 1237-1251.
[19] Maleknejad, K., Hashemizadeh, E. and Basirat, B. A Bernstein operational matrix approach for solving a system of high order linear Volterra-Fredholm integro-differential equations. Math. Comput. Model. 55 (2012) 1363-1372.
[20] Maleknejad, K., Hashemizadeh, E. and Ezzati, R. New approach to the numerical solution of Volterra integral equations by using Bernsteins approximation. Commun. Nonlinear Sci. Numer. Simul. 16 (2011) 647-655.
[21] Mandal, B. N. and Bhattacharya, S. Numerical solution of some classes of integral equations using Bernstein polynomials. Appl. Math. Comput. 190 (2007) 707-716.
[22] Yousefi, S. A. and Behroozifar, M. Operational matrices of Bernstein polynomials and their applications. Int.J. Syst. Sci. 41 (6) (2010) 709-716.
[23] Jabari Sabeg, D., Ezzati, R. and Maleknejad, K. Solving Two-Dimensional Integeral Equations of Fractional Order by Using Operational Mtrix of Two-Dimentional Shifted Legendre Polynomials. Nonlinear Dynamics and Systems Theory 18 (3) (2018) 297-306.
[24] Vityuk, A. N. and Golushkov, A. V. Existence of solutions of systems of partial differential equations of fractional order. Nonlinear Oscil. 7 (2004) 318-325.
[25] Abbas S., and Benchohra, M. Fractional order integral equations of two independent variables. Appl. Math. Comput. 227 (2014) 755-761.
[26] Kreyszig, E. Introductory Functional Analysis with Applications. New York: John Wiley and Sons, 1978.
[27] Kilicman, A. and Al Zhour, Z.A.A. Kronecker operational matrices for fractional calculus and some applications. Appl. Math. Comput. 187 (1) (2007) 250-265.
[28] Pandey, R.K., Bhardwaj, A. and Syam, M.I. An efficient method for solving fractional differential equations using Bernstein polynomials. Journal of Fractional Calculus and Applications 5 (1) (2014) 129-145.
[29] Asgari, M. and Ezzati, R. Using operational matrix of two-dimensional Bernstein polynomials for solving two-dimensional integral equations of fractional order. Appl. Math. Comput. 307 (2017) 290-298.
[30] Najafalizadeh, S. and Ezzati, R. Numerical methods for solving two-dimensional nonlinear integral equations of fractional order by using two-dimensional block pulse operational matrix. Appl. Math. Comput. 280 (20) (2016) 46-56.

# Singular Analysis of Reduced ODEs of Rotating Stratified Boussinesq Equations Through the Mirror Transformations 

B.S. Desale ${ }^{1 *}$ and K.D. Patil ${ }^{2}$<br>${ }^{1}$ Department of Mathematics, University of Mumbai, Kalina, Santacruz (East), Mumbai 400 098, India.<br>${ }^{2}$ Department of Mathematics, School of Mathematical Sciences, North Maharashtra University, Jalgaon 425 001, India.

Received: March 11, 2018; Revised: December 13, 2018


#### Abstract

In this paper we have considered the system of six coupled non-linear ordinary differential equations (ODEs), which arose in the reduction of uniformly stratified fluid contained in a rotating rectangular box of dimension $L \times L \times H$ which is completely integrable if the Rayleigh number $R a=0$. In our investigations, we have shown that there exits a regular mirror system near movable singularities of these integrable ODEs. Moreover, we have used the mirror system to prove the convergence of Laurent series solutions obtained by the Painlevé method.


Keywords: mirror transformation; mirror system; Painlevé test.
Mathematics Subject Classification (2010): 37K10, 34M55.

## 1 Introduction

In general, we believed that the differential system is integrable due to some sort of underlying linear structure(s). But, when it comes to this concept, it is never clear what does it mean. On the other hand the integrability of nonlinear system is quite ambiguous. In this connection many mathematicians started to work over the investigation of integrability of nonlinear system. In 1889, Sophie Kowalevski [12] proved the complete integrability of the system of ordinary differential equations (ODEs) governing the motion of a spinning top moving under the influence of gravity. In her study, she was seeking

[^2]analytic solutions whose singularities are movable poles. This was done by substituting a Frobenius series into the system of ODEs. Then, few years later, that is in 1897, Paul Painlevé [6] classified first and second order algebraic differential equations whose solutions exist in the complex domain and are devoid of movable essential singularities or movable branch points. ODEs possessing this property are said to be of the Painlevé type. Painlevé test in view of partial differential equations is generally known as WTC (Weiss-Tabor-Carnevale, [7]) test, which is further modified by S. Kichenassamy and G. K. Srinivasan [3]. So far various properties are considered as indicator of integrability: solitons, the Lax pair, the Bäcklund transformations, the underlying Hamiltonian formulations, Hirota's bilinear representation, etc. The relation between these properties has yet to be understood.

In 1999-2000, Hu J. and Yan M. [8, 9] introduced the mirror transformation, which is a new tool used in the singularity analysis of ODEs. With the help of this method we constructed the mirror system of given PDEs or ODEs successfully; we could focus commonly at the singularity structure and symplectic structure of the Hamiltonian system for each principle balance in the Painlevé test. Further to this study, Hu et al [11] proved that the mirror transformation is canonical for finite-dimensional Hamiltonian systems. Furthermore, in 2001 Yee [13] showed that linearization of the mirror systems near movable poles provides the possibility to construct the associated Backlund transformations. In continuous development of mirror transformations in 2011, Tat-Leung Yee [14] extended the mirror method with perturbations which was utilized for finer analysis of certain nonlinear equations possessing negative Fuchsian indices.

In connection with the basin scale dynamics, Maas [5] has considered the flow of fluid contained in a rectangular basin of dimension $L \times L \times H$, which is temperature stratified with fixed zeroth order moment of mass and heat. The container is assumed to be steady, uniform rotation of an f-plane. With this assumption Maas [5] reduces the rotating stratified Boussinesq equation to a beautiful six coupled system of ODEs. Srinivasan et al. [4] extended this work and gave a detail mathematical analysis of the reduced system of six coupled ODEs. Furthermore, Desale and Patil [2] tested the system of six coupled ODEs (5) for complete integrability using the Painlevé test. Also, they investigated the case of non-integrability for $R a \neq 0$ and thereby they have obtained weak solutions (in the form of logarithmic psi-series) in the different branches of leading order.

In this paper we have successfully implemented the mirror transformations and constructed the mirror system of (5) for $R a=0$ which is regular near movable singularity. Further, with the help of mirror transformation, we have proved that the Laurent series obtained by using the Painlevé test are convergent. In the following section we imploy the mirror transformation to find the mirror system of ideal rotating stratified Boussinesq equations.

## 2 Mirror System of Six Coupled Non-Linear ODEs

Consider the rotating stratified Boussinesq equations (see Majda [1], p. 1)

$$
\begin{align*}
\frac{D \overrightarrow{\mathbf{v}}}{D t}+f\left(\hat{\mathbf{e}_{\mathbf{3}}} \times \overrightarrow{\mathbf{v}}\right) & =-\nabla p+\nu(\Delta \overrightarrow{\mathbf{v}})-\frac{g \tilde{\rho}}{\rho_{b}} \hat{\mathbf{e}_{\mathbf{3}}}, \\
\operatorname{div} \overrightarrow{\mathbf{v}} & =0  \tag{1}\\
\frac{D \tilde{\rho}}{D t} & =\kappa \Delta \tilde{\rho}
\end{align*}
$$

where $\overrightarrow{\mathbf{v}}$ denotes the velocity field, $\rho$ is the density which is the sum of constant reference density $\rho_{b}$ and perturbation density $\tilde{\rho}, p$ is the pressure, $g$ is the acceleration due to gravity that points in $-\hat{\mathbf{e}_{3}}$ direction, $f$ is the rotation frequency of earth, $\nu$ is the coefficient of viscosity, $\kappa$ is the coefficient of heat conduction and $\frac{D}{D t}=\frac{\partial}{\partial t}+(\overrightarrow{\mathbf{v}} \cdot \nabla)$ is a convective derivative. For more about the rotating stratified Boussinesq equations one may see Majda [1]. Maas [5] reduces the system of equations (1) to the following system of six coupled ODEs:

$$
\begin{align*}
\operatorname{Pr}^{-1} \frac{d \overrightarrow{\mathbf{w}}}{d t}+f^{\prime} \hat{\mathbf{e}_{\mathbf{3}}} \times \overrightarrow{\mathbf{w}} & =\hat{\mathbf{e}_{\mathbf{3}}} \times \overrightarrow{\mathbf{b}}-\left(w_{1}, w_{2}, r w_{3}\right)+\hat{T} \overrightarrow{\mathbf{T}} \\
\frac{d \overrightarrow{\mathbf{b}}}{d t}+\overrightarrow{\mathbf{b}} \times \overrightarrow{\mathbf{w}} & =-\left(b_{1}, b_{2}, \mu b_{3}\right)+R a \overrightarrow{\mathbf{F}} . \tag{2}
\end{align*}
$$

In these equations, $\overrightarrow{\mathbf{b}}=\left(b_{1}, b_{2}, b_{3}\right)$ is the center of mass, $\overrightarrow{\mathbf{w}}=\left(w_{1}, w_{2}, w_{3}\right)$ is the basin averaged angular momentum vector, $\overrightarrow{\mathbf{T}}$ is the differential momentum, $\overrightarrow{\mathbf{F}}$ are buoyancy fluxes, $f^{\prime}=f / 2 r_{h}$ is the earth rotation, $r=r_{v} / r_{h}$ is the friction ( $r_{v, h}$ are Rayleigh damping coefficients), $R a$ is the Rayleigh number, $\operatorname{Pr}$ is the Prandtl number, $\mu$ is the diffusion coefficient and $\hat{T}$ is the magnitude of the wind stress torque.

Neglecting diffusive and viscous terms, Maas [5] considers the dynamics of an ideal rotating, uniformly stratified fluid in response to forcing. He assumes this to be due solely to differential heating in the meridional $(y)$ direction. $\overrightarrow{\mathbf{F}}=(0,1,0)$, the wind effect is neglected, i.e. $\overrightarrow{\mathbf{T}}=0$. For the Prandtl number $\operatorname{Pr}$, equal to one, the system of equations (2) reduces to the following ideal rotating, uniformly stratified system of six coupled ODEs

$$
\begin{align*}
\frac{d \overrightarrow{\mathbf{w}}}{d t} & =-f^{\prime} \hat{\mathbf{e}_{\mathbf{3}}} \times \overrightarrow{\mathbf{w}}+\hat{\mathbf{e}_{\mathbf{3}}} \times \overrightarrow{\mathbf{b}} \\
\frac{d \overrightarrow{\mathbf{b}}}{d t} & =-\overrightarrow{\mathbf{b}} \times \overrightarrow{\mathbf{w}}+R a \overrightarrow{\mathbf{F}} \tag{3}
\end{align*}
$$

The system of ODEs (3) can be written component wise as

$$
\begin{align*}
& \dot{w_{1}}=f^{\prime} w_{2}-b_{2}, \quad \dot{w_{2}}=-f^{\prime} w_{1}+b_{1}, \quad \dot{w_{3}}=0, \\
& \dot{b_{1}}=w_{2} b_{3}-w_{3} b_{2}, \quad \dot{b_{2}}=w_{3} b_{1}-w_{1} b_{3}+R a, \quad \dot{b_{3}}=w_{1} b_{2}-w_{2} b_{1} . \tag{4}
\end{align*}
$$

Since $\dot{w_{3}}=0$, this gives $w_{3}=$ constant $=k_{1}$. Consequently, we have the following system of ODEs:

$$
\begin{align*}
& \dot{w_{1}}=f^{\prime} w_{2}-b_{2}, \quad \dot{w_{2}}=-f^{\prime} w_{1}+b_{1}, \\
& \dot{b_{1}}=w_{2} b_{3}-k_{1} b_{2}, \quad \dot{b_{2}}=k_{1} b_{1}-w_{1} b_{3}+R a, \quad \dot{b_{3}}=w_{1} b_{2}-w_{2} b_{1} . \tag{5}
\end{align*}
$$

In our earlier study [2], we have shown that the system of ODEs (5) is completely integrable provided that $R a=0$ and we have determined the solutions in the form of Laurent series with the help of the Painlevé method. Now our aim is to determine the mirror system of (5) and its solutions in the following form:

$$
\begin{align*}
& w_{1}(t)=\theta^{-m_{1}}, \quad \theta^{\prime}=l_{0}+l_{1} \theta+l_{2} \theta^{2}+l_{3} \theta^{3}+l_{4} \theta^{4}+\cdots, \\
& w_{2}(t)=\theta^{-m_{2}}\left(w_{20}+w_{21} \theta+w_{22} \theta^{2}+w_{23} \theta^{3}+w_{24} \theta^{4}+\cdots\right), \\
& b_{1}(t)=\theta^{-m_{3}}\left(b_{10}+b_{11} \theta+b_{12} \theta^{2}+b_{13} \theta^{3}+b_{14} \theta^{4}+\cdots\right),  \tag{6}\\
& b_{2}(t)=\theta^{-m_{4}}\left(b_{20}+b_{21} \theta+b_{22} \theta^{2}+b_{23} \theta^{3}+b_{24} \theta^{4}+\cdots\right), \\
& b_{3}(t)=\theta^{-m_{5}}\left(b_{30}+b_{31} \theta+b_{32} \theta^{2}+b_{33} \theta^{3}+b_{34} \theta^{4}+\cdots\right),
\end{align*}
$$

where $\theta=t-t_{0}$ and $t_{0}$ is an arbitrary position of singularity. We found that there were several possible cases of dominant balance of the system (5) similar to those in the Painlevé test. Among the several possible cases of principle dominant balance we have obtained the singular solution only in the following case of principle dominant balance:

$$
\begin{equation*}
\dot{w_{1}}=-b_{2}, \quad \dot{w_{2}}=b_{1}, \quad \dot{b_{1}}=w_{2} b_{3}, \quad \dot{b_{2}}=-w_{1} b_{3}, \quad \dot{b_{3}}=w_{1} b_{2}-w_{2} b_{1} \tag{7}
\end{equation*}
$$

and the exponent with this principle dominant balance are as follows:

$$
\begin{equation*}
m_{1}=m_{2}=-1, \quad m_{3}=m_{4}=m_{5}=-2 . \tag{8}
\end{equation*}
$$

Since $w_{1}, w_{2}$ are of order 1 near the movable singularity, we can introduce the indicial normalization $w_{1}(t)=\theta^{-1}$ and try to calculate the formal $\theta$ - series of (6) with $m_{2}=$ $-1, m_{3}=m_{4}=m_{5}=-2$. Since the system (5) is autonomous, the coefficients appearing in the series given by (5) are to be constant. Substituting the values of exponents from (8) into the equations (6) and then substituting these series into the system (5) and hence equating the like powers of $\theta$ on both sides, we obtain the following equations in leading order coefficients:

$$
\begin{align*}
& l_{0}=b_{20}, \quad-w_{20} l_{0}=b_{10}, \quad-2 b_{10} l_{0}=w_{20} b_{30}  \tag{9}\\
& 2 b_{20} l_{0}=b_{30}, \quad-2 b_{30} l_{0}=b_{20}-w_{20} b_{10}
\end{align*}
$$

Solving equations (9), we find two possible branches of leading order coefficients which are as follows:

$$
\begin{equation*}
l_{0}=r_{1}^{\prime}, w_{20}= \pm \sqrt{-1-4 r_{1}^{\prime 2}}, b_{10}=\mp r_{1}^{\prime} \sqrt{-1-4 r_{1}^{\prime 2}}, b_{20}=r_{1}^{\prime}, b_{30}=2 r_{1}^{\prime 2} \tag{10}
\end{equation*}
$$

where $r_{1}^{\prime}$ is an arbitrary constant.
Definition 2.1 The leading exponents $m_{1}, m_{2}, m_{3}, m_{4}, m_{5}$ for system of ODEs (5) are Fuchsian, if the $m_{*}$-weighted degree of the right-hand side of $(5)$ is $\leq m_{i}+1$.

The $m_{*}$-weighted degree of polynomial in $w_{1}, w_{2}, b_{1}, b_{2}, b_{3}$ is found by taking the degree of $w_{i}^{\prime} s, i=1,2, b_{i}^{\prime} s, i=1,2,3$ to be $m_{i}, i=1,2,3,4,5$. And we verified that the exponents $m_{i}$ 's, $i=1,2,3,4,5$ are Fuchsian for the system (5).

Remark 2.1 Since all leading order coefficients given by (10) are nonzero, the selection of leading exponents is natural and these exponents satisfy the Fuchsian condition.

So far in the employment of mirror transformations we have completed the two steps of algorithm, that is, we have determined leading order coefficients in principle dominant balance and exponents. Now, in the following section we will implement the third step of the algorithm and determine the resonances in the following way.

### 2.1 Resonances

Now we substitute the assumed $\theta$-series (6) with the values of exponents given by (8) into the system of ODEs (5) and after doing some algebraic calculations we specify the following recursive relations to determine the coefficients $w_{1 j}, w_{2 j}, b_{1 j}, b_{2 j}$ and $b_{3 j}$ for $j=1,2,3, \ldots$ which are valid for $j \geq 2$ :

$$
M(j)\left(\begin{array}{c}
l_{j}  \tag{11}\\
w_{2 j} \\
b_{1 j} \\
b_{2 j} \\
b_{3 j}
\end{array}\right)=\left(\begin{array}{c}
A_{j} \\
B_{j} \\
C_{j} \\
D_{j} \\
E_{j}
\end{array}\right),
$$

where

$$
\begin{align*}
& A_{j}=f^{\prime} w_{2(j-1)}, \quad B_{j}=-\sum_{k=1}^{j-1} l_{k} w_{2(j-k)} \\
& C_{j}=-k_{1} b_{2(j-1)}+\sum_{k=1}^{j-1} w_{2 k} b_{3(j-k)}-\sum_{k=1}^{j-1} l_{k} b_{1(j-k)}  \tag{12}\\
& D_{j}=k_{1} b_{1(j-1)}-\sum_{k=1}^{j-1} l_{k} b_{2(j-k)} \\
& E_{j}=-\sum_{k=1}^{j-1} w_{2 k} b_{1(j-k)}-\sum_{k=1}^{j-1} l_{k} b_{3(j-k)}
\end{align*}
$$

and matrix $M(j)$ is

$$
M(j)=\left(\begin{array}{ccccc}
-1 & 0 & 0 & 1 & 0  \tag{13}\\
-w_{20} & (j-1) l_{0} & -1 & 0 & 0 \\
-2 b_{10} & -b_{30} & (j-2) l_{0} & 0 & -w_{20} \\
-2 b_{20} & 0 & 0 & (j-2) l_{0} & 1 \\
-b_{30} & b_{10} & w_{20} & -1 & (j-2) l_{0}
\end{array}\right)
$$

The above recursive relations $(11,12)$ determine the unknown expansion coefficients uniquely unless the determinant of matrix $M(j)$ is zero. Those values of $j$ at which the determinant of matrix $M(j)$ vanishes are called the resonances. Here, we observe that for both possible branches of leading order coefficients given in equations (10), the resonances are $j=0,2,3,4$. Since $j=0$ is the resonance, one of the variable in (10) appears to be a resonance parameter, say $l_{0}=r_{1}^{\prime}$, and we should replace it by $\overline{r_{1}}$ (where $\overline{r_{1}}=\sqrt{-4-k_{2}^{2}}$, the arbitrary constant $k_{2}$ is the resonance parameter in the Painlevé test [2]), which satisfies the condition $\overline{r_{1}}-m_{1}=r_{1}^{\prime}$, that is, ${\overline{r_{1}}}^{-1}=r_{1}^{\prime}$. Let us denote by $k_{2}=r_{1}$ the resonance parameter, and hence we have $\overline{r_{1}}=\sqrt{-4-r_{1}^{2}}$. Now, we refresh the leading order coefficients given by (10) as follows:

$$
\begin{align*}
& l_{0}=\left(\sqrt{-4-r_{1}^{2}}\right)^{-1}, w_{20}= \pm \frac{r_{1}}{\sqrt{-4-r_{1}^{2}}}, b_{10}=\mp \frac{r_{1}}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}}  \tag{14}\\
& b_{20}=\left(\sqrt{-4-r_{1}^{2}}\right)^{-1}, b_{30}=\frac{2}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}}
\end{align*}
$$

### 2.2 Compatibility conditions

Further, we need to check the compatibility conditions for each resonance $j=2,3,4$. We will do this for the first branch.
Case I: Consider the leading order coefficients

$$
\begin{align*}
& l_{0}=\left(\sqrt{-4-r_{1}^{2}}\right)^{-1}, w_{20}=\frac{r_{1}}{\sqrt{-4-r_{1}^{2}}}, b_{10}=-\frac{r_{1}}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}}, \\
& b_{20}=\left(\sqrt{-4-r_{1}^{2}}\right)^{-1}, b_{30}=\frac{2}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}} . \tag{15}
\end{align*}
$$

- Compatibility condition at $j=1$.

As $j=1$ is not resonance, we get the unique solution. Since the recursion relations (11, 12) remain valid when $j \geq 2$, we directly substitute the equations (15) into the
equations (6) and then into (5). After that, equating the like powers of $\theta$ on both sides of the resulting expansion, we obtain the system of linear equations which determine the coefficients $l_{1}, w_{21}, b_{11}, b_{21}$ and $b_{31}$ uniquely as

$$
\begin{align*}
& l_{1}=\frac{\left(-f^{\prime}+k_{1}\right) r_{1}}{\sqrt{-4-r_{1}^{2}}}, \quad w_{21}=\frac{2\left(f^{\prime}-k_{1}\right)}{\sqrt{-4-r_{1}^{2}}}, \\
& b_{11}=\frac{4 f^{\prime}+k_{1} r_{1}^{2}}{r_{1}^{2}+4}, \quad b_{21}=\frac{k_{1} r_{1}}{\sqrt{-4-r_{1}^{2}}}, \quad b_{31}=\frac{2\left(f^{\prime}-k_{1}\right) r_{1}}{4+r_{1}^{2}} . \tag{16}
\end{align*}
$$

- Compatibility condition at the resonance $j=2$.

Now $j=2$ is a resonance so that one of the coefficients in the computation of the system (11) at this level is independent. Let $b_{32}$ be independent and let $b_{32}=r_{2}$ (the arbitrary coefficient), where $r_{2}$ is the second resonance parameter so that the values of coefficients are given in terms of $r_{2}$, which are as follows:

$$
\begin{align*}
& l_{2}=\frac{\left(r_{2}-f^{\prime} k_{1}\right)}{2} \sqrt{-4-r_{1}^{2}}, \quad w_{22}=0 \\
& b_{12}=\frac{r_{1}}{2}\left(f^{\prime} k_{1}-r_{2}\right), \quad b_{22}=\frac{1}{2}\left[r_{2} \sqrt{-4-r_{1}^{2}}+\frac{f^{\prime}\left(4 f^{\prime}+k_{1} r_{1}^{2}\right)}{\sqrt{-4-r_{1}^{2}}}\right], \quad b_{32}=r_{2} . \tag{17}
\end{align*}
$$

## - Compatibility condition at the resonance $j=3$.

To check the compatibility condition at $j=3$, we substitute the equations $(15,16,17)$ into the system of ODEs (5), then we obtain a system of linear equations. While solving that linear system, we found the variable $b_{23}$ to be independent. Now assign the arbitrary value to $b_{23}$, say $b_{23}=r_{3}$, and solving the corresponding system we obtain the following solution. At this level of resonance, we have the third resonance parameter $r_{3}$ :

$$
\begin{align*}
& l_{3}=r_{3}, w_{23}=\frac{-r_{3}}{r_{1}}, b_{13}=-\frac{1}{\sqrt{-4-r_{1}^{2}}}\left(r_{1} r_{3}+\frac{2 r_{3}}{r_{1}}\right),  \tag{18}\\
& b_{23}=r_{3}, b_{33}=\frac{r_{3}}{\sqrt{-4-r_{1}^{2}}}
\end{align*}
$$

- Compatibility condition at the resonance $j=4$.

Now $j=4$ is the fourth resonance and solving the system (11) for $j=4$ involves the resonance parameter, say $r_{4}$. Solving the system (11) for this value of $j$, we obtain the following solution with $b_{24}$ as an arbitrary constant with value $r_{4}$ :

$$
\begin{align*}
& l_{4}=r_{4}+\frac{f^{\prime} r_{3}}{r_{1}}, \quad w_{24}=\left(k_{1}-f^{\prime}\right) r_{3}, \\
& b_{14}=\frac{1}{\sqrt{-4-r_{1}^{2}}}\left[-r_{1} r_{4}+\left(-2 f^{\prime}+k_{1}\right) r_{3}\right], \quad b_{24}=r_{4}  \tag{19}\\
& b_{34}=\frac{\left(f^{\prime}-k_{1}\right)\left(2+r_{1}^{2}\right) r_{3}}{r_{1} \sqrt{-4-r_{1}^{2}}}
\end{align*}
$$

Substituting all the values of coefficients $l_{j}, w_{2 j}, b_{1 j}, b_{2 j}$ and $b_{3 j}$ for $j=0,1,2,3,4 \ldots$ into the equations (6), we get

$$
\begin{align*}
\theta^{\prime}= & \frac{1}{\sqrt{-4-r_{1}^{2}}}+\frac{\left(-f^{\prime}+k_{1}\right) r_{1}}{\sqrt{-4-r_{1}^{2}}} \theta+\frac{1}{2} \sqrt{-4-r_{1}^{2}}\left(r_{2}-f^{\prime} k_{1}\right) \theta^{2}+r_{3} \theta^{3} \\
+ & \left(r_{4}+\frac{f^{\prime} r_{3}}{r_{1}}\right) \theta^{4}+\cdots, \\
w_{2}(t)= & \theta^{-1}\left[\frac{r_{1}}{\sqrt{-4-r_{1}^{2}}}+\frac{2\left(f^{\prime}-k_{1}\right)}{\sqrt{-4-r_{1}^{2}}} \theta-\frac{r_{3}}{r_{1}} \theta^{3}+\left(k_{1}-f^{\prime}\right) r_{3} \theta^{4}+\cdots\right], \\
b_{1}(t)= & \theta^{-2}\left[-\frac{r_{1}}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}}+\left(\frac{4 f^{\prime}+k_{1} r_{1}^{2}}{r_{1}^{2}+4}\right) \theta+\frac{r_{1}}{2}\left(f^{\prime} k_{1}-r_{2}\right) \theta^{2}-\frac{1}{\sqrt{-4-r_{1}^{2}}}\right. \\
& \left.\left(r_{1} r_{3}+\frac{2 r_{3}}{r_{1}}\right) \theta^{3}+\frac{1}{\sqrt{-4-r_{1}^{2}}}\left(-r_{1} r_{4}+\left(-2 f^{\prime}+k_{1}\right) r_{3}\right) \theta^{4}+\cdots\right], \\
b_{2}(t)= & \theta^{-2}\left[\frac{1}{\sqrt{-4-r_{1}^{2}}}+\left(\frac{k_{1} r_{1}}{\sqrt{-4-r_{1}^{2}}}\right) \theta+\frac{1}{2}\left(r_{2} \sqrt{-4-r_{1}^{2}}+\frac{f^{\prime}\left(4 f^{\prime}+k_{1} r_{1}^{2}\right)}{\sqrt{-4-r_{1}^{2}}}\right) \theta^{2}\right. \\
+ & \left.r_{3} \theta^{3}+r_{4} \theta^{4}+\cdots\right], \\
b_{3}(t)= & \theta^{-2}\left[\frac{2}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}}-\left(\frac{2\left(f^{\prime}-k_{1}\right) r_{1}}{4+r_{1}^{2}}\right) \theta+r_{2} \theta^{2}+\frac{r_{3}}{\sqrt{-4-r_{1}^{2}}} \theta^{3}\right. \\
+ & \left.\frac{\left(f^{\prime}-k_{1}\right)\left(2+r_{1}^{2}\right) r_{3}}{r_{1} \sqrt{-4-r_{1}^{2}}} \theta^{4}+\cdots\right] . \tag{20}
\end{align*}
$$

We have just finished the primary calculations of the system (11) and we have determined the resonance parameters, say $r_{1}, r_{2}, r_{3}$ and $r_{4}$. In the following subsection we obtain the mirror transformations and consequently, we determine the mirror system of (5).

### 2.3 Mirror system

In this subsection we will develop the mirror transformations by which we transform the system (5) to its mirror system. Thereby, we discuss the regularity of it.

Now the important step towards determining the mirror system is to introduce a new variable in which we develop the mirror system. Let us introduce the new variables $\xi_{1}, \xi_{2}, \xi_{3}$ and $\xi_{4}$ in the Laurent $\theta$-series of $w_{2}, b_{1}, b_{2}$ and $b_{3}$ by successively truncating the expansion at the free parameters (resonance parameters) $r_{1}, r_{2}, r_{3}$ and $r_{4}$. Now we begin to truncate the $\theta$-series of $w_{2}$ at the first resonance parameter $r_{1}$ by introducing the variable $\xi_{1}$ as

$$
\begin{equation*}
w_{2}(t)=\theta^{-1} \xi_{1} \tag{21}
\end{equation*}
$$

where

$$
\begin{equation*}
\xi_{1}=\frac{r_{1}}{\sqrt{-4-r_{1}^{2}}}+\frac{2\left(f^{\prime}-k_{1}\right)}{\sqrt{-4-r_{1}^{2}}} \theta-\frac{r_{3}}{r_{1}} \theta^{3}+\left(k_{1}-f^{\prime}\right) r_{3} \theta^{4}+\cdots \tag{22}
\end{equation*}
$$

We convert this into

$$
\begin{equation*}
r_{1}=\xi_{1} \overline{r_{1}}-2\left(f^{\prime}-k_{1}\right) \theta+\frac{r_{3}}{\xi_{1}} \theta^{3}-r_{3}\left(f^{\prime}-k_{1}\right)\left(\frac{2}{\xi_{1}^{2} \overline{r_{1}}}+\overline{r_{1}}\right) \theta^{4}+\cdots \tag{23}
\end{equation*}
$$

Upon substituting the value of $r_{1}$ in $b_{1}$, we get

$$
\begin{align*}
b_{1}(t) & =-\frac{\xi_{1}}{\overline{r_{1}}} \theta^{-2}+\left(\frac{-2 f^{\prime}-2 k_{1}}{{\overline{r_{1}}}^{2}}-k_{1} \xi_{1}^{2}\right) \theta^{-1}+\left[\frac{1}{2}\left(f^{\prime} k_{1}-r_{2}\right) \xi_{1} \overline{r_{1}}\right. \\
& \left.+\frac{4 k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)}{\overline{r_{1}}}\right]+\left[\frac{-3 r_{3}}{\xi_{1}{\overline{r_{1}}}^{2}}-\frac{4 k_{1}\left(f^{\prime}-k_{1}\right)^{2}}{{\overline{r_{1}}}^{2}}-\left(f^{\prime} k_{1}-r_{2}\right)\left(f^{\prime}-k_{1}\right)\right.  \tag{24}\\
& \left.-\xi_{1} r_{3}\right] \theta+\left[\frac{-2 r_{3}\left(f^{\prime}-k_{1}\right)}{\xi_{1}^{2}{\overline{r_{1}}}^{3}}-\xi_{1} r_{4}+\frac{\left(f^{\prime} r_{3}-4 k_{1} r_{3}\right)}{\overline{r_{1}}}\right] \theta^{2} .
\end{align*}
$$

Next we proceed to cut the $\theta$-series of $b_{1}$ at $r_{2}$ by introducing the second variable, say $\xi_{2}$ :

$$
\begin{equation*}
b_{1}(t)=-\frac{\xi_{1}}{\overline{r_{1}}} \theta^{-2}+\left(\frac{-2 f^{\prime}-2 k_{1}}{{\overline{r_{1}}}^{2}}-k_{1} \xi_{1}^{2}\right) \theta^{-1}+\xi_{2} \tag{25}
\end{equation*}
$$

where

$$
\begin{align*}
\xi_{2} & =\left[\frac{1}{2}\left(f^{\prime} k_{1}-r_{2}\right) \xi_{1} \overline{r_{1}}+\frac{4 k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)}{\overline{r_{1}}}\right]+\left[\frac{-3 r_{3}}{\overline{\xi_{1}} \overline{r_{1}^{2}}}-\frac{4 k_{1}\left(f^{\prime}-k_{1}\right)^{2}}{{\overline{r_{1}}}^{2}}\right. \\
& \left.-\left(f^{\prime} k_{1}-r_{2}\right)\left(f^{\prime}-k_{1}\right)-\xi_{1} r_{3}\right] \theta+\left[\frac{-2 r_{3}\left(f^{\prime}-\overline{k_{1}}\right)}{\xi_{1}^{2} \overline{r_{1}^{3}}}-\xi_{1} r_{4}\right.  \tag{26}\\
& \left.+\frac{\left(f^{\prime} r_{3}-4 k_{1} r_{3}\right)}{\overline{r_{1}}}\right] \theta^{2}+\cdots .
\end{align*}
$$

From the $\theta$-series of $\xi_{2}$, we have

$$
\begin{align*}
r_{2} & =f^{\prime} k_{1}-\frac{2 \xi_{2}}{\xi_{1} \overline{r_{1}}}+\frac{8 k_{1}\left(f^{\prime}-k_{1}\right)}{{\overline{r_{1}}}^{2}}+\frac{2}{\xi_{1} \overline{r_{1}}}\left[\frac{-3 r_{3}}{\xi_{1}{\overline{r_{1}}}^{2}}+\frac{4 k_{1}\left(f^{\prime}-k_{1}\right)^{2}}{\bar{r}_{1}^{2}}-\frac{2 \xi_{2}\left(f^{\prime}-k_{1}\right)}{\xi_{1} \overline{r_{1}}}\right. \\
& \left.-\xi_{1} r_{3}\right] \theta-\frac{2}{\xi_{1} \overline{r_{1}}}\left[\frac{8 r_{3}}{\xi_{1}^{2}{\overline{r_{1}}}^{3}}\left(f^{\prime}-k_{1}\right)-\frac{8 k_{1}\left(f^{\prime}-k_{1}\right)^{3}}{\xi_{1}{\overline{r_{1}}}^{3}}+\frac{4 \xi_{2}\left(f^{\prime}-k_{1}\right)^{2}}{\xi_{1}^{2}{\overline{r_{1}}}^{2}}\right. \\
& \left.+\frac{\left(f^{\prime}+2 k_{1}\right) r_{3}}{\overline{r_{1}}}+\xi_{1} r_{4}\right] \theta^{2}+\cdots . \tag{27}
\end{align*}
$$

Now, we substitute the value of $r_{2}$ into $\theta$-series of $b_{2}$ and consequently, we update it. And then after cutting this series at the third resonance parameter $r_{3}$, we obtain the $\theta$-series of $b_{2}$ as follows:

$$
\begin{align*}
b_{2}(t) & =\frac{1}{\overline{r_{1}}} \theta^{-2}+k_{1} \xi_{1} \theta^{-1}+\left[\frac{2 k_{1}\left(f^{\prime}-k_{1}\right)}{\overline{r_{1}}}+\frac{1}{2} f^{\prime} k_{1} \overline{r_{1}}+\frac{1}{2} f^{\prime} k_{1} \xi_{1}^{2} \overline{r_{1}}+\frac{2 f^{\prime 2}}{\overline{r_{1}}}\right. \\
& \left.-\frac{\xi_{2}}{\xi_{1}}\right]+\xi_{3} \theta \tag{28}
\end{align*}
$$

where

$$
\begin{align*}
\xi_{3} & =\left[\frac{-3 r_{3}}{\xi_{1}^{2} \overline{r_{1}^{2}}}+\frac{4 k_{1}\left(f^{\prime}-k_{1}\right)^{2}}{\xi_{1}{\overline{r_{1}}}^{2}}-\frac{2 \xi_{2}\left(f^{\prime}-k_{1}\right)}{\xi_{1}^{2} \overline{r_{1}}}-2 f^{\prime} k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)\right]+\left[\frac{-8 r_{3}\left(f^{\prime}-k_{1}\right)}{\xi_{1}^{3} \overline{\bar{r}_{1}^{3}}}\right. \\
& \left.+\frac{8 k_{1}\left(f^{\prime}-k_{1}\right)^{3}}{\xi_{1}^{3}{\overline{r_{1}}}^{3}}-\frac{4 \xi_{2}\left(f^{\prime}-k_{1}^{2}\right)}{\xi_{1}^{3}{\overline{r_{1}}}^{2}}-\frac{\left(f^{\prime}+k_{1}\right) r_{3}}{\xi_{1} \overline{r_{1}}}+\frac{2 f^{\prime} k_{1}\left(f^{\prime}-k_{1}^{2}\right)}{\overline{r_{1}}}\right] \theta+\cdots . \tag{29}
\end{align*}
$$

From the $\theta$-series of $\xi_{3}$ we have

$$
\begin{align*}
r_{3} & =-\frac{\xi_{1}^{2}{\overline{r_{1}}}^{2} \xi_{3}}{3}+\frac{4}{3} k_{1}\left(f^{\prime}-k_{1}\right)^{2} \xi_{1}-\frac{2}{3} \xi_{2} \overline{r_{1}}\left(f^{\prime}-k_{1}\right)-\frac{2}{3} f^{\prime} k_{1} \xi_{1}^{3}{\overline{r_{1}}}^{2}\left(f^{\prime}-k_{1}\right) \\
& -\left[-\frac{8}{9}\left(f^{\prime}-k_{1}\right) \xi_{1} \xi_{3} \overline{r_{1}}+\frac{8 k_{1}\left(f^{\prime}-k_{1}\right)^{3}}{9 \overline{r_{1}}}-\frac{4 \xi_{2}\left(f^{\prime}-k_{1}\right)^{2}}{9 \xi_{1}}-\frac{22}{9}\left(f^{\prime}-k_{1}\right)^{2} f^{\prime} k_{1} \xi_{1}^{2} \overline{r_{1}}\right. \\
& -\frac{1}{9}\left(f^{\prime}+k_{1}\right) \xi_{1}^{3} \xi_{3}{\overline{r_{1}}}^{3}+\frac{4}{9}\left(f^{\prime}+k_{1}\right)\left(f^{\prime}-k_{1}\right)^{2} k_{1} \xi_{1}^{2} \overline{r_{1}} \\
& \left.-\frac{2}{9}\left(f^{\prime}+k_{1}\right)\left(f^{\prime}-k_{1}\right) \xi_{1} \xi_{2}{\overline{r_{1}}}^{2}-\frac{2}{9} f^{\prime} k_{1}\left(f^{\prime}+k_{1}\right)\left(f^{\prime}-k_{1}\right) \xi_{1}^{4}{\overline{r_{1}}}^{3}\right] \theta+\cdots . \tag{30}
\end{align*}
$$

Similarly, we truncate the $\theta$ series of $b_{3}$ at the resonance parameter $r_{4}$ and we obtain the following $\theta$-series:

$$
\begin{align*}
b_{3}(t) & =\frac{2}{\overline{r_{1}^{2}}} \theta^{-2}-\frac{2\left(f^{\prime}-k_{1}\right) \xi_{1}}{\overline{r_{1}}} \theta^{-1}+\left[\frac{4\left(f^{\prime}-k_{1}\right)\left(f^{\prime}+k_{1}\right)}{\overline{r_{1}} 2}-\frac{2 \xi_{2}}{\xi_{1} \overline{r_{1}}}+f^{\prime} k_{1}\right] \\
& +\left[\frac{2}{\overline{r_{1}}} \xi_{3}+\frac{4 f^{\prime} k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)}{\overline{r_{1}}}+\frac{\xi_{1}^{2} \xi_{3} \overline{r_{1}}}{3}-\frac{4 k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)^{2}}{3 \overline{r_{1}}}+\frac{2}{3} \xi_{2}\left(f^{\prime}-k_{1}\right)\right. \\
& \left.+\frac{2}{3} f^{\prime} k_{1} \xi_{1}^{3} \overline{r_{1}}\left(f^{\prime}-k_{1}\right)\right] \theta+\xi_{4} \theta^{2} . \tag{31}
\end{align*}
$$

Hence, we have

$$
\begin{align*}
\xi_{4}= & \frac{2}{9} \xi_{1} \xi_{3}\left(7 k_{1}-4 f^{\prime}\right)-\frac{4 k_{1}}{9{\overline{r_{1}}}^{2}}\left(f^{\prime}-k_{1}\right)^{2}\left(7 f^{\prime}+8 k_{1}\right)+\frac{4}{9 \xi_{1} \overline{r_{1}}}\left(f^{\prime}-k_{1}\right) \\
& \left(4 k_{1}-f^{\prime}\right) \xi_{2}+\frac{4}{9}\left(f^{\prime}-k_{1}\right)^{2}\left(-10 f+k_{1}\right) k_{1} \xi_{1}^{2}-\frac{2}{9}\left(2 f^{\prime}-k_{1}\right) \xi_{1}^{3} \xi_{3}{\overline{r_{1}}}^{2}  \tag{32}\\
- & \frac{2}{9}\left(f^{\prime}-k_{1}\right)\left(2 f^{\prime}-k_{1}\right) \xi_{1} \xi_{2} \overline{r_{1}}-\frac{4}{9} f^{\prime} k_{1} \xi_{1}^{4} \bar{r}_{1}^{2}\left(f^{\prime}-k_{1}\right)\left(2 f^{\prime}-k_{1}\right)-\frac{2}{\overline{r_{1}}} r_{4} \\
+ & \frac{4}{3} k_{1} \xi_{1}^{2}\left(f^{\prime}-k_{1}\right)\left(f^{\prime 2}-2 f^{\prime} k_{1}+k_{1}^{2}\right)+\cdots .
\end{align*}
$$

Using (21), (25), (28) and (31) with $w_{1}=\theta^{-1}$, we get the change of variables $\left(w_{1}, w_{2}, b_{1}, b_{2}, b_{3}\right) \longleftrightarrow\left(\theta, \xi_{1}, \xi_{2}, \xi_{3}, \xi_{4}\right)$. The following is the conversion of given system into the mirror system in terms of the new variables $\theta, \xi_{1}, \xi_{2}, \xi_{3}$ and $\xi_{4}$ :

$$
\begin{align*}
\theta^{\prime} & =\frac{1}{\bar{r}_{1}}+\left(k_{1}-f^{\prime}\right) \xi_{1} \theta+\left[\frac{2\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)}{\overline{r_{1}}}+\frac{1}{2} f^{\prime} k_{1} \overline{r_{1}}\left(1+\xi_{1}^{2}\right)-\frac{\xi_{2}}{\xi_{1}}\right] \theta^{2} \\
& +\xi_{3} \theta^{3}, \\
\xi_{1}^{\prime} & =\left[-\left(1+\xi_{1}^{2}\right) f^{\prime}-\frac{2\left(f^{\prime}+k_{1}\right)}{\bar{r}_{1}^{2}}\right]+\left[\frac{2\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)}{\overline{r_{1}}}+\frac{1}{2} f^{\prime} k_{1} \overline{r_{1}} \xi_{1}\left(1+\xi_{1}^{2}\right)\right] \theta \\
& +\xi_{1} \xi_{3} \theta^{2}, \\
\xi_{2}^{\prime} & =\left[\frac{\left(-1-\xi_{1}^{2}\right)\left(f^{\prime}+k_{1}\right)}{\overline{r_{1}}}-\frac{4\left(f^{\prime}+k_{1}\right)}{\overline{r_{1}}}\right] \theta^{-2}+\left[\frac{2 \xi_{1}}{\overline{r_{1}^{2}}}\left(2 f^{\prime 2}-4 k_{1}^{2}-3 k_{1} f^{\prime}\right)\right. \\
& \left.-\frac{f^{\prime} k_{1} \xi_{1}}{2}\left(3+5 \xi_{1}^{2}\right)-k_{1}^{2} \xi_{1}-\left(k_{1}-f^{\prime}\right) k_{1} \xi_{1}^{3}\right] \theta^{-1}+\left[-\frac{4 k_{1} \xi_{1}^{2}\left(f^{\prime}-k_{1}\right)^{2}}{3 \overline{r_{1}}}-\frac{f^{\prime} k_{1}^{2} \xi_{1}^{4} \overline{r_{1}}}{6}\right. \\
& +\frac{\xi_{1} \xi_{3}+5 f^{\prime 2} \xi_{1}^{2} k_{1}-3 f^{\prime} k_{1}^{2} \xi_{1}^{2}-3 k_{1}^{2} f^{\prime}-3 f^{\prime 2} k_{1}+2 k_{1}^{3}\left(1-\xi_{1}^{2}\right)}{\overline{r_{1}}-\frac{f^{\prime} k_{1}^{2} \overline{r_{1}}}{2}+\frac{k_{1} \xi_{2}}{\xi_{1}}} \\
& +\frac{1}{3}\left(\xi_{1}^{3} \xi_{3} \overline{r_{1}}+\xi_{1} \xi_{2}\left(2 f^{\prime}+k_{1}\right)+2 f^{\prime 2} k_{1} \xi_{1}^{4} \overline{r_{1}}\right)-\frac{4\left(f^{\prime}+k_{1}\right)\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)}{\overline{r_{1}^{3}}} \\
& \left.+\frac{2 \xi_{2}\left(f^{\prime}+k_{1}\right)}{\xi_{1} \overline{r_{1}^{2}}}\right]+\left[\xi_{1} \xi_{4}+k_{1} \xi_{3}\left(\xi_{1}^{2}-1\right)-\frac{2 \xi_{3}\left(f^{\prime}+k_{1}\right)}{\overline{r_{1}^{2}}}\right] \theta, \tag{33}
\end{align*}
$$

$$
\begin{align*}
& \xi_{3}{ }^{\prime}=\left[\frac{\left(-1-\xi_{1}^{2}\right)\left(f^{\prime}+k_{1}\right)}{\xi_{1} \overline{r_{1}}}-\frac{4\left(f^{\prime}+k_{1}\right)}{\xi_{1}{\overline{r_{1}}}^{3}}\right] \theta^{-3}+\left[\frac{-2 k_{1} f^{\prime}-8 k_{1}^{2}+4 f^{\prime 2}}{{\overline{r_{1}}}^{2}}-k_{1}^{2}\left(\xi_{1}^{2}+1\right)\right. \\
& \left.-\frac{1}{2} f^{\prime} k_{1}\left(\xi_{1}^{2}+1\right)\right] \theta^{-2}+\left[\frac{\xi_{3}+3 f^{\prime 2} k_{1} \xi_{1}+3 f^{\prime} k_{1}^{2} \xi_{1}}{\overline{r_{1}}}+k_{1} \xi_{2}+\frac{1}{2} f^{\prime} k_{1}^{2} \xi_{1}^{3} \overline{r_{1}}-\xi_{3}\right. \\
& +f^{\prime 2} k_{1} \xi_{1} \overline{r_{1}}\left(\xi_{1}^{2}+1\right)-\frac{3 k_{1}^{2} f^{\prime}+3 f^{\prime 2} k_{1}-2 k_{1}^{3}\left(1-\xi_{1}^{2}\right)}{\xi_{1} \overline{r_{1}}}-\frac{f^{\prime} k_{1}^{2} \overline{r_{1}}}{2 \xi_{1}}+\frac{k_{1} \xi_{2}+\left(1+\xi_{1}^{2}\right) f^{\prime} \xi_{2}}{\xi_{1}^{2}} \\
& \left.-\frac{4\left(f^{\prime}+k_{1}\right)\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)}{\xi_{1}{\overline{r_{1}}}^{3}}+\frac{4 \xi_{2}\left(f^{\prime}+k_{1}\right)}{\xi_{1}^{2} \bar{r}_{1}^{2}}\right] \theta^{-1}+\left[\frac{-2 f^{\prime} k_{1} \xi_{1}^{2} \overline{r_{1}}\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)}{\overline{r_{1}}}\right. \\
& -\frac{1}{2} f^{\prime 2} k_{1}^{2}{\overline{r_{1}}}^{2} \xi_{1}^{2}\left(1+\xi_{1}^{2}\right)+\frac{k_{1} \xi_{3}\left(\xi_{1}^{2}-1\right)}{\xi_{1}}-\frac{2 \xi_{3}\left(f^{\prime}+k_{1}\right)}{\xi_{1}{\overline{r_{1}}}^{2}}-\frac{2 \xi_{2}}{\xi_{1} \overline{r_{1}}}\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right) \\
& \left.-\frac{f^{\prime} k_{1} \overline{r_{1}} \xi_{2}\left(1+\xi_{1}^{2}\right)}{2 \xi_{1}}\right]+\left[-f^{\prime} k_{1} \xi_{1}^{2} \xi_{3} \overline{r_{1}}-\frac{\xi_{2} \xi_{3}}{\overline{r_{1}}}\right] \theta, \\
& \xi{ }^{\prime}=\left[\frac{\xi_{1}^{2}+1}{\overline{r_{1}}}+\frac{4}{\overline{\bar{r}_{1}^{3}}}\right] \theta^{-5}+\left[k_{1} \xi_{1}\left(1+\xi_{1}^{2}\right)+\frac{4 \xi_{1}}{{\overline{r_{1}}}^{2}}\left(-f^{\prime}+2 k_{1}\right)+\frac{4 \xi_{1}}{3 \overline{r_{1}^{2}}}\left(f^{\prime}+k_{1}\right)\right. \\
& \left.+\frac{\xi_{1}\left(\xi_{1}^{2}+1\right)\left(f^{\prime}+k_{1}\right)}{3}\right] \theta^{-4}+\left[\frac{f^{\prime} k_{1} \overline{r_{1}}}{2}+\frac{2 f^{\prime} k_{1} \xi_{1}^{2} \overline{r_{1}}}{3}-\frac{\xi_{2}}{\xi_{1}}+\frac{4}{\overline{r_{1}^{3}}}\left(f^{\prime 2}-k_{1}^{2}\right)\right. \\
& +\frac{1}{\overline{r_{1}}}\left(4 f^{\prime} k_{1}-2 k_{1}^{2}+2 k_{1}^{2} \xi_{1}^{2}-2 f^{\prime} k_{1} \xi_{1}^{2}\right)-\frac{4 \xi_{2}}{\xi_{1} \overline{r_{1}^{2}}}+\frac{1}{6} f^{\prime} k_{1} \xi_{1}^{4} \overline{r_{1}}+\frac{8\left(f^{\prime 2}-k_{1}^{2}\right)}{3 \overline{r_{1}^{3}}} \\
& \left.+\frac{1}{3 \overline{r_{1}}}\left(2 k_{1} f^{\prime} \xi_{1}^{2}+6 k_{1}^{2} \xi_{1}^{2}-2 f^{\prime 2} \xi_{1}^{2}-2\left(k_{1}^{2}-f^{\prime 2}\right)\right)+\frac{\xi_{1}^{2} \overline{r_{1}} k_{1}^{2}}{3}\left(\xi_{1}^{2}+1\right)\right] \theta^{-3} \\
& +\left[\frac{2}{3} f^{\prime} k_{1}^{2} \xi_{1}\left(1+\xi_{1}^{2} \overline{r_{1}}-4 \xi_{1}^{2}\right)-\frac{1}{6} f^{\prime} k_{1}^{2} \xi_{1}{\overline{r_{1}}}^{2}\left(\xi_{1}^{4}-1\right)-\frac{2}{3} f^{\prime 2} k_{1} \xi_{1}^{3}\left(1+\overline{r_{1}}\right)\right. \\
& -\frac{1}{3} f^{\prime 2} k_{1} \xi_{1}^{3}{\overline{r_{1}}}^{2}\left(1+\xi_{1}^{2}\right)-\frac{1}{3} f^{\prime} \xi_{2}\left(2+\overline{r_{1}}\right)+\frac{1}{3} k_{1} \xi_{2}\left(2-\overline{r_{1}}\right)-\frac{1}{3} \xi_{1}^{2} \xi_{2} \overline{r_{1}}\left(f^{\prime}+k_{1}\right) \\
& +\frac{1}{\overline{\bar{r}_{1}^{2}}}\left(4 \xi_{3}+4 f^{\prime 2} k_{1} \xi_{1}-12 f^{\prime} k_{1}^{2} \xi_{1}\right)+\frac{1}{\overline{r_{1}}}\left(2 \xi_{2}\left(f^{\prime}-2 k_{1}\right)-4 f^{\prime} k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)\right) \\
& +\frac{1}{3 \overline{r_{1}^{2}}}\left(20 f^{\prime} k_{1} \xi_{1}\left(f^{\prime}+k_{1}\right)-28 k_{1}^{3} \xi_{1}-4 \xi_{1} f^{\prime 3}\right)+\frac{1}{3}\left(\xi_{1}^{2} \xi_{3}-4 k_{1}^{3} \xi_{1}\right) \\
& \left.+\frac{1}{3 \overline{r_{1}}}\left(-2 k_{1} \xi_{2}+4 k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)^{2}\right)\right] \theta^{-2}+\left[2 f^{\prime 3} \xi_{1}^{2} k_{1} \overline{r_{1}}\left(1+\xi_{1}^{2}\right)\right. \\
& -f^{\prime 2} k_{1}^{2} \xi_{1}^{2} \overline{r_{1}}\left(1+\frac{1}{9} \xi_{1}^{2}\right)+\frac{1}{\overline{r_{1}}}\left(-8 f^{\prime} k_{1}^{3} \xi_{1}^{2}+12 f^{\prime 3} k_{1} \xi_{1}^{2}+4 f^{\prime 3} k_{1}-4 f^{\prime 2} k_{1}^{2}\right) \\
& -\frac{4 \xi_{2}\left(f^{\prime 2}-k_{1}^{2}\right)}{3 \xi_{1}{\overline{r_{1}}}^{2}}-\frac{2\left(f^{\prime}-k_{1}\right) k_{1} \xi_{2}}{3 \xi_{1}}+\frac{8 f^{\prime} k_{1}\left(f^{\prime 2}-k_{1}^{2}\right)}{{\overline{r_{1}}}^{3}}+\frac{1}{9} f^{\prime} k_{1} \xi_{1}^{4} \overline{r_{1}}\left(-7 k_{1}^{2}+4 f^{\prime 2}\right) \\
& +\frac{1}{6} f^{\prime 2} k_{1}^{2} \xi_{1}^{4} \bar{r}_{1}^{3}\left(1+\xi_{1}^{2}\right)+\frac{1}{3} \xi_{1}^{3} \xi_{3} \overline{r_{1}}\left(f^{\prime}-\frac{1}{3} k_{1}\right)+\xi_{1} \xi_{3} \overline{r_{1}}\left(2+\frac{1}{3} k_{1}\right) \\
& +\frac{4}{3} k_{1} \xi_{1} \xi_{2}\left(2 f^{\prime}-k_{1}\right)+\frac{2}{3} f^{\prime 2} \xi_{1} \xi_{2}+\frac{1}{2} f^{\prime} k_{1} \overline{r_{1}} \xi_{1}\left(\xi_{2}+\xi_{1}^{2} \xi_{3}\right)+2 \xi_{4} \xi_{3} \overline{r_{1}} \\
& +\quad f^{\prime} k_{1}^{2} \bar{r}_{1}\left(f^{\prime}-k_{1}\right)+\frac{1}{3 \bar{r}_{1}}\left(4 \xi_{1} \xi_{3}\left(f^{\prime}+2 k_{1}\right)+2 f^{\prime 3} k_{1}\left(1-\frac{17}{3} \xi_{1}^{2}\right)+2 f^{\prime} k_{1}^{3}\left(-7+\xi_{1}^{2}\right)\right. \\
& \left.\left.+8 f^{\prime 2} k_{1}^{2}\left(1+2 \xi_{1}^{2}\right)+4 k_{1}^{4}\left(1-\frac{5}{3} \xi_{1}^{2}\right)+\frac{8 f^{\prime 2}\left(f^{\prime 2}-k_{1}^{2}\right)}{3 \overline{r_{1}^{3}}}\right)\right] \theta^{-1}+\left[\frac { 1 } { { \overline { r _ { 1 } } } ^ { 2 } } \left(2 \xi_{2} \xi_{3}\right.\right. \\
& \left.-8 f^{\prime} k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)\right)+\frac{1}{3{\overline{r_{1}^{2}}}^{2}}\left(8 k_{1}\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)\left(f^{\prime}-k_{1}\right)^{2} \xi_{1}\right. \\
& \left.+4 \xi_{3}\left(f^{\prime-} k_{1}^{2}\right)\right)+\frac{1}{3}\left(-f^{\prime} k_{1} \xi_{1}^{2} \xi_{3}{\overline{r_{1}}}^{2}+2 k_{1}^{2} f^{\prime} \xi_{1}\left(f^{\prime}-k_{1}\right)^{2}\left(1+\xi_{1}^{2}\right)-\xi_{3}\left(\xi_{1}^{2}-1\right)\right. \\
& \left.-2\left(f^{\prime}-k_{1}\right) \xi_{1} \xi_{4}\right)-\frac{\xi_{2} \xi_{3}}{\xi_{1}}+2 \xi_{1}^{2} \xi_{3}\left(f^{\prime} k_{1}+\xi_{2}\right)-f^{\prime 2} k_{1}^{2} \xi_{1}\left(f^{\prime}-k_{1}\right)\left(1+\xi_{1}^{2}\right) \\
& \left.\left(2+\xi_{1}^{2}{\overline{r_{1}}}^{2}\right)-4 \xi_{1}^{2}\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)\left(\xi_{3}+f^{\prime} k_{1} \xi_{1}\left(f^{\prime}-k_{1}\right)\right)-2 \xi_{1} \xi_{4}\left(f^{\prime}-k_{1}\right)\right] \\
& +\left[-\frac{2 \xi_{1}^{2} \xi_{3} \overline{r_{1}}}{3}+\frac{4 k_{1} \xi_{1} \xi_{3}\left(f^{\prime}-k_{1}\right)\left(-2 f^{\prime}-k_{1}\right)}{3 \overline{r_{1}}}-\frac{4 \xi_{4}\left(k_{1} f^{\prime}-k_{1}^{2}+f^{\prime 2}\right)}{\overline{r_{1}}}\right. \\
& \left.-2 f^{\prime} k_{1} \xi_{1}^{3} \xi_{3} \overline{r_{1}}-f^{\prime} k_{1} \xi_{4} \overline{r_{1}}\left(1+\xi_{1}^{2}\right)+\frac{2 \xi_{2} \xi_{4}}{\xi_{1}}\right] \theta-2 \xi_{3} \xi_{4} \theta^{2} . \tag{34}
\end{align*}
$$

By similar calculations, we can find the mirror system for the following branch of leading order coefficients:

$$
\begin{align*}
& l_{0}=\left(\sqrt{-4-r_{1}^{2}}\right)^{-1}, \quad w_{20}=-\frac{r_{1}}{\sqrt{-4-r_{1}^{2}}}, \\
& b_{10}=\frac{r_{1}}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}}, \quad b_{20}=\left(\sqrt{-4-r_{1}^{2}}\right)^{-1}, \quad b_{30}=\frac{2}{\left(\sqrt{-4-r_{1}^{2}}\right)^{2}} . \tag{35}
\end{align*}
$$

The mirror system obtained so far for the present case of leading order coefficient is regular if and only if the following condition are satisfied:

$$
\begin{equation*}
\xi_{1}=\frac{\sqrt{-4-\overline{r_{1}^{2}}}}{\overline{r_{1}}}, \xi_{2}=-\frac{26 k_{1}^{2} \xi_{1}}{9 \overline{r_{1}}}, f^{\prime}=k_{1}, \xi_{3}=0 \tag{36}
\end{equation*}
$$

The most prominent thing for the singularity analysis is that the system is regular near $\theta=0$, which corresponds to movable singularity of the system of six coupled ODEs (5).

## 3 Alternative Approach of the Convergence of Laurent Series in Painlevé Test

The convergence of Laurent series solution obtained by the Painlevé test is guaranteed by Kichenassamy and Littman [4]. But here we are going to present an alternative approach of the convergence of these series by making use of the mirror system and the Cauchy-Kowalevski theorem.

An ideal rotating, uniformly stratified system of six coupled ODEs (5) is completely integrable for the Rayleigh number $R a=0$. For $R a=0$, the Painlevé test produces the following formal solution of ODEs (5) for the first case of leading order coefficients:

$$
\begin{align*}
w_{1}(t) & =\sqrt{-4-k_{2}^{2}} \tau^{-1}+\frac{\left(f^{\prime}-k_{1}\right) k_{2}}{2}+\frac{\sqrt{-4-k_{2}^{2}}}{2}\left(-k_{3}+f^{\prime} k_{1}\right) \tau \\
& +\left[-\frac{k_{4}}{2}+\frac{f^{\prime} k_{2}}{4}\left(-k_{3}+f^{\prime} k_{1}\right)\right] \tau^{2} \\
& +\left\{-\frac{k_{5}}{3}+\frac{f^{\prime} \sqrt{-4-k_{2}^{2}}}{12 k_{2}}\left[f^{\prime} k_{2}\left(k_{3}-f^{\prime} k_{1}\right)+2 k_{4}\right]\right\} \tau^{3} \\
& +\sum_{j=5}^{\infty} w_{1 j} \tau^{j-1}, \\
w_{2}(t) & =k_{2} \tau^{-1}+\left[\frac{\sqrt{-4-k_{2}^{2}}}{2}\left(-f^{\prime}+k_{1}\right)\right]+\frac{\left(-k_{3} k_{2}+f^{\prime} k_{2} k_{1}\right)}{2} \tau  \tag{37}\\
& +\sqrt{-4-k_{2}^{2}}\left[\frac{k_{4}}{2 k_{2}}+\frac{f^{\prime}}{4}\left(k_{3}-f^{\prime} k_{1}\right)\right] \tau^{2} \\
& +\left[\frac{-k_{5} k_{2}}{3 \sqrt{-4-k_{2}^{2}}}+\frac{f^{\prime}}{12}\left(f^{\prime} k_{2} k_{3}+2 k_{4}-f^{\prime 2} k_{2} k_{1}\right)\right] \tau^{3}+\sum_{j=5}^{\infty} w_{1 j} \tau^{j-1}, \\
b_{1}(t) & =-k_{2} \tau^{-2}+f^{\prime} \sqrt{-4-k_{2}^{2}} \tau^{-1}+\frac{\left(-k_{2} k_{3}+f^{\prime 2} k_{2}\right)}{2}+\frac{k_{4} \sqrt{-4-k_{2}^{2}}}{k_{2}} \tau \\
& -\frac{k_{5} k_{2}}{\sqrt{-4-k_{2}^{2}} \tau^{2}+\sum_{j=5}^{\infty} b_{1 j} \tau^{j-2},}
\end{align*}
$$

$$
\begin{align*}
b_{2}(t) & =\sqrt{-4-k_{2}^{2}} \tau^{-2}+f^{\prime} k_{2} \tau^{-1}+\left[\frac{\sqrt{-4-k_{2}^{2}}}{2}\left(k_{3}-f^{\prime 2}\right)\right]+k_{4} \tau+k_{5} \tau^{2} \\
& +\sum_{j=5}^{\infty} b_{2 j} \tau^{j-2} \\
b_{3}(t) & =2 \tau^{-2}+k_{3}-\frac{4 k_{5}}{3 \sqrt{-4-k_{2}^{2}}} \tau^{2}  \tag{38}\\
& -\frac{1}{6 k_{2}}\left(f^{\prime 2} k_{2} k_{3}-3 k_{2} k_{3}^{2}+2 f^{\prime} k_{4}-f^{\prime 3} k_{2} k_{1}+3 f^{\prime} k_{2} k_{3} k_{1}-6 k_{4} k_{1}\right) \tau^{2} \\
& +\sum_{j=5}^{\infty} b_{3 j} \tau^{j-2} .
\end{align*}
$$

The above Laurent series contains the five arbitrary constant $w_{30}=k_{1}, k_{2}, k_{3}, k_{4}$ and $k_{5}$. Here $\tau=t-t_{0}$ and $t_{0}$ is an arbitrary position of singularity in complex domain. As we see, the above Laurent series has a movable pole type singularity, and using the Painlevé method we conclude that the above Laurent series (37) and (38) are convergent for small $\tau$; and this convergence is guaranteed by Kichenassamy and Littman [4]. But for an alternative approach, we convert these series into an initial value problem for the mirror system (33) and (34). For this purpose we substitute the formal Laurent series (37) and (38) into the mirror transformation $w_{1}=\theta^{-1}$, (21), (25), (28) and (31). After simplification, we obtain the following formal power series for $\theta, \xi_{1}, \xi_{2}, \xi_{3}$ and $\xi_{4}$ :

$$
\begin{align*}
\theta & =\left(\sqrt{-4-k_{2}^{2}}\right)^{-1} \tau-\frac{\left(f^{\prime}-k_{1}\right) k_{2}}{\left(\sqrt{-4-k_{2}^{2}}\right)^{2}} \tau^{2}+\frac{1}{4\left(\sqrt{-4-k_{2}^{2}}\right)^{3}} \\
& +\left[\frac{\left(-8 k_{3}-2 k_{2}^{2} k_{3}+8 f^{\prime} k_{1}+f^{\prime 2} k_{2}^{2}+k_{1}^{2} k_{2}^{2}\right) \tau^{3}}{2\left(\sqrt{-4-k_{2}^{2}}\right)^{2}}\left(k_{4}+\left(-k_{3}+f^{\prime} k_{1}\right)\left(\frac{f^{\prime} k_{2}}{2}-k_{1} k_{2}\right)\right)-\frac{\left(f^{\prime}-k_{1}\right)^{3} k_{2}^{3}}{8\left(\sqrt{-4-k_{2}^{2}}\right)^{4}}\right] \tau^{4} \\
& +\left[\frac{k_{5}}{3\left(\sqrt{-4-k_{2}^{2}}\right)^{2}}+\frac{1}{\sqrt{-4-k_{2}^{2}}}\left(-\frac{f^{\prime 2}}{12}\left(k_{3}-f^{\prime} k_{1}\right)-\frac{f^{\prime} k_{4}}{6 k_{2}}+\frac{\left(-k_{3}+f^{\prime} k_{1}\right)^{2}}{4}\right)\right. \\
+ & \frac{1}{2\left(\sqrt{-4-k_{2}^{2}}\right)^{3}}\left(-k_{4}\left(f^{\prime}-k_{1}\right) k_{2}+\frac{\left(f^{\prime}-k_{1}\right) k_{2}^{2}}{4}\left(-k_{3}+f^{\prime} k_{1}\right)\left(3 k_{1}-f^{\prime}\right)\right) \\
& \left.+\frac{\left(f^{\prime}-k_{1}\right)^{4} k_{2}^{4}}{\left(16 \sqrt{-4-k_{2}^{2}}\right)^{5}}\right] \tau^{5}+\ldots, \\
\xi_{1} & \frac{k_{2}}{\sqrt{-4-k_{2}^{2}}}-\frac{2\left(f^{\prime}-k_{1}\right)}{4+k_{2}^{2}} \tau-\frac{k_{2}\left(f^{\prime}-k_{1}\right)^{2}}{\left(-4-k_{2}^{2}\right)^{\frac{3}{2}}} \tau^{2}+\left[\frac { 1 } { ( - 4 - k _ { 2 } ^ { 2 } ) } \left(\frac{k_{2} k_{4}}{2 \sqrt{-4-k_{2}^{2}}}\right.\right. \\
& \left.-\frac{1}{4}\left(f^{\prime} k_{1}-k_{3}\right) k_{1} k_{2}^{2}+\frac{1}{8}\left(-f^{\prime}+k_{1}\right)\left(8 f^{\prime} k_{1}+f^{\prime 2} k_{2}^{2}+k_{1}^{2} k_{2}^{2}\right)\right) \\
& +\frac{k_{4}}{\left.2 k_{2}-f^{\prime 2} k_{1}+\frac{k_{1} k_{3}}{4}-\frac{\left(f^{\prime}-k_{1}\right)^{3} k_{2}^{4}}{8\left(-4-k_{2}^{2}\right)^{2}}\right] \tau^{3}+\ldots,} \\
\xi_{2} & =\frac{k_{2}}{2}\left(f^{\prime 2}-3 k_{3}+2 f^{\prime} k_{1}\right)+\frac{1}{\left(-4-k_{2}^{2}\right)}\left(f^{\prime}-k_{1}\right) k_{2}\left(\frac{1}{4} k_{2}^{2}\left(f^{\prime}+k_{1}\right)+2\left(f^{\prime}+2 k_{1}\right)\right) \\
& +\left[\frac { 1 } { \sqrt { - 4 - k _ { 2 } ^ { 2 } } } \left(-\frac{k_{2} k_{4}}{2}+\frac{1}{4}\left(f^{\prime} k_{1}-k_{3}\right)\left(4 f^{\prime}-k_{1}\right) k_{2}^{2}+2 f^{\prime 2} k_{1}-3 k_{3}\left(f^{\prime}-k_{1}\right)\right.\right. \\
& \left.\left.-\frac{1}{4} f^{\prime} k_{1} k_{2}^{2}\left(f^{\prime}-k_{1}\right)\right)+\sqrt{-4-k_{2}^{2}}\left(\frac{k_{4}}{2 k_{2}}-f^{\prime 2} k_{1}+\frac{k_{1} k_{3}}{4}\right)+\frac{4 k_{1}\left(f^{\prime}-k_{1}\right)^{2}}{\left(-4-k_{2}^{2}\right)^{\frac{3}{2}}}\right] \tau+\ldots,
\end{align*}
$$

$$
\begin{align*}
& \xi_{3}=\sqrt{-4-k_{2}^{2}}\left(\frac{3}{2} k_{4}-\frac{2 k_{4}}{k_{2}}-\frac{k_{4} k_{2}}{2}+\frac{5 f^{\prime 2} k_{1}}{k_{2}}-\frac{k_{1} k_{3}}{2}+\frac{1}{2} f^{\prime 2} k_{1} k_{2}+\frac{1}{4} f^{\prime} k_{1}^{2} k_{2}-\frac{f^{\prime 3}}{k_{2}}\right. \\
& \left.+\frac{2 f^{\prime} k_{1}^{2}}{k_{2}}\right)+\frac{\left(f^{\prime}-k_{1}\right)}{\sqrt{-4-k_{2}^{2}}}\left[-2 f^{\prime} k_{1} k_{2}+\frac{1}{2} k_{2}\left(k_{1}^{2}-f^{\prime 2}\right)-\frac{3}{8} f^{\prime} k_{1} k_{2}^{3}+\frac{1}{4} k_{1}^{2} k_{2}^{3}\right. \\
& \left.-\frac{4 f^{\prime}}{k_{2}}\left(f^{\prime}-k_{1}\right)\right]+\left\{-\frac{3}{8} k_{1}^{2} k_{2}^{2} k_{3}+\frac{1}{2} f^{\prime} k_{2} k_{4}-\frac{3}{8} f^{\prime 2} k_{2}^{2} k_{3}+\frac{5}{8} f^{\prime 3} k_{1} k_{2}^{2}-\frac{15}{8} k_{1} k_{2} k_{4}\right. \\
& -\frac{1}{2} f^{\prime 2} k_{1}^{2}-2 f^{\prime} k_{1}^{3}+\frac{3}{2} f^{\prime 3} k_{1}+\frac{7}{2} f^{\prime} k_{1} k_{3}-\frac{3}{2} k_{1}^{2} k_{3}-f^{\prime 2} k_{3}-\frac{5}{8} f^{\prime 2} k_{1}^{2} k_{2}^{2} \\
& +f^{\prime} k_{1} k_{2}^{2} k_{3}-f^{\prime} k_{1}\left(f^{\prime}-k_{1}\right)+\frac{1}{4} k_{1}^{3} k_{2}^{2}\left(f^{\prime}-\frac{1}{2} k_{1}^{2}\right)+\frac{1}{k_{2}}\left[2 f^{\prime 4}-6 f^{\prime} k_{3}\left(f^{\prime}-2 k_{1}\right)\right. \\
& \left.-6 k_{1}^{2}\left(f^{\prime 2}+k_{3}+4 f^{\prime} k_{1}^{3}+f^{\prime} k_{4}+\frac{2 f^{\prime 3} k_{1}}{k_{2}}+\frac{6 k_{3}\left(f^{\prime}-k_{1}\right)^{2}}{k_{2}}-k_{1} k_{4}-\frac{2 f^{\prime 2} k_{1}^{2}}{k_{2}}\right)\right] \\
& +\left(-4-k_{2}^{2}\right)\left[-\frac{5}{12} f^{\prime 2} k_{3}+\frac{f^{\prime} k_{4}}{6 k_{2}}-\frac{13}{12} f^{\prime 3} k_{1}-\frac{1}{4} f^{\prime} k_{1} k_{3}+\frac{1}{2}\left(k_{3}^{2}+f^{\prime 3} k_{1}\right)-\frac{k_{1} k_{4}}{k_{2}}\right. \\
& \left.+2 f^{\prime 2} k_{1}^{2}-\frac{k_{1}^{2}}{2}\left(4 f^{\prime 2}-k_{3}\right)-\frac{1}{k_{2}^{3}} k_{4}\left(k_{1}-2 f^{\prime}\right)+\frac{2 f^{\prime 2} k_{1}}{k_{2}^{2}}\left(f^{\prime}-k_{1}\right)+\frac{k_{1}^{2} k_{3}}{2 k_{2}^{2}}\right] \\
& +\sqrt{-4-k_{2}^{2}}\left[\frac{5}{3} k_{5}+\frac{1}{k_{2}}\left(6 f^{\prime 2} k_{1}+2 k_{1} k_{3}-3 f^{\prime} k_{3}\right)-k_{4}+\frac{7}{4} f^{\prime 2} k_{1} k_{2}-\frac{2 k_{4}}{k_{2}^{2}}\right. \\
& \left.-f^{\prime} k_{2} k_{3}\right]+\frac{1}{\left(-4-k_{2}^{2}\right)}\left[k_{1}\left(f^{\prime}-k_{1}\right)^{3}\left(4-\frac{8}{k_{2}^{2}}+\frac{k_{2}^{4}}{8}+\frac{1}{2} k_{2}^{2}\right)+\frac{4\left(f^{\prime}-k_{1}\right)^{3}}{k_{2}}\right. \\
& \left.\left.\left(\frac{1}{4} k_{2}^{2}\left(f^{\prime}+k_{1}\right)+2\left(f^{\prime}+2 k_{1}\right)\right)\right]\right\} \tau+\ldots, \\
& \xi_{4}=\frac{5}{18} k_{5} \sqrt{-4-k_{2}^{2}}+\frac{2}{3}\left(f^{\prime}-k_{1}\right)^{2}\left(2 f^{\prime}+k_{1}\right) k_{1}-12 f^{\prime} k_{1}\left(f^{\prime}-k_{1}\right)^{2} k_{2}^{2} \\
& -\frac{1}{\sqrt{-4-k_{2}^{2}}}\left(f^{\prime}-k_{1}\right)^{3}\left(f^{\prime}+k_{1}\right) k_{2}^{2}+\frac{1}{3}\left(f^{\prime}-k_{1}\right)^{2} \sqrt{-4-k_{2}^{2}}+\frac{\left(f^{\prime}+k_{1}\right)^{2} k_{2}^{2}}{3\left(4+k_{2}^{2}\right)} \\
& {\left[12 f^{\prime} k_{1}+f^{\prime} k_{1} k_{2}^{2}-3\left(f^{\prime}-k_{1}\right)^{2} k_{2}^{2}\right]+\frac{1}{3} k_{2}\left(f^{\prime}-k_{1}\right)^{2}\left(\frac{8}{k_{2}^{3}}-\frac{2}{k_{2}}+\frac{1}{2} k_{2}\right.} \\
& \left.-\frac{2\left(4+k_{2}^{2}\right)\left(f^{\prime} k_{1}-k_{3}\right)}{k_{2}\left(f^{\prime}-k_{1}\right)^{2}}\right)\left[\frac{f^{\prime 2}+2 f^{\prime} k_{1}-\left(f^{\prime}-k_{1}\right)\left(f^{\prime}\left(8+k_{2}^{2}\right)+k_{1}\left(16+k_{2}^{2}\right)\right)}{2\left(4+k_{2}^{2}\right)}-3 k_{3}\right] \\
& +2\left(f^{\prime} k_{1}-k_{3}\right)\left[3\left(f^{\prime}-k_{1}\right)^{2}-2\left(f^{\prime 2}-k_{1}^{2}\right) \sqrt{-4-k_{2}^{2}}\right]+\frac{7}{2}\left(4+k_{2}^{2}\right) \\
& \left(f^{\prime} k_{1}-k_{3}\right)^{2}+\frac{1}{4}\left(f^{\prime}-k_{1}\right)^{2} k_{2}^{2} k_{3}-\frac{5}{2}\left(f^{\prime}-k_{1}\right) k_{2}\left(f^{\prime 2} k_{1} k_{2}-f^{\prime} k_{2} k_{3}-2 k_{4}\right) \\
& +\frac{72\left(f^{\prime}-k_{1}\right) k_{2}\left(f^{\prime 2} k_{1} k_{2}-f^{\prime} k_{2} k_{3}-2 k_{4}\right)}{\sqrt{-4-k_{2}^{2}}}+12\left(f^{\prime}-k_{1}\right)\left(-4-k_{2}^{2}\right)\left[-2 f^{\prime 2} k_{1}\right. \\
& \left.+\frac{1}{2} k_{1} k_{3}+\frac{k_{4}}{k_{2}}-\frac{k_{2}^{4}\left(f^{\prime}-k_{1}\right)^{4}}{\left(4+k_{2}^{2}\right)^{2}}+\frac{k_{2}^{2}\left(-f^{\prime 2}-f^{\prime} k_{1}^{2}+k_{1}^{3}+k_{1}\left(f^{\prime 2}+2 k_{3}\right)\right)+4 k_{2} k_{4}}{\sqrt{-4-k_{2}^{2}}}\right] \\
& +\ldots \text {, } \tag{40}
\end{align*}
$$

Thus, we have the mirror system (33) and (34) with the following initial data

$$
\begin{align*}
\theta(0) & =0, \quad \xi_{1}(0)=\frac{k_{2}}{\sqrt{-4-k_{2}^{2}}}, \\
\xi_{2}(0) & =\frac{k_{2}}{2}\left(-3 k_{3}+f^{\prime 2}+2 f^{\prime} k_{1}\right)+\frac{1}{\left(-4-k_{2}^{2}\right)}\left(f^{\prime}-k_{1}\right) k_{2}\left[\frac{1}{4} k_{2}^{2}\left(f^{\prime}+k_{1}\right)\right. \\
& \left.+2\left(f^{\prime}+2 k_{1}\right)\right], \\
\xi_{3}(0) & =\sqrt{-4-k_{2}^{2}}\left(\frac{3}{2} k_{4}-\frac{2 k_{4}}{k_{2}}-\frac{k_{4} k_{2}}{2}+\frac{5 f^{\prime 2} k_{1}}{k_{2}}-\frac{k_{1} k_{3}}{2}+\frac{1}{2} f^{\prime 2} k_{1} k_{2}+\frac{1}{4} f^{\prime} k_{1}^{2} k_{2}\right. \\
& \left.-\frac{f^{\prime 3}}{k_{2}}+\frac{2 f^{\prime} k_{1}^{2}}{k_{2}}\right)+\frac{\left(f^{\prime}-k_{1}\right)}{\sqrt{-4-k_{2}^{2}}}\left[\left(-2 f^{\prime} k_{1} k_{2}+\frac{1}{2} k_{2}\left(k_{1}^{2}-f^{\prime 2}\right)-\frac{3}{8} f^{\prime} k_{1} k_{2}^{3}\right.\right. \\
& \left.+\frac{1}{4} k_{1}^{2} k_{2}^{3}-\frac{4 f^{\prime}}{k_{2}}\left(f^{\prime}-k_{1}\right)\right], \\
\xi_{4}(0) & =\frac{5}{18} k_{5} \sqrt{-4-k_{2}^{2}}+\frac{2}{3}\left(f^{\prime}-k_{1}\right)^{2}\left(2 f^{\prime}+k_{1}\right) k_{1}-12 f^{\prime} k_{1}\left(f^{\prime}-k_{1}\right)^{2} k_{2}^{2}+\ldots . \tag{41}
\end{align*}
$$

Now we are ready to show the convergence of (37)and (38) by using the Cauchy theorem [10, p.150-151]. From the differential equations (33) and (34) and the initial conditions (41) we see that the coefficients of variable in (33), (34) and initial value conditions (41) are analytic functions provided that $r_{1}=k_{2} \neq \pm 2 i$. Thus, the initial value problem (33) and (34) with initial conditions (41) has unique analytic solutions which are convergent in the neighbourhood of $\theta=0$.

Substituting the series (39) and (40) back into $w_{1}=\theta^{-1},(21),(25),(28)$ and (31), we obtain the convergent power series for $w_{1}, w_{2}, b_{1}, b_{2}$ and $b_{3}$ which was not just formal. Furthermore, with some computation we see that these series are exactly (37) and (38). Therefore, we come to the conclusion that the Laurent series (37) and (38) are convergent. Thus, we summarise these results in terms of the following theorem.

Theorem 3.1 For the principal Laurent series solution of the ideal rotating, uniformly stratified system of six coupled ODEs (3), there is a change of variables of the form (6) such that the system of ODEs (3) is transformed into a regular system of ODEs (33) and (34) for the new variables $\left(\theta, \xi_{1}, \xi_{2}, \xi_{3}, \xi_{4}\right)$. Further, the Laurent series (37) and (38) in the principle dominant balance are converted into the power series (39) and (40) with initial data (41) which are the analytic functions in terms of new variables and thus, the series solutions (39)with (40) are convergent in the neighbourhood of $\theta=0$.

## 4 Conclusion

The reduced system of ODEs (3) which arose in the reduction of uniformly stratified fluid contained in the rotating box of dimension $L \times L \times H$ is completely integrable if the Rayleigh number $R a=0$. By taking $R a=0$, we have obtained the mirror system for both possible branches of leading ordered coefficients of system (3). The main feature in the singularity analysis is that the mirror system is regular near $\theta=0$, which corresponds to the movable singularity of the system (3) provided (36) holds. Also, we have shown that the formal Laurent series solutions arising from successful application of the Painlevé test to the system of ODEs (3) are convergent.

## References

[1] Majada, A.J. and Shefter, M.G. Elementary stratified flows with instability at large Richardson number. J. Fluid Mechanics 376 (1998) 319-350.
[2] Desale, B.S. and Patil, K.D. Painlevé test to a reduced system of six coupled nonlinear ODEs. Nonlinear Dynamics and System Theory 10(4) (2010) 349-361.
[3] Kichenassamy, S. and Srinivasan, G. The structure of WTC expansions and applications. Journal of Phys. A: Math. Gen. 28 (1995) 1977-2004.
[4] Kichenassamy, S. and Littman, W. Blow-up surfaces for nonlinear wave equations I. Commun. in PDE 18 (1993) 431-452.
[5] Maas, L.R.M. Theory Of basin scale dynamics of a stratified rotating fluid. Surveys in Geophysics. 25 (2004) 249-279.
[6] Paul, P. Sur les équations différentielles du Second Ordre et dórdre supérieur dont l'integrale générale est uniform. Acta Math. 25 (1902) 1-86.
[7] Weiss, J., Tabor, M. and Carnevale, G. The Painlevé property for partial differential equations. Journal of Math. Phys. 24 (1983) 522-526.
[8] Hu, J. and Yan, M. Singularity analysis for integrable systems by their mirrors. Nonlinearity 12 (1999) 1531-1543.
[9] Hu, J. and Yan, M. The mirror systems of integrable equations. Studies in App. Maths. 104 (2000) 67-90.
[10] Hu, J. and Yan, M. Local analyticity of solutions in the Painlevé test. In: Proc. of the workshop on nonlinearity, integrability and all that: twenty years after needs' 79 (2000) 146-152.
[11] Hu, J., Yan, M. and and Yee, T. Mirror transformations of hamiltonian system. Physica D. 152 (2001) 110-123.
[12] Kowalevski, S. Sur le problè me de la rotation dún corps solide autour dún point fixe. Acta Math. 12 (1889) 177-232.
[13] Yee, T. L. Linearization of mirror transform. Journal of Nonlinear Mathematical Physics $\mathbf{9}$ (2001) 234-242.
[14] Yee, T. L. A new perturbative approach in nonlinear singularity analysis. Journal of Mathematics and Statistics 7(3) (2011) 249-254.

# Real Time Analysis of Signed Marked Random Measures with Applications to Finance and Insurance 

Jewgeni H. Dshalalow* and Kizza M. Nandyose<br>Department of Mathematics, Florida Institute of Technology Melbourne, FL 32901, USA

Received: June 17, 2018; Revised: December 9, 2018


#### Abstract

This paper deals with stationary and independent increments processes in real time initiated in [14] embellishing it to a two-dimensional signed random measure with position dependent marking. The real-valued component of the associated marked point process is non-monotone presenting an analytical challenge. We manage to investigate various characteristics of that component, including the $n$th drop or a sharp surge that find applications to finance (like option trading) and risk theory. The need for time sensitive feature of our study (i.e., an analytical association with real time parameter $t$ ) allows stochastic control implementation in sharp contrast with time insensitive analysis in the present literature. We proceed with the classical approach of fluctuation analysis of a particle running through a random grid of a convex set that the particle is trying to escape. We find the distribution of the first passage time and its location in space.


Keywords: random walk; independent and stationary increments processes; fluctuations of stochastic processes; marked point processes; first passage time; signed marked random measures; time sensitive analysis.

Mathematics Subject Classification (2010): 60G50, 60G51, 60G52, 60G55, 60G57, 60K05, 60K35, 60K40, 60G25, 90B18, 90B10, 90B15, 90B25.

## 1 Introduction

In many scientific, financial, and game theoretic processes, timing is of at most importance and a main strategic issue. Several studies have been done on the first passage time in fluctuation theory and their applications to queuing, stochastic games, seismology, and finance (cf. [1,2,8-10,11,12,13,15,16,19,22-24,27,30]). Fluctuation theory pertains to the behavior of an underlying process around a critical threshold and more generally, when a process escapes from a fixed manifold. The time when that passage takes place is referred

[^3]to as the first passage time. Another critical value of that situation is the new location of the process upon its escape. Besides the original topics mentioned above, fluctuation theory has become a stand-alone subject in numerous articles appeared through the decades of intense research, cf. [3-7,17,20,21,29,31].

In our most recent paper [15], we worked with time sensitive functionals of the same entities but under real time observation of a monotone process. We dealt with nonnegative random measures and increment processes. In this paper we study a class of signed marked random measures $(\mathcal{A}, \Pi, \mathcal{T})=\sum_{n=0}^{\infty}\left(X_{n}, \pi_{n}\right) \varepsilon_{t_{n}}$ with position dependent marking, on a filtered probability space $\left(\Omega, \mathcal{F}, \mathcal{F}_{t}, P\right)$. Marks $X_{n}$ 's are non-negative, while marks $\pi_{n}$ 's are real-valued, with the support counting random measure $\sum_{n=0}^{\infty} \varepsilon_{t_{n}}$. This is a significant upgrade from [15], because not only is yet another component added, but it is non-monotone. Studies of non-monotone components are very few in the literature on fluctuations. Most prominent of them was by Lajos Takács [30]. However, the results in [30] were not tractable.

As in the theory of fluctuations, we focus on the behavior of $(\mathcal{A}, \Pi, \mathcal{T})$ around a fixed threshold $M>0$ with respect to its first component $\mathcal{A}$, referred to as an active component. With

$$
\begin{equation*}
A_{n}=X_{0}+X_{1}+\ldots+X_{n} \tag{1.1}
\end{equation*}
$$

we have $\left\{A_{n}\right\}$ monotone non-decreasing, whereas

$$
\begin{equation*}
P_{n}=\pi_{0}+\pi_{1}+\ldots+\pi_{n} \tag{1.2}
\end{equation*}
$$

is non-monotone, as $\pi_{k}$ 's are real-valued marks. Our interest is in an extreme behavior of the marginal process $(\Pi, \mathcal{T})=\sum_{n=0}^{\infty} \pi_{n} \varepsilon_{t_{n}}$ that is difficult to analyze due to the non-monotone nature of its marks. For that reason we introduce active mark $X_{n}$ being nonnegative and integer-valued that is to oversee $\pi_{n}$. For instance, we might be curious when the process $(\Pi, \mathcal{T})$ changes its monotonicity or when it experiences its first extreme drop or a surge. For example, we set $X_{0}=X_{1}=\ldots=X_{n-1}=0, X_{n}=1$, if $\pi_{0}>$ $a, \pi_{1}>a, \ldots, \pi_{n-1}>a$, and $\pi_{n} \leq a$. In the general case, the increments $X_{i}$ 's need not be constant, but they can be random variables with particular marginal distributions. For a fixed positive integer $M$, we define the exit index as

$$
\begin{equation*}
\nu:=\inf \left\{n=0,1, \ldots: A_{n} \geq M\right\} . \tag{1.3}
\end{equation*}
$$

Then, $t_{\nu}$ is called the first passage time of process $(\mathcal{A}, \Pi, \mathcal{T})$. It is the first epoch when the crossing of $M$ occurs. Obviously, $t_{\nu}$ is a stopping time relative to filtration $\mathcal{F}_{t}$. The respective excess values of $A_{\nu}$ and $P_{\nu}$ representing active and passive components, $\mathcal{A}$ and $\Pi$, respectively, are also of interest. We further assume that A1 the increments $\left\{X_{n}, \pi_{n}, \Delta_{n}=t_{n}-t_{n-1}\right\}$ for $n=0,1,2, \ldots, t_{-1}=0$, of the process $(\mathcal{A}, \Pi, \mathcal{T})$ are independent (position dependent marking), that is, $X_{n}$ and $\pi_{n}$ are dependent only on $\Delta_{n}$. A2 for $n=1,2, \ldots,\left\{X_{n}, \pi_{n}, \Delta_{n}\right\}$ are identically distributed.

Associated with $(\mathcal{A}, \Pi, \mathcal{T})$ is the "time sensitive counting" process

$$
\begin{equation*}
\left(N_{t}, \Pi_{t}\right)=(\mathcal{A}, \Pi)[0, t]=\sum_{n=0}^{\infty}\left(X_{n}, \pi_{n}\right) \varepsilon_{t_{n}}[0, t], t \geq 0 \tag{1.4}
\end{equation*}
$$

We will be interested in the value of $\left(N_{t}, \Pi_{t}\right)$ of some $t$ enclosed between $t_{\nu-1}$ and $t_{\nu}$ providing us with the information about $(\mathcal{A}, \Pi, \mathcal{T})$ between two key reference points as well as $\left(N_{t}, \Pi_{t}\right)$ for $t \in\left[0, \tau_{\nu}\right)$ (that we will discuss later on, in Section 5).

So we target the joint Laplace- and Fourier-Stieltjes transform of the above r.v.'s:

$$
\begin{array}{r}
\Phi_{\nu}(t)=E z^{N_{t}} e^{-i \eta I_{t}} e^{-i \varphi P_{\nu-1}} u^{A_{\nu-1}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta_{0} t_{\nu-1}-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t),  \tag{1.5}\\
\quad\|z\| \leq 1,\|u\| \leq 1,\|v\| \leq 1, \operatorname{Re} \vartheta_{0} \geq 0, \operatorname{Re\vartheta } \geq 0, \eta \in \mathbb{R}, \varphi \in \mathbb{R}, \phi \in \mathbb{R} .
\end{array}
$$

Note that because we manage to observe the process in real time, i.e., upon $t_{0}, t_{1}, t_{2}, \ldots$ (meaning that there are no changes between those epochs), it raises a question about a need in the continuous time interpolation. Indeed, in some past work (cf. Dshalalow and White [17]) when a process was observed over arbitrary time epochs (i.e., unrelated to $\left.t_{0}, t_{1}, t_{2}, \ldots\right)$, its continuous revival made perfect sense. In our case, however, it is more about associating the point process $t_{0}, t_{1}, t_{2}, \ldots$, especially the reference points $t_{\nu-1}, t_{\nu}$, with time $t$, than anything else. Its very obvious benefit is to know about the process over time related intervals like $[0, t]$ which was impossible with time insensitive versions. From a practical stand point, observing the process over arbitrary time epochs is more realistic than in real time. However, whenever it is possible to render, its second benefit lies in far more tractable results compared to delayed observations that additionally require the named point process to be Poisson or alike. Furthermore, we also obtain explicit characteristics of the continuous time parameter process in interval $\left[0, t_{\nu}\right)$ giving us a broad spectrum of information about process $N_{t}$. The associated functional will read

$$
\begin{equation*}
E z^{N_{t}} e^{-i \eta \Pi_{t}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[0, t_{\nu}\right)}(t),\|z\| \leq 1,\|v\| \leq 1, \operatorname{Re} \vartheta \geq 0, \eta \in \mathbb{R}, \phi \in \mathbb{R} \tag{1.6}
\end{equation*}
$$

Back to the random measure $(\mathcal{A}, \Pi, \mathcal{T})$, we recall that the passive component $\Pi$ is realvalued making this random measure signed. Studies related to signed random measures have previously been done in various topological and stochastic analysis contexts. In [19] Hellmund extended the idea of completely random measures to completely random signed measure and gave a characterization of this class of signed random measures. He demonstrated that the classes of Lévy random measures (utilized in Lévy adaptive regression kernel models) and Lévy bases (utilized in spatio-temporal modeling) are natural extensions of completely random signed measures and that independence is a fundamental concept in defining Lévy random measures and Lévy bases. Other concepts related to signed random measures are in the work by Smorodina and Faddeev [29] who studied symmetric stable signed measures and showed that they are limit measures of sums of independent random variables.

Various applications of fluctuation theory that we explore can also be found in stochastic signals such as time continuous readings for automated seizure detection and quantification using EEGs, heart attack activity monitoring through detection by EKGs, real time blood pressure monitoring, and the stock market. In this paper, we illustrate the applicability of our study by expounding on the case of stock prices. We are able to predict the time of the first drop of a stock (or first increase if we short it) at $t_{\nu}$ and thus, the highest price at $t_{\nu-1}$ at which we can sell it at that point in time.

Our model also applies to the classical risk problem originally posed by Filip Lundberg (see [27]). Assume that an insurance company starts at zero with the initial capital $u$ and let the premium be a linear function with a constant premium rate $c$, so that the premium income of the company at time $t$ is $u+c t$. Assume that the aggregate claims form a marked point process $\mathcal{Y}=\sum_{k=0}^{\infty} Y_{k} \varepsilon_{t_{k}}$, with $t_{k}$ being the time of the $k$ th claim and $Y_{k}$ - the amount of claim. Now Lundberg postulated that $\mathcal{Y}$ was a marked Poisson process with position independent marking. We relax either condition by assuming that
neither is $\mathcal{Y}$ Poisson, nor is it with position independent marking. If $\Delta_{k}=t_{k}-t_{k-1}$, we have $c \Delta_{k}$ premiums' increase from $t_{k-1}$ to $t_{k}$. The mark $\pi_{k}=c \Delta_{k}-Y_{k}$ is the change of company's asset from $t_{k-1}$ to $t_{k}$. Now,

$$
\Pi=\sum_{k=0}^{\infty} \pi_{k} \varepsilon_{t_{k}}
$$

is a purely signed marked random measure and

$$
P_{t}=\Pi[0, t]
$$

is the process describing the asset changes of the insurance company on interval $[0, t]$. Notice that $P_{t}$ does not give us the true value of the company's asset at time $t$, because $P_{t}$ is a piecewise constant interpolation of the true asset value process

$$
R_{t}=u+c t-\sum_{k=0}^{\infty} Y_{k} \varepsilon_{t_{k}}[0, t]
$$

known as the risk process. They coincide upon times $t_{0}, t_{1}, t_{2}, \ldots$ which is exactly what we need. Our process $(\mathcal{A}, \Pi, \mathcal{T})$ is defined through the active component

$$
X_{k}= \begin{cases}0, & \pi_{k}>0 \\ 1, & \pi_{k} \leq 0\end{cases}
$$

So we are interested in the moment when $P_{k}$ becomes negative or zero for the first time (which would trigger $X_{k}=1$ ). Thus, $\pi_{0}, \pi_{1}, \ldots, \pi_{\nu-1}$ are positive, while $\pi_{\nu}$ is negative or zero. $\left\{t_{\nu_{n}}\right\}$ is the embedded sequence of consecutive drops of $P_{t}$. Then obviously, the risk process $R_{t}$ will become negative or zero only upon one of the epochs $\left\{t_{\nu_{n}}\right\}$, known as the ruin time of $R_{t}$.

Let $\mathcal{F}_{t}$ be the natural filtration with respect to the risk process $R_{t}$. Then, $\left\{t_{\nu_{n}}\right\}$ is a sequence of stopping times relative $\mathcal{F}_{t}$ that are also locally strong Markov points, that is either $R_{t}$ and $P_{t}$ have a locally strong Markov property at each point $t_{\nu_{n}}$. Therefore, $R_{t}$ and $P_{t}$ conditionally regenerate upon these epochs. We can slightly modify $P_{t}$ to make it semi-regenerative with respect to $\left\{t_{\nu_{n}}\right\}$.

While a further discussion on the risk process and its study as a semi-regenerative process is beyond the scope of this paper, the time of the first or the second or the $n$th drop of the risk process is of interest for statistics purposes and it is often raised by insurance companies.

We continue this paper in Section 2 through a further formalism of our model and introduce basics of discrete operational calculus earlier developed by Dshalalow [6,7] and Dshalalow and Iwezulu [13]. In Section 3, we use the method of stochastic decomposition previously developed in Dshalalow and Nandyose [15] and Dshalalow and White [17,18], only now embellished for non-monotone components. We establish a key formula for the functional $\Phi_{\nu}(t)$ of (1.5) that we claim is analytically tractable. This claim is justified throughout Section 4 in a number of examples and special cases. We conclude our paper in Section 5 with time sensitive analysis where time $t$ runs interval $\left[0, \tau_{\nu}\right)$ and find the joint transform of $N_{t}, P_{t}, N_{\nu}$, and the first passage time $t_{\nu}$ in a fully closed form.

## 2 Formalism and Notation

We now return to the functional $\Phi_{\nu}$. Note that we do not know the distribution of the random vector $\left(A_{\nu}-A_{\nu-1}, t_{\nu}-t_{\nu-1}\right)$ nor is the latter independent of $\left(A_{\nu-1}, t_{\nu-1}\right)$. The remedy for this predicament is the use of stochastic expansion that will include several steps. In the first step, we introduce the auxiliary sequence $\{\nu(p)\}$ of exit indices relative to the sequence $\{0,1, \ldots\}$ of thresholds to be crossed by $A_{n}$, of which $\nu=\nu(M-1)$ was introduced in (1.3). Namely, let

$$
\begin{equation*}
\nu(p)=\inf \left\{n=0,1, \ldots: A_{n}>p\right\}, p=0,1, \ldots \tag{2.1}
\end{equation*}
$$

With $p$ fixed, we have the sequence of functionals

In our second step, we apply to $\Phi_{\nu(p)}$ of (2.2) the transformation $D_{p}$ defined as

$$
\begin{equation*}
D_{p}\{f(p)\}(x):=\sum_{p=0}^{\infty} x^{p} f(p)(1-x),\|x\|<1 \tag{2.3}
\end{equation*}
$$

where $f$ is a real-valued function with the domain $\mathbb{N}_{0}=\{0,1, \ldots\}$. The inverse of $D_{p}$ is the so-called $\mathcal{D}$-operator previously introduced in Dshalalow [6,7]:

$$
\mathcal{D}_{x}^{k} \varphi(x, y)=\left\{\begin{array}{cc}
\lim _{x \rightarrow 0} \frac{1}{k!} \frac{\partial^{k}}{\partial x^{k}}\left[\frac{1}{1-x} \varphi(x, y)\right], & k \geq 0  \tag{2.4}\\
0, & k<0
\end{array}\right.
$$

From $\Phi_{\nu(p)}(t)=\sum_{n=0}^{\infty} \Phi_{\nu(p)}(t) \mathbf{1}_{\{v(p)=n\}}$, we have

$$
\begin{gathered}
\Phi(t, x):=D_{p}\left[\Phi_{\nu(p)}(t)\right](x)=\sum_{n=0}^{\infty} \Phi_{\nu(p)}(t) D_{p} \mathbf{1}_{\{\nu(p)=n\}}(x) \\
=\sum_{n=0}^{\infty} \Phi_{\nu(p)=n}(t) D_{p} \mathbf{1}_{\{\nu(p)=n\}}(x),
\end{gathered}
$$

with

$$
\begin{equation*}
\Phi_{\nu(p)=n}(t)=E z^{N_{t}} u^{A_{n-1}} e^{-i \eta \Pi_{t}} e^{-i \varphi P_{n-1}} v^{A_{n}} e^{-i \phi P_{n}} e^{-\vartheta_{0} t_{n-1}-\vartheta t_{n}} \mathbf{1}_{\left\{t_{n-1} \leq t<t_{n}\right\}}=F_{n}(t) . \tag{2.5}
\end{equation*}
$$

From $\mathbf{1}_{\{v(p)=n\}}=\mathbf{1}_{\left\{A_{n-1} \leq p\right\}} \mathbf{1}_{\left\{A_{n}>p\right\}}$,

$$
\begin{aligned}
D_{p} \mathbf{1}_{\{v(p)=n\}}(x) & =(1-x) \sum_{p=0}^{\infty} x^{p} \mathbf{1}_{\left\{A_{n-1} \leq p\right\}} \mathbf{1}_{\left\{A_{n}>p\right\}} \\
& =(1-x) \sum_{p=A_{n-1}}^{A_{n}-1} x^{p} \\
=(1-x)\left(\sum_{p=0}^{A_{n}-1} x^{p}-\sum_{p=0}^{A_{n-1}-1} x^{p}\right) & =(1-x)\left(\frac{1-x^{A_{n}}}{1-x}-\frac{1-x^{A_{n-1}}}{1-x}\right)=x^{A_{n-1}}-x^{A_{n}}
\end{aligned}
$$

that yields

$$
\begin{gather*}
\Phi(t, x)=\sum_{n=0}^{\infty} F_{n}(t)\left(x^{A_{n-1}}-x^{A_{n}}\right) \\
=\sum_{n=0}^{\infty}\left[F_{n}\left(u x, v, z, \vartheta_{0}, \vartheta, t\right)-F_{n}\left(u, v x, z, \vartheta_{0}, \vartheta, t\right)\right], \text { where } A_{-1}=0 . \tag{2.6}
\end{gather*}
$$

Finally, applying the Laplace transform to $\Phi(t, x)$ of (2.6) we have

$$
\begin{equation*}
\Phi^{*}(\theta, x)=\int_{t=0}^{\infty} e^{-\theta t} \Phi(t, x) d t=\sum_{n=0}^{\infty}\left[F_{n}^{*}\left(u x, v, z, \vartheta_{0}, \vartheta, t\right)-F_{n}^{*}\left(u, v x, z, \vartheta_{0}, \vartheta, t\right)\right] \tag{2.7}
\end{equation*}
$$

Now functionals $F_{n}$ and their transforms $F_{n}^{*}$ are subject to our scrutiny in Section 3 .

## 3 Analysis of $F_{n}$

With $n=1,2, \ldots$, we work on

$$
\begin{equation*}
F_{n}(t)=E z^{N_{t}} u^{A_{n-1}} e^{-i \eta I_{t}} e^{-i \varphi P_{n-1}} v^{A_{n}} e^{-i \phi P_{n}} e^{-\vartheta_{0} t_{n-1}-\vartheta t_{n}} \mathbf{1}_{\left\{t_{n-1} \leq t<t_{n}\right\}} \tag{3.1}
\end{equation*}
$$

(defined in (2.5)). (3.1) can be brought to the expression

$$
\begin{gather*}
F_{n}(t)=E\left[(z u v)^{A_{n-1}} e^{-i(\eta+\varphi+\phi) P_{n-1}} e^{-\left(\vartheta_{0}+\vartheta\right) t_{n-1}} v^{A_{n}-A_{n-1}}\right. \\
\times e^{-i \phi\left(P_{n}-P_{n-1}\right)} e^{-\vartheta\left(t_{n}-t_{n-1}\right)} \mathbf{1}_{\left.\left\{t_{n-1} \leq t<t_{n}\right\}\right]} \\
=E(z u v)^{A_{n-1}} e^{-i(\eta+\varphi+\phi) P_{n-1}} e^{-\left(\vartheta_{0}+\vartheta\right) t_{n-1}} v^{X_{n}} e^{-i \phi \pi_{n}} e^{-\vartheta \Delta_{n}} \mathbf{1}_{\left\{t_{n-1} \leq t<t_{n}\right\}}, n=1,2 \ldots \tag{3.2}
\end{gather*}
$$

The Laplace transform of $F_{n}$ with the expectation unfolded reads

$$
\begin{gathered}
F_{n}^{*}(\theta)=\int_{t=0}^{\infty} e^{-\theta t} F_{n}(t) d t \\
=\sum_{k=0}^{\infty}(z u v)^{k} \sum_{j=0}^{\infty} v^{j} \int_{p=-\infty}^{\infty} e^{-i(\eta+\varphi+\phi) p} \int_{s \geq 0} e^{-\left(\vartheta_{0}+\vartheta\right) s} e^{-\theta s} \\
\quad \times \int_{q=-\infty}^{\infty} e^{-i \phi q} \int_{\delta \geq 0} e^{-\vartheta \delta} \int_{t-s=0}^{\delta} e^{-\theta(t-s)} d t \\
\times P_{A_{n-1} \otimes P_{n-1} \otimes t_{n-1} \otimes X_{n} \otimes \pi_{n} \otimes \Delta_{n}}(k, j, d p, d s, d q, d \delta) \\
=\frac{1}{\theta} \sum_{k=0}^{\infty}(z u v)^{k} \sum_{j=0}^{\infty} v^{j} \int_{p=-\infty}^{\infty} e^{-i(\eta+\varphi+\phi) p} \int_{s \geq 0} e^{-\left(\vartheta_{0}+\vartheta+\theta\right) s} \int_{q=-\infty}^{\infty} e^{-i \phi q} \int_{\delta \geq 0} e^{-\vartheta \delta} \\
\times P_{A_{n-1} \otimes P_{n-1} \otimes t_{n-1} \otimes X_{n} \otimes \pi_{n} \otimes \Delta_{n}}(k, j, d p, d s, d q, d \delta) \\
-\frac{1}{\theta} \sum_{k=0}^{\infty}(z u v)^{k} \sum_{j=0}^{\infty} v^{j} \int_{p=-\infty}^{\infty} e^{-i(\eta+\varphi+\phi) p} \int_{s \geq 0} e^{-\left(\vartheta_{0}+\vartheta+\theta\right) s} \int_{q=-\infty}^{\infty} e^{-i \phi q} \int_{\delta \geq 0} e^{-(\vartheta+\theta) \delta} \\
\times P_{A_{n-1} \otimes P_{n-1} \otimes t_{n-1} \otimes X_{n} \otimes \pi_{n} \otimes \Delta_{n}}(k, j, d p, d s, d q, d \delta)
\end{gathered}
$$

due to independence of $A_{n-1} \otimes P_{n-1} \otimes t_{n-1}$ and $X_{n} \otimes \pi_{n} \otimes \Delta_{n}$

$$
=\frac{1}{\theta} E\left[(z u v)^{A_{n-1}} e^{-i(\eta+\varphi+\phi) P_{n-1}} e^{-\left(\vartheta_{0}+\vartheta+\theta\right) t_{n-1}}\right]
$$

$$
\begin{gather*}
\times\left[E v^{X_{n}} e^{-i \phi \pi_{n}} e^{-\vartheta \Delta_{n}}-E v^{X_{n}} e^{-i \phi \pi_{n}} e^{-(\vartheta+\theta) \Delta_{n}}\right] \\
=\frac{1}{\theta} \Gamma_{n-1}\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)[\gamma(v, \phi, \vartheta)-\gamma(v, \phi, \vartheta+\theta)], \tag{3.3}
\end{gather*}
$$

where

$$
\begin{gather*}
\Gamma_{n-1}\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right) \\
=\gamma_{0}\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right) \gamma^{n-1}\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right) \text { for } n \geq 1 \tag{3.4}
\end{gather*}
$$

and

$$
\begin{equation*}
\gamma_{0}(u, \varphi, \vartheta)=E u^{X_{0}} e^{-i \varphi \pi_{0}} e^{-\vartheta t_{0}}, \gamma(u, \varphi, \vartheta)=E u^{X_{k}} e^{-i \varphi \pi_{k}} e^{-\vartheta \Delta_{k}}, k=1,2, \ldots \tag{3.5}
\end{equation*}
$$

Summing up $F_{n}$ for all $n=1,2, \ldots$, with (3.3-3.4) in mind, we formally arrive at the expression

$$
\begin{gather*}
\sum_{n=1}^{\infty} F_{n}^{*}(\theta)=\frac{1}{\theta} \gamma_{0}\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)  \tag{3.6}\\
\times[\gamma(v, \phi, \vartheta)-\gamma(v, \phi, \vartheta+\theta)] \frac{1}{1-\gamma\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)}
\end{gather*}
$$

To warrant the convergence of the geometric series $\sum_{n=1}^{\infty} F_{n}^{*}(\theta)$, in the proposition below, we show that the norm $\left\|\gamma\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)\right\|<1$.

Proposition 3.1 The series

$$
\begin{aligned}
& \sum_{n=1}^{\infty} F^{*}(\theta)=\sum_{n=1}^{\infty} \int_{t=0}^{\infty} e^{-\theta t} E\left[(z u v)^{A_{n-1}} e^{-i(\eta+\varphi+\phi) P_{n-1}}\right. \\
&\left.\times e^{-\left(\vartheta_{0}+\vartheta\right) t_{n-1}} v^{X_{n}} e^{-i \phi \pi_{n}} e^{-\vartheta \Delta_{n}} \mathbf{1}_{\left\{t_{n-1} \leq t<t_{n}\right\}}\right] d t
\end{aligned}
$$

converges to

$$
\begin{aligned}
& \sum_{n=1}^{\infty} F_{n}^{*}(\theta)= \\
& \frac{1}{\theta} \gamma_{0}\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right) \\
& \times {[\gamma(v, \phi, \vartheta)-\gamma(v, \phi, \vartheta+\theta)] \frac{1}{1-\gamma\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)}, }
\end{aligned}
$$

with

$$
\left\|\gamma\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)\right\|<1
$$

provided one of the following conditions is met:

$$
\operatorname{Re} \vartheta_{0}>0, \text { or } \operatorname{Re} \vartheta>0, \text { or } \operatorname{Re} \theta>0 \text { or }\|u\|<1, \text { or }\|v\|<1, \text { or }\|z\|<1
$$

Proof. The first part of the proposition is due to the above steps that formally ended in formula (3.6). Inequality (3.7) holds due to the following arguments:

$$
\begin{gathered}
\left\|\gamma\left(u v z, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)\right\| \leq E\left\|(u v z)^{X_{1}} e^{-i(\eta+\varphi+\phi) \pi_{n}} e^{-\left(\vartheta_{0}+\vartheta+\theta\right) \Delta_{1}}\right\| \\
=\sum_{k=0}^{\infty}\|u v z\|^{k} \int_{t=0}^{\infty} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right) t} P_{X_{1} \otimes \Delta_{1}}(k, d t) \\
=\int_{t=0}^{\infty} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right) t} P_{X_{1} \otimes \Delta_{1}}(0, d t)+\sum_{k=1}^{\infty}\|u v z\|^{k} \int_{t=0}^{\infty} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right) t} P_{X_{1} \otimes \Delta_{1}}(k, d t)
\end{gathered}
$$

$$
\begin{aligned}
& =\int_{t=0}^{1} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right) t} P_{X_{1} \otimes \Delta_{1}}(0, d t)+\int_{t=1}^{\infty} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right) t} P_{X_{1} \otimes \Delta_{1}}(0, d t) \\
& \quad+\sum_{k=1}^{\infty}\|u v z\|^{k} \int_{t=0}^{1} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right) t} P_{X_{1} \otimes \Delta_{1}}(k, d t) \\
& \quad+\sum_{k=1}^{\infty}\|u v z\|^{k} \int_{t=1}^{\infty} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right) t} P_{X_{1} \otimes \Delta_{1}}(k, d t) \\
& \leq \int_{t=0}^{1} P_{X_{1} \otimes \Delta_{1}}(0, d t)+e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right)} \int_{t=1}^{\infty} P_{X_{1} \otimes \Delta_{1}}(0, d t) \\
& +\|u v z\| \sum_{k=1}^{\infty} \int_{t=0}^{1} P_{X_{1} \otimes \Delta_{1}}(k, d t)+\|u v z\| \sum_{k=1}^{\infty} e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right)} \int_{t=1}^{\infty} P_{X_{1} \otimes \Delta_{1}}(k, d t)
\end{aligned}
$$

since $\|u v z\| \geq\|u v z\|^{k}$ for $\|u v z\| \leq 1$ and $k>1$. Let

$$
\begin{gathered}
a:=\int_{t=0}^{1} P_{X_{i} \otimes \Delta_{i}}(0, d t), b:=\int_{t=1}^{\infty} P_{X_{i} \otimes \Delta_{i}}(0, d t) \\
c:=\sum_{k=1}^{\infty} \int_{t=0}^{1} P_{X_{i} \otimes \Delta_{i}}(k, d t), d:=\sum_{k=1}^{\infty} \int_{t=1}^{\infty} P_{X_{i} \otimes \Delta_{i}}(k, d t) .
\end{gathered}
$$

Then clearly, $a+b+c+d=1$ and thus,

$$
a+e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right)} b+\|u v z\| c+\|u v z\| e^{-\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right)} d<1
$$

whenever $\|u v z\|<1$ or $\operatorname{Re}\left(\vartheta_{0}+\vartheta+\theta\right)>0$ and we are done with the proof.
We continue with $F_{n}$ for $n=0 . F_{0}$ is the functional of the underlying process on interval $\left[0, t_{0}\right)$. With $N_{t}=\Pi_{t}=A_{-1}=P_{-1}=t_{-1}=0$ we have

$$
\begin{gathered}
F_{0}(t)=E z^{N_{t}} u^{A_{n-1}} e^{-i \eta \Pi_{t}} e^{-i \varphi P_{n-1}} v^{A_{n}} e^{-i \phi P_{n}} e^{-\vartheta_{0} t_{n-1}-\vartheta t_{n}} \mathbf{1}_{\left\{0 \leq t<t_{0}\right\}} \\
=E v^{A_{0}} e^{-i \phi P_{0}} e^{-\vartheta t_{0}} \mathbf{1}_{\left[0, t_{0}\right)}(t)
\end{gathered}
$$

The following is easy to prove.
Proposition 3.2 Let $F_{0}(t)=E v^{A_{0}} e^{-i \phi P_{0}} e^{-\vartheta t_{0}} \mathbf{1}_{\left[0, t_{0}\right)}(t)$. Then

$$
\begin{equation*}
F_{0}^{*}(\theta)=\frac{1}{\theta}\left[\gamma_{0}(v, \phi, \vartheta)-\gamma_{0}(v, \phi, \vartheta+\theta)\right] . \tag{3.7}
\end{equation*}
$$

With Proposition 3.2, we can augment the series $\sum_{n=1}^{\infty} F_{n}^{*}$ of formula (3.6) to include $F_{0}^{*}$ :

$$
\begin{align*}
\sum_{n=0}^{\infty} F_{n}^{*}(\theta)= & \frac{1}{\theta}\left[\gamma_{0}(v, \phi, \vartheta)-\gamma_{0}(v, \phi, \vartheta+\theta)\right]+\frac{1}{\theta} \gamma_{0}\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right) \\
& \times[\gamma(v, \phi, \vartheta)-\gamma(v, \phi, \vartheta+\theta)] \frac{1}{1-\gamma\left(z u v, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)} \tag{3.8}
\end{align*}
$$

From (2.7) and (3.8) we arrive at

$$
\begin{gather*}
\Phi^{*}(\theta, x)=\sum_{n=0}^{\infty}\left[F_{n}^{*}\left(u x, v, z, \vartheta_{0}, \vartheta, t\right)-F_{n}^{*}\left(u, v x, z, \vartheta_{0}, \vartheta, t\right)\right] \\
=\frac{1}{\theta}\left[\gamma_{0}(v, \phi, \vartheta)-\gamma_{0}(v, \phi, \vartheta+\theta)\right]-\frac{1}{\theta}\left[\gamma_{0}(v x, \phi, \vartheta)-\gamma_{0}(v x, \phi, \vartheta+\theta)\right] \\
+\frac{1}{\theta} \gamma_{0}\left(z u v x, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right) \frac{1}{1-\gamma\left(z u v x, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)} \\
\quad \times[\gamma(v, \phi, \vartheta)-\gamma(v x, \phi, \vartheta)+\gamma(v x, \phi, \vartheta+\theta)-\gamma(v, \phi, \vartheta+\theta)] \tag{3.9}
\end{gather*}
$$

The Laplace transform $\Phi_{\nu}^{*}(\theta)=\int_{t=0}^{\infty} e^{-\theta t} \Phi_{\nu}(t) d t$ of the functional

$$
\Phi_{\nu}(t)=E z^{N_{t}} e^{-i \eta \Pi_{t}} e^{-i \varphi P_{\nu-1}} u^{A_{\nu-1}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta_{0} t_{\nu-1}-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)
$$

can be extracted from $\Phi^{*}(\theta, x)$ of (4.9) using the $\mathcal{D}$-operator.
The entire effort in this section can be reduced to the following.
Theorem 3.1 Let $\Phi_{\nu}(\theta)$ denote the Laplace transform of the functional

$$
\begin{gather*}
\Phi_{\nu}(t)=E z^{N_{t}} e^{-i \eta \Pi_{t}} e^{-i \varphi P_{\nu-1}} u^{A_{\nu-1}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta_{0} t_{\nu-1}-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)  \tag{3.10}\\
\|z\| \leq 1,\|u\| \leq 1,\|v\| \leq 1, \operatorname{Re} \vartheta_{0} \geq 0, \operatorname{Re} \vartheta \geq 0, \eta, \varphi, \phi \in \mathbb{R}
\end{gather*}
$$

Then, with $\|u\|<1$, or $\|v\|<1$, or $\|z\|<1$, or $\operatorname{Re} \vartheta_{0}>0$, or $\operatorname{Re} \vartheta>0$, or $\operatorname{Re} \theta>0$,

$$
\begin{gather*}
\Phi_{\nu}^{*}(\theta)  \tag{3.11}\\
=\mathcal{D}_{x}^{M-1}\left\{\frac{1}{\theta}\left[\gamma_{0}(v, \phi, \vartheta)-\gamma_{0}(v, \phi, \vartheta+\theta)\right]-\frac{1}{\theta}\left[\gamma_{0}(v x, \phi, \vartheta)-\gamma_{0}(v x, \phi, \vartheta+\theta)\right]\right. \\
+ \\
\quad \frac{1}{\theta} \gamma_{0}\left(z u v x, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right) \frac{1}{1-\gamma\left(z u v x, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)}  \tag{3.12}\\
\quad \times[\gamma(v, \phi, \vartheta)-\gamma(v x, \phi, \vartheta)+\gamma(v x, \phi, \vartheta+\theta)-\gamma(v, \phi, \vartheta+\theta)]\} .
\end{gather*}
$$

## 4 Applications to Option Trading

For an illustration, consider the following special case. Suppose that we observe a constantly fluctuating stock price of some company over the times $t_{0}=0, t_{1}, t_{2}, \ldots$ that starts off at time zero with a price $\pi_{0}$.

## Case 1. Observation of process $P_{i}$ upon the first drop.

1a. Suppose we are interested in the characteristics of the process around the period when the stock price drops for the first time. Because the stock prices cannot be modeled by a monotone process, we have the observed prices upon t's as the passive component, and introduce the active component

$$
X_{n}= \begin{cases}0, & \pi_{n} \geq 0  \tag{4.1}\\ 1, & \pi_{n}<0\end{cases}
$$

Suppose $\pi_{0}$ is a nonnegative r.v. with some specified distribution and let $X_{0}=\tau_{0}=0$.
So, $\gamma_{0}(z, \phi, \theta)=E e^{-i \phi \pi_{0}}($ innotation $)=\gamma_{0}(\phi)$.
Next, with $M=1$ according to our assumption about the first drop, formula (3.12) further reduces to

$$
\begin{align*}
& \theta \Phi_{\nu}^{*}(\theta)=\gamma_{0}(\eta+\varphi+\phi) \frac{1}{1-\gamma\left(0, \eta+\varphi+\phi, \vartheta_{0}+\vartheta+\theta\right)} \\
& \quad \times[\gamma(v, \phi, \vartheta)-\gamma(0, \phi, \vartheta)+\gamma(0, \phi, \vartheta+\theta)-\gamma(v, \phi, \vartheta+\theta)] \tag{4.2}
\end{align*}
$$

Because the active component is merely auxiliary, we are less interested in any information about $N_{t}, A_{\nu-1}, A_{\nu}$, as well as $P_{\nu-1}, t_{\nu-1}$, so we set $z=u=v=1$ and $\varphi=\vartheta_{0}=0$ restricting the Laplace transform of $\Phi_{\nu}$ to the marginal transform

$$
\begin{align*}
& \int_{t=0}^{\infty} e^{-\theta t} E e^{-i \eta \Pi_{t}} e^{-i \phi P_{\nu}} e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t \\
& \quad=\frac{1}{\theta} \gamma_{0}(\eta+\phi) \frac{1}{1-\gamma(0, \eta+\phi, \vartheta+\theta)} \\
& \times[\gamma(1, \phi, \vartheta)-\gamma(0, \phi, \vartheta)+\gamma(0, \phi, \vartheta+\theta)-\gamma(1, \phi, \vartheta+\theta)], \tag{4.3}
\end{align*}
$$

where

$$
\gamma(z, \phi, \theta)=E z^{X_{1}} e^{-i \phi \pi_{1}} e^{-\Delta_{1} \theta} \text { and } \gamma(0, \phi, \theta)=\left.E z^{X_{1}} e^{-i \phi \pi_{1}} e^{-\Delta_{1} \theta}\right|_{z=0}
$$

From

$$
\left.E z^{X_{1}}\right|_{z=0}=P\left\{X_{1}=0\right\}+\left.z P\left\{X_{1}=1\right\}\right|_{z=0}=P\left\{X_{1}=0\right\}=E \mathbf{1}_{\left\{X_{1}=0\right\}}=E \mathbf{1}_{\left\{\pi_{1} \geq 0\right\}}
$$

we have

$$
\gamma(0, \phi, \theta)=E \mathbf{1}_{\left\{\pi_{1} \geq 0\right\}} e^{-i \phi \pi_{1}} e^{-\Delta_{1} \theta}
$$

Suppose now that $\Delta$ 's and $\pi$ 's are independent, that is, the observation epochs and stock price changes are independent. This may not always apply, but it would simplify establishing of $\gamma$. Then

$$
\gamma(0, \phi, \theta)=E \mathbf{1}_{\left\{\pi_{1} \geq 0\right\}} e^{-i \phi \pi_{1}} E e^{-\Delta_{1} \theta}=E \mathbf{1}_{\left\{\pi_{1} \geq 0\right\}} e^{-i \phi \pi_{1}} \frac{\gamma}{\gamma+\theta},
$$

if the observation epochs occur according to a Poisson point process of intensity $\gamma$. Our next assumption is that the marginal distribution of $\pi_{1}$ is Laplace with parameter $\mu$ and zero shift. That being said, the PDF of $\pi_{1}$ is

$$
\begin{equation*}
f_{\pi_{1}}(x)=\frac{1}{2} \mu e^{-\mu|x|}, x \in \mathbb{R} \tag{4.4}
\end{equation*}
$$

Then

$$
\gamma(0, \phi, 0)=E 1_{\left\{\pi_{1} \geq 0\right\}} e^{-i \phi \pi_{1}}=\int_{x=0}^{\infty} e^{-i \phi x} \frac{1}{2} \mu e^{-\mu x} d x=\frac{1}{2} \frac{\mu}{\mu+i \phi}
$$

Because $E e^{-i \phi \pi_{1}}=E e^{-i \phi \pi_{1}}\left(\mathbf{1}_{\left\{\pi_{1} \geq 0\right\}}+\mathbf{1}_{\left\{\pi_{1}<0\right\}}\right)$, we have

$$
\begin{aligned}
& E e^{-i \phi \pi_{1}}=\frac{1}{2} \frac{\mu}{\mu+i \phi}+\int_{x=-\infty}^{0} e^{-i \phi x} \frac{1}{2} \mu e^{\mu x} d x \\
= & \frac{1}{2} \frac{\mu}{\mu+i \phi}+\frac{1}{2} \frac{\mu}{\mu-i \phi}=\frac{1}{2} \mu \frac{2 \mu}{\mu^{2}+\phi^{2}}=\frac{\mu^{2}}{\mu^{2}+\phi^{2}}
\end{aligned}
$$

Thus,

$$
\gamma(1, \phi, \theta)=E e^{-i \phi \pi_{1}} E e^{-\Delta_{1} \theta}=E e^{-i \phi \pi_{1}} \frac{\gamma}{\gamma+\theta}=\frac{\mu^{2}}{\mu^{2}+\phi^{2}} \frac{\gamma}{\gamma+\theta}
$$

Next the following two further marginals are of interest.
(i) With $\eta=\phi=0$ in (4.3), the functional

$$
\begin{gather*}
\int_{t=0}^{\infty} e^{-\theta t} E e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t \\
=\frac{1}{\theta} \frac{1}{1-\gamma(0,0, \vartheta+\theta)}[\gamma(1,0, \vartheta)-\gamma(0,0, \vartheta)+\gamma(0,0, \vartheta+\theta)-\gamma(1,0, \vartheta+\theta)] \tag{4.5}
\end{gather*}
$$

represents the Laplace transform of the first passage time $t_{\nu}$ 's marginal functional at the first drop with the time $t$ falling between the pre-first passage time $t_{\nu-1}$ and $t_{\nu}$. Here

$$
\begin{gathered}
\gamma(1,0, \vartheta+\theta)=\frac{\gamma}{\gamma+\vartheta+\theta} \\
\gamma(0, \phi, \theta)=E \mathbf{1}_{\left\{\pi_{1} \geq 0\right\}} e^{-i \phi \pi_{1}} E e^{-\Delta_{1} \theta}=\frac{1}{2} \frac{\mu}{\mu+i \phi} \frac{\gamma}{\gamma+\theta} \\
\gamma(0,0, \vartheta)=\frac{1}{2} \frac{\gamma}{\gamma+\vartheta} .
\end{gathered}
$$

Therefore,

$$
\begin{gather*}
\int_{t=0}^{\infty} e^{-\theta t} E e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t \\
=\left(1+\frac{\gamma}{\gamma+2(\vartheta+\theta)}\right) \frac{\gamma}{2} \frac{1}{(\gamma+\vartheta)(\gamma+\vartheta+\theta)} \tag{4.6}
\end{gather*}
$$

implying that the inverse of the Laplace transform is

$$
\begin{equation*}
E e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=\mathcal{L}_{\theta}^{-1}\left\{\int_{t=0}^{\infty} e^{-\theta t} E e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t\right\}=\frac{\gamma}{2(\gamma+\vartheta)} e^{-\frac{t}{2}(\gamma+2 \vartheta)} \tag{4.7}
\end{equation*}
$$

(ii) With $\eta=\vartheta=0$ in (4.3), we have the Laplace transform of the $P_{\nu}$ 's marginal functional upon the first passage time $t_{\nu}$ jointly with the time $t$ running between $t_{\nu-1}$ and $t_{\nu}$.

$$
\begin{gather*}
\int_{t=0}^{\infty} e^{-\theta t} E e^{-i \phi P_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t \\
=\frac{1}{\theta} \gamma_{0}(\phi) \frac{1}{1-\gamma(0, \phi, \theta)}[\gamma(1, \phi, 0)-\gamma(0, \phi, 0)+\gamma(0, \phi, \theta)-\gamma(1, \phi, \theta)] . \tag{4.8}
\end{gather*}
$$

Because

$$
\gamma_{0}(\phi)=e^{-i \phi p_{0}}
$$

(assuming the initial price $\pi_{0}=p_{0}$ a.s. where $p_{0}$ is a constant)
and

$$
\begin{gathered}
\gamma(1, \phi, \theta)=\frac{\mu^{2}}{\mu^{2}+\phi^{2}} \frac{\gamma}{\gamma+\theta}, \\
\gamma(0, \phi, \theta)=E 1_{\left\{\pi_{1} \geq 0\right\}} e^{-i \phi \pi_{1}} E e^{-\Delta_{1} \theta}=\frac{1}{2} \frac{\mu}{\mu+i \phi} \frac{\gamma}{\gamma+\theta}, \\
1-\gamma(0, \phi, \theta)=\frac{2(\mu+i \phi)(\gamma+\theta)-\mu \gamma}{2(\mu+i \phi)(\gamma+\theta)},
\end{gathered}
$$

and

$$
\frac{1}{1-\gamma(0,0, \vartheta+\theta)}=1+\frac{\mu \gamma}{2(\mu+i \phi)(\gamma+\theta)-\mu \gamma}
$$

we have that

$$
\begin{gather*}
\int_{t=0}^{\infty} e^{-\theta t} E e^{-i \phi P_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t \\
=\frac{1}{\theta} e^{-i \phi p_{0}}\left(1+\frac{\mu \gamma}{2(\mu+i \phi)(\gamma+\theta)-\mu \gamma}\right) \\
\times\left[\frac{\mu^{2}}{\mu^{2}+\phi^{2}}-\frac{1}{2} \frac{\mu}{\mu+i \phi}+\frac{1}{2} \frac{\mu}{\mu+i \phi} \frac{\gamma}{\gamma+\theta}-\frac{\mu^{2}}{\mu^{2}+\phi^{2}} \frac{\gamma}{\gamma+\theta}\right] \\
=e^{-i \phi p_{0}}\left[\frac{\mu^{2}}{\mu^{2}+\phi^{2}}-\frac{1}{2} \frac{\mu}{\mu+i \phi}\right]\left(1+\frac{\mu \gamma}{2(\mu+i \phi)(\gamma+\theta)-\mu \gamma}\right) \frac{1}{\gamma+\theta} . \tag{4.9}
\end{gather*}
$$

Thus,

$$
\begin{gather*}
E e^{-i \phi P_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=\mathcal{L}_{\theta}^{-1}\left\{\int_{t=0}^{\infty} e^{-\theta t} E e^{-i \phi P_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)\right\} \\
=\left[\frac{\mu^{2}}{\mu^{2}+\phi^{2}}-\frac{1}{2} \frac{\mu}{\mu+i \phi}\right] e^{-\left(\frac{\gamma t}{2}\left(\frac{\mu+2 i \phi}{\mu+i \phi}\right)+i \phi p_{0}\right)} \tag{4.10}
\end{gather*}
$$

and

$$
\begin{gather*}
E P_{\nu} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=i \lim _{\phi \rightarrow 0} \frac{\partial}{\partial \phi} E e^{-i \phi P_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=\frac{1}{2 \mu}\left(\frac{\gamma t}{2}+\mu p_{0}-1\right) e^{-\frac{\gamma t}{2}}  \tag{4.11}\\
E P_{\nu}^{2} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=-\lim _{\phi \rightarrow 0} \frac{\partial^{2}}{\partial \phi^{2}} E e^{-i \phi P_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) \\
=\frac{1}{2 \mu^{2}}\left(2+\left(\frac{\gamma t}{2}\right)^{2}+\left(\mu p_{0}\right)^{2}+2 \mu p_{0} \frac{\gamma t}{2}-2 \mu p_{0}\right) e^{-\frac{\gamma t}{2}} . \tag{4.12}
\end{gather*}
$$

So

$$
\begin{equation*}
E \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=P\left\{t_{\nu-1} \leq t<t_{\nu}\right\}=\mathcal{L}_{\theta}^{-1}\left\{\int_{t=0}^{\infty} e^{-\theta t} E \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t\right\}=\frac{e^{-\frac{\gamma}{2} t}}{2} \tag{4.13}
\end{equation*}
$$

1b. One could be interested in when the passive component drops lower than $R$, for some $R<0$. Thus the active component reads now

$$
X_{n}= \begin{cases}0, & \pi_{n} \geq R,  \tag{4.14}\\ 1, & \pi_{n}<R .\end{cases}
$$

With $M=1$ assumed and because

$$
\left.E z^{X_{1}}\right|_{z=0}=P\left\{X_{1}=0\right\}+\left.z P\left\{X_{1}=1\right\}\right|_{z=0}=P\left\{X_{1}=0\right\}=E \mathbf{1}_{\left\{X_{1}=0\right\}}=E \mathbf{1}_{\left\{\pi_{1} \geq R\right\}}
$$

we have

$$
\begin{aligned}
& \gamma(0, \phi, \theta)=E \mathbf{1}_{\left\{\pi_{1} \geq R\right\}} e^{-i \phi \pi_{1}} e^{-\Delta_{1} \theta}=E \mathbf{1}_{\left\{\pi_{1} \geq R\right\}} e^{-i \phi \pi_{1}} \frac{\gamma}{\gamma+\theta} \\
& \gamma(0, \phi, 0)=E \mathbf{1}_{\left\{\pi_{1} \geq R\right\}} e^{-i \phi \pi_{1}}=\int_{x=R}^{0} e^{-i \phi x} \frac{1}{2} \mu e^{\mu x} d x+\int_{x=0}^{\infty} e^{-i \phi x} \frac{1}{2} \mu e^{-\mu x} d x \\
&=\frac{1}{2} \frac{\mu}{\mu-i \phi}\left[1-e^{(\mu-i \phi) R}\right]+\frac{1}{2} \frac{\mu}{\mu+i \phi}
\end{aligned}
$$

Since

$$
E e^{-i \phi \pi_{1}}=E e^{-i \phi \pi_{1}}\left(\mathbf{1}_{\left\{\pi_{1} \geq R\right\}}+\mathbf{1}_{\left\{\pi_{1}<R\right\}}\right),
$$

we have

$$
E e^{-i \phi \pi_{1}}=\frac{1}{2} \frac{\mu}{\mu-i \phi}\left[1-e^{(\mu-i \phi) R}\right]+\frac{1}{2} \frac{\mu}{\mu+i \phi}+\int_{x=-\infty}^{R} e^{-i \phi x} \frac{1}{2} \mu e^{\mu x} d x=\frac{\mu^{2}}{\mu^{2}+\phi^{2}}
$$

Thus,

$$
\gamma(1, \phi, \theta)=E e^{-i \phi \pi_{1}} E e^{-\Delta_{1} \theta}=E e^{-i \phi \pi_{1}} \frac{\gamma}{\gamma+\theta}=\frac{\mu^{2}}{\mu^{2}+\phi^{2}} \frac{\gamma}{\gamma+\theta}
$$

and with $\eta=\phi=0=\vartheta$ in (4.3), the functional

$$
\begin{gathered}
\int_{t=0}^{\infty} e^{-\theta t} E \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t \\
=\frac{1}{\theta} \frac{1}{1-\gamma(0,0, \theta)}[\gamma(1,0,0)-\gamma(0,0,0)+\gamma(0,0, \theta)-\gamma(1,0, \theta)]=e^{\mu R} \frac{1}{2 \theta+\gamma e^{\mu R}}
\end{gathered}
$$

and

$$
\begin{gather*}
E \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=P\left\{t_{\nu-1} \leq t<t_{\nu}\right\} \\
=\mathcal{L}_{\theta}^{-1}\left\{\int_{t=0}^{\infty} e^{-\theta t} E \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t\right\}=\frac{1}{2} e^{-\left(\frac{\gamma t e^{\mu R}}{2}-\mu R\right)} \tag{4.15}
\end{gather*}
$$

which reduces to (4.13) when $R=0$.

## Case 2. Observation of process $P_{i}$ upon general $M$ th drop.

2a. For the general threshold level $M$ (when the stock price drops $M$ th times), since the active process increments $X_{n}$ are Bernoulli with $p=0.5$ due to the symmetric Laplace PDF of $\pi_{n}$ defined in (4.4) above with zero shift and with

$$
\begin{gather*}
E \mathbf{1}_{\left(t_{\nu-1}, t_{\nu}\right]}(t)=\left.\Phi_{\nu}(t)\right|_{z, v, u, \vartheta=1, \eta, \varphi, \phi, \vartheta_{0}, \vartheta=0} \\
\Phi_{\nu}^{*}(\theta)=\int_{t=0}^{\infty} e^{-\theta t} E \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t) d t \\
=\mathcal{D}_{x}^{M-1} \frac{1}{\theta} \gamma_{0}(0) \frac{1}{1-\gamma(x, 0, \theta)} \times[\gamma(1,0,0)-\gamma(x, 0,0)+\gamma(x, 0, \theta)-\gamma(1,0, \theta)] \tag{4.16}
\end{gather*}
$$

where

$$
\gamma(1,0, \theta)=\frac{\gamma}{\gamma+\theta}, \quad \gamma(x, 0,0)=\frac{1+x}{2}, \gamma(x, 0, \theta)=\left(\frac{1+x}{2}\right) \frac{\gamma}{\gamma+\theta} .
$$

Therefore,

$$
\begin{gathered}
\Phi_{\nu}^{*}(\theta)=\mathcal{D}_{x}^{M-1} \frac{1}{\theta} \frac{2(\gamma+\theta)}{\gamma+2 \theta-\gamma x}\left[1-\frac{1+x}{2}+\frac{1+x}{2} \frac{\gamma}{\gamma+\theta}-\frac{\gamma}{\gamma+\theta}\right] \\
=\mathcal{D}_{x}^{M-1} \frac{1}{\theta} \frac{2(\gamma+\theta)}{\gamma+2 \theta-\gamma x}\left[\frac{1-x}{2} \frac{\theta}{\gamma+\theta}\right] \\
=\mathcal{D}_{x}^{M-1}\left\{\frac{1}{\gamma+2 \theta-\gamma x}\right\}-\mathcal{D}_{x}^{M-2}\left\{\frac{1}{\gamma+2 \theta-\gamma x}\right\}=\frac{1}{(\gamma+2 \theta)}\left(\frac{\gamma}{\gamma+2 \theta}\right)^{M-1}=\frac{\gamma^{M-1}}{(\gamma+2 \theta)^{M}} .
\end{gathered}
$$

So

$$
\begin{equation*}
E \mathbf{1}_{\left(t_{\nu-1}, t_{\nu}\right]}(t)=P\left\{t_{\nu-1} \leq t<t_{\nu}\right\}=\mathcal{L}_{\theta}^{-1}\left\{\Phi_{\nu}(\theta)\right\}=\frac{1}{2} \frac{\left(\frac{\gamma t}{2}\right)^{M-1}}{(M-1)!} e^{-\frac{\gamma}{2} t} \tag{4.17}
\end{equation*}
$$

2b. Next we obtain the result for $E \mathbf{1}_{\left(t_{\nu-1}, t_{\nu}\right]}(t)$ for general $M$ and general shift parameter $a$ in our model such that

$$
f_{\pi_{1}}(x)=\frac{1}{2} \mu e^{-\mu|x-a|}, x \in \mathbb{R} .
$$

After some algebra we have

$$
\begin{aligned}
\gamma(0, \phi, 0)=E 1_{\left\{\pi_{1} \geq 0\right\}} e^{-i \phi \pi_{1}} & =\int_{x=0}^{a} e^{-i \phi x} \frac{1}{2} \mu e^{\mu(x-a)} d x+\int_{x=a}^{\infty} e^{-i \phi x} \frac{1}{2} \mu e^{-\mu(x-a)} d x \\
= & \frac{1}{2} \mu \frac{2 \mu e^{-i \phi a}-e^{-\mu a}(\mu+i \phi)}{\mu^{2}+\phi^{2}}, \\
\gamma(0, \phi, \theta) & =\frac{1}{2} \mu \frac{2 \mu e^{-i \phi a}-e^{-\mu a}(\mu+i \phi)}{\mu^{2}+\phi^{2}} \frac{\gamma}{\gamma+\theta}
\end{aligned}
$$

and

$$
\begin{gathered}
\gamma(0, \phi, 0)=\frac{1}{2} \mu \frac{2 \mu e^{-i \phi a}-e^{-\mu a}(\mu+i \phi)}{\mu^{2}+\phi^{2}}, \\
E e^{-i \phi \pi_{1}}=\frac{2 \mu^{2} e^{-i \phi a}}{\mu^{2}+\phi^{2}}-\frac{1}{2} \frac{\mu}{\mu^{2}+\phi^{2}}\left[(\mu+i \phi) e^{-\mu a}+(\mu-i \phi) e^{\mu a}\right] \\
\gamma(1, \phi, \theta)=\left[\frac{2 \mu^{2} e^{-i \phi a}}{\mu^{2}+\phi^{2}}-\frac{1}{2} \frac{\mu}{\mu^{2}+\phi^{2}}\left[(\mu+i \phi) e^{-\mu a}+(\mu-i \phi) e^{\mu a}\right]\right] \frac{\gamma}{\gamma+\theta} \\
\gamma(x, \phi, \theta)=(p+q x)\left[\frac{2 \mu^{2} e^{-i \phi a}}{\mu^{2}+\phi^{2}}-\frac{1}{2} \frac{\mu}{\mu^{2}+\phi^{2}}\left[(\mu+i \phi) e^{-\mu a}+(\mu-i \phi) e^{\mu a}\right]\right] \frac{\gamma}{\gamma+\theta} .
\end{gathered}
$$

Hence

$$
\begin{gathered}
\Phi_{\nu}^{*}(\theta)=\int_{t=0}^{\infty} e^{-\theta t} E \mathbf{1}_{\left(t_{\nu-1}, t_{\nu}\right]}(t) d t \\
=\mathcal{D}_{x}^{M-1} \frac{1}{\theta} \gamma_{0}(0) \frac{1}{1-\gamma(x, 0, \theta)} \times[\gamma(1,0,0)-\gamma(x, 0,0)+\gamma(x, 0, \theta)-\gamma(1,0, \theta)] \\
=\mathcal{D}_{x}^{M-1}\left\{\frac{1}{\theta} \frac{(\gamma+\theta)}{\gamma+\theta-p \gamma(2-\cosh (\mu a))-q \gamma(2-\cosh (\mu a)) x}\right.
\end{gathered}
$$

$$
\begin{gather*}
\times[(2-\cosh (\mu a))-(p+q x)(2-\cosh \mu a))+(p+q x)(2-\cosh (\mu a)) \frac{\gamma}{\gamma+\theta} \\
=\frac{\gamma^{M-1}(q(2-\cosh (\mu a)))^{M}}{(\gamma+\theta-p \gamma(2-\cosh (\mu a)))^{M}} \tag{4.18}
\end{gather*}
$$

by the $\mathcal{D}$-operator inversion formulas from [12].

$$
\begin{gather*}
E \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)=P\left\{t_{\nu-1} \leq t<t_{\nu}\right\}=\mathcal{L}_{\theta}^{-1}\left\{\Phi_{\nu}^{*}(\theta)\right\} \\
=\gamma^{M-1}(q(2-\cosh (\mu a)))^{M} \frac{t^{M-1}}{(M-1)!} e^{-(\gamma-p \gamma(2-\cosh (\mu a))) t} \\
=q(2-\cosh (\mu a)) \frac{(\gamma q(2-\cosh (\mu a)) t)^{M-1}}{(M-1)!} e^{-(\gamma-p \gamma(2-\cosh (\mu a))) t} \tag{4.19}
\end{gather*}
$$

Notice that when $a=0$ (in the symmetric case), (4.19) reduces to (4.17) and the value of $\mu$ is irrelevant given it is finite.

## 5 Continuous Time Parameter Process on Interval [0, $t_{\nu}$ )

Now consider the functional of passive process $P$ being observed over the period $\left[0, t_{\nu}\right)$, jointly with the active process $A_{\nu}$, the first passage time $t_{\nu}$, and the counting processes $N_{t}$ and $\Pi_{t}$. The functional satisfies the formula:

$$
\begin{aligned}
& \hat{\Phi}_{\nu}(t)=E z^{N_{t}} e^{-i \eta \Pi_{t}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[0, t_{\nu}\right)}(t) \\
= & \sum_{k=0}^{\infty} E z^{N_{t}} e^{-i \eta \Pi_{t}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[0, t_{\nu}\right)}(t) \mathbf{1}_{\{\nu=k\}} .
\end{aligned}
$$

Since $\sum_{j=0}^{\nu} E \mathbf{1}_{\left[t_{\nu-j-1}, t_{\nu-j}\right)}(t)=E \mathbf{1}_{\left[0, t_{\nu}\right)}(t)$,

$$
\begin{aligned}
\hat{\Phi}_{\nu}(t)=\sum_{k=0}^{\infty} \sum_{j=0}^{k} E\left[z^{A_{k-j-1}} v^{A_{k-j-1}} v^{\sum_{i=k-j}^{k} X_{i}} e^{-i \eta \Pi_{k-j-1}-i \phi P_{k-j-1}}\right. \\
\left.\times e^{-i \phi \sum_{i=k-j}^{k} \pi_{i}} e^{-\vartheta t_{k-j-1}} e^{-\vartheta \sum_{i=k-j}^{k} \Delta_{i}} \mathbf{1}_{\left[t_{k-j-1}, t_{k-j}\right)}(t)\right],
\end{aligned}
$$

and applying the transformation $D_{p}$ to $\hat{\Phi}_{\nu}(t)$ we have:

$$
\begin{aligned}
& D_{p}\left[\hat{\Phi}_{\nu}(t)\right](x)=\sum_{k=0}^{\infty} \sum_{j=0}^{k} F_{j k}(t) x^{X_{k-j+1}+\cdots+X_{k-1}} \\
& \times E(v x)^{X_{k-j+1+\cdots+X_{k-1}} e^{-i \phi\left(\pi_{k-j+1}+\cdots+\pi_{k-1}\right)} e^{-\vartheta\left(\Delta_{k-j+1}+\cdots+\Delta_{k-1}\right)}} \begin{array}{l}
\times E\left(1-x^{X_{k}}\right) e^{-i \phi \pi_{k}} e^{-\vartheta \Delta_{k}} v^{X_{k}}
\end{array}
\end{aligned}
$$

where

$$
\begin{gathered}
F_{j k}(t)= \\
E(z v x)^{A_{k-j-1}} e^{-i(\eta+\phi) P_{k-j-1}} e^{-\vartheta t_{k-j-1}} \mathbf{1}_{\left[t_{k-j-1}, t_{k-j}\right)}(t)(v x)^{X_{k-j}} e^{-i \phi\left(\pi_{k-j}\right)} e^{-\vartheta\left(\Delta_{k-j}\right)},
\end{gathered}
$$

$$
\tilde{F}(t)=E z^{A} e^{-i \eta P} e^{-\vartheta T} \mathbf{1}_{[T, T+\Delta)}(t) v^{X} e^{-i \phi \pi} e^{-\vartheta \Delta}
$$

under the assumptions that random vectors $A \otimes P \otimes T$ and $X \otimes \pi \otimes \Delta$ are independent. Then

$$
\begin{aligned}
\tilde{F}^{*}(\theta) & =\sum_{r} z^{r} \sum_{m} v^{m} \int_{p} e^{-i \eta p} \int_{w} e^{-i \phi w} \int_{s \geq 0} e^{-\vartheta s} e^{-\theta s} . \\
& \times \frac{1}{\theta} \int_{\delta}\left(e^{-\vartheta \delta}-e^{-(\vartheta+\theta) \delta}\right) P_{A \otimes P \otimes T \otimes X \otimes \pi \otimes \Delta}(r, m, d p, d s, d w, d \delta)
\end{aligned}
$$

and because $A \otimes P \otimes T$ and $X \otimes \pi \otimes \Delta$ are independent,

$$
=\frac{1}{\theta} E\left[z^{A} e^{-i \eta P} e^{-(\vartheta+\theta) T}\right][\gamma(v, \phi, \vartheta)-\gamma(v, \phi, \vartheta+\theta)] .
$$

Thus

$$
\begin{equation*}
F_{j k}^{*}(\theta)=\frac{1}{\theta} \gamma_{0} \delta^{1 j-1}\left[\delta-\delta^{1}\right] \gamma^{k-j-1}\left[\delta^{1}-\delta^{13}\right] \tag{5.1}
\end{equation*}
$$

(i) $\quad \sum_{k>0} \sum_{j=1}^{k-1} F_{j k}^{*}(\theta)=\frac{1}{\theta} \gamma_{0} \Psi \delta \sum_{k>0} \gamma^{k-2} \sum_{j=1}^{k-1}\left(\frac{\delta^{1}}{\gamma}\right)^{j-1}=\frac{1}{\theta} \gamma_{0} \frac{\Psi \delta}{(1-\gamma)\left(1-\delta^{1}\right)}$,
with notation $\gamma:=\gamma(z v x, \eta+\phi, \vartheta)$ and $\gamma_{0}:=\gamma_{0}(z v x, \eta+\phi, \vartheta)$, and further

$$
\begin{gathered}
\delta^{1}=\gamma(v x, \phi, \vartheta), \delta_{0}^{1}=\gamma_{0}(v x, \phi, \vartheta), \delta=\gamma(v, \phi, \vartheta) \\
\delta^{3}=\gamma(v, \phi, \vartheta+\theta), \delta^{13}=\gamma(v x, \phi, \vartheta+\theta), \delta_{0}=\gamma_{0}(v, \phi, \vartheta), \delta_{0}^{13}=\gamma_{0}(v x, \phi, \vartheta+\theta), \\
\Gamma \delta=\delta-\delta^{3}-\delta^{1}+\delta^{13}, \Lambda \delta=\frac{\Psi \delta}{1-\delta^{1}}+\Gamma \delta, \Psi \delta=\left(\delta-\delta^{1}\right)\left(\delta^{1}-\delta^{13}\right)
\end{gathered}
$$

(ii) Consider $j=k=0 . A_{-1}=t_{-1}=P_{-1}=0$ for $t \in\left[0, t_{0}\right)$ and $N_{t}=A_{-1}=\Pi_{t}=0$.

$$
\begin{gather*}
F_{00}(t)=E \mathbf{1}_{\left[0, t_{0}\right)}(t) e^{-\vartheta t_{0}} v^{A_{0}} e^{-i \phi P_{0}}\left(1-x^{A_{0}}\right) . \\
F_{00}^{*}(\theta)=\sum_{r} v^{r} \int_{p} e^{-i \phi p} \int_{s} e^{-\vartheta s} \int_{t=0}^{s} e^{-\theta t} d t P_{A_{0} \otimes P_{0} \otimes t_{0}}(r, d p, d s) \\
=\sum_{r} v^{r} \int_{p} e^{-i \phi p} \int_{s} e^{-\vartheta s} \frac{1}{\theta}\left[e^{-\vartheta s}-e^{-(\vartheta+\theta) s}\right] P_{A_{0} \otimes P_{0} \otimes t_{0}}(r, d p, d s)=\frac{1}{\theta} \Gamma \delta_{0} . \tag{5.3}
\end{gather*}
$$

(iii) Consider $j=0, k>0$.

$$
\begin{gathered}
F_{0 k}(t)=E z^{N_{t}} v^{A_{k}} e^{-i \eta P_{k-1}} e^{-i \phi P_{k}} e^{-\vartheta t_{k}} \mathbf{1}_{\left[t_{k-1}, t_{k}\right)}(t)\left(x^{A_{k-1}}-x^{A_{k}}\right) \\
=E(z v x)^{A_{k-1}} e^{-i(\eta+\phi) P_{k-1}} e^{-\vartheta t_{k-1}} \mathbf{1}_{\left[t_{k-1}, t_{k}\right)}(t) v^{X_{k}}\left(1-x^{X_{k}}\right) e^{-i \phi \pi_{k}} e^{-\vartheta \Delta_{k}} \\
F_{0 k}^{*}(\theta)=\int_{t} e^{-\theta t} \sum_{r}(z v x)^{r} \int_{p} e^{-i(\eta+\phi) p} \sum_{m} v^{m} \int_{q} e^{-i \phi q} \int_{s} e^{-\vartheta s} \\
\times \int_{\delta} e^{-\vartheta \delta} \mathbf{1}_{[s, s+\delta)}(t) d t P_{A_{k-1} \otimes P_{k-1} \otimes T_{k-1} \otimes X_{k} \otimes \pi_{k} \otimes \Delta_{k}}(r, d p, d s, m, d q, d \delta) \\
=\sum_{r}(z v x)^{r} \int_{p} e^{-i(\eta+\phi) p} \sum_{m} v^{m} \int_{q} e^{-i \phi q} \int_{s} e^{-\vartheta s} e^{-\theta s}
\end{gathered}
$$

$$
\begin{gathered}
\times \int_{t-s=0}^{\delta} e^{-\theta(t-s)} d t P_{A_{k-1} \otimes P_{k-1} \otimes T_{k-1} \otimes X_{k} \otimes \pi_{k} \otimes \Delta_{k}}(r, d p, d s, m, d q, d \delta) \\
=\sum_{r}(z v x)^{r} \int_{p} e^{-i(\eta+\phi) p} \sum_{m} v^{m} \int_{q} e^{-i \phi q} \int_{\delta}\left[e^{-\vartheta \delta}-e^{-(\vartheta+\theta) \delta}\right] \\
\times P_{A_{k-1} \otimes P_{k-1} \otimes X_{k} \otimes \pi_{k} \otimes \Delta_{k}}(r, d p, m, d q, d \delta) \\
=\frac{1}{\theta} \gamma^{k-1}(z v x, \eta+\phi, \vartheta) \gamma_{0}(z v x, \eta+\phi, \vartheta) \\
\quad \times[\gamma(v, \phi, \vartheta)-\gamma(v, \phi, \vartheta+\theta)-\gamma(v x, \phi, \vartheta)+\gamma(v x, \phi, \vartheta+\theta)]
\end{gathered}
$$

and

$$
\begin{gather*}
\sum_{k>0} F_{0 k}^{*}(\theta)=\frac{1}{\theta} \gamma_{0}(z v x, \eta+\phi, \vartheta) \frac{1}{1-\gamma(z v x, \eta+\phi, \vartheta)} \\
\times[\gamma(v, \phi, \vartheta)-\gamma(v, \phi, \vartheta+\theta)-\gamma(v x, \phi, \vartheta)+\gamma(v x, \phi, \vartheta+\theta)]=\frac{1}{\theta} \frac{\gamma_{0}}{1-\gamma} \Gamma \delta \tag{5.4}
\end{gather*}
$$

$(i v) \quad$ Consider $j=k>0 . F_{k k}(t)=E \mathbf{1}_{\left[0, t_{0}\right)}(t) e^{-\vartheta t_{k}} v^{A_{k}} e^{-i \phi P_{k}}\left(x^{A_{k-1}}-x^{A_{k}}\right)$

$$
\begin{aligned}
&=E \mathbf{1}_{\left[0, t_{0}\right)}(t) e^{-\vartheta t_{k}}(v x)^{A_{0}} e^{-i \phi P_{0}} e^{-\vartheta t_{0}}(v x)^{X_{1}+\cdots+X_{k-1}} e^{-i \phi\left(\pi_{1}+\cdots+\pi_{k-1}\right)} e^{-\vartheta\left(\Delta_{1}+\cdots+\Delta_{k-1}\right)} \\
& \times\left[v^{X_{k}}-(v x)^{X_{k}}\right] e^{-i \phi \pi_{k}} e^{-\vartheta \Delta_{k}} \\
&=E \mathbf{1}_{\left[0, t_{0}\right)}(t) e^{-\vartheta t_{k}}(v x)^{A_{0}} e^{-i \phi P_{0}} e^{-\vartheta t_{0}} \gamma^{k-1}(v x, \phi, \vartheta)[\gamma(v, \phi, \vartheta)-\gamma(v x, \phi, \vartheta)]
\end{aligned}
$$

So,

$$
\begin{equation*}
\sum_{k>0} F_{k k}^{*}(\theta)=\frac{1}{\theta} \frac{\Psi \delta_{0}}{1-\delta^{1}} \tag{5.5}
\end{equation*}
$$

Altogether, from $(i)$ through $(i v)$ we have

$$
\begin{gather*}
\stackrel{\Phi}{\nu}_{*}^{(\theta)=\int_{t=0}^{\infty} e^{-\theta t} \Phi_{\nu}(t) d t=} \mathcal{D}_{x}^{M-1}\left\{\sum_{k>0} \sum_{j=1}^{k-1} F_{j k}^{*}(\theta)+F_{00}^{*}(t)+\sum_{k>0} F_{0 k}^{*}(\theta)+\sum_{k>0} F_{k k}^{*}(\theta)\right\} \\
=\mathcal{D}_{x}^{M-1}\left\{\frac{1}{\theta}\left(\Lambda \delta_{0}+\frac{\gamma_{0}}{1-\gamma} \Lambda \delta\right)\right\} \tag{5.6}
\end{gather*}
$$

where $\Lambda \alpha=\Gamma \alpha+\frac{\Psi \alpha}{1-\delta^{1}}$ and $\alpha=\delta$ or $\delta_{0}$. The Laplace inverse of (5.6) will permit the recovery of $\hat{\Phi}_{\nu}(t)$.

## 6 Conclusion

In this paper we study a class of signed marked random measures $(\mathcal{A}, \Pi, \mathcal{T})=$ $\sum_{n=0}^{\infty}\left(X_{n}, \pi_{n}\right) \varepsilon_{t_{n}}$ with position dependent marking, on a filtered probability space $\left(\Omega, \mathcal{F}, \mathcal{F}_{t}, P\right)$. We target the critical behavior of the underlying stochastic process about a fixed threshold in the context of time sensitivity. The latter means that all related characteristics, such as first passage time and the location of the process upon crossing the threshold relate to deterministic time $t \geq 0$. The major benefit of this study is to utilize stochastic control over the process that must traditionally be considered on time
interval $[0, t], t \geq 0$. Using and further embellishing fluctuation theory, we find explicitly the functionals

$$
\Phi_{\nu}(t)=E z^{N_{t}} e^{-i \eta \Pi_{t}} e^{-i \varphi P_{\nu-1}} u^{A_{\nu-1}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta_{0} t_{\nu-1}-\vartheta t_{\nu}} \mathbf{1}_{\left[t_{\nu-1}, t_{\nu}\right)}(t)
$$

and

$$
\hat{\Phi}_{\nu}(t)=E z^{N_{t}} e^{-i \eta \Pi_{t}} e^{-i \phi P_{\nu}} v^{A_{\nu}} e^{-\vartheta t_{\nu}} \mathbf{1}_{\left[0, t_{\nu}\right)}(t)
$$

with respect to time $t \in\left[\tau_{\nu-1}, \tau_{\nu}\right)$ and $t \in\left[0, \tau_{\nu}\right)$, respectively. These functionals describe the status of underlying processes $N_{t}=\sum_{n=0}^{\infty} X_{n} \varepsilon_{t_{n}}[0, t]$ and $\Pi_{t}=\sum_{n=0}^{\infty} \pi_{n} \varepsilon_{t_{n}}[0, t]$, along with other characteristics like the values of these processes upon the crossing as well as just prior to crossing the threshold.

We discuss various applications to the finance (stock option trading) and risk theory. A number of special cases and examples demonstrate analytic tractability of the results obtained.

## References

[1] Al-Matar, N. and Dshalalow, J.H. A game-theoretic approach in single-server queues with maintenance. Time sensitive analysis. Commun. in Appl. Nonlin. Analysis 17 (1) (2010) 65-92.
[2] Al-Matar, N. and Dshalalow, J.H. Time sensitive functionals in classes of queues with sequential maintenance. Stochastic Models 27 (2011) 687-704.
[3] Bingham, N.H. Random walk and fluctuation theory. In: Handbook of Statistics (Eds. D.N. Shanbhag and C.R. Rao), Vol. 19, 2001, Elsevier Science, 171-213.
[4] Borokov, A.A. On the first passage time for one class of processes with independent increments. Theor. Probab. Appl. 10 (1964) 331-334.
[5] Coutin, L. and Dorobantu, D. First passage time law for some Levy processes with compound Poisson: Existence of a density. Bernoulli 17 (4) (2011) 1127-1135
[6] Dshalalow, J.H., First excess level of vector processes. J. Appl. Math. Stoch. Anal. 7 (3) (1994) 457-464.
[7] Dshalalow, J.H. On the level crossing of multidimensional delayed renewal processes. In: Special Issue Stochastic Systems, Journ. Appl. Math. Stoch. Anal. 10 (4) (1997) 355-361.
[8] Dshalalow, J.H. Fluctuations of recurrent processes and their application to the stock market. Stoch. Anal. Appl. 22 (1) (2004) 67-79.
[9] Dshalalow, J.H. On exit times of a multivariate random walk with some applications to finance. Nonlinear Analysis 63 (2005) 569-577.
[10] Dshalalow, J.H. Random walk analysis in antagonistic stochastic games. Stochastic Analysis and Applications (26) (2008) 738-783.
[11] Dshalalow, J.H. On multivariate antagonistic marked point processes. Math. and Comp. Modeling 49 (2009) 432-452.
[12] Dshalalow, J.H. Stochastic Processes. Lecture Notes, FIT, Melbourne, FL, 2015.
[13] Dshalalow, J.H. and Iwezulu, K. Discrete versus continuous operational calculus in antagonistic stochastic games. São Paulo Journal of Math. Sci., Springer, NY, 11 (2017) 471-489.
[14] Dshalalow, J.H., Iwezulu, K. and White, R.T. Discrete operational calculus in delayed stochastic games. Neural, Parallel, and Scientific Computations 24 (2016) 55-64.
[15] Dshalalow, J.H. and Nandyose, K.M. Continuous time interpolation of monotone marked random measures and their applications. Neural, Parallel, and Scientific Computations $\mathbf{2 4}$ (2018) 119-141.
[16] Dshalalow, J.H. and White, R.T. On reliability of stochastic networks. Neural, Parallel, and Scientific Computations 21 (2013) 141-160.
[17] Dshalalow, J.H. and White, R.T. On strategic defense in stochastic networks. Stochastic Analysis and Applications 32 (2014) 365-396.
[18] Dshalalow, J.H. and White, R.T. Time sensitive analysis of independent and stationary increment processes. Journal of Mathematical Analysis and Applications 443 (2016) 817833.
[19] Hellmund, G. Completely random signed measures. Statisitcs and Probability Letters 79 (2009) 894-898.
[20] Hida, T. (Editor), Mathematical Approach to Fluctuations: Astronomy, Biology and Quantum Dynamics: Proceedings of the Iias Workshop: Kyoto, Japan, May 18-21, 1992, World Scientific Publishers, 1995.
[21] Kadankova, T.V. On the distribution of the moment of the first exit time from an interval and value of overjump through borders interval for the processes with independent increments and random walk. Random Operators and Stochastic Equations 13 (3) (2005) 219-244.
[22] Kadankova, T.V. Exit, passage, and crossing times and overshoots for a Poisson compound process with an exponential component. Theor. Probability and Math. Statist. 75 (2007) 23-29.
[23] Kyprianou, A.E. and Pistorius, M.R. Perpetual options and Canadization through fluctuation theory. Ann. Appl. Prob. 13 (3) (2003) 1077-1098.
[24] Mellander, E., Vredin, A, and Warne, A. Stochastic trends and economic fluctuations in a small open economy. J. Applied Econom. 7 (4) (1992) 369-394.
[25] Muzy, J., Delour1, J., and Bacry, E. Modelling fluctuations of financial time series: from cascade process to stochastic volatility model. Eur. Phys. J. B 17 (2000) 537-548.
[26] Redner, S. A Guide to First-Passage Processes. Cambridge University Press, Cambridge, 2001.
[27] Schmidli, H. Risk Theory. Springer Actuarial, Cham, Switzerland, 2017.
[28] Shinozuka, M. and Wu, W-F. On the first passage problem and its application to earthquake engineering. Proceedings of Ninth World Conference on Earthquake Engineering, August 29, (VIII) 1988, Tokyo-Kyoto, Japan.
[29] Smorodina, N. and Faddeev M., The Lvy-Khinchin representation of the one class of signed stable measures and some of its applications. Acta Appl. Math. 110 (2010) 1289-1308.
[30] Takács, L. On fluctuations of sums of random variables, in Studies in Probability and Ergodic Theory. In: Advances in Mathematics; Supplementary Studies, Vol. 2 (G.-C. Rota, Ed.) (1978) 45-93.
[31] Yin, C., Wen, Y., Zong, Z., and Shen, Y. The first passage time problem for mixedexponential jump processes with applications in insurance and finance. Abstract and Applied Analysis (2014), 9 pages.
[32] Zolotarev, V.M. The first passage time of a level and the behavior at infnity for a class of processes with independent increments. Theor. Probab. Appl. 9 (1964) 653-664.

# Decentralized Stabilization for a Class of Nonlinear Interconnected Systems Using SDRE Optimal Control Approach 

A. Feydi, S. Elloumi and N. Benhadj Braiek*<br>Advanced System Laboratory (Laboratoire des Systèmes Avancés - LSA), Tunisia Polytechnic School-EPT, University of Carthage. BP 743, 2078, La Marsa, Tunisia.

Received: September 9, 2018; Revised: December 10, 2018


#### Abstract

This paper presents a new approach to assure the decentralized optimal control of interconnected nonlinear systems based on the decentralized statedependent riccati equation (SDRE). To remedy the problem of persistent stability in other works, we based our approach on the foundations of the Lyapunov theory. It allows developing a new sufficient condition to guarantee the global asymptotic stability of the systems under study. We conducted a simulation of this new control method on a numerical example. It demonstrated its efficiency and the sufficiency of the new stability conditions.


Keywords: decentralized optimal control; state-dependent Riccati equation (SDRE); interconnected nonlinear systems; Lyapunov theory; Kronecker product.

Mathematics Subject Classification (2010): 93D15, 34D23, 93 A 14.

## 1 Introduction

In recent years, the modern dynamical systems are getting more complex, highly interconnected, and mutually interdependent. This change is caused either by physical attributes, and/or a multitude of information and communication network constraints [1-3]. The important dimension and complexity of these large-scale systems often require a hierarchical decentralized architecture to analyze and control these systems [4-10]. Since these complex dynamic systems can be characterized by an interconnection between many subsystems, possible control strategies are generally based on a decentralized approach. The

[^4]advantage of such method is to reduce the complexity and therefore make the implementation of the control law more feasible.

In fact, the decentralized control refers to a control design with local decisions. These decisions are based only on local information of the subsystems. This method is given considerable interest because it brings up significant solutions for the traditional control approach limitations such as the implementation constraints, cost and reliability considerations especially for large-scale systems.

Optimal control of nonlinear systems is one of the most challenging subjects in control theory. Indeed, the classical problems of optimal control are based on the solution of the Hamilton-Jacobi equation (HJE) [11,12]. The solution to the HJE is a function of the state of the nonlinear system which makes it possible to characterize the quadratic optimal law of control sought under some hypotheses. However, in most cases it is impossible to solve it analytically, and despite recent progress, unsolved problems still exist and researchers often complain about the very limited applicability of contemporary theories because of conditions imposed on the system. This has led to numerous methods proposed in the literature for obtaining a suboptimal state feedback control law for the general case of nonlinear dynamic systems [13,14].

The SDRE approach is one of the methods applied in the determination of a suboptimal quadratic control based on the solution of a state-dependent Riccati equation. This strategy provides an efficient algorithm for nonlinear state feedback control synthesis while retaining the nonlinearities of the complex dynamic system, thanks to the flexibility of the state-dependent weighting matrices [15, 16]. This approach, proposed by Pearson [17] and later extended by Wernli and Cook [18], was studied independently by Mracek and Cloutier [19]. It should be pointed out that, although it is a relatively simplified and practical technique for controlling nonlinear systems, the SDRE approach involves problems that deserve to be treated with great attention, in particular the stability problem of the system controller [20,21]. Elloumi and Benhadj Braiek [22,23] have developed a sufficient condition for the stability of nonlinear system with optimal control based on SDRE approach. In this paper, we extend this work to the case of large scale interconnected systems. In this direction we carried out the synthesis of decentralized optimal control law based on the SDRE technique. This approach aims to minimize a performance criterion in order to compute decentralized optimal control gains when some sufficient conditions developed using the Lyapunov theory are verified.

The rest of the paper is organized as follows: the second section is devoted to the description of the systems under study and the formulation of the problem. In the third section, we present the decentralized optimal control law based on the SDRE approach. The fourth section treats the stability of the system in question using the quadratic Lyapunov function. The simulation results are set out in the fifth section to illustrate the applicability of the developed approach. Finally, conclusions are drawn and future scope of study is outlined.

## 2 Description of the System Under Study and Problem Formulation

A nonlinear system can be described by the interconnection of subsystems as follows:

$$
\left\{\begin{array}{l}
\dot{x}_{i}=f_{i}\left(\left(x_{i}, x_{j}\right), u_{i}(t), t\right), \quad i \neq j  \tag{1}\\
y_{i}=h_{i}\left(x_{i}\right), \quad i=1, \ldots, n, \quad j=1, \ldots, n
\end{array}\right.
$$

where $x_{i}(t) \in \mathbb{R}^{n_{i}}, u_{i} \in \mathbb{R}^{m_{i}}$ and $y_{i} \in \mathbb{R}^{p_{i}}$ are, respectively, the state, the control and the output of the $i^{\text {th }}$ subsystem.
$f_{i}\left(x_{i}, x_{j}\right)$ and $h_{i}\left(x_{i}\right)$ are nonlinear functions of the state. Through the statedependent coefficient (SDC) factorization, system designers can represent the nonlinear equations of the system under consideration as linear structures with state-dependent coefficients. Thus, the following procedure is similar to the optimal linear control (LQR) method, except that all matrices may depend on the states. Based on this concept, the state space equation for the nonlinear interconnected subsystem can be expressed as a linear-like state-space equation using direct SDC factorization as:

$$
\left\{\begin{array}{l}
\dot{x}_{i}(t)=A_{i}\left(x_{i}\right) x_{i}(t)+B_{i}\left(x_{i}\right) u_{i}(t)+\sum_{j=1, j \neq i}^{n} H_{i j}\left(x_{i}, x_{j}\right) x_{j}(t)  \tag{2}\\
y_{i}(t)=C_{i}\left(x_{i}\right) x_{i}(t), \quad i=1, \ldots, n
\end{array}\right.
$$

where $A_{i}\left(x_{i}\right)$ is the characteristic matrix that depends on the state of the $i^{t h}$ subsystem, $B_{i}\left(x_{i}\right)$ is the control vector of the $i^{t h}$ subsystem, $C_{i}\left(x_{i}\right)$ is the state-dependent observation matrix of the $i^{t h}$ subsystem and $H_{i j}\left(x_{i}, x_{j}\right)$ is the state -dependent interconnection matrix between the $i^{t h}$ and the $j^{\text {th }}$ subsystem.

The global interconnected system can be defined by the following compact form:

$$
\left\{\begin{array}{l}
\dot{x}=A(x) x+B(x) u+H(x) x  \tag{3}\\
y=C(x) x
\end{array}\right.
$$

with
$x^{T}=\left[x_{1}^{T}, x_{2}^{T}, \ldots, x_{n}^{T}\right]$ being the state vector of the overall system; $x \in \mathbb{R}^{n}, n=\sum_{i=1}^{n} n_{i} ;$
$u^{T}=\left[u_{1}^{T}, u_{2}^{T}, \ldots, u_{n}^{T}\right]$ being the control vector of the overall system,
$A(x)=\operatorname{diag}\left[A_{i}\left(x_{i}\right)\right], B(x)=\operatorname{diag}\left[B_{i}\left(x_{i}\right)\right]$ and $C(x)=\operatorname{diag}\left[C_{i}\left(x_{i}\right)\right]$.
$H(x)$ is the global interconnection matrix given as follows:

$$
H(x)=\left(\begin{array}{cccc}
0 & H_{12}(x) & \cdots & H_{1 n}(x)  \tag{4}\\
H_{21}(x) & 0 & \cdots & H_{2 n}(x) \\
\vdots & \vdots & \ddots & \vdots \\
H_{n 1}(x) & \cdots & \cdots & 0
\end{array}\right)
$$

Our contribution consists in the application of a decentralized optimal control via the SDRE approach to nonlinear interconnected systems. We based on solving the decentralized state-dependent Riccati equations to obtain the local control gains. The synthesis of a decentralized control for the system in question is detailed in the following section.

## 3 Decentralized State-Dependent Riccati Regulation Theory

The decentralized state-dependent Riccati equation technique is a nonlinear control design method for the direct construction of nonlinear sub-optimal feedback controllers. The determination of such decentralized control is based on considering the decoupled subsystem, expressed as follows:

$$
\left\{\begin{array}{l}
\dot{x}_{i}=A_{i}\left(x_{i}\right) x_{i}+B_{i}\left(x_{i}\right) u_{i}, \quad i=1, \ldots, n  \tag{5}\\
y_{i}=C_{i}\left(x_{i}\right) x_{i}
\end{array}\right.
$$

Note that $A_{i}\left(x_{i}\right)$ is not a unique matrix because there could be many possible choices in the direct (SDC) factorization. For this subsystem, the SDRE technique finds an input $u_{i}(t)$ that approximately minimizes the following performance criterion:

$$
\begin{equation*}
J_{i}=\frac{1}{2} \int_{0}^{\infty}\left(x_{i}^{T} Q_{i}\left(x_{i}\right) x_{i}+u_{i}^{T} R_{i}\left(x_{i}\right) u_{i}\right) d t \tag{6}
\end{equation*}
$$

where $Q_{i}\left(x_{i}\right) \in \mathbb{R}^{\left(n_{i} \times n_{i}\right)}$ and $R_{i}\left(x_{i}\right) \in \mathbb{R}^{\left(m_{i} \times m_{i}\right)}$ are symmetric, positive definite matrices. $x_{i}^{T} Q_{i}\left(x_{i}\right) x_{i}$ is a measure of the control accuracy and $u_{i}^{T} R_{i}\left(x_{i}\right) u_{i}$ is a measure of the control effort.

### 3.1 Existence of a control solution

The SDRE feedback control provides a similar approach as the algebraic Riccati equation (ARE) for LQR problems to the nonlinear regulation problem for the decoupled nonlinear subsystem (5) with cost functional (6). Indeed, once a SDC form has been found, the SDRE approach is reduced to solving a LQR problem at each sampling instant.

To guarantee the existence of such controller, the conditions in the following definitions must be satisfied [19].

- Definition 3.1: $A_{i}\left(x_{i}\right)$ is a controllable (stabilizable) parametrization of the nonlinear subsystem for a given region if $\left[A_{i}\left(x_{i}\right), B_{i}\left(x_{i}\right)\right]$ are pointwise controllable (stabilizable) in the linear sense for all $x_{i}$ in that region.
- Definition 3.2: $A_{i}\left(x_{i}\right)$ is an observable (detectable) parametrization of the nonlinear subsystem for a given region if $\left[C_{i}\left(x_{i}\right), A_{i}\left(x_{i}\right)\right]$ are pointwise observable (detectable) in the linear sense for all $x_{i}$ in that region.

Given these standing assumption, the state feedback decentralized controller is obtained in the following form:

$$
\begin{equation*}
u_{i}\left(x_{i}\right)=-K_{i}\left(x_{i}\right) x_{i} \tag{7}
\end{equation*}
$$

and the state feedback decentralized gain for minimizing (6) is

$$
\begin{equation*}
K_{i}\left(x_{i}\right)=R_{i}^{-1}\left(x_{i}\right) B_{i}^{T}\left(x_{i}\right) P_{i}\left(x_{i}\right), \tag{8}
\end{equation*}
$$

where $P_{i}\left(x_{i}\right)$ is the unique symmetric positive-definite solution of the decentralized state dependent Riccati equation (SDRE)

$$
\begin{align*}
& A_{i}^{T}\left(x_{i}\right) P_{i}\left(x_{i}\right)+P_{i}\left(x_{i}\right) A_{i}\left(x_{i}\right)  \tag{9}\\
& \quad-P_{i}\left(x_{i}\right) B_{i}\left(x_{i}\right) R_{i}^{-1}\left(x_{i}\right) B_{i}^{T}\left(x_{i}\right) P_{i}\left(x_{i}\right)+C_{i}^{T}\left(x_{i}\right) Q_{i}\left(x_{i}\right) C_{i}\left(x_{i}\right)=0 .
\end{align*}
$$

Remark 3.1: It is important to note that the existence of the decentralized optimal control for a particular parametrization of the subsystem is not guaranteed. Furthermore, there may be an infinite number of parametrizations of the subsystem, therefore the choice of parametrization is very important. The other factor which may determine the existence of a solution is the $Q_{i}\left(x_{i}\right)$ and $R_{i}\left(x_{i}\right)$ weighting matrices in the state dependent Riccati equation (9).

Remark 3.2. The greatest advantage of the state-dependent Riccati equation approach is that physical intuition is always present and the designer can directly control the performance by tuning the weighting matrices $Q_{i}\left(x_{i}\right)$ and $R_{i}\left(x_{i}\right)$. In other words, via the SDRE, the design flexibility of LQR formulation is directly translated to control the nonlinear interconnected systems. Moreover, $Q_{i}\left(x_{i}\right)$ and $R_{i}\left(x_{i}\right)$ are not only allowed to be constant, but can also vary as functions of states. In this way, different modes of behavior can be imposed in different regions of the state-space [21].

### 3.2 Optimality of the SDRE regulation

As $x_{i} \rightarrow 0, A_{i}\left(x_{i}\right) \rightarrow \partial f_{i}(0) / \partial x_{i}$ which implies that $P_{i}\left(x_{i}\right)$ approaches the linear ARE at the origin. Furthermore, the SDRE control solution asymptotically approaches the optimal control as $x_{i} \rightarrow 0$ and away from the origin the SDRE control is arbitrarily close to the optimal feedback. Hence the SDRE approach yields an asymptotically optimal feedback solution.

Let the Hamiltonian be defined by the following expression:

$$
\begin{equation*}
H_{i}\left(x_{i}, u_{i}, \lambda_{i}\right)=\frac{1}{2}\left[x_{i}^{T} Q_{i}\left(x_{i}\right) x_{i}+u_{i}^{T} R_{i}\left(x_{i}\right) u_{i}\right]+\lambda_{i}^{T}\left[A_{i}\left(x_{i}\right) x_{i}+B_{i}\left(x_{i}\right) u_{i}\right] \tag{10}
\end{equation*}
$$

Mracek and Cloutier developed the necessary conditions for the optimality of a general nonlinear regulator, that is the regulator governed by (5) and (6), and then extend these results to determine the optimality of the SDRE approach [19].

Theorem 1. For the general multivariable nonlinear SDRE control case (i.e., $n>1$ ), the SDRE nonlinear feedback solution and its associated state satisfy the first necessary condition for optimality $\partial H_{i} / \partial u_{i}=0$ of the nonlinear optimal regulator problem defined by (5) and (6). Additionally, the second necessary condition for optimality $\dot{\lambda}_{i}=-\partial H_{i} / \partial x_{i}$ is asymptotically satisfied at a quadratic rate.

Proof. Pontryagin'S maximum principle states that necessary conditions for optimality are

$$
\begin{equation*}
\frac{\partial H_{i}}{\partial u_{i}}=0, \quad \dot{\lambda}_{i}=-\frac{\partial H_{i}}{\partial x_{i}}, \quad \dot{x}_{i}=\frac{\partial H_{i}}{\partial \lambda_{i}} \tag{11}
\end{equation*}
$$

where $H_{i}$ is the Hamiltonian. Using (7) yields

$$
\begin{equation*}
\frac{\partial H_{i}}{\partial u_{i}}=B_{i}^{T}\left(x_{i}\right)\left[\lambda_{i}-P_{i}\left(x_{i}\right) x_{i}\right] \tag{12}
\end{equation*}
$$

and $\lambda_{i}$, the adjoint vector for the system, satisfies

$$
\begin{equation*}
\lambda_{i}=P_{i}\left(x_{i}\right) x_{i} \tag{13}
\end{equation*}
$$

and the first optimality condition (12) is satisfied identically for the nonlinear regulator problem. With the Hamiltonian defined in (10), the second necessary condition becomes

$$
\begin{align*}
\dot{\lambda}_{i}=-x_{i}^{T} & \left(\frac{\partial A_{i}\left(x_{i}\right)}{\partial x_{i}}\right)^{T} \lambda_{i}-u_{i}^{T}\left(\frac{\partial B_{i}\left(x_{i}\right)}{\partial x_{i}}\right)^{T} \lambda_{i}-Q_{i}\left(x_{i}\right) x_{i}  \tag{14}\\
& -\frac{1}{2} x_{i}^{T} \frac{\partial Q_{i}\left(x_{i}\right)}{\partial x_{i}} x_{i}-\frac{1}{2} u_{i}^{T} \frac{\partial R_{i}\left(x_{i}\right)}{\partial x_{i}} u_{i} .
\end{align*}
$$

Taking the time derivative of (13) yields

$$
\begin{equation*}
\dot{\lambda}_{i}=\dot{P}_{i}\left(x_{i}\right) x_{i}+P_{i}\left(x_{i}\right) \dot{x}_{i} . \tag{15}
\end{equation*}
$$

Substituting this result, along with (5), (7) and (14) into (9) leads to the SDRE necessary condition for optimality

$$
\begin{align*}
\dot{P}_{i}\left(x_{i}\right) x_{i} & +\frac{1}{2} x_{i}^{T} P_{i}\left(x_{i}\right) B_{i}\left(x_{i}\right) R_{i}^{-1}\left(x_{i}\right) \frac{\partial R_{i}\left(x_{i}\right)}{\partial x_{i}} R_{i}^{-1}\left(x_{i}\right) B_{i}^{T}\left(x_{i}\right) P_{i}\left(x_{i}\right) x_{i} \\
& +x_{i}^{T}\left(\frac{\partial A_{i}\left(x_{i}\right)}{\partial x_{i}}\right)^{T} P_{i}\left(x_{i}\right) x_{i}+\frac{1}{2} x_{i}^{T} \frac{\partial Q_{i}\left(x_{i}\right)}{\partial x_{i}} x_{i}  \tag{16}\\
& -x_{i}^{T} P_{i}\left(x_{i}\right) B_{i}\left(x_{i}\right) R_{i}^{-1}\left(x_{i}\right)\left(\frac{\partial B_{i}\left(x_{i}\right)}{\partial x_{i}}\right)^{T} P_{i}\left(x_{i}\right) x_{i}=0 .
\end{align*}
$$

Hence, whenever (16) is satisfied, the closed-loop SDRE solution satisfies all the firstorder necessary conditions for an extremum of the cost functional.

## 4 Stability Study

In this section, we study the asymptotic stability of interconnected system based on the Lyapunov theory [10]. We begin with the stability study of each subsystem, thereafter we deal with the development of a sufficient condition to assure the asymptotic stability of the overall interconnected nonlinear system.

### 4.1 Stability of a decoupled nonlinear subsystem

Stability of SDRE systems is still an open problem. Local stability results are presented by Cloutier, D'souza and Mracek in the case when the closed-loop coefficient matrix is assumed to have a special structure.

The authors in [22,23] presented the optimal control solution for nonlinear subsystem using the SDRE method. The asymptotic stability of decoupled subsystem (5) with SDRE feedback control is guaranteed provided that

$$
\begin{align*}
& M_{i}\left(x_{i}\right)=-C_{i}^{T}\left(x_{i}\right) Q_{i}\left(x_{i}\right) C_{i}\left(x_{i}\right)-P_{i}\left(x_{i}\right) B_{i}\left(x_{i}\right) R_{i}^{-1}\left(x_{i}\right) B_{i}^{T}\left(x_{i}\right) P_{i}\left(x_{i}\right) \\
& -\left(I_{n} \otimes x_{i}^{T} P_{i}\left(x_{i}\right) B_{i}\left(x_{i}\right) R_{i}^{-1}\left(x_{i}\right) B_{i}^{T}\left(x_{i}\right)\right) \frac{\partial P_{i}\left(x_{i}\right)}{\partial x_{i}}+\left(I_{n} \otimes\left(x_{i}^{T} A_{i}^{T}\left(x_{i}\right)\right) \frac{\partial P_{i}\left(x_{i}\right)}{\partial x_{i}}\right. \tag{17}
\end{align*}
$$

is negative definite for all $x_{i} \in \mathbb{R}^{n_{i}}$.
Now, to guarantee the asymptotic stability of the overall interconnected system (3), we carry out a stability study of interconnected system (2) with the decentralized control (7) as depicted in the following subsection.

### 4.2 Stability of a global interconnected system

In this paragraph, we present our contribution which consists in developing a sufficient condition to assure the asymptotic stability of the overall interconnected nonlinear system (3) with the decentralized control law (7). This study is based on the quadratic Lyapunov function

$$
\begin{equation*}
V(x)=x^{T} P(x) x \tag{18}
\end{equation*}
$$

where $P(x)=\operatorname{diag}\left[P_{i}\left(x_{i}\right)\right]$.
The global asymptotic stability of the equilibrium state $(x=0)$ of system (3) is ensured when the time derivative $\dot{V}(x)$ of $V(x)$ is negative define for all $x \in \mathbb{R}^{n}$,

$$
\begin{equation*}
\dot{V}(x)=\dot{x}^{T} P(x) x+x^{T} \frac{d P(x)}{d t} x+x^{T} P(x) \dot{x} \tag{19}
\end{equation*}
$$

The use of expression (19) and the following equality:

$$
\begin{equation*}
\frac{d P(x)}{d t}=\left(I_{n} \otimes \dot{x}^{T}\right) \frac{\partial P(x)}{\partial x} \tag{20}
\end{equation*}
$$

yields

$$
\begin{align*}
& \dot{V}(x)=x^{T}\left[A^{T}(x) P(x)+P(x) A(x)\right] x+x^{T}\left[H^{T}(x) P(x)+P(x) H(x)\right] x \\
& -2 x^{T}\left[P(x) B(x) R^{-1}(x) B^{T}(x) P(x)\right] x+x^{T}\left(I_{n} \otimes \dot{x}^{T}\right) \frac{\partial P(x)}{\partial x} x, \tag{21}
\end{align*}
$$

then

$$
\begin{align*}
& \dot{V}(x)=x^{T}\left[A^{T}(x) P(x)+H^{T}(x) P(x)\right. \\
& \left.+P(x) A(x)+P(x) H(x)-2 P(x) B(x) R^{-1}(x) B^{T}(x) P(x)\right] x  \tag{22}\\
& +x^{T}\left[\left(I_{n} \otimes\left(x^{T} A^{T}(x)+x^{T} H^{T}(x)-x^{T} P(x) B(x) R^{-1}(x) B^{T}(x)\right)\right) \frac{\partial P(x)}{\partial x}\right]
\end{align*}
$$

where $\otimes$ is the Kronecker product notation whose definition and properties are detailed in the appendix. Using the state-dependent Riccati equation (9), expression (22) can be simplified as follows:

$$
\begin{align*}
& \dot{V}(x)=x^{T}\left[-C^{T}(x) Q(x) C(x)-P(x) L(x) P(x)\right] x \\
& +x^{T}\left[H^{T}(x) P(x)+P(x) H(x)\right] x+x^{T}\left[\left(I_{n} \otimes\left(x^{T} A^{T}(x)+x^{T} H^{T}(x)\right)\right) \frac{\partial P(x)}{\partial x}\right] x  \tag{23}\\
& -x^{T}\left[\left(I_{n} \otimes\left(x^{T} P(x) B(x) R^{-1}(x) B^{T}(x)\right)\right) \frac{\partial P(x)}{\partial x}\right] x,
\end{align*}
$$

where $L(x)=B(x) R^{-1}(x) B^{T}(x), \forall x \in \mathbb{R}^{n}$.
To ensure the asymptotic stability of the overall systems (3) with the decentralized optimal control law (7), $\dot{V}(x)$ should be negative definite, which is equivalent to $M(x)$ being negative definite, with

$$
\begin{align*}
M(x)= & -C^{T}(x) Q(x) C(x)-P(x) B(x) R^{-1}(x) B^{T}(x) P(x) \\
& -\left(I_{n} \otimes x^{T} P(x) B(x) R^{-1}(x) B^{T}(x)\right) \frac{\partial P(x)}{\partial x}  \tag{24}\\
& +\left(I_{n} \otimes x^{T} A^{T}(x)+x^{T} H^{T}(x)\right) \frac{\partial P(x)}{\partial x}+P(x) H(x)+H^{T}(x) P(x) .
\end{align*}
$$

We need to simplify the manipulation of matrix $M(x)$ by expressing $\partial P(x) / \partial x$ in terms of $P(x)>0, \forall x \in \mathbb{R}^{n}$. When deriving the SDRE (9) with respect to the state vector $x \in \mathbb{R}^{n}$, we get the following expression:

$$
\begin{align*}
& \frac{\partial P(x)}{\partial x} A(x)+\left(I_{n} \otimes P(x)\right) \frac{\partial A(x)}{\partial x}+\frac{\partial A^{T}(x)}{\partial x} P(x)+\left(I_{n} \otimes A^{T}(x)\right) \frac{\partial P(x)}{\partial x}+\frac{\partial \Phi(x)}{\partial x} \\
& -\frac{\partial P(x)}{\partial x} L(x) P(x)-\left(I_{n} \otimes P(x) L(x)\right) \frac{\partial P(x)}{\partial x}-\left(I_{n} \otimes P(x)\right) \frac{\partial L(x)}{\partial x} P(x)=0 \tag{25}
\end{align*}
$$

with $\Phi(x)=C^{T}(x) Q(x) C(x)$, which gives

$$
\begin{equation*}
\left[I_{n} \otimes A^{T}(x)-I_{n} \otimes(P(x) L(x))\right] \frac{\partial P(x)}{\partial x}+\frac{\partial P(x)}{\partial x}[A(x)-L(x) P(x)]=W(x) \tag{26}
\end{equation*}
$$

with

$$
\begin{equation*}
W(x)=\left(I_{n} \otimes P(x)\right) \frac{\partial L(x)}{\partial x} P(x)-\left(I_{n} \otimes P(x)\right) \frac{\partial A(x)}{\partial x}-\frac{\partial A^{T}(x)}{\partial x} P(x)-\frac{\partial \Phi(x)}{\partial x} . \tag{27}
\end{equation*}
$$

To simplify the partial derivative expression $\partial P(x) / \partial x$ we use the functions $V e c$ and mat and their properties defined in this paper appendix; so (26) becomes

$$
\begin{array}{r}
V e c\left(\frac{\partial P(x)}{\partial x}\right)=\left[I_{n} \otimes\left(I_{n} \otimes A^{T}(x)+I_{n} \otimes P(x) L(x)\right)\right.  \tag{28}\\
+ \\
\left.(A(x)-L(x) P(x)) \otimes I_{n}\right]^{-1} V e c(W(x))
\end{array}
$$

which leads to

$$
\begin{align*}
\frac{\partial P(x)}{\partial x}= & \operatorname{mat}_{\left(n^{2}, n\right)}\left[\left(I _ { n } \otimes \left[I_{n} \otimes A(x)+I_{n} \otimes A^{T}(x)\right.\right.\right.  \tag{29}\\
& \left.\left.\left.-2\left(I_{n} \otimes L(x) P(x)\right)\right]\right)^{-1} \operatorname{Vec}(W(x))\right]
\end{align*}
$$

Therefore, we can state the following result.
Theorem 2. The overall system (3) is globally asymptotically stabilizable by the optimal decentralized control law (7), with the cost function (6) if the matrix $M(x)$ defined by (24) is negative definite for all $x \in \mathbb{R}^{n}$.

## 5 Simulation Results

In this section we will illustrate the performance of the decentralized SDRE approach, discussed in the previous paragraph, by a numerical example. We consider a nonlinear interconnected system defined by the following two subsystems of state equations:

$$
\left\{\begin{array}{l}
\sum 1:\left\{\begin{array}{l}
\dot{x}_{11}=-2 x_{11}+x_{11} x_{12}, \\
\dot{x}_{12}=x_{13}+x_{12} x_{11}+x_{22}^{2} x_{21} \\
\dot{x}_{13}=u_{1}+x_{13}^{2}\left(x_{12} x_{11}+x_{11}^{2}\right)+x_{22} x_{21}
\end{array}\right.  \tag{30}\\
\sum 2:\left\{\begin{array}{l}
\dot{x}_{21}=-x_{21}+x_{22}^{2}, \\
\dot{x}_{22}=x_{6}+\left(x_{12}^{2} x_{11}+x_{22}^{2} x_{21}\right) \\
\dot{x}_{23}=u_{2}+x_{23}^{2}\left(x_{12} x_{11}^{2}+x_{22} x_{21}^{2}\right)+x_{23}^{2} x_{21}^{2}
\end{array}\right.
\end{array}\right.
$$

with

- $x_{1}=\left[\begin{array}{lll}x_{11} & x_{12} & x_{13}\end{array}\right]^{T}, x_{2}=\left[\begin{array}{lll}x_{21} & x_{22} & x_{23}\end{array}\right]^{T}$ being the state vectors of subsystems $\sum 1$ and $\sum 2$,
- $u=\left[\begin{array}{ll}u_{1} & u_{2}\end{array}\right]^{T}$ being the inputs of the interconnected nonlinear system.

We solve equation (9) with

$$
\begin{gather*}
Q_{1}\left(x_{1}\right)=Q_{2}\left(x_{2}\right)=\left(\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right),  \tag{31}\\
R_{1}\left(x_{1}\right)=R_{2}\left(x_{2}\right)=0.1 \tag{32}
\end{gather*}
$$

For interconnected nonlinear systems (30), we choose the following (SDC) parametrization:

$$
A_{1}\left(x_{1}\right)=\left(\begin{array}{ccc}
-2 & x_{11} & 0 \\
x_{12} & 0 & 1 \\
x_{13}^{2} x_{12} & 0 & x_{13} x_{11}^{2}
\end{array}\right), A_{2}\left(x_{2}\right)=\left(\begin{array}{ccc}
-1 & x_{22} & 0 \\
0 & x_{22} x_{21} & 1 \\
x_{23}^{2} x_{22} x_{21}+x_{23}^{2} x_{21}^{2} & 0 & 0
\end{array}\right) .
$$

The control matrices are given as follows:

$$
B_{1}\left(x_{1}\right)=B_{2}\left(x_{2}\right)=\left(\begin{array}{l}
0 \\
0 \\
1
\end{array}\right)
$$

The interconnection matrices between subsystem 1 and subsystem 2 are expressed as follows:

$$
H_{12}\left(x_{2}\right)=\left(\begin{array}{ccc}
0 & 0 & 0 \\
x_{22}^{2} & 0 & 0 \\
x_{22} & 0 & 0
\end{array}\right), \quad H_{21}\left(x_{1}, x_{2}\right)=\left(\begin{array}{ccc}
0 & 0 & 0 \\
x_{12}^{2} & 0 & 0 \\
x_{23}^{2} x_{12} & 0 & 0
\end{array}\right) .
$$

The controllability matrices, respectively, for subsystem 1 and subsystem 2 are given as follows:

$$
\begin{align*}
\zeta_{1}\left(x_{1}\right) & =\left[\begin{array}{lll}
B_{1}\left(x_{1}\right) & A_{1}\left(x_{1}\right) B_{1}\left(x_{1}\right) & A_{1}^{2}\left(x_{1}\right) B_{1}\left(x_{1}\right)
\end{array}\right] \\
& =\left(\begin{array}{ccc}
0 & 0 & x_{11} \\
0 & 1 & x_{13} x_{11}^{2} \\
1 & x_{13} x_{11}^{2} & x_{13}^{2} x_{11}^{4}
\end{array}\right),  \tag{33}\\
\zeta_{2}\left(x_{2}\right) & =\left[\begin{array}{lll}
B_{2}\left(x_{2}\right) & \left.A_{2}\left(x_{2}\right) B_{2}\left(x_{2}\right) A_{2}^{2}\left(x_{2}\right) B_{2}\left(x_{2}\right)\right]
\end{array}\right. \\
& =\left(\begin{array}{ccc}
0 & 0 & x_{22} \\
0 & 1 & x_{22} x_{21} \\
1 & 0 & 0
\end{array}\right) . \tag{34}
\end{align*}
$$

$\zeta_{1}\left(x_{1}\right), \zeta_{2}\left(x_{2}\right)$ have a full order rank for all $x_{i}$, which can justify the good choice of (SDC) parametrization. Now, we referring to equation (9), we can write the following decentralized state-dependent Riccati equations:

$$
\left\{\begin{array}{l}
P_{1}\left(x_{1}\right) A_{1}\left(x_{1}\right)+A_{1}^{T}\left(x_{1}\right) P_{1}\left(x_{1}\right)+Q_{1}\left(x_{1}\right)  \tag{35}\\
-P_{1}\left(x_{1}\right) B_{1}\left(x_{1}\right) R_{1}^{-1}\left(x_{1}\right) B_{1}^{T}\left(x_{1}\right) P_{1}\left(x_{1}\right)=0 \\
P_{2}\left(x_{2}\right) A_{2}\left(x_{2}\right)+A_{2}^{T}\left(x_{2}\right) P_{2}\left(x_{2}\right)+Q_{2}\left(x_{2}\right) \\
-P_{2}\left(x_{2}\right) B_{2}\left(x_{2}\right) R_{2}^{-1}\left(x_{2}\right) B_{2}^{T}\left(x_{2}\right) P_{2}\left(x_{2}\right)=0
\end{array}\right.
$$

The decentralized optimal control are expressed as follows:

$$
u_{1}\left(x_{1}\right)=-0.1\left(\begin{array}{lll}
0 & 0 & 1
\end{array}\right) P_{1}\left(x_{1}\right)\left(\begin{array}{l}
x_{11}  \tag{36}\\
x_{12} \\
x_{13}
\end{array}\right)
$$

and

$$
u_{2}\left(x_{2}\right)=-0.1\left(\begin{array}{ccc}
0 & 0 & 1
\end{array}\right) P_{2}\left(x_{2}\right)\left(\begin{array}{l}
x_{21}  \tag{37}\\
x_{22} \\
x_{23}
\end{array}\right) .
$$

## - Numerical simulation:

Figure 1 (respectively Figure 2) shows the behavior of the first states variables $x_{11}, x_{12}$ and $x_{13}$, (respectively, the second states variables $x_{21}, x_{21}$ and $x_{23}$ of interconnected system (30) controlled by the decentralized control laws illustrated in Figure 3. Initial conditions were taken as follows: $x_{11}(0)=x_{13}(0)=x_{21}(0)=x_{22}(0)=0.1, x_{12}(0)=x_{23}(0)=0$.


Figure 1: Closed loop reponses of $x_{1}$.


Figure 2: Closed loop reponses of $x_{2}$.


Figure 3: Decentralized control signals evolution.
We can note a satisfactory stabilization of state variables which converge into the origin point confirming the asymptotic stability of the controlled interconnected system using the decentralized SDRE approach.

## 6 Conclusion

In this paper, we have considered the method for feedback control of nonlinear interconnected systems using the decentralized state-dependent Riccati equation. This decentralized optimal approach is based on the solution of algebraic Riccati equation. Our first result was to determine and prove sufficient conditions that guarantee the global asymptotic stability of the overall interconnected system. We have then run some numerical simulations on a third order system. As expected, these simulations have shown the aptitude of the SDRE approach to be implemented easily and to give satisfactory result in terms of performance for a wide class of nonlinear interconnected systems. One of the possible perspectives that we can consider as a continuity of this research would be to investigate an optimal control for interconnected nonlinear systems via approximate methods.

## Appendix

We recall hereafter the useful mathematical notations and properties concerning the Kronecker tensor product used in this paper.

## A.1. Kronecker product:

The Kronecker product of $A(p \times q)$ and $B(r \times s)$ denoted by $A \otimes B$ is the $(p r \times q s)$ matrix defined by $[24,25]$

$$
A \otimes B=\left(\begin{array}{ccc}
a_{11} B & \ldots & a_{1 q} B  \tag{38}\\
\vdots & \ddots & \vdots \\
a_{p 1} B & \cdots & a_{p q} B
\end{array}\right)
$$

## A.2. Vec-function:

Vec-function is a linear algebra tool which is important in the multidimensional regression matrix representation. This operator is defined as follows [24, 25] :

$$
A=\left(A_{1} A_{2} \ldots A_{n}\right) ; \quad \operatorname{Vec}(A)=\left(\begin{array}{c}
A_{1}  \tag{39}\\
A_{2} \\
\vdots \\
A_{n}
\end{array}\right)
$$

where $\forall i \in\{1, \ldots, n\}, A_{i}$ is a vector of $\mathbb{R}^{m}$.
We recall the following useful rule of this function, given as follows:

$$
\begin{equation*}
V e c(E . A . C)=\left(C^{T} \otimes E\right) V e c(A) \tag{40}
\end{equation*}
$$

## A.3. Mat function :

An important matrix-valued linear function of a vector, denoted by mat ${ }_{(n, m)}($.$) ,$ was defined in [24, 25] as follows: if $V$ is a vector of dimension $p=n . m$, then $M=m a t_{(n, m)}(V)$ is the $(n \times m)$ matrix verifying $V=V e c(M)$.

## References

[1] Yucelen, T., Medagam, P.V. and Pourboghrat, F. Nonlinear quadratic optimal control for cascaded multilevel static compensators. In: 39th North American Power Symposium, Las Cruces, NM, USA, (2007) 523-527.
[2] Bahuguna, D. and Muslim, M. A study of nonlocal history-valued retarded differential equations using analytic semigroups. Nonlinear Dynamics and Systems Theory 6 (1) (2006) 63-75.
[3] Gruyitch, L.T. Consistent Lyapunov Methodology: Non-Differentiable Non-Linear Systems. Nonlinear Dynamics and Systems Theory 1(1) (2001) 1-22.
[4] Elloumi, S., Abidi, B. and Benhadj Braiek, N. Decentralized optimal controller design for multimachine power systems using successive approximation approach. Journal of the Franklin Institute 350 (2013) 2994-3010.
[5] Guang-Hongand, Y. and Si-Ying, Z. Decentralized robust control interconnected systems with time-varying uncertainties. Automatica 32 (1996) 1603-1608.
[6] Hou, H., Nian, X., Xu, S., Sun, M. and Xiong, H. Robust decentralized control for largescale web-winding systems: A linear matrix inequality approach. Transactions of the Institute of Measurement and Control 39 (7)(2017) 953-964.
[7] Li, Y. and Tong, S. Adaptive neural networks decentralized FTC design for nonstrictfeedback nonlinear interconnected large-scale systems against actuator faults. IEEE transactions on neural networks and learning systems 45 (2016) 138-149.
[8] Elloumi, S. and Benhadj Braiek, N. A decentralized stabilization approach of a class of nonlinear polynomial interconnected systems-application for a large scale power system. Nonlinear Dynamics and Systems Theory 12(2) (2012) 159-172.
[9] Elloumi, S. and Benhadj Braiek, N. Robust decentralized control for multimachine power systems-the LMI approach. In: IEEE International Conference on Systems, Man and Cybernetics 6 IEEE, (2002).
[10] Chibani, A., Chadli, M. and Benhadj Braiek, N. A sum of squares approach for polynomial fuzzy observer design for polynomial fuzzy systems with unknown inputs. International Journal of Control, Automation and Systems 14(1) (2016) 323-330.
[11] Huang, Y. and Lu, W.M. Nonlinear optimal control: Alternatives to Hamilton-Jacobi equation. In: Proceedings of the 35th IEEE Conference on Decision and Control, Kobe, Japan, Japan (1996) 3942-3947.
[12] Crandall, M.G., Evans, L.C. and Lions, P.L. Some properties of viscosity solutions of Hamilton-Jacobi equations. Transactions of the American Mathematical Society 282 (1984) 487-502.
[13] Won, C.H. and Biswas, S. Optimal control using an algebraic method for control affine nonlinear systems. International Journal of Control 80 (2007) 1491-1502.
[14] Wernli, A. and Cook, G. Suboptimal control for the nonlinear quadratic regulator problem. Automatica 11 (1975) 75-84.
[15] Primbs, J.A., Nevistić, V. and Doyle, J.C. Nonlinear optimal control: A control Lyapunov function and receding horizon perspective. Asian Journal of Control 1 (1) (1999) 14-24.
[16] Kamocki, R. Necessary and sufficient optimality conditions for fractional nonhomogeneous Roesser model. Optimal Control Applications and Methods 37 (4) (2016) 574-589.
[17] Pearson, J.D. Approximation methods in optimal control. Journal of Electronics and Control 13 (1962) 453-469.
[18] Wernli, A. and Cook, G. Suboptimal control for the nonlinear quadratic regulator problem. Automatica 11(1) (1975) 75-84.
[19] Mracek, C.P. and Cloutier, J.R. Control designs for the nonlinear benchmark problem via the state-dependent Riccati equation method. International Journal of Robust and Nonlinear Control 8 (4-5) (1998) 401-433.
[20] Cimen, T. State-dependent Riccati equation (SDRE) control: A survey. IFAC Proceedings 41(2) (2008) 3761-3775.
[21] Banks, H.T., Kwon, H.D., Toivanen, J.A. and Tran, H.T. A state-dependent Riccati equation-based estimator approach for HIV feedback control. Optimal Control Applications and Methods 27 (2006) 93-121.
[22] Elloumi, S. and Benhadj Braiek, N. On Feedback Control Techniques of Nonlinear Analytic Systems. Journal of Applied Research and Technology 12(2014) 500-513.
[23] Elloumi, S., Sansa, I. and Benhadj Braiek, N. On the stability of optimal controlled systems with SDRE approach. In: 9th International Multi-Conference on Systems, Signals and Devices (SSD), Chemnitz, Germany (2012) 1-5.
[24] Bouafoura, M.K., Moussi, O. and Benhadj Braiek, N. A fractional state space realization method with block pulse basis. Signal Processing 91(3) (2011) 492-497.
[25] Brewer, J.W. Kroneker products and matrix calculus in system theory. IEEE Transaction on Circuits and Systems (CAS) 25 (1978) 772-781.

# Analysis and Adaptive Control Synchronization of a Novel 3-D Chaotic System 

F. Hannachi *<br>Department of Management Sciences, University of Tebessa, (12002), Algeria

Received: August 28, 2018; Revised: December 8, 2018


#### Abstract

In this paper, a new 3D chaotic system is introduced. Basic dynamical characteristics and properties of this new chaotic system are studied, namely the equilibrium points and their stability, the Lyapunov exponent, Lyapunov exponent spectrum and the Kaplan-Yorke dimension. Also, we derive new control results via the adaptive control method based on Lyapunov stability theory and the adaptive control theory of this new chaotic system with unknown parameters. The results are validated by numerical simulation using Matlab.


Keywords: chaotic system; strange attractor; Lyapunov exponent; Lyapunov stability theory; adaptive control; synchronization.
Mathematics Subject Classification (2010): 37B55, 34C28, 34D08, 37B25, 37D45, 93C40, 93D05.

## 1 Introduction

In mathematics and physics, chaos theory deals with the behavior of certain nonlinear dynamical systems that under certain conditions exhibit a phenomenon known as chaos, which is characterised by a sensitivity to initial conditions [1]. Chaos as an important nonlinear phenomenon has been studied in mathematics, engineering and many other disciplines. Since Lorenz discovered a three-dimensional autonomous chaotic system [2], many other systems have been introduced and analysed, we mention the Chen, Rössler and Lü systems $[3,4,5]$. After that hyperchaotic systems were constructed using many different methods. The synchronization of two chaotic systems was introduced in the work of Pecora and Carroll [6], then many different methodologies have been developed for synchronization of chaotic systems such as the OGY method [7], active contol method [8], sliding mode control [9], backstepping control [10], function projective method [11], adaptive control [12-14], etc.

In this work, a new chaotic system is introduced and we derive new control results via the adaptive control method based on Lyapunov stability theory and the adaptive control theory for this new chaotic system with unknown parameters. The results are validated by numerical simulation using Matlab.

[^5]
### 1.1 Description of the novel chaotic system

In this research work, we propose a new 3D chaotic system with two quadratic nonlinearities, which is given in the system form as

$$
\left\{\begin{array}{l}
\frac{d x_{1}}{d t}=a\left(x_{2}-x_{1}\right)  \tag{1}\\
\frac{d x_{2}}{d t}=c x_{1}+x_{1} x_{3} \\
\frac{d x_{3}}{d t}=-x_{1} x_{2}+b\left(x_{1}-x_{3}\right)
\end{array}\right.
$$

where $a, b, c$ are positive reals parameters. In the first part of this paper, we shall show that the system (1) is chaotic when the system parameters $a, b$ and $c$ take the values:

$$
\begin{equation*}
a=13, b=2.5, c=50 \tag{2}
\end{equation*}
$$

### 1.2 Basic properties

In this section, some basic properties of system (1) are given. We start with the equilibrum points of the system and check their stability at the initial values of the parameters $a, b, c$.

### 1.3 Equilibrum points

Putting equations of system (1) equal to zero, i.e.

$$
\begin{equation*}
a\left(x_{2}-x_{1}\right)=0, \quad c x_{1}+x_{1} x_{3}=0, \quad-x_{1} x_{2}+b\left(x_{1}-x_{3}\right)=0 \tag{3}
\end{equation*}
$$

gives the three equilibrium points

$$
\begin{equation*}
p_{0}=(0,0,0), \quad p_{1,2}=\left(\frac{1}{2} b \mp \frac{1}{2} \sqrt{4 b c+b^{2}}, \frac{1}{2} b \mp \frac{1}{2} \sqrt{4 b c+b^{2}},-c\right) . \tag{4}
\end{equation*}
$$

### 1.4 Stability

In order to check the stability of the equilibrum points we derive the Jacobian matrix at a point $p(x, y, z)$ of the system (1)

$$
J(p)=\left(\begin{array}{ccc}
-a & a & 0  \tag{5}\\
c+z & 0 & x \\
b-y & -x & -b
\end{array}\right)
$$

For $p_{0}$, we obtain $J\left(p_{0}\right)=\left(\begin{array}{ccc}-a & a & 0 \\ c & 0 & 0 \\ b & 0 & -b\end{array}\right)$, with the characteristic polynomial equation $\lambda^{3}+(a+b) \lambda^{2}+(a b-a c) \lambda-a b c=0$, which has three eigenvalues

$$
\begin{equation*}
\lambda_{1}=19.811, \lambda_{2}=-2.5, \lambda_{3}=-32.811 \tag{6}
\end{equation*}
$$

Since all the eigenvalues are real, the Hartma-Grobman theorem implies that $p_{0}$ is a saddle point which is unstable according to the Lyapunov theorem of stability.

By the same method, the eigenvalues of the Jacobian at $p_{1}$ are:

$$
\begin{equation*}
\lambda_{1}=0.99385-12.895 i, \lambda_{2}=0.99385+12.895 i, \lambda_{3}=-17.488 \tag{7}
\end{equation*}
$$

The eigenvalues of the Jacobian at $p_{2}$ are:

$$
\begin{equation*}
\lambda_{1}=0.76322-14.634 i, \lambda_{2}=0.76322+14.634 i, \lambda_{3}=-17.026 \tag{8}
\end{equation*}
$$

Then $p_{1}$ and $p_{2}$ are two unstable saddle-foci because none of the eigenvalues have zero real part and $\lambda_{1}, \lambda_{2}$ are complex.

### 1.5 Dissipativity

A dissipative dynamical system satisfies the condition

$$
\begin{equation*}
\nabla \cdot V=\frac{\partial \dot{x}}{\partial x}+\frac{\partial \dot{y}}{\partial y}+\frac{\partial \dot{z}}{\partial z}<0 . \tag{9}
\end{equation*}
$$

In the case of the system (1), we have

$$
\begin{equation*}
\nabla \cdot V=-(a+b) \tag{10}
\end{equation*}
$$

For $a=13, b=2.5, c=50$ we obtain $\nabla \cdot V=-15.5<0$, and threfore dissipativity condition holds for this system. Also,

$$
\begin{equation*}
\frac{d V}{d t}=e^{-(a+b)}=1.8554 \times 10^{-7} \tag{11}
\end{equation*}
$$

Then the volume of the attractor decreases by a factor of 0.00000018554 .

## 2 Lyapunov Exponents and Kaplan-Yorke Dimension

Lyapunov exponents are used to measure the exponential rates of divergence and convergence of nearby trajectoiries, which is an important characterstic to judge whether the system is chaotic or not. The existence of at least one positive Lyapunov exponent implies that the system is chaotic.

For the chosen parameter values (2), the Lyapunov exponents of the novel chaotic system (1) are obtained using Matlab as:

$$
\begin{equation*}
L_{1}=1.4375, L_{2}=-0.000166417, L_{3}=-16.9373 \tag{12}
\end{equation*}
$$

The Lyapunov exponents spectrum is shown in Fig. 1.
Since the spectrum of Lyapunov exponents (13) has a positive term $L_{1}$, it follows that the novel 3-D chaotic system (1) is chaotic. The Kaplan-Yorke dimension of system (1) is calculated as

$$
\begin{equation*}
D_{K L}=2+\frac{L_{1}+L_{2}}{\left|L_{3}\right|}=2.0849 \tag{13}
\end{equation*}
$$

## 3 Adaptive Control of the Novel 3-D Chaotic System

This section describes an adaptive design of a globally stabilizing feedback controller for the chaotic system (1) with unknown parameters. The design is carried out using the adaptive control theory and Lyapunov stability theory.


Figure 1: Lyapunov exponents spectrum.


Figure 2: Projection of the strange attractor of the system (1) into the (z; x)-plane.

A controlled chaotic system of (1) is given by

$$
\left\{\begin{align*}
\frac{d x_{1}}{d t} & =a\left(x_{2}-x_{1}\right)+u_{1}  \tag{14}\\
\frac{d x_{2}}{d t} & =c x_{1}+x_{1} x_{3}+u_{2} \\
\frac{d x_{3}}{d t} & =-x_{1} x_{2}+b\left(x_{1}-x_{3}\right)+u_{3}
\end{align*}\right.
$$

where $a, b, c$ are unknown constant parameters, and $u_{1}, u_{2}, u_{3}$ are adaptive controllers to be found using the states $x_{1}, x_{2}, x_{3}$ and estimates $a_{1}(t), b_{1}(t), c_{1}(t)$ of the unknown parameters $a, b, c$, respectively.

We take the adaptive control law defined by

$$
\left\{\begin{array}{l}
u_{1}=-a_{1}(t)\left(x_{2}-x_{1}\right)-k_{1} x_{1}  \tag{15}\\
u_{2}=-c_{1}(t) x_{1}-x_{1} x_{3}-k_{1} x_{2} \\
u_{3}=x_{1} x_{2}-b_{1}(t)\left(x_{1}-x_{3}\right)-k_{3} x_{3}
\end{array}\right.
$$

where $k_{1}, k_{2}, k_{3}$ are positive gain constants.
Substituting (15) into (14), we obtain the closed-loop control system as

$$
\left\{\begin{array}{l}
\frac{d x_{1}}{d t}=\left(a-a_{1}(t)\right)\left(x_{2}-x_{1}\right)-k_{1} x_{1}  \tag{16}\\
\frac{d x_{2}}{d t}=\left(c-c_{1}(t)\right) x_{1}-k_{2} x_{2} \\
\frac{d x_{3}}{d t}=\left(b-b_{1}(t)\right)\left(x_{1}-x_{3}\right)-k_{3} x_{3}
\end{array}\right.
$$

We define the parameter estimation errors as

$$
\begin{equation*}
e_{a}(t)=a-a_{1}(t), \quad e_{c}(t)=c-c_{1}(t), \quad e_{b}(t)=b-b_{1}(t) \tag{17}
\end{equation*}
$$

By using (17), we rewrite the closed-loop system (16) as

$$
\left\{\begin{array}{l}
\frac{d x_{1}}{d t}=e_{a}(t)\left(x_{2}-x_{1}\right)-k_{1} x_{1}  \tag{18}\\
\frac{d x_{2}}{d t}=e_{c}(t) x_{1}-k_{2} x_{2} \\
\frac{d x_{3}}{d t}=e_{b}(t)\left(x_{1}-x_{3}\right)-k_{3} x_{3}
\end{array}\right.
$$

Differentiating (17) with respect to $t$, we obtain

$$
\left\{\begin{align*}
\frac{d e_{a}(t)}{d t} & =-\frac{d a_{1}(t)}{d t}  \tag{19}\\
\frac{d e_{c}(t)}{d t} & =-\frac{d c_{1}(t)}{d t} \\
\frac{d e_{b}(t)}{d t} & =-\frac{d b_{1}(t)}{d t}
\end{align*}\right.
$$

To find an update law for the parameter estimates, we shall use the Lyapunov stability theory. We consider the quadratic Lyapunov function given by

$$
\begin{equation*}
V\left(x_{1}, x_{2}, x_{3}, e_{a}, e_{b}, e_{c}\right)=\frac{1}{2}\left(x_{1}^{2}+x_{2}^{2}+x_{3}^{2}+e_{a}^{2}+e_{b}^{2}+e_{c}^{2}\right) \tag{20}
\end{equation*}
$$

which is a positive definite function on $\mathbb{R}^{6}$.
Differentiating $V$ along the trajectories of the systems (18) and (19), we obtain the following:
$\dot{V}=-\sum_{i=1}^{3} k_{i} x_{i}^{2}+e_{a}\left(x_{1} x_{2}-x_{1}^{2}-\frac{d a_{1}(t)}{d t}\right)+e_{b}\left(x_{1} x_{3}-x_{3}^{2}-\frac{d b_{1}(t)}{d t}\right)+e_{c}\left(x_{1} x_{2}-\frac{d c_{1}(t)}{d t}\right)$.
In view of (21), we take the parameter update law as follows

$$
\left\{\begin{array}{l}
\frac{d a_{1}(t)}{d t}=x_{1} x_{2}-x_{1}^{2}  \tag{22}\\
\frac{d b_{1}(t)}{d t}=x_{1} x_{3}-x_{3}^{2} \\
\frac{d c_{1}(t)}{d t}=x_{1} x_{2}
\end{array}\right.
$$

Theorem 3.1 The 3-D novel chaotic system (14) with unknown parameters is globally and exponentially stabilized by the adaptive feedback control law (15) and the parameter update law (22), where $k_{1}, k_{2}, k_{3}$ are positive constants 3.1.

Proof. Substituting the parameter update law (21) into (20), we obtain the time derivative of $V$ as:

$$
\begin{equation*}
\dot{V}=-k_{1} x_{1}^{2}-k_{2} x_{2}^{2}-k_{3} x_{3}^{2} \tag{23}
\end{equation*}
$$

which is a negative definite function on $\mathbb{R}^{6}$. By the direct method of Lyapunov [15], it follows that $x_{1}, x_{2}, x_{3}, e_{a}, e_{b}, e_{c}$ are globally exponentially stable.

### 3.1 Numerical simulations

We used the classical fourth-order Runge-Kutta method with the step size $h=10^{-8}$ to solve the system of differential equations (14) and (22), when the adaptive control law (15) is applied.

The parameter values of the novel 3-D chaotic system (14) are chosen as in the chaotic case (2). The positive gain constants are taken as $k_{i}=3$, for $i=1,2,3$.

The initial conditions of the novel chaotic system (14) are chosen as $x_{1}(0)=$ $6.4, x_{2}(0)=-4.7, x_{3}(0)=2.5$. Furthermore, as initial conditions of the parameter estimates of the unknown parameters, we have chosen: $a_{1}(0)=2.5, b_{1}(0)=5.3, c_{1}(0)=4.8$.

In Figs. 3-4, the exponential convergence of the controlled states $x_{1}(t), x_{2}(t), x_{3}(t)$ and the time-history of the parameter estimates $a_{1}(t) ; b_{1}(t) ; c_{1}(t)$ are depicted, when the adaptive control law (15) and parameter update law (22) are implemented.


Figure 3: Exponential convergence of the controlled states $x_{1}(t) ; x_{2}(t) ; x_{3}(t)$.

## 4 Adaptive Synchronization of the Identical Novel 3-D Chaotic Systems

In this section, we derive an adaptive control law for globally and exponentially synchronizing the identical novel 3-D chaotic systems with unknown system parameters. Thus, the master system is given by the novel chaotic system dynamics

$$
\left\{\begin{array}{l}
\frac{d x_{1}}{d t}=a\left(x_{2}-x_{1}\right)  \tag{24}\\
\frac{d x_{2}}{d t}=c x_{1}+x_{1} x_{3} \\
\frac{d x_{3}}{d t}=-x_{1} x_{2}+b\left(x_{1}-x_{3}\right)
\end{array}\right.
$$

Also, the slave system is given by the novel chaotic system dynamics

$$
\left\{\begin{array}{l}
\frac{d y_{1}}{d t}=a\left(y_{2}-y_{1}\right)+u_{1}  \tag{25}\\
\frac{d y_{2}}{d t}=c y_{1}+y_{1} y_{3}+u_{2} \\
\frac{d y_{3}}{d t}=-y_{1} y_{2}+b\left(y_{1}-y_{3}\right)+u_{3}
\end{array}\right.
$$



Figure 4: Time-history of the parameter estimates $a_{1}(t) ; b_{1}(t) ; c_{1}(t)$.

In (24) and (25), the system parameters $a, b, c$ are unknown and the design goal is to find an adaptive feedback controls $u_{1}, u_{2}, u_{3}$ using the states $x_{1}, x_{2}, x_{3}$ and estimates $a_{1}(t), b_{1}(t), c_{1}(t)$ of the unknown parameters $a, b, c$, respectively. The synchronization error between the novel chaotic systems (24) and (25) is defined as

$$
\begin{equation*}
e_{1}=y_{1}-x_{1}, \quad e_{2}=y_{2}-x_{2}, \quad e_{3}=y_{3}-x_{3} \tag{26}
\end{equation*}
$$

Then (26) implies

$$
\left\{\begin{array}{l}
\dot{e}_{1}=\dot{y}_{1}-\dot{x}_{1},  \tag{27}\\
\dot{e}_{2}=\dot{y}_{2}-\dot{x}_{2}, \\
\dot{e}_{3}=\dot{y}_{3}-\dot{x}_{3} .
\end{array}\right.
$$

Thus, the synchronization error dynamics is obtained as

$$
\left\{\begin{array}{l}
\dot{e}_{1}=a\left(e_{2}-e_{1}\right)+u_{1}  \tag{28}\\
\dot{e}_{2}=c e_{1}+y_{1} y_{3}-x_{1} x_{3}+u_{2} \\
\dot{e}_{3}=b\left(e_{1}-e_{3}\right)-y_{1} y_{2}+x_{1} x_{2}+u_{3}
\end{array}\right.
$$

We take the adaptive control law defined by

$$
\left\{\begin{array}{l}
u_{1}=-a_{1}\left(e_{2}-e_{1}\right)-k_{1} e_{1},  \tag{29}\\
u_{2}=-c_{1} e_{1}-y_{1} y_{3}+x_{1} x_{3}-k_{2} e_{2} \\
u_{3}=-b_{1}\left(e_{1}-e_{3}\right)+y_{1} y_{2}-x_{1} x_{2}-k_{3} e_{3}
\end{array}\right.
$$

where $k_{1}, k_{2}, k_{3}$ are positive gain constants.
Substituting (29) into (28), we obtain the closed-loop error dynamics as

$$
\left\{\begin{array}{l}
\dot{e}_{1}=\left(a-a_{1}\right)\left(e_{2}-e_{1}\right)-k_{1} e_{1}  \tag{30}\\
\dot{e}_{2}=\left(c-c_{1}\right) e_{1}-k_{2} e_{2} \\
\dot{e}_{3}=\left(b-b_{1}\right)\left(e_{1}-e_{3}\right)-k_{3} e_{3}
\end{array}\right.
$$

The parameter estimation errors are defined as

$$
\begin{equation*}
e_{a}(t)=a-a_{1}(t), \quad e_{c}(t)=c-c_{1}(t), \quad e_{b}(t)=b-b_{1}(t) . \tag{31}
\end{equation*}
$$

Differentiating (31) with respect to $t$, we obtain

$$
\left\{\begin{array}{l}
\frac{d e_{a}(t)}{d t}=-\frac{d a_{1}(t)}{d t},  \tag{32}\\
\frac{d e_{c}(t)}{d t}=-\frac{d c_{1}(t)}{d t}, \\
\frac{d e_{b}(t)}{d t}=-\frac{d b_{1}(t)}{d t} .
\end{array}\right.
$$

By using (31), we rewrite the closed-loop system (30) as

$$
\left\{\begin{array}{l}
\dot{e}_{1}=e_{a}\left(e_{2}-e_{1}\right)-k_{1} e_{1}  \tag{33}\\
\dot{e}_{2}=e_{c} e_{1}-k_{2} e_{2} \\
\dot{e}_{3}=e_{b}\left(e_{1}-e_{3}\right)-k_{3} e_{3}
\end{array}\right.
$$

We consider the quadratic Lyapunov function given by

$$
\begin{equation*}
V\left(x_{1}, x_{2}, x_{3}, e_{a}, e_{b}, e_{c}\right)=\frac{1}{2}\left(x_{1}^{2}+x_{2}^{2}+x_{3}^{2}+e_{a}^{2}+e_{b}^{2}+e_{c}^{2}\right) \tag{34}
\end{equation*}
$$

which is a positive definite function on $\mathbb{R}^{6}$.
Differentiating $V$ along the trajectories of the systems (33) and (32), we obtain the following:
$\dot{V}=-\sum_{i=1}^{3} k_{i} e_{i}^{2}+e_{a}\left(e_{1} e_{2}-e_{1}^{2}-\frac{d a_{1}(t)}{d t}\right)+e_{b}\left(e_{1} e_{3}-e_{3}^{2}-\frac{d b_{1}(t)}{d t}\right)+e_{c}\left(e_{1} e_{2}-\frac{d c_{1}(t)}{d t}\right)$.
In view of (35), we take the parameter update law as follows:

$$
\left\{\begin{array}{l}
\frac{d a_{1}(t)}{d t}=e_{1} e_{2}-e_{1}^{2}  \tag{36}\\
\frac{d b_{1}(t)}{d t}=e_{1} e_{3}-e_{3}^{2} \\
\frac{d c_{1}(t)}{d t}=e_{1} e_{2}
\end{array}\right.
$$

Substituting (36) into (35), we get

$$
\begin{equation*}
\dot{V}=-\sum_{i=1}^{3} k_{i} e_{i}^{2} \tag{37}
\end{equation*}
$$

which is a negative definite function on $\mathbb{R}^{3}$. Hence, by the Lyapunov stability theory [15], it follows that $e_{i}(t) \longrightarrow 0$ as $t \longrightarrow \infty$ for $i=1,2,3$. Hence, we have proved the following theorem.

Theorem 4.1 The novel 3-D chaotic systems (24) and (25) with unknown parameters are globally and exponentially synchronized for all initial conditions by the adaptive feedback control law (29) and the parameter update law (36), where $k_{1}, k_{2}, k_{3}$ are positive constants 4.1.


Figure 5: Synchronization of the states $x_{1}(t)$ and $y_{1}(t)$.


Figure 6: Synchronization of the states $x_{2}(t)$ and $y_{2}(t)$.

### 4.1 Numerical simulations

We used the classical fourth-order Runge-Kutta method with the step size $h=10^{-8}$ to solve the system of differential equations (24), (25) and (36), when the adaptive control law (29) is applied.

The parameter values of the novel 3-D chaotic system (24) are chosen as in the chaotic case (2). The positive gain constants are taken as $k_{i}=4$, for $i=1,2,3$.

The initial conditions for the master system (24) are chosen as $x_{1}(0)=5, x_{2}(0)=$


Figure 7: Synchronization of the states $x_{3}(t)$ and $y_{3}(t)$.


Figure 8: Time-history of the synchronization errors $e_{1}(t), e_{2}(t), e_{3}(t)$.
$-3, x_{3}(0)=-10$ and those for the slave system (25) are chosen as $y_{1}(0)=14, y_{2}(0)=$ $10, y_{3}(0)=5$. Furthermore, as initial conditions of the parameter estimates of the unknown parameters, we have chosen $a_{1}(0)=10, b_{1}(0)=15, c_{1}(0)=20$. In Figs. 5-7, the synchronization of the states of the master system (24) and slave system (25) is depicted, when the adaptive control law (29) and parameter update law (36) are implemented. In Fig. 8, the time-history of the synchronization errors $e_{1}(t), e_{2}(t), e_{3}(t)$ is depicted.

## 5 Conclusion

In this paper, a new chaotic system is introduced. Basic properties of this system are studied, namely, the equilibrum points and their stability, the Lyapunov exponent and the Kaplan-Yorke dimension. Moreover, adaptive control schemes have been proposed to stabilize and synchronize such two new chaotic systems. Numerical simulations using MATLAB have been made to illustrate our results for the new chaotic system with unknown parameters.

## Acknowledgment

The author would like to thank the editor in chief and the referees for their valuable suggestions and comments.

## References

[1] Ott, E. Chaos in Dynamical Systems. Cambridge University Press, Cambridge, 2002.
[2] Lorenz, E.N. Deterministic nonperiodic flow. Journal of the Atmospheric Sciences 20 (5) (1963) 130-141.
[3] Chen, G. and Ueta, T. Yet another chaotic attractor. International Journal of Bifurcation and Chaos 9 (7) (1999) 1465-1466.
[4] Lü, J. and Chen, G. A new chaotic attractor coined. International Journal of Bifurcation and Chaos 12 (3) (2002) 659-661.
[5] Rössler, O. An equation for continuous chaos. Physics Letters A 57 (5) (1976) 397-398.
[6] Pecora, L. and Carroll, T. Synchronization in chaotic systems. Physical Review Letters 64 (8) (1990) 821-824.
[7] Grebogi, C. and Lai, Y. C. Controlling chaotic dynamical systems. Systems and control letters 31 (5) (1997) 307-312.
[8] Ho, M. and Hung, Y. Synchronization of two different chaotic systems using generalized active control. Physics Letters A 301 (5) (2002) 424-428.
[9] Sun, J., Shen, Y., Wang, X. et al. Finite time combination-combination synchronization of four different chaotic systems with unknown parameters via sliding mode control. Nonlinear Dynamics 76 (1) (2014) 383-397.
[10] Xiao-Qun, W. and Jun-An, L. Parameter identification and backstepping control of uncertain Lü system. Chaos, Solitons and Fractals 18 (1) (2003) 721-729.
[11] Othman, A. A., Noorani, M. S. M. and Al-Sawalha, M. M. Function projective dual synchronization of chaotic systems with uncertain parameters. Nonlinear Dynamics and Systems Theory 17 (2) (1963) 193-204.
[12] Adloo, H. and Roopaei, M. Review article on adaptive synchronization of chaotic systems with unknown parameters. Nonlinear Dynamics 65 (1) (2011) 141-159.
[13] Vaidyanathan, S., Vollos, C., Pham, V . and Madhavan, K. Analysis, adaptive control and synchronization of a novel 4-D hyperchaotic hyperjerk system and its SPICE implementation. Archives of Control Sciences 25 (1) (2003) 135-158.
[14] Vaidyanathan, S. and Vollos, C. Analysis and adaptive control of a novel 3-D conservative noequilibrium chaotic system. Archives of Control Sciences 25 (3) (2015) 333-353.
[15] Hahn, W. The Stability of Motion. Springer, Berlin, New York, 1967.

# Mathematical Model of $C_{d}$ for Circular Cylinder Using Two Passive Controls at $\operatorname{Re}=5000$ 

C. Imron ${ }^{1 *}$ C.J. Kumalasari ${ }^{1}$ B. Widodo ${ }^{2}$ and T.Y. Yuwono ${ }^{2}$<br>${ }^{1}$ Department of Mathematics, Faculty of Mathematics, Computation and Data Science, Institute of Technology Sepuluh Nopember, Surabaya, Indonesia<br>${ }^{2}$ Department of Mechanical Engineering, Faculty of Industrial Technology, Institute of Technology Sepuluh Nopember, Surabaya, Indonesia

Received: December 8, 2017; Revised: December 10, 2018


#### Abstract

This study focuses on two passive controls. Passive control is the addition of a small object to an object to reduce the drag force of the object. In this case, two passive controls are placed in front of and in the rear of the main object. The distance between the main object and the two passive controls varies and the Reynolds number used is 5000 . The main object is a circular cylinder, and its passive control in front is a cylinder of type- $I$ at the distance $\mathrm{S} / \mathrm{D}=0.6,1.2 ; 1.8 ; 2.4 ; 3.0$ and in the rear is an elliptical or circular cylinder at the distance $\mathrm{T} / \mathrm{D}=0,6 ; 0,9 ; 1,2 ; 1,5 ; 1,8$ and 2,1 . In this study, we want to find an effective distance of the main object to two passive controls so that the drag coefficient of the main object is minimal compared to that with non-passive control or with one passive control in front. In addition, a mathematical model of the drag coefficient of circular cylinders with two passive controls at $\operatorname{Re}=5000$ will be obtained.


Keywords: passive control; drag coefficient; cylinder.

Mathematics Subject Classification (2010): 58D30, 65C20.

[^6]
## 1 Introduction

Today, so many people are racing to create new technologies. Technological advancement is growing rapidly. Technology is actually a way and effort to improve the quality of human life [1]. New technologies can be created by conducting ongoing research, where the new technology is expected to change the behavior of users of these new technologies. Research related to fluid flow can be done by experiment or simulation. Study of the flow of fluids through objects with the aim of reducing the drag force of most objects is a paramount concern of the researchers.

Some researchers used one passive control placed in front of various shapes, such as cylindrical cylinders, type- $I$ cylinders, type- $D$ cylinders etc. Circular cylinders, elliptical cylinders or other shapes are commonly used objects for designing industrial chimneys, offshore and flyover structures and others. In this case, the design process should allow for the geometrical shape of the object because it affects the value of the drag coeffcient, so that for different geometric shapes the drag coeffcient values are also different. At the interaction between the fluid flow and the object the resulting fluid flow across a single object or multiple grouped objects will produce different flow characteristics.

In this study, we consider a boundary layer because it is seen that the liquid that flows through the surface of the object comes with the flow of particles around it. Basically, the boundary layer is an increase in shear stress which will affect the flow velocity in each layer [14]. The surface of the object will move slowly due to the friction force, so that the particle flow velocity around the object will be zero. While the other particles will interact, the velocity of the flow away from the object will be faster. This is due to increased shear stress.

There are some studies that use boundary layer concept, and the concept of the boundary layer can help to find the answer to the effect of shear stress having a very important role in flow characteristics around the object [2]. The research, among others, has been conducted on the flow of fluids through an object, such as a single cylindrical circular object [3], or a modified cylinder such as a cylinder of type- $I$ or a cylinder of type$D[4,5]$ and a study has been conducted on a fluid stream through more than one object, i.e. fluid flow through more than one cylinder of various sizes and configurations, fluid flow through a circular cylinder with tandem configuration [6-9] and eliptical cylinders with their side configurations $[10,11]$.

The existence of a drag force occurs when an object is bypassed by a fluid. In this case, the drag force is influenced by several parameters, one of which is the drag coefficient. One way to reduce the drag force on the objects bypassed is to add a smaller object in front of the main object called the passive control. The addition of passive control is carried out to reduce the coefficient by $48 \%$ [6], also one can find a mathematical model for a circular cylinder with two passive controls with the Reynolds number 5000. The cylinder of type- $I$ is a circular cylinder obtained by cutting the left and right ends at a certain angle, so that the cylinder is shaped like I. The best cutting edge is $53^{\circ}$, this is because the wake occured is wider than that at the other angle, forming also a wider and more annoying strong flow on the object wall.

In this study, we will get a mathematical model for a circular cylinder with two passive controls with the Reynolds number when these two passive controls effectively decrease the drag coefficient. The Reynolds number used is $\operatorname{Re}=5000$. Two passive controls will be used, the passive control in front is the cylinder of type- $I$ and the passive control of type- $I$ is placed perpendicular to the flow, while the passive control in the rear
is landscape. The distance between the passive control in front and the circular cylinder is varying, as well as the distance between the passive control in the rear and the circular cylinder.

## 2 Numerical Method

The previously described problem can be solved by using the unstable incompressible fluid equation and the Navier-Stokes equation:

$$
\begin{gather*}
\frac{\partial \mathbf{v}}{\partial t}+\nabla \cdot \mathbf{v} \mathbf{v}=-\nabla P+\frac{1}{\operatorname{Re}} \nabla^{2} \mathbf{v}  \tag{1}\\
\nabla \cdot \mathbf{v}=0 \tag{2}
\end{gather*}
$$

where Re is the Reynolds number, $\mathbf{v}$ is the velocity, and P is the pressure. The NavierStokes equation can be solved by using SIMPLE algorithms and numerical methods. The first thing to do is to give the initial value for each variable. By ignoring the pressure components, we will find the velocity component of the momentum equation, so equation (1) becomes

$$
\begin{equation*}
\frac{\partial \mathbf{v}}{\partial t}=-\nabla \cdot \mathbf{v v}+\frac{1}{\operatorname{Re}} \nabla^{2} \mathbf{v} \tag{3}
\end{equation*}
$$

by using the finite difference method, we have

$$
\begin{gathered}
\left(f_{x}\right)_{i}=\frac{2 f_{i+1}+3 f_{i}-6 f_{i-1}+f_{i-2}}{6 \mathrm{dx}} \quad \text { and } \quad\left(f_{y}\right)_{j}=\frac{2 f_{j+1}+3 f_{j}-6 f_{j-1}+f_{j-2}}{6 \mathrm{dx}}, \\
\left(f_{x x}\right)_{i}=\frac{f_{i+1}-2 f_{i}+f_{i-1}}{\mathrm{dx}^{2}} \quad \text { and } \quad\left(f_{y} y\right)_{j}=\frac{f_{j+1}-2 f_{j}+f_{j-1}}{\mathrm{dx}^{2}}
\end{gathered}
$$

and afterwards

$$
\begin{equation*}
\frac{\partial \mathbf{v}}{\partial t}=\frac{\mathbf{v}^{* *}-\mathbf{v}^{*}}{\Delta t}=-\nabla P \tag{4}
\end{equation*}
$$

because of equation (2), then equation (4) becomes

$$
\begin{equation*}
\frac{\nabla \cdot \mathbf{v}^{*}}{\Delta t}=-\Delta P \tag{5}
\end{equation*}
$$

by using SOR (Successive Over Relaxation)

$$
\begin{equation*}
\left(P_{n}\right)_{i, j}=(1-\epsilon)\left(P_{n-1}\right)_{i, j}+\epsilon\left(P_{n}\right)_{i, j} \tag{6}
\end{equation*}
$$



Figure 1: Design of the research system.


Figure 2: Schematic of two passive controls and a circular cylinder.

## 3 Main Result

Our research system is 10D 20D, where D is the diameter of the circular cylinder, placed at the distance of 4 D from the front of the system and in the center of the system, as shown in Figure 1.

In this study, we used two passive controls. The first passive control is a cylinder of type- $I$ placed in front of a circular cylinder at varying distance, i.e. $\mathrm{S} / \mathrm{D}=0.6,1,2,1,8,2,4$ and 3.0. The second passive control are circular cylindrical and elliptical cylinders. The second passive control is placed in the rear of the circular cylinder at varying distance, i.e. $\mathrm{T} / \mathrm{D}=0.6,0,9,1,2,1,5,1,8$ and 2.1 as shown in Figure 2.

### 3.1 Drag coefficient

The drag coeffcient of a single circular cylinder has been obtained by using the simulation program, the results are compared with experimental results and other simulation programs. We calculated that the drag coefficient of a single cylinder with $\operatorname{Re}=100$ is 1.356, while other researchers, with the same Reynolds number, have obtained: Zulhidayat has 1.4 and Five has 1.39 [12]. In this paper we will simulate a circular cylinder
with two passive controls, and the Reynolds number used is 5000. The drag coefficient for a circular cylinder with $\operatorname{Re}=5000$ is 1.51 .

| $\mathrm{S} / \mathrm{D}$ | 0.6 | 1.2 | 1.8 | 2.4 | 3.0 |
| :---: | :---: | :---: | :---: | :---: | :---: |
| $C_{D} 5000$ | 1.455 | 1.273 | 1.221 | 1.224 | 1.216 |

Table 1: Cd of a circular cylinder for $\mathrm{Re}=5000$ with difference $\mathrm{S} / \mathrm{D}$.

Table 1 presents data on the drag coeffcient of a circular cylinder with a passive control, the cylinder of type- $I$, located at the front at varying distance. From the table it is clear that for the Reynold number $\operatorname{Re}=5000$, the best distance to get the minimum drag coeffcient is $\mathrm{S} / \mathrm{D}=1.8$ or $\mathrm{S} / \mathrm{D}=3.0$ with a drag coeffcient of 1.221 or 1.216 . The value of the drag coeffcient is still smaller than the drag coeffcient without passive control.

| $C_{D O}$ | $\mathrm{~S} / \mathrm{D}$ |  |  |  |  |
| :---: | :---: | :---: | :---: | :---: | :---: |
| $\mathrm{T} / \mathrm{D}$ | 0.6 | 1.2 | 1.8 | 2.4 | 3.0 |
| 0.6 | 1.116 | 1.012 | 0.973 | 0.992 | 0.987 |
| 0.9 | 1.205 | 1.043 | 1.015 | 1.007 | 1.008 |
| 1.2 | 1.169 | 1.014 | 0.977 | 0.990 | 0.986 |
| 1.5 | 1.412 | 1.277 | 0.916 | 1.265 | 1.245 |
| 1.8 | 1.557 | 1.284 | 1.225 | 1.222 | 1.195 |
| 2.1 | 1.401 | 1.384 | 1.220 | 1.209 | 1.191 |

Table 2: $C_{D}$ of a circular cylinder for $\mathrm{Re}=5000$ with difference S/D.

The drag coefficient of a circular cylinder with two passive controls at the front and in the rear. Passive control in front of the circular cylinder is the cylinder of type- $I$, while the passive control behind the circular cylinder is a small circular cylinder. The data on the drag coefficient with the Reynolds number $\mathrm{Re}=5000$ and the configuration as above, can be seen in Table 2. It appears that the passive control behind has a significant effect on the drag coeffcient, since the drag coeffcient is still smaller than that without passive control. The minimum drag coefficient of the configuration is 0.916 , this occurs at S/D $=1.8$ and $\mathrm{T} / \mathrm{D}=1.5$.

### 3.2 Mathematical Model

In this case, the simulation result of the drag coefficient with two passive controls is interpolated to obtain the mathematical model. By using the bilinear interpolation approach one can make a mathematical model of the drag coefficient. Bilinear interpolation is the development of linear interpolation of two variables [13]. In this study we use the 2nd order bilinear interpolation. In this case, the variables used are $(x, y)=(T / D, S /$ D). By taking the nine points of drag data that have been obtained from the simulation results in Table 2 we can get the interpolation formulation. Therefore, nine polynomial equations and nine unknown coefficients can be obtained. The polynomial interpolation function can be written as follows:

$$
\begin{equation*}
f(x, y)=a_{00}+a_{01} y+a_{02} y^{2}+a_{10} x+a_{11} x y+a_{12} x y^{2}+a_{20} x^{2}+a_{21} x^{2} y+a_{22} x^{2} y^{2} \tag{7}
\end{equation*}
$$

Taking data from Table 3 and substituting ( $x, y$ ) $=(T / D, S / D)$ into $f(x, y)$ we find the unknown coefficients. Therefore, we can obtain the mathematical model of the drag coefficient as follows:

$$
\begin{align*}
E(x, y)= & 0.0275 x^{2} y^{2}-0.0240 x^{2} y+0.0590 x^{2}-0.1713 x y^{2}+0.1792 x y-0.2958 x \\
& +5.5913 y^{2}-0.848 y+1.5542 \tag{8}
\end{align*}
$$

| $f(T / D, S / D)$ | $\mathrm{S} / \mathrm{D}$ |  |  |
| :---: | :---: | :---: | :---: |
| $\mathrm{T} / \mathrm{D}$ | 0.6 | 1.8 | 3.0 |
| 0.6 | 1.116 | 0.973 | 0.987 |
| 1.2 | 1.169 | 0.977 | 0.986 |
| 1.8 | 1.557 | 1.225 | 1.195 |

Table 3: Nine drag data results for a circular cylinder

The error in the above mathematical model is calculated using an absolute error as follows:

$$
\begin{equation*}
e(x, y)=|E(x, y)-f(x, y)| . \tag{9}
\end{equation*}
$$



Figure 3: Graphic plot of Table 2 in Matlab.


Figure 4: Comparison of the graphic plot of Table 2 and the graphic plot of Table 3.

If we simulate the original data in Table 2 as shown in Figure 3 and compare it with the simulation result by using bilinear interpolation shown in Figure 4 then it appears that the smallest absolute error is $\mathrm{S} / \mathrm{D}=0.6,1.8,3.0, \mathrm{~T} / \mathrm{D}=0.6,1.2,1.8$ with the value of $\mathrm{Cd}=1.116,1.169,1.557,0.973,0.977,1.225,0.987,0.986,1.195$ and also the obtained largest absolute error is in $\mathrm{S} / \mathrm{D}=2.4, \mathrm{~T} / \mathrm{D}=1.5$, with the value of $\mathrm{Cd}=$ 1.265. In other words, the error will not exceed the point of 0.2282 .

### 3.3 Wake

In this study, the velocity data at the distance $6 \mathrm{D}, 8.5 \mathrm{D}$ and 11 D from the center of the circular cylinder or at the distance 10D , 12.5D and 15D from the front of the system, are shown in Figure 1. In both passive controls with $\operatorname{Re}=5000$ there is a wake. Also, it can be seen for the drag coefficient of the main circular cylinder that there is a significant decrease. In this case there is a decrease in the drag coefficient which affects the magnitude of the average velocity behind the circular cylinder.

It appears that Table 4 shows that a decrease in flow velocity behind the circular
cylinder correspons to the decrease in the drag coefficient. In addition, the flow velocity near the circular cylinder (i.e., 6D from the center of the circular cylinder) will increase as the distance moves farther away from the center of the circular cylinder and will return equally to the speed without passive control.

| Re | Single | 1 PC | \% | 2 PC | \% |
| :---: | :---: | :---: | :---: | :---: | :---: |
| 5000 | 1.51 | 1.216 | 19.47 | 0.916 | 39.34 |

Table 4: $C_{D}$ for $R_{e}=5000$.

## 4 Conclusion

By using the bilinear interpolation approach one can make a mathematical model of the drag coefficient. Bilinear interpolation is the development of linear interpolation of two variables. In this study we use the 2 nd order bilinear interpolation. Thus, a mathematical model can be formed for $C_{d}$ of a circular cylinder using two passive controls at $\mathrm{Re}=$ 5000. The mathematical model can be written as follows:

$$
\begin{align*}
E(x, y)= & 0.0275 x^{2} y^{2}-0.0240 x^{2} y+0.0590 x^{2}-0.1713 x y^{2}+0.1792 x y-0.2958 x \\
& +5.5913 y^{2}-0.848 y+1.5542 \tag{10}
\end{align*}
$$

In addition, a reduction of the drag coefficient in a circular cylinder can be done by adding passive control. Passive control can be placed in front and/or behind. The drag coefficient can be reduced by up to $40 \%$ if using passive control in the form of the cylinder of type- $I$ and an ellipse-shaped cylindrical back control rather than a passive control drag coefficient. This is reinforced by the decreasing flow velocity behind the circular cylinder.

## Acknowledgment

We appreciate the support of the head of the Research and Community Service of the Institute of Technology Sepuluh Nopember (LPPM ITS) Surabaya, Indonesia, on behalf of the Ministry of Technology and Higher Education (Kemenristekdikti), who has given us a research grant under the contract No. 010/SP2H/LT/DRPM/IV/2017, April 20th, 2017, and the opportunity to disseminate our research by this paper.

## References

[1] [online] Available : http://ariantiyoulie.blogspot.co.id/2014/01/01. December, 2017.
[2] Widodo, B. The Influence of Hydrodynamics on Pollutant Dispersion in the River. International Journal of Contemporary Mathematical Sciences (IJCMS) ISSN 1312-7586. HIKARI Ltd Journals and Books Publisher, Bulgaria, 2012, 2229-2234.
[3] Ladjedel, A.O. Yahiaoui, B.T. Adjlout, C.L. and Imine, D.O. Experimental and Numerical Studies of Drag Reduction on Circular Cylinder. World Academy of Sciences. Engineering and Technology, Bulgaria, 2011, 357-361.
[4] Igarashi, T. and Shiba, Y. Drag Reduction for D-Shape and I-Shape Cylinders (Aerodynamics Mechanism of ReductionDrag). JSME International Journal (Series B). 2006, 1036-1042.
[5] Triyogi, Y. and Wawan, A.W. Flow Characteristics Around a D-Type Cylinder Near a Plane Wall. In: Regional Conferences on Mechanical and Aerospace Technology. Bali, 2010.
[6] Bouak, F. and Lemay, J. Passive Control of the Aerodynamic Forces Acting on a Circular Cylinder. Experimental Thermal and Fluid Science (1998) 112-121.
[7] Tsutsui, T. and Igarashi, T. Drag Reduction of a Circular Sylinder in an Air-Stream. Journal of Wind Engineering and Industrial Aerodynamics (2002) 527-541.
[8] Triyogi Y. and Nuh, M. Using of a Bluff Body Cut from a Circular Cylinder as Passive Control to Reduce Aerodynamics Forces on a Circular Cylinder. In: The International Conference on Fluid and Thermal energy Conversion. Bali, Indonesia, 2003.
[9] Lee, Sang-Joon. Lee, Sang-Ik. and Park, Cheol-Woo. Reducing the Drag on a Circular Cylinder by Upstream Installation of a Small Control Rod. Fluid Dynamic Research (2004) 233-250.
[10] Imron, C. and Yunus, M. Mathematical Modeling of Pressure on Cylindrical Ellipse Using Side-by-Side Configuration. International Journal of Computing Science and Applied Mathematics 1 (1) (2015) 10-11.
[11] Imron, C. and Apriliani, E. Effect of Major Axis Length to the Pressure on Ellips. International Journal of Computing Science and Applied Mathematics 2 (3) (2016) 41-44.
[12] Widodo, B. Yuwono, T. and Imron, C. The influence of distance between passive control and circular cylinder on wake. Journal of Physics Conference Series 890 (1) (2017) 012053.
[13] The Wikipedia website [Online]. Available: https://en.wikipedia.org/wiki/Bilinear ${ }_{i}$ nter polation. 29 October, 2017.
[14] Widodo, B. Yuwono, T. and Imron, C. The effectiveness of passive control to reduce the drag coefficient. Journal of Physics Conference Series 890 (1) (2017) 012044.


# A Variety of New Solitary-Solutions for the Two-mode Modified Korteweg-de Vries Equation 

A. Jaradat ${ }^{1}$, M.S.M. Noorani ${ }^{1}$, M. Alquran ${ }^{2 *}$, and H.M. Jaradat ${ }^{3}$<br>${ }^{1}$ School of Mathematical Sciences, University of Kebangsaan Malaysia, Bangi, Malaysia<br>${ }^{2}$ Department of Mathematics and Statistics, Jordan University of Science and Technology, P.O.Box (3030), Irbid (22110), Jordan<br>${ }^{3}$ Department of Mathematics, Al al-Bayt University, Mafraq, Jordan

Received: October 13, 2018; Revised: December 12, 2018


#### Abstract

In this paper, we studied the nonlinear two-mode modified Korteweg-de Vries (TMmKdV) equation. We derived multiple singular soliton solutions to this new version of KdV equation by using the simplified form of Hirota's direct method. Also, kink and periodic solutions are extracted by using the tanh-expansion and the sine-cosine function methods. Finally, graphical analysis is conducted to show some physical features regarding TMmKdV equation.


Keywords: two-mode mKdV; Hirota bilinear method; sine-cosine function method; multiple singular solutions; kink and periodic solutions.

Mathematics Subject Classification (2010): 35C08, 74J35.

## 1 Introduction

Sergei V. Korsunsky [1] was the first who established the nonlinear two-mode Kortewegde Vries (TMKdV) equation which reads

$$
\begin{align*}
& w_{t t}+\left(a_{1}+a_{2}\right) w_{x t}+a_{1} a_{2} w_{x x}+\left(\left(\lambda_{1}+\lambda_{2}\right) \frac{\partial}{\partial t}+\left(\lambda_{1} a_{2}+\lambda_{2} a_{1}\right) \frac{\partial}{\partial x}\right) w w_{x}  \tag{1}\\
& +\left(\left(\mu_{1}+\mu_{2}\right) \frac{\partial}{\partial t}+\left(\mu_{1} a_{2}+\mu_{2} a_{1}\right) \frac{\partial}{\partial x}\right) w_{x x x}
\end{align*}
$$

where $w(x, t)$ is a field function representing the height of the free water surface above a flat bottom, $a_{1}$ and $a_{2}$ are the phase velocities, $\mu_{1}$ and $\mu_{2}$ are the dispersion parameters,

[^7]$\lambda_{1}$ and $\lambda_{2}$ are the parameters of nonlinearity.
The modified Korteweg-de Vries (mKdV) equation for the one-dimensional propagation of solitary waves in a fluid is given by
\[

$$
\begin{equation*}
w_{t}+\alpha w_{x x x}+\beta w^{2} w_{x}=0 \tag{2}
\end{equation*}
$$

\]

which is a generalized model in ocean dynamics, nonlinear lattice and plasma physics. In this paper we reconstruct and study the two-mode modified Korteweg-de Vries equation which describes the propagation of two wave modes of the same orientation. Now, the two-mode modified Korteweg-de Vries (TMmKdV) equation in a scaled-form reads

$$
\begin{align*}
& w_{t t}+\left(a_{1}+a_{2}\right) w_{x t}+a_{1} a_{2} w_{x x}+\left(\beta\left(\lambda_{1}+\lambda_{2}\right) \frac{\partial}{\partial t}+\beta\left(\lambda_{1} a_{2}+\lambda_{2} a_{1}\right) \frac{\partial}{\partial x}\right) w^{2} w_{x}  \tag{3}\\
& +\left(\alpha\left(\mu_{1}+\mu_{2}\right) \frac{\partial}{\partial t}+\alpha\left(\mu_{1} a_{2}+\mu_{2} a_{1}\right) \frac{\partial}{\partial x}\right) w_{x x x}
\end{align*}
$$

where $a_{1}, a_{2}, \lambda_{1}, \lambda_{2}, \mu_{1}, \mu_{2}$ are some real numbers, $w(x, t)$ is a field function, $a_{1}$ and $a_{2}$ are the phase velocities, $\mu_{1}$ and $\mu_{2}$ are the dispersion parameters, $\lambda_{1}$ and $\lambda_{2}$ are the parameters of nonlinearity. Note that $a_{1}, a_{2}$ are considered to be distinct and $x, t$ $\in(-\infty, \infty)$. Now we suggest the changes of variable by using the transformations [1-5]:

$$
\begin{aligned}
T & =\left(\mu_{1}+\mu_{2}\right)^{\frac{-1}{2}} t \\
X & =\left(\mu_{1}+\mu_{2}\right)^{\frac{-1}{2}}\left(x-a_{0} t\right) \\
a_{0} & =\frac{a_{1}+a_{2}}{2} \\
W & =\left(\lambda_{1}+\lambda_{2}\right)^{\frac{1}{2}} w
\end{aligned}
$$

Therefore, equation (3) reduces to TMmKdV equation in a scaled form as

$$
\begin{equation*}
W_{T T}-a^{2} W_{X X}+\left(\beta \frac{\partial}{\partial T}-\beta \lambda a \frac{\partial}{\partial X}\right) W^{2} W_{x}+\left(\alpha \frac{\partial}{\partial T}-\alpha \mu a \frac{\partial}{\partial X}\right) W_{X X X} \tag{4}
\end{equation*}
$$

where

$$
\begin{aligned}
a & =\frac{a_{1}-a_{2}}{2} \\
\lambda & =\frac{\lambda_{2}-\lambda_{1}}{\lambda_{2}+\lambda_{1}},|\lambda| \leq 1 \\
\mu & =\frac{\mu_{2}-\mu_{1}}{\mu_{2}+\mu_{1}},|\mu| \leq 1
\end{aligned}
$$

where $|\lambda| \leq 1,|\mu| \leq 1$ and $a$ is defined above. Note that when $a=0$, by integrating with respect to $t$, the two-mode modified Korteweg-de Vries equation (4) is reduced to the standard modified Korteweg-de Vries equation (2).

Finally, for more details about generating two-mode equations and physical features such models possess, we recommend for the readers the following references [6-14].

## 2 Multiple Soliton Solutions

In this section, we apply the simplified bilinear method [15-20], to find single soliton solutions and multiple soliton solutions for TMmKdV equation. First, we substitute

$$
W(X, T)=e^{\varepsilon_{i}}, \quad \varepsilon_{i}(X, T)=h_{i} X-\omega_{i} T
$$

into the linear terms of (4) and solve the resulting equation to obtain the dispersion relation

$$
\begin{equation*}
\omega_{i}=\frac{\alpha h_{i}^{3} \pm h_{i} \sqrt{\alpha^{2} h_{i}^{4}+4 \alpha \mu a h_{i}^{2}+4 a^{2}}}{2} \tag{5}
\end{equation*}
$$

As a result $\varepsilon_{i}$ becomes

$$
\begin{equation*}
\varepsilon_{i}(X, T)=h_{i} X-\frac{\alpha h_{i}^{3} \pm h_{i} \sqrt{\alpha^{2} h_{i}^{4}+4 \alpha \mu a h_{i}^{2}+4 a^{2}}}{2} T, \quad i=1,2, \ldots \tag{6}
\end{equation*}
$$

Second, we propose the solutions of (4) in the form

$$
\begin{equation*}
W(X, T)=R\left(\arctan \left(\frac{h(X, T)}{k(X, T)}\right)\right)_{X}=R \frac{h_{X} k-k_{X} h}{h^{2}+k^{2}} . \tag{7}
\end{equation*}
$$

The auxiliary functions $h(X, T)$ and $k(X, T)$ for single-soliton solution are given by

$$
\left\{\begin{array}{c}
h(X, T)=e^{\varepsilon_{1}(X, T)}=e^{h_{1} X-\frac{\alpha h_{1}^{3} \pm h_{1} \sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}}{2}} T  \tag{8}\\
k(X, T)=1 .
\end{array}\right.
$$

Substituting (7) and (8) into (4) and solving for $R$, we get

$$
\begin{equation*}
R= \pm 2 \sqrt{\frac{6 \alpha}{\beta}} . \tag{9}
\end{equation*}
$$

Under the constraint condition $\lambda=\mu$, the single soliton solution is given by

$$
\begin{align*}
W(X, T) & =2 h_{1} \sqrt{\frac{6 \alpha}{\beta}} \frac{e^{h_{1} X-\frac{\alpha h_{1}^{3} \pm h_{1}}{\sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}}} T}{1+e^{2 h_{1} X-\left(\alpha h_{1}^{3} \pm h_{1} \sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}\right) T}} \\
& =h_{1} \sqrt{\frac{6 \alpha}{\beta}} \operatorname{sech}\left(\varepsilon_{1}(X, T)\right), \tag{10}
\end{align*}
$$

where

$$
\varepsilon_{1}(X, T)=h_{1} X-\frac{\alpha h_{1}^{3} \pm h_{1} \sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}}{2} T .
$$

To find the two-soliton solution, we assume

$$
\begin{align*}
h(X, T) & =e^{\varepsilon_{1}}+e^{\varepsilon_{2}} \\
& =e^{h_{1} X-\frac{\alpha h_{1}^{3} \pm h_{1} \sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}}{2}} T+e^{h_{2} X-\frac{\alpha h_{2}^{3} \pm h_{2} \sqrt{\alpha^{2} h_{2}^{4}+4 \alpha \mu a h_{2}^{2}+4 a^{2}}}{2}} T \\
k(X, T) & =1-c_{12} e^{h_{1} X-\frac{\alpha h_{1}^{3} \pm h_{1} \sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}}{2}} T+h_{2} X-\frac{\alpha h_{2}^{3} \pm h_{2} \sqrt{\alpha^{2} h_{2}^{4}+4 \alpha \mu a h_{2}^{2}+4 a^{2}}}{2} T . \tag{11}
\end{align*}
$$

Substituting (7) and (11) into (4) and solving for $c_{12}$, we see that the constraint condition of two soliton solutions exists only if $\lambda=\mu= \pm 1$ and the phase shift $c_{12}$ is obtained by

$$
\begin{equation*}
c_{12}=\frac{\left(h_{1}-h_{2}\right)^{2}}{\left(h_{1}+h_{2}\right)^{2}} \tag{12}
\end{equation*}
$$

and this can be generalized as

$$
\begin{equation*}
c_{i j}=\frac{\left(h_{i}-h_{j}\right)^{2}}{\left(h_{i}+h_{j}\right)^{2}}, \quad 1 \leq i<j \leq 3 \tag{13}
\end{equation*}
$$

To get the two-soliton solutions for (4), we substitute (11) and (12) into (7) and use $\lambda=\mu=1$. As a result, we get

$$
\begin{align*}
U(X, T) & =\frac{h_{1} e^{h_{1} X-r_{1} T}\left(1+\frac{\left(h_{1}-h_{2}\right)^{2}}{\left(h_{1}+h_{2}\right)^{2}} e^{2 h_{2} X-\left(\alpha h_{2}^{3} \pm\left(\alpha h_{2}^{3}+2 a\right)\right) T}\right) \sqrt{\frac{6 \alpha}{\beta}}}{\left(\frac{\left(h_{1}-h_{2}\right)^{2}}{\left(h_{1}+h_{2}\right)^{2}} e^{h_{1} X-r_{1} T+h_{2} X-r_{2} T}-1\right)^{2}+\left(e^{h_{1} X-r_{1} T}+e^{h_{2} X-r_{2} T}\right)^{2}} \\
& +\frac{h_{2} e^{h_{2} X-r_{2} T}\left(1+\frac{\left(h_{1}-h_{2}\right)^{2}}{\left(h_{1}+h_{2}\right)^{2}} e^{2 h_{1} X-\left(\alpha h_{1}^{3} \pm\left(\alpha h_{1}^{3}+2 a\right)\right) T}\right) \sqrt{\frac{6 \alpha}{\beta}}}{\left(\frac{\left(h_{1}-h_{2}\right)^{2}}{\left(h_{1}+h_{2}\right)^{2}} e^{h_{1} X-r_{1} T+h_{2} X-r_{2} T}-1\right)^{2}+\left(e^{h_{1} X-r_{1} T}+e^{h_{2} X-r_{2} T}\right)^{2}}, \tag{14}
\end{align*}
$$

where

$$
\begin{aligned}
& r_{1}=\frac{\alpha h_{1}^{3} \pm\left(\alpha h_{1}^{3}+2 a\right)}{2} \\
& r_{2}=\frac{\alpha h_{2}^{3} \pm\left(\alpha h_{2}^{3}+2 a\right)}{2}
\end{aligned}
$$

For the three-soliton solutions, we use

$$
\left\{\begin{array}{c}
h(X, T)=e^{\varepsilon_{1}}+e^{\varepsilon_{2}}+e^{\varepsilon_{3}}+c_{123} e^{\epsilon_{1}+\epsilon_{2}+\varepsilon_{3}}  \tag{15}\\
k(X, T)=1-c_{12} e^{\epsilon_{1}+\epsilon_{2}}-c_{13} e^{\epsilon_{1}+\varepsilon_{3}}-c_{23} e^{\epsilon_{2}+\varepsilon_{3}}
\end{array}\right.
$$

where $c_{i j}$ are given in (13). Substituting (7) and (15) into (4) and solving for $c_{123}$ under the constraint condition $\lambda=\mu= \pm 1$, we find

$$
c_{123}=c_{12} c_{13} c_{23}
$$

Finally, we reach to the fact that TMmKdV equation given in (4) has $N$-soliton solutions under the constraint condition $\lambda=\mu= \pm 1$ which can be obtained for finite $N$, where $N \geq 3$.

## 3 Singular Soliton Solutions

In this section we construct a multiple singular-soliton solution for (4) where the solution is assumed to be of the form

$$
\begin{equation*}
W(X, T)=R \ln \left(\frac{h(X, T)}{k(X, T)}\right)_{X}=R \frac{k h_{X}-h k_{X}}{k h} . \tag{16}
\end{equation*}
$$

The dispersion relation as in the previous section is given by

$$
\omega_{i}=\frac{\alpha h_{i}^{3} \pm h_{i} \sqrt{\alpha^{2} h_{i}^{4}+4 \alpha \mu a h_{i}^{2}+4 a^{2}}}{2}
$$

and hence $\varepsilon_{i}(X, T)=h_{i} X-\frac{\alpha h_{i}^{3} \pm h_{i} \sqrt{\alpha^{2} h_{i}^{4}+4 \alpha \mu a h_{i}^{2}+4 a^{2}}}{2} T, i=1,2, \ldots$.

For the singular one-soliton solution, we consider

$$
\begin{equation*}
h(X, T)=1+e^{\varepsilon_{1}(X, T)}, \quad k(X, T)=1-e^{\varepsilon_{1}(X, T)} . \tag{17}
\end{equation*}
$$

Substituting (17) and (16) into (4) and solving for $R$, we get

$$
\begin{equation*}
R= \pm \sqrt{-6 \alpha / \beta} \tag{18}
\end{equation*}
$$

Under the constraint condition $\lambda=\mu$, the single-soliton solution is given by

$$
\begin{align*}
W(X, T) & =-2 h_{1} \sqrt{\frac{-6 \alpha}{\beta}} \frac{e^{h_{1} X-\frac{\alpha h_{1}^{3} \pm h_{1} \sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}}{2}} T}{e^{2 h_{1} X-\left(\alpha h_{1}^{3} \pm h_{1} \sqrt{\alpha^{2} h_{1}^{4}+4 \alpha \mu a h_{1}^{2}+4 a^{2}}\right) T}-1} \\
& =-h_{1} \sqrt{\frac{-6 \alpha}{\beta}} \operatorname{csch}\left(\varepsilon_{1}(X, T)\right) . \tag{19}
\end{align*}
$$

To obtain singular two-soliton solution, we set

$$
\left\{\begin{array}{c}
h(X, T)=1+e^{\varepsilon_{1}(X, T)}+e^{\varepsilon_{2}(X, T)}+c_{12} e^{\varepsilon_{1}(X, T)+\varepsilon_{2}(X, T)}  \tag{20}\\
k(x, t)=1-e^{\varepsilon_{1}(X, T)}-e^{\varepsilon_{2}(X, T)}+c_{12} e^{\varepsilon_{1}(X, T)+\varepsilon_{2}(X, T)}
\end{array}\right.
$$

Substituting (18), (20) and (16) into (4) and solving for $c_{12}$ lead to the two soliton solutions only if $\lambda=\mu= \pm 1$ and the same phase shift $c_{12}$ obtained in (12) and hence $c_{i j}$ given by (13).

To construct the singular three-soliton solution, we set

$$
\begin{align*}
h(X, T) & =1+e^{\varepsilon_{1}(X, T)}+e^{\varepsilon_{2}(X, T)}+e^{\epsilon_{3}(X, T)}+c_{12} e^{\varepsilon_{1}(X, T)+\varepsilon_{1}(X, T)}+c_{23} e^{\varepsilon_{2}(X, T)+\varepsilon_{3}(X, T)} \\
& +c_{13} e^{\varepsilon_{1}(X, T)+\varepsilon_{3}(X, T)}+c_{123} e^{\varepsilon_{1}(X, T)+\varepsilon_{2}(X, T)+\varepsilon_{3}(X, T)} \\
k(X, T) & =1-e^{\varepsilon_{1}(X, T)}-e^{\varepsilon_{2}(X, T)}-e^{\epsilon_{3}(X, T)}+c_{12} e^{\varepsilon_{1}(X, T)+\varepsilon_{1}(X, T)}+c_{23} e^{\varepsilon_{2}(X, T)+\varepsilon_{3}(X, T)} \\
& +c_{13} e^{\varepsilon_{1}(X, T)+\varepsilon_{3}(X, T)}-c_{123} e^{\varepsilon_{1}(X, T)+\varepsilon_{2}(X, T)+\varepsilon_{3}(X, T)} . \tag{21}
\end{align*}
$$

Repeating the same previous steps, we reach to the same fact that the three single-soliton solutions exists only under the constraint condition $\lambda=\mu= \pm 1$.

## 4 Solitary Ansatze Methods

In this part, we introduce in brief two methods, the tanh-technique and the sine-cosine function method to solve the problem (4).

### 4.1 The tanh method

The tanh technique [21-26] suggests the following solution

$$
\begin{equation*}
W(\zeta)=S(Y)=\sum_{i=0}^{M} b_{i} Y^{i} \tag{22}
\end{equation*}
$$

where $Y=\tanh (\delta \zeta)$. The index $M$ can be determined by a balance procedure. Once we have $M$, we collect all coefficients of powers of $Y$ in the resulting equation and set them to zero. Finally, we solve the obtained algebraic system to retrieve the values of the required coefficients $b_{i}$.

Now, we consider a new variable $\zeta=X-\gamma T$ to reduce (4) into the following differential equation

$$
\begin{equation*}
\left(\gamma^{2}-a^{2}\right) W-\frac{\beta}{3}(\gamma+\lambda a) W^{3}-\alpha(\gamma+\mu a) W^{\prime \prime}=0 \tag{23}
\end{equation*}
$$

where $W=W(\zeta)$ and the prime denotes the ordinary derivative. By a blanching procedure for equation (23), the value of the parameter $M$ is equal to 1 and thus $W(\zeta)=A+B \tanh (\delta \zeta)$. Substituting this proposed solution in (23) yields the following algebraic system:

$$
\begin{align*}
& 0=-A a^{2}-\frac{1}{3} \lambda A^{3} a \beta-\frac{1}{3} A^{3} \beta \gamma+A \gamma^{2} \\
& 0=-B a^{2}-\lambda A^{2} B a \beta-A^{2} B \beta \gamma+B \gamma^{2}+2 \mu B a \alpha \delta^{2}+2 B \alpha \gamma \delta^{2} \\
& 0=-\lambda A B^{2} a \beta-A B^{2} \beta \gamma \\
& 0=-\frac{1}{3} \lambda B^{3} a \beta-\frac{1}{3} B^{3} \beta \gamma-2 \mu B a \alpha \delta^{2}-2 B \alpha \gamma \delta^{2} \tag{24}
\end{align*}
$$

Solving the above system produces the following two-wave solution

$$
\begin{equation*}
W(X, T)= \pm \frac{\sqrt{-6 \alpha \delta^{2}((-1+\lambda \mu) a+(\lambda-\mu) \gamma)}}{\sqrt{\beta\left(\left(-1+\lambda^{2}\right) a+2(-\lambda+\mu) \alpha \delta^{2}\right)}} \tanh (\delta(X-\gamma T)) \tag{25}
\end{equation*}
$$

with $\gamma=\left(-\alpha \delta^{2} \pm \sqrt{a^{2}-2 \mu a \alpha \delta^{2}+\alpha^{2} \delta^{4}}\right)$. If the tanh-function is replaced by cothfunction in (25), a new solution will be obtained.

### 4.2 The sine-cosine method

The sine-cosine technique $[24,25,27-31]$ assumes the solution of (23) in the form of

$$
\begin{equation*}
W(\zeta)=A \sin ^{B}(\delta \zeta) \tag{26}
\end{equation*}
$$

or

$$
\begin{equation*}
W(\zeta)=A \cos ^{B}(\delta \zeta) \tag{27}
\end{equation*}
$$

To determine the values of $A, B, \gamma$ and $\delta$, we substitute (26) in (23) to get

$$
\begin{align*}
0 & =\left(A \mu B a \alpha \delta^{2}-A \mu B^{2} a \alpha \delta^{2}+A B \alpha \gamma \delta^{2}-A B^{2} \alpha \gamma \delta^{2}\right) \sin ^{B-2}(\delta z)  \tag{28}\\
& -\left(A a^{2}+A \gamma^{2}+A \mu B^{2} a \alpha \delta^{2}+A B^{2} \alpha \gamma \delta^{2}\right) \sin ^{B}(\delta z)-\left(\frac{1}{3} \lambda A^{3} a \beta-\frac{1}{3} A^{3} \beta \gamma\right) \sin ^{3 B}(\delta z)
\end{align*}
$$

Now, equating the exponents $B-2$ and $3 B$ in (28) and setting the coefficients of same power to zero, produce the following two-wave solution

$$
\begin{equation*}
W(X, T)=\frac{\sqrt{-6 \alpha \delta^{2}((-1+\lambda \mu) a+(\lambda-\mu) \gamma)}}{\sqrt{\beta\left(\left(-1+\lambda^{2}\right) a+(-\lambda+\mu) \alpha \delta^{2}\right)}} \csc (\delta(X-\gamma T)) \tag{29}
\end{equation*}
$$

with $\gamma=\frac{1}{2}\left(-\alpha \delta^{2} \pm \sqrt{4 a^{2}-4 \mu a \alpha \delta^{2}+\alpha^{2} \delta^{4}}\right)$.
Finally, by using the cosine-function method (27), another two-wave solution will be obtained being the same as given in (29) but with csc replaced by sec.

## 5 Numerical Example

In this section, we study some physical features of the solution of TMmKdV equation given in (25). In Figure 1, increasing the phase velocity $a$ leads to a gradual increase in the space between the two-waves of TMmKdV equation. In Figure 2, decreasing of the nonlinearity parameter $\lambda$ leads to interaction of the two-waves of the TMmKdV equation.


Figure 1: Behaviors of two-waves in (25) at the increasing phase velocity: $a=1,3,5$ respectively. The assigned values for the other parameters are $\delta=\gamma=1, \lambda=\frac{1}{2}, \mu=\frac{1}{4}, \alpha=-1$, $\beta=1$.


Figure 2: Behaviors of two-waves in (25) at the decreasing nonlinearity parameter: $\lambda=$ $-\frac{1}{2}, 0, \frac{1}{2}$ respectively. The assigned values for the other parameters are $\delta=\gamma=1, s=1$, $\mu=\frac{1}{4}, \alpha=-1, \beta=1$.

## 6 Conclusion

In this paper we studied the solutions of the scaled TMmKdV equation which reads

$$
W_{T T}-a^{2} W_{X X}+\left(\beta \frac{\partial}{\partial T}-\beta \lambda a \frac{\partial}{\partial X}\right) W^{2} W_{x}+\left(\alpha \frac{\partial}{\partial T}-\alpha \mu a \frac{\partial}{\partial X}\right) W_{X X X}
$$

We used three different methods, the simplified bilinear method, the tanh-technique and the sine-cosine function method. The following findings are observed in this work.

- When $\lambda=\mu= \pm 1$, TMmKdV equation admits multiple-soliton solutions by means of the simplified bilinear method.
- For arbitrary $\lambda$ and $\mu$, periodic solutions are obtained for TMmKdV equation by using the sine-cosine method.
- For arbitrary $\lambda$ and $\mu$, kink solutions are obtained for TMmKdV equation by using the tanh-expansion method.


## Acknowledgment

Authors would like to thank the editor and the anonymous referees for their in-depth reading and insightful comments on an earlier version of this paper.

## References

[1] Korsunsky, S.V. Soliton solutions for a second-order KdV equation. Phys. Lett. A. 185 (1994) 174-176.
[2] Lee, C.T. and Liu, J.L. A Hamiltonian model and soliton phenomenon for a two-mode KdV equation. Rocky Mountain J. Math. 41 (4) (2011) 1273-1289.
[3] Lee, C.C., Lee, C.T., Liu, J.L. and Huang, W.Y. Quasi-solitons of the two-mode Kortewegde Vries equation. Eur. Phys. J. Appl. Phys. 52 (2010) 11301.
[4] Wazwaz, A.M. Multiple soliton solutions and other exact solutions for a two-mode KdV equation. Math. Methods Appl. Sci. 40 (6) (2017) 1277-1283.
[5] Wazwaz, A.M. A Two-Mode Burgers Equation of Weak Shock Waves in a Fluid: Multiple Kink Solutions and Other Exact Solutions. Int. J. Appl. Comput. Math. (2016), DOI 10.1007/s40819-016-0302-4.
[6] Wazwaz, A.M. Two-mode Sharma-Tasso-Olver equation and two-mode fourth-order Burgers equation: Multiple kink solutions. Alexandria Eng. J. (2017), https://doi.org/10.1016/j.aej.2017.04.003.
[7] Alquran, M. and Jarrah, A. Jacobi elliptic function solutions for a two-mode KdV equation. Journal of King Saud University-Science (2017), https://doi.org/10.1016/j.jksus.2017.06.010.
[8] Syam, M., Jaradat, H.M. and Alquran, M. A study on the two-mode coupled modified Korteweg-de Vries using the simplified bilinear and the trigonometric-function methods. Nonlinear Dynamics 90 (2) (2017) 1363-1371.
[9] Jaradat, H.M., Syam, M. and Alquran, M. A two-mode coupled Korteweg-de Vries: multiple-soliton solutions and other exact solutions. Nonlinear Dynamics 90 (1) (2017) 371-377.
[10] Alquran, M. Jaradat, H.M. and Syam, M. A modified approach for a reliable study of new nonlinear equation: two-mode Korteweg-de Vries-Burgers equation. Nonlinear Dynamics 91 (3) (2018) 1619-1626.
[11] Jaradat, I., Alquran, M. and Ali, M. A numerical study on weak-dissipative two-mode perturbed Burgers and Ostrovsky models: right-left moving waves. Eur. Phys. J. Plus. 133: 164 (2018).
[12] Jaradat, I., Alquran, M., Momani, S. and Biswas, A. Dark and singular optical solutions with dual-mode nonlinear Schrodinger's equation and Kerr-law nonlinearity. Optik 172 (2018) 822-825.
[13] Alquran, M. and Yassin, O. Dynamism of two-modes parameters on the field function for third-order dispersive Fisher: application for fibre optics. Optical and Quantum Electronics 50 (9) (2018): 354.
[14] Yassin, O. and Alquran, M. Constructing new Solutions for some types of two-mode nonlinear equations. Appl. Math. Inf. Sci. 12 (2) (2018) 361-367.
[15] Wazwaz, A.M. Gaussian solitary wave solutions for nonlinear evolution equations with logarithmic nonlinearities. Nonlinear Dynamics 83 (2016) 591-596.
[16] Wazwaz, A.M. Multiple kink solutions for two coupled integrable $(2+1)$-dimensional systems. Applied Mathematics Letters 58 (2016) 1-6.
[17] Hirota, R. Exact solutions of the Korteweg-de Vries equation for multiple collisions of solitons. Phys. Rev. Lett. 27 (1971) 1192-1194.
[18] Hirota R., Exact N-soliton solutions of a nonlinear wave equation. J. Math. Phys. 14 (1973) 805-809.
[19] Jaradat, H.M, Awawdeh, F., Al-Shara', S., Alquran, M. and Momani, S. Controllable dynamical behaviors and the analysis of fractal burgers hierarchy with the full effects of inhomogeneities of media. Romanian Journal of Physics 60 (3-4) (2015) 324-343.
[20] Alquran, M., Jaradat, H.M., Al-Shara', S. and Awawdeh, F. A New Simplified Bilinear Method for the N-Soliton Solutions for a Generalized FmKdV Equation with TimeDependent Variable Coefficients. International Journal of Nonlinear Sciences and Numerical Simulation 16 (6) (2015) 259-269.
[21] Wazwaz, A.M. A variety of distinct kinds of multiple soliton solutions for a (3+1)dimensional nonlinear evolution equation. Math. Meth. Appl. Sci. 36 (3) (2013) 349-357.
[22] Maliet, W. and Heremanm, W. The tanh method: I. Exact solutions of nonlinear evolution and wave equations. Physica Scripta 54 (1996) 563-568.
[23] Alquran, M. and Al-Khaled, K. Mathematical methods for a reliable treatment of the $(2+1)$-dimensional Zoomeron equation. Mathematical Sciences 6 (12) (2012).
[24] Alquran, M. and Al-Khaled, K. The tanh and sine-cosine methods for higher order equations of Korteweg-de Vries type. Physica Scripta 84(2011) 025010 (4pp).
[25] Alquran, M. and Al-Khaled, K. Sinc and solitary wave solutions to the generalized Benjamin-Bona-Mahony-Burgers equations. Physica Scripta 83(2011) 065010 (6pp).
[26] Wazwaz, A.M. Peakons and soliton solutions of newly developed Benjamin-Bona-Mahonylike equations. Nonlinear Dynamics and Systems Theory 15 (2) (2015) 209-220.
[27] Alquran, M., Jarrah, A. and Krishnan, E.V. Solitary wave solutions of the phi-four equation and the breaking soliton system by means of Jacobi elliptic sine-cosine expansion method. Nonlinear Dynamics and Systems Theory 18 (3) (2018) 233-240.
[28] Alquran, M. Solitons and periodic solutions to nonlinear partial differential equations by the sine-cosine method. Appl. Math. Inf. Sci. 6 (1) (2012) 85-88.
[29] Ali, A.H.A., Soliman, A.A. and Raslan, K.R. Soliton solution for nonlinear partial differential equations by cosine-function method. Phys. Lett. A. 368 (2007) 299-304.
[30] Alquran, M. and Qawasmeh, A. Classifications of solutions to some generalized nonlinear evolution equations and systems by the sine-cosine method. Nonlinear Studies 20 (2) (2013) 261-270.
[31] Alquran, M., Al-Omary, R. and Katatbeh, Q. New explicit solutions for homogeneous KdV equations of third order by trigonometric and hyperbolic function methods. Applications and Applied Mathematics AAM 7 (1) (2012) 211-225.

# Existence of Renormalized Solutions for Some Strongly Parabolic Problems in Musielak-Orlicz-Sobolev Spaces 

A. Talha*, A. Benkirane and M.S.B. Elemine Vall<br>Laboratory LAMA, Faculty of Sciences Dhar El Mahraz University Sidi Mohamed Ben Abdellah, P.O. Box 1796, Atlas Fez, Morocco

Received: December 19, 2016; Revised: December 13, 2018


#### Abstract

In this work, we prove an existence result of renormalized solutions in Musielak-Orlicz-Sobolev spaces for a class of nonlinear parabolic equations with two lower order terms and $L^{1}$-data.


Keywords: parabolic problems, Musielak-Orlicz space, renormalized solutions.
Mathematics Subject Classification (2010): 46E35, 35K15, 35K20, 35K60.

## 1 Introduction

We consider the following nonlinear parabolic problem:

$$
(\mathcal{P})\left\{\begin{array}{l}
\frac{\partial u}{\partial t}-\operatorname{div}(a(x, t, u, \nabla u)+\Phi(u))+g(x, t, u, \nabla u)=f \quad \text { in } Q \\
u=0 \quad \text { on } \partial Q=\partial \Omega \times[0, T] \\
u(x, 0)=u_{0} \quad \text { on } \quad \Omega
\end{array}\right.
$$

where $A(u)=-\operatorname{div}(a(x, t, u, \nabla u))$ is an operator of Leray-Lions type, the lower order term $\Phi \in C^{0}\left(\mathbb{R}, \mathbb{R}^{N}\right), g$ is a nonlinearity term which satisfies the growth and the sign condition and the data $f$ belong to $L^{1}(Q)$. Under these assumptions the term $\operatorname{div}(\Phi(u))$ may not exist in the distributions sense, since the function $\Phi(u)$ does not belong to $\left(L_{\mathrm{loc}}^{1}(Q)\right)^{N}$.

In the setting of classical Sobolev spaces, the existence of a weak solution for the problem $(\mathcal{P})$ has been proved in [10] in the case of $\Phi \equiv g \equiv 0$. It is well known that this weak solution is not unique in general (see [16] for a counter-example in the stationary case).

[^8]In order to obtain well-posedness for this type of problems the notion of renormalized solution has been introduced by Lions and DiPerna [12] for the study of Boltzmann equation (see also Lions [13] for a few applications to fluid mechanics models). This notion was then adapted to the elliptic version by Boccardo et al. [11]. At the same time, the equivalent notion of entropy solutions has been developed independently by Bénilan et al. [5] for the study of nonlinear elliptic problems.

The existence and uniqueness of a renormalized solution has been proved by D. Blanchard and F. Murat [8] in the case where $a(x, t, s, \xi)$ is independent of $s$, with $\Phi \equiv 0$ and $g \equiv 0$, by D. Blanchard, F. Murat and H. Redwane [9] with the large monotonicity on $a$. For measure data, $u=b(x, u)$ and $\Phi \equiv 0$, the existence of renormalized solution for the problem ( $\mathcal{P}$ ) has been proved by Y. Akdim et al.[3] in the framework of weighted Sobolev space, by L. Aharouch, J. Bennouna and A. Touzani [1], and by A. Benkirane and J. Bennouna [6] in the Orlicz spaces and degenerated spaces.

In the Musielak framework, the existence of a weak solution for the problem ( $\mathcal{P}$ ) has been proved by M.L. Ahmed Oubeid, A. Benkirane and M. Sidi El Vally in [2] where $\Phi \equiv 0$, the existence of entropy solutions for the problem $(\mathcal{P})$ has been studied by A. Talha, A. Benkirane and M.S.B. Elemine Vall in [19].

As an example of equations to which the present result can be applied, we give

$$
\frac{\partial u}{\partial t}-\operatorname{div}\left(\frac{m(x,|\nabla u|)}{|\nabla u|} \cdot \nabla u+u|u|^{\sigma}\right)+\frac{\operatorname{sign}(u)}{1+u^{2}} \varphi(x,|\nabla u|)=f \in L^{1}(Q),
$$

where $m$ is the derivative of $\varphi$ with respect to $t$.

## 2 Preliminaries

### 2.1 Musielak-Orlicz-Sobolev spaces.

Let $\Omega$ be an open set in $\mathbb{R}^{N}$ and let $\varphi$ be a real-valued function defined in $\Omega \times \mathbb{R}_{+}$, and satisfying the following conditions:
a) $\varphi(x, \cdot)$ is an N-function,
b) $\varphi(\cdot, t)$ is a measurable function.

The function $\varphi$ is called a Musielak-Orlicz function. For a Musielak-orlicz function $\varphi$ we put $\varphi_{x}(t)=\varphi(x, t)$ and we associate its nonnegative reciprocal function $\varphi_{x}^{-1}$, with respect to $t$ that is $\varphi_{x}^{-1}(\varphi(x, t))=\varphi\left(x, \varphi_{x}^{-1}(t)\right)=t$. The Musielak-orlicz function $\varphi$ is said to satisfy the $\Delta_{2}$-condition if for some $k>0$ and a non negative function $h$ integrable in $\Omega$, we have

$$
\begin{equation*}
\varphi(x, 2 t) \leq k \varphi(x, t)+h(x) \text { for all } x \in \Omega \text { and } t \geq 0 \tag{1}
\end{equation*}
$$

When (1) holds only for $t \geq t_{0}>0$; then $\varphi$ is said to satisfy the $\Delta_{2}$-condition near infinity.

Let $\varphi$ and $\gamma$ be two Musielak-orlicz functions. We say that $\gamma$ grows essentially less rapidly than $\varphi$ at 0 (resp. near infinity), and we write $\gamma \prec \prec \varphi$, if for every positive constant $c$ we have

$$
\left.\lim _{t \longrightarrow 0}\left(\sup _{x \in \Omega} \frac{\gamma(x, c t)}{\varphi(x, t)}\right)=0, \quad \text { (resp. } \quad \lim _{t \longrightarrow \infty}\left(\sup _{x \in \Omega} \frac{\gamma(x, c t)}{\varphi(x, t)}\right)=0\right)
$$

We define the functional $\rho_{\varphi, \Omega}(u)=\int_{\Omega} \varphi(x,|u(x)|) d x$, where $u: \Omega \longrightarrow \mathbb{R}$ is a Lebesgue measurable function.

We define the Musielak-Orlicz space (the generalized Orlicz spaces) by

$$
L_{\varphi}(\Omega)=\left\{u: \Omega \longrightarrow \mathbb{R} \text { measurable } / \rho_{\varphi, \Omega}\left(\frac{|u(x)|}{\lambda}\right)<+\infty, \text { for some } \lambda>0\right\}
$$

For a Musielak-Orlicz function we put: $\psi(x, s)=\sup _{t \geq 0}\{s t-\varphi(x, t)\} . \psi$ is called the Musielak-Orlicz function complementary to $\varphi$ in the sense of Young with respect to the variable $s$. In the space $L_{\varphi}(\Omega)$ we define the following two norms:

$$
\|u\|_{\varphi, \Omega}=\inf \left\{\lambda>0 / \int_{\Omega} \varphi\left(x, \frac{|u(x)|}{\lambda}\right) d x \leq 1\right\}
$$

which is called the Luxemburg norm and the so-called Orlicz norm by

$$
\||u|\|_{\varphi, \Omega}=\sup _{\|v\|_{\psi} \leq 1} \int_{\Omega}|u(x) v(x)| d x
$$

where $\psi$ is the Musielak-Orlicz function complementary to $\varphi$. These two norms are equivalent [14]. The closure in $L_{\varphi}(\Omega)$ of the bounded measurable functions with compact support in $\bar{\Omega}$ is denoted by $E_{\varphi}(\Omega)$.

We say that a sequence of functions $u_{n} \in L_{\varphi}(\Omega)$ is modular convergent to $u \in L_{\varphi}(\Omega)$ if there exists a constant $k>0$ such that $\lim _{n \rightarrow \infty} \rho_{\varphi, \Omega}\left(\frac{u_{n}-u}{k}\right)=0$.

For any fixed nonnegative integer $m$ we define

$$
W^{m} L_{\varphi}(\Omega)=\left\{u \in L_{\varphi}(\Omega): \forall|\alpha| \leq m, D^{\alpha} u \in L_{\varphi}(\Omega)\right\}
$$

and

$$
W^{m} E_{\varphi}(\Omega)=\left\{u \in E_{\varphi}(\Omega): \forall|\alpha| \leq m, D^{\alpha} u \in E_{\varphi}(\Omega)\right\}
$$

where $\alpha=\left(\alpha_{1}, \ldots, \alpha_{n}\right)$ with nonnegative integers $\alpha_{i},|\alpha|=\left|\alpha_{1}\right|+\ldots+\left|\alpha_{n}\right|$ and $D^{\alpha} u$ denote the distributional derivatives. The space $W^{m} L_{\varphi}(\Omega)$ is called the Musielak-Orlicz-Sobolev space. Let

$$
\bar{\rho}_{\varphi, \Omega}(u)=\sum_{|\alpha| \leq m} \rho_{\varphi, \Omega}\left(D^{\alpha} u\right) \text { and }\|u\|_{\varphi, \Omega}^{m}=\inf \left\{\lambda>0: \bar{\rho}_{\varphi, \Omega}\left(\frac{u}{\lambda}\right) \leq 1\right\} .
$$

For $u \in W^{m} L \varphi(\Omega)$ these functionals are a convex modular and a norm on $W^{m} L_{\varphi}(\Omega)$, respectively, and the pair $\left(W^{m} L \varphi(\Omega),\| \|_{\varphi, \Omega}^{m}\right)$ is a Banach space if $\varphi$ satisfies the following condition [14] :

$$
\begin{equation*}
\text { there exists a constant } c>0 \text { such that } \inf _{x \in \Omega} \varphi(x, 1) \geq c \text {. } \tag{2}
\end{equation*}
$$

The space $W^{m} L_{\varphi}(\Omega)$ will always be identified with a subspace of the product $\prod_{|\alpha| \leq m} L_{\varphi}(\Omega)=\Pi L_{\varphi}$, this subspace is $\sigma\left(\Pi L_{\varphi}, \Pi E_{\psi}\right)$ closed. We denote by $\mathcal{D}(\Omega)$ the space of infinitely smooth functions with compact support in $\Omega$ and by $\mathcal{D}(\bar{\Omega})$ ) the restriction of $\mathcal{D}\left(\mathbb{R}^{N}\right)$ on $\Omega$. Let $W_{0}^{m} L_{\varphi}(\Omega)$ be the $\sigma\left(\Pi L_{\varphi}, \Pi E_{\psi}\right)$ closure of $\mathcal{D}(\Omega)$ in $W^{m} L_{\varphi}(\Omega)$. Let $W^{m} E_{\varphi}(\Omega)$ be the space of functions $u$ such that $u$ and its distribution derivatives up to order $m$ lie in $E_{\varphi}(\Omega)$, and $W_{0}^{m} E_{\varphi}(\Omega)$ is the (norm) closure of $\mathcal{D}(\Omega)$ in $W^{m} L_{\varphi}(\Omega)$.

The following spaces of distributions will also be used:

$$
W^{-m} L_{\psi}(\Omega)=\left\{f \in \mathcal{D}^{\prime}(\Omega) ; f=\sum_{|\alpha| \leq m}(-1)^{|\alpha|} D^{\alpha} f_{\alpha} \text { with } f_{\alpha} \in L_{\psi}(\Omega)\right\}
$$

and

$$
W^{-m} E_{\psi}(\Omega)=\left\{f \in \mathcal{D}^{\prime}(\Omega) ; f=\sum_{|\alpha| \leq m}(-1)^{|\alpha|} D^{\alpha} f_{\alpha} \text { with } f_{\alpha} \in E_{\psi}(\Omega)\right\}
$$

We say that a sequence of functions $u_{n} \in W^{m} L_{\varphi}(\Omega)$ is modular convergent to $u \in$ $W^{m} L_{\varphi}(\Omega)$ if there exists a constant $k>0$ such that $\lim _{n \rightarrow \infty} \bar{\rho}_{\varphi, \Omega}\left(\frac{u_{n}-u}{k}\right)=0$.

The inhomogeneous Musielak-Orlicz-Sobolev spaces of order 1 are defined as follows:

$$
W^{1, x} L_{\varphi}(Q)=\left\{u \in L_{\varphi}(Q): \forall|\alpha| \leq 1 D_{x}^{\alpha} u \in L_{\varphi}(Q)\right\}
$$

and

$$
W^{1, x} E_{\varphi}(Q)=\left\{u \in E_{\varphi}(Q): \forall|\alpha| \leq 1 D_{x}^{\alpha} u \in E_{\varphi}(Q)\right\}
$$

The last space is a subspace of the first one, and both are Banach spaces under the norm $\|u\|=\sum_{|\alpha| \leq m}\left\|D_{x}^{\alpha} u\right\|_{\varphi, Q}$. We have the following complementary system

$$
\left(\begin{array}{lc}
W_{0}^{1, x} L_{\varphi}(Q) & F \\
W_{0}^{1, x} E_{\varphi}(Q) & F_{0}
\end{array}\right),
$$

$F$ being the dual space of $W_{0}^{1, x} E_{\varphi}(Q)$. It is also, except for an isomorphism, the quotient of $\Pi L_{\psi}$ by the polar set $W_{0}^{1, x} E_{\varphi}(Q)^{\perp}$, and will be denoted by $F=W^{-1, x} L_{\psi}(Q)$ and it is shown that

$$
W^{-1, x} L_{\psi}(Q)=\left\{f=\sum_{|\alpha| \leq 1} D_{x}^{\alpha} f_{\alpha}: f_{\alpha} \in L_{\psi}(Q)\right\}
$$

This space will be equipped with the usual quotient norm $\|f\|=\inf \sum_{|\alpha| \leq 1}\left\|f_{\alpha}\right\|_{\psi, Q}$, where the inf is taken on all possible decompositions $f=\sum_{|\alpha| \leq 1} D_{x}^{\alpha} f_{\alpha}, \quad f_{\alpha} \in L_{\psi}(Q)$.

The space $F_{0}$ is then given by

$$
F_{0}=\left\{f=\sum_{|\alpha| \leq 1} D_{x}^{\alpha} f_{\alpha}: f_{\alpha} \in E_{\psi}(Q)\right\}=W^{-1, x} E_{\psi}(Q)
$$

Let us give the following lemma which will be needed later.
Lemma 2.1 [7]. Let $\Omega$ be a bounded Lipschitz domain in $\mathbb{R}^{N}$ and let $\varphi$ and $\psi$ be two complementary Musielak-Orlicz functions which satisfy the following conditions:
i) There exists a constant $c>0$ such that $\inf _{x \in \Omega} \varphi(x, 1) \geq c$,
ii) There exists a constant $A>0$ such that for all $x, y \in \Omega$ with $|x-y| \leq \frac{1}{2}$ we have

$$
\begin{equation*}
\frac{\varphi(x, t)}{\varphi(y, t)} \leq t^{\left(\frac{A}{\log \left(\frac{1}{|x-y|}\right.}\right)}, \quad \forall t \geq 1 \tag{3}
\end{equation*}
$$

iii)

$$
\begin{equation*}
\text { If } D \subset \Omega \text { is a bounded measurable set, then } \int_{D} \varphi(x, 1) d x<\infty \text {. } \tag{4}
\end{equation*}
$$

$i \boldsymbol{v})$ There exists a constant $C>0$ such that $\psi(x, 1) \leq C$ a.e in $\Omega$.
Under these assumptions, $\mathcal{D}(\Omega)$ is dense in $L_{\varphi}(\Omega)$ with respect to the modular topology, $\mathcal{D}(\Omega)$ is dense in $W_{0}^{1} L_{\varphi}(\Omega)$ for the modular convergence and $\mathcal{D}(\bar{\Omega})$ is dense in $W^{1} L_{\varphi}(\Omega)$ for the modular convergence.

Consequently, the action of a distribution $S$ in $W^{-1} L_{\psi}(\Omega)$ on an element $u$ of $W_{0}^{1} L_{\varphi}(\Omega)$ is well defined. It will be denoted by $\langle S, u\rangle$.

Lemma 2.2 (Poincaré inequality) [18] Let $\varphi$ be a Musielak-Orlicz function which satisfies the assumptions of Lemma 2.1, suppose that $\varphi(x, t)$ decreases with respect to one
of coordinates of $x$. Then, there exists a constant $c>0$ depending only on $\Omega$ such that

$$
\begin{equation*}
\int_{\Omega} \varphi(x,|u(x)|) d x \leq \int_{\Omega} \varphi(x, c|\nabla u(x)|) d x, \quad \forall u \in W_{0}^{1} L_{\varphi}(\Omega) \tag{5}
\end{equation*}
$$

## 3 Assumptions and Main Result

Let $\Omega$ be a bounded open set on $\mathbb{R}^{N}$ satisfying the segment property and $T>0$, we denote $Q=\Omega \times[0, T]$, and let $\varphi$ and $\gamma$ be two Musielak-Orlicz functions such that $\gamma \prec \prec \varphi$ and $\varphi$ satisfies the conditions of Lemma 2.2. Let $A: D(A) \subset W_{0}^{1, x} L_{\varphi}(Q) \longrightarrow W^{-1, x} L_{\psi}(Q)$ be a mapping given by $A(u)=-\operatorname{div}(a(x, t, u, \nabla u))$, where $a: a(x, t, s, \xi): \Omega \times[0, t] \times \mathbb{R} \times$ $\mathbb{R}^{N} \longrightarrow \mathbb{R}^{N}$ is a Carathéodory function satisfying, for a.e $(x, t) \in Q$ and for all $s \in \mathbb{R}$ and all $\xi, \xi^{\prime} \in \mathbb{R}^{N}, \xi \neq \xi^{\prime}$,

$$
\begin{gather*}
|a(x, t, s, \xi)| \leq \beta\left(c(x, t)+\psi_{x}^{-1} \varphi(x, \nu|\xi|)\right)  \tag{6}\\
\left(a(x, t, s, \xi)-a\left(x, t, s, \xi^{\prime}\right)\right)\left(\xi-\xi^{\prime}\right)>0  \tag{7}\\
a(x, t, s, \xi) \cdot \xi \geq \alpha \varphi(x,|\xi|) \tag{8}
\end{gather*}
$$

where $c(x, t)$ is a positive function, $c(x, t) \in E_{\psi}(Q)$ and $\beta, \nu, \alpha \in \mathbb{R}_{+}^{*}$.
Let $g: \Omega \times[0, t] \times \mathbb{R} \times \mathbb{R}^{N} \longrightarrow \mathbb{R}^{N}$ be a Caratheodory function satisfying for a.e. $(x, t) \in \Omega \times[0, t]$ and $\forall s \in \mathbb{R}, \xi \in \mathbb{R}^{N}$,

$$
\begin{gather*}
|g(x, t, s, \xi)| \leq b(|s|)\left(c_{2}(x, t)+\varphi(x,|\xi|)\right),  \tag{9}\\
g(x, t, s, \xi) s \geq 0 \tag{10}
\end{gather*}
$$

where $c_{2}(x, t) \in L^{1}(Q)$ and $b: \mathbb{R}^{+} \longrightarrow \mathbb{R}^{+}$is a continuous and nondecreasing function. Furthermore, let

$$
\begin{equation*}
\Phi \in C^{0}\left(\mathbb{R}, \mathbb{R}^{N}\right) \tag{11}
\end{equation*}
$$

$$
\begin{equation*}
f \in L^{1}(Q) \text { and } u_{0} \text { is an element of } L^{1}(Q) \tag{12}
\end{equation*}
$$

For $\ell>0$ we define the truncation at height $\ell: T_{\ell}: \mathbb{R} \longrightarrow \mathbb{R}$ by

$$
T_{\ell}(s)=\left\{\begin{array}{cc}
s & \text { if }|s| \leq \ell  \tag{13}\\
\ell \frac{s}{|s|} & \text { if }|s|>\ell
\end{array}\right.
$$

The definition of a renormalized solution for $\operatorname{problem}(\mathcal{P})$ can be stated as follows.
Definition 3.1 A measurable function u defined on $Q$ is a renormalized solution of Problem ( $\mathcal{P}$ ) if

$$
\begin{equation*}
\int_{\{(x, t) \in Q ; m \leq|u(x, t)| \leq m+1\}} a(x, t, u, \nabla u) \cdot \nabla u d x d t \longrightarrow 0 \text { as } m \longrightarrow \infty \tag{14}
\end{equation*}
$$

and if, for every function $S$ in $W^{2, \infty}(\mathbb{R})$ which is piecewise $C^{1}$ and such that $S^{\prime}$ has a compact support, we have

$$
\begin{align*}
& \frac{\partial S(u)}{\partial t}-\operatorname{div}\left(a(x, t, u, \nabla u) S^{\prime}(u)\right)+S^{\prime \prime}(u) a(x, t, u, \nabla u) \cdot \nabla u \\
& -\operatorname{div}\left(\Phi(u) S^{\prime}(u)\right)+S^{\prime \prime}(u) \Phi(u) \cdot \nabla u+g(x, t, u, \nabla u) S^{\prime}(u)=f S^{\prime}(u) \text { in } \mathcal{D}^{\prime}(Q) \\
& S(u)(t=0)=S\left(u_{0}\right) \text { in } \Omega \tag{16}
\end{align*}
$$

We will prove the following existence theorem.
Theorem 3.1 Assume that (6) to (11) hold true. Then, there exists a renormalized solution $u$ of problem ( $\mathcal{P}$ ) in the sense of Definition 3.1.

Proof. The proof of Theorem 3.1 is divided into five steps.
Step 1: Approximate problem. Let consider us the following approximate problem

$$
\left(\mathcal{P}_{n}\right)\left\{\begin{array}{l}
\frac{\partial u_{n}}{\partial t}-\operatorname{div}\left(a\left(x, t, u_{n}, \nabla u_{n}\right)+\Phi_{n}\left(u_{n}\right)\right)+g_{n}\left(x, t, u_{n}, \nabla u_{n}\right)=f_{n} \quad \text { in } \mathcal{D}^{\prime}(Q), \\
u_{n}=0 \quad \text { on } \quad \partial \Omega \times(0, T), \\
u_{n}(t=0)=u_{0 n} \quad \text { on } \Omega,
\end{array}\right.
$$

where $\left(f_{n}\right) \in L^{1}(Q)$ is a sequence of smooth functions such that $f_{n} f_{n} \rightarrow$ $f$ in $L^{1}(Q) f$ in $L^{1}(Q), \Phi_{n}(s)=\Phi\left(T_{n}(s)\right)$ and $g_{n}(x, t, s, \xi)=T_{n}(g(x, t, s, \xi))$. Note that $g_{n}(x, t, s, \xi) s \geq 0,\left|g_{n}(x, t, s, \xi)\right| \leq|g(x, t, s, \xi)|$ and $\left|g_{n}(x, t, s, \xi)\right| \leq n$. Since $\Phi$ is continuous, we have $\Phi\left(T_{n}(s)\right) \leq c_{n}$, then the problem $\left(\mathcal{P}_{n}\right)$ has at least one solution $u_{n} \in W_{0}^{1, x} L_{\varphi}(Q)$ (see e.g. [2]).

Step 2: A priori estimates. We take $T_{\ell}\left(u_{n}\right) \chi_{(0, \tau)}$ as a test function in $\left(\mathcal{P}_{n}\right)$, we get for every $\tau \in(0, T)$

$$
\begin{gather*}
\int_{\Omega} \widehat{T}_{\ell}\left(u_{n}(\tau)\right) d x+\int_{Q_{\tau}} a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right) \cdot \nabla T_{\ell}\left(u_{n}\right) d x d t+\int_{Q_{\tau}} \Phi_{n}\left(u_{n}\right) \cdot \nabla T_{\ell}\left(u_{n}\right) d x d t \\
\quad=\int_{Q_{\tau}} f_{n} T_{\ell}\left(u_{n}\right) d x d t-\int_{Q_{\tau}} g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) T_{\ell}\left(u_{n}\right) d x d t+\int_{\Omega} \widehat{T}_{\ell}\left(u_{0 n}\right) d x \tag{17}
\end{gather*}
$$

where

$$
\widehat{T}_{\ell}(s)=\int_{0}^{s} T_{\ell}(\sigma) d \sigma=\left\{\begin{array}{c}
\frac{s^{2}}{2}, \quad \text { if }|s| \leq \ell  \tag{18}\\
\ell|s|-\frac{s^{2}}{2}, \quad \text { if }|s|>\ell
\end{array}\right.
$$

The Lipshitz character of $\Phi_{n}$ and the Stokes formula together with the boundary condition $u_{n}=0$ on $(0, T) \times \partial \Omega$ make it possible to obtain

$$
\begin{equation*}
\int_{Q_{\tau}} \Phi_{n}\left(u_{n}\right) \cdot \nabla T_{\ell}\left(u_{n}\right) d x d t=0 \tag{19}
\end{equation*}
$$

Due to the definition of $\widehat{T}_{\ell}$ and (12) we have

$$
\begin{equation*}
0 \leq \int_{\Omega} \widehat{T}_{\ell}\left(u_{0 n}\right) d x \leq \ell \int_{\Omega}\left|u_{0 n}\right| d x \leq \ell\left\|u_{0}\right\|_{L^{1}(\Omega)} \tag{20}
\end{equation*}
$$

Using the same argument as in [15], we can see that

$$
\begin{equation*}
\int_{Q} g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) d x d t \leq C_{g} \tag{21}
\end{equation*}
$$

Here and below $C_{i}$ denotes positive constants not depending on $n$ and $\ell$. By using (12), (19), (20), (21) we can deduce from (17) that

$$
\begin{equation*}
\int_{\Omega} \widehat{T}_{\ell}\left(u_{n}(\tau)\right) d x+\int_{Q_{\tau}} a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right) \cdot \nabla T_{\ell}\left(u_{n}\right) d x d t \leq \ell C_{0} \tag{22}
\end{equation*}
$$

By using (22), (7) and the fact that $\widehat{T}_{\ell}\left(u_{n}\right) \geq 0$, we deduce that

$$
\begin{equation*}
\int_{Q_{\tau}} \varphi\left(x,\left|\nabla T_{\ell}\left(u_{n}\right)\right|\right) d x d t \leq \frac{1}{\alpha} \int_{Q_{\tau}} a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right) \cdot \nabla T_{\ell}\left(u_{n}\right) d x d t \leq \ell C_{1}, \tag{23}
\end{equation*}
$$

we deduce from the above inequality (22) that

$$
\begin{equation*}
\int_{\Omega} \widehat{T}_{\ell}\left(u_{n}(\tau)\right) d x \leq \ell C_{0}, \text { for almost any } \tau \text { in }(0, T) \tag{24}
\end{equation*}
$$

On the other hand, thanks to Lemma 2.2, there exists a constant $\lambda>0$ depending only on $\Omega$ such that

$$
\begin{equation*}
\int_{Q_{\tau}} \varphi(x,|v|) d x d t \leq \int_{Q_{\tau}} \varphi(x, \lambda|\nabla v|) d x d t, \quad \forall v \in W_{0}^{1} L_{\varphi}(\Omega) . \tag{25}
\end{equation*}
$$

Taking $v=\frac{T_{\ell}\left(u_{n}\right)}{\lambda}$ in (25) and using (23), one has

$$
\begin{equation*}
\int_{Q_{\tau}} \varphi\left(x, \frac{\left|T_{\ell}\left(u_{n}\right)\right|}{\lambda}\right) d x d t \leq \ell C_{1} \tag{26}
\end{equation*}
$$

Then we deduce by using (26), that

$$
\begin{align*}
\operatorname{meas}\left\{\left|u_{n}\right|>\ell\right\} & \leq \frac{1}{\inf _{x \in \Omega} \varphi\left(x, \frac{\ell}{\lambda}\right)} \int_{Q_{\tau}} \varphi\left(x, \frac{1}{\lambda}\left|T_{\ell}\left(u_{n}\right)\right|\right) d x d t \\
& \leq \frac{C_{1} \ell}{\inf _{x \in \Omega} \varphi\left(x, \frac{\ell}{\lambda}\right)} \quad \forall n, \quad \forall \ell \geq 0 . \tag{27}
\end{align*}
$$

By using the definition of $\varphi$, we can deduce

$$
\begin{equation*}
\lim _{\ell \longrightarrow \infty}\left(\operatorname{meas}\left\{(x, t) \in Q_{\tau}:\left|u_{n}\right|>\ell\right\}\right)=0 \tag{28}
\end{equation*}
$$

uniformly with respect to $n$. Moreover, we have from (26) that $T_{\ell}\left(u_{n}\right)$ is bounded in $W_{0}^{1, x} L_{\varphi}(Q)$ for every $\ell>0$. Consider now in $C^{2}(\mathbb{R})$ a nondecreasing function $\zeta_{\ell}(s)=s$ for $|s| \leq \frac{\ell}{2}$ and $\zeta_{\ell}(s)=\ell \operatorname{sign}(s)$. Multiplying the approximating equation by $\zeta_{\ell}^{\prime}\left(u_{n}\right)$, we obtain

$$
\begin{aligned}
& \frac{\partial\left(\zeta_{\ell}\left(u_{n}\right)\right)}{\partial t}=\operatorname{div}\left(a\left(x, t, u_{n}, \nabla u_{n}\right) \zeta_{\ell}^{\prime}\left(u_{n}\right)\right)-\zeta_{\ell}^{\prime \prime}\left(u_{n}\right) a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} \\
& +\operatorname{div}\left(\Phi_{n}\left(u_{n}\right) \zeta_{\ell}^{\prime}\left(u_{n}\right)\right)-\zeta_{\ell}^{\prime \prime}\left(u_{n}\right) \Phi_{n}\left(u_{n}\right) \cdot \nabla u_{n}-g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) \zeta_{\ell}^{\prime}\left(u_{n}\right)+f_{n} \zeta_{\ell}^{\prime}\left(u_{n}\right)
\end{aligned}
$$

in the sense of distributions. Thanks to (26) and the fact that $\zeta_{\ell}^{\prime}$ has a compact support, $\zeta_{\ell}^{\prime}\left(u_{n}\right)$ is bounded in $W_{0}^{1, x} L_{\varphi}(Q)$ while its time derivative $\frac{\partial\left(\zeta_{\ell}\left(u_{n}\right)\right)}{\partial t}$ is bounded in $W_{0}^{-1, x} L_{\varphi}(Q)+L^{1}(Q)$, hence Corollary 4.5 of [2] allows us to conclude that $\zeta_{\ell}\left(u_{n}\right)$ is compact in $L^{1}(Q)$. Due to the choice of $\zeta_{\ell}$, we conclude that for each $\ell$, the sequence $T_{\ell}\left(u_{n}\right)$ converges almost everywhere in $Q$. Therefore, following [8,9,15], we can see that there exists a measurable function $u \in L^{\infty}\left(0, T ; L^{1}(\Omega)\right)$ such that for every $\ell>0$ and a subsequence, not relabeled,

$$
\begin{equation*}
u_{n} \rightarrow u \text { a. e. in } Q, \tag{29}
\end{equation*}
$$

and

$$
\begin{gather*}
T_{\ell}\left(u_{n}\right) \rightharpoonup T_{\ell}(u) \text { weakly in } W_{0}^{1, x} L_{\varphi}(Q) \text { for } \sigma\left(\Pi L_{\varphi}, \Pi E_{\psi}\right),  \tag{30}\\
\text { strongly in } L^{1}(Q) \text { and a. e. in } Q .
\end{gather*}
$$

Now we shall to prove the boundness of $\left(a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right)\right)_{n}$ in $\left(L_{\psi}(Q)\right)^{N}$.
Let $\phi \in\left(E_{\varphi}(Q)\right)^{N}$ with $\|\phi\|_{\varphi, Q}=1$. In view of the monotonicity of a one easily has,

$$
\begin{equation*}
\int_{Q}\left[a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right)-a\left(x, t, T_{\ell}\left(u_{n}\right), \phi\right)\right]\left[\nabla T_{\ell}\left(u_{n}\right)-\phi\right] d x d t \geq 0 \tag{31}
\end{equation*}
$$

which gives

$$
\begin{align*}
\int_{Q} a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right) \cdot \phi d x d t \leq & \int_{Q} a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right) \cdot \nabla T_{\ell}\left(u_{n}\right) d x d t  \tag{32}\\
& +\int_{Q} a\left(x, t, T_{\ell}\left(u_{n}\right), \phi\right) \cdot\left[\nabla T_{\ell}\left(u_{n}\right)-\phi\right] d x d t
\end{align*}
$$

Using (6) and (23), we easily see that

$$
\begin{equation*}
\int_{Q} a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right) \cdot \phi d x d t \leq C_{3} \tag{33}
\end{equation*}
$$

And so, we conclude that $\left(a\left(x, t, T_{\ell}\left(u_{n}\right), \nabla T_{\ell}\left(u_{n}\right)\right)\right)_{n}$ is a bounded sequence in $\left(L_{\psi}(Q)\right)^{N}$. Now, we prove that

$$
\begin{equation*}
\lim _{m \longrightarrow \infty} \lim _{n \longrightarrow \infty} \int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} d x d t=0 . \tag{34}
\end{equation*}
$$

Using in $\left(\mathcal{P}_{n}\right)$ the test function $v=T_{1}\left(u_{n}-T_{m}\left(u_{n}\right)\right)$, we obtain

$$
\begin{align*}
& \left\langle\frac{\partial u_{n}}{\partial t}, v\right\rangle+\int_{\left\{m \leq u_{n} \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} d x d t+\int_{Q} g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) v d x d t \\
& +\int_{Q} \operatorname{div}\left[\int_{0}^{u_{n}} \Phi_{n}(r) T_{1}^{\prime}\left(u_{n}-T_{m}\left(u_{n}\right)\right) d r\right] d x d t=\int_{Q} f_{n} v d x d t . \tag{35}
\end{align*}
$$

By using $\int_{0}^{u_{n}} \Phi_{n}(r) T_{1}^{\prime}\left(u_{n}-T_{m}\left(u_{n}\right)\right) d r \in W_{0}^{1, x} L_{\varphi}(Q)$ and the Stokes formula, we get

$$
\begin{align*}
& \int_{\Omega} U_{n}^{m}\left(u_{n}(T)\right) d x+\int_{\left\{m \leq u_{n} \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} d x d t  \tag{36}\\
& \leq \int_{Q}\left(\left|f_{n}+g_{n}\left(x, t, u_{n}, \nabla u_{n}\right)\right|\right)\left|T_{1}\left(u_{n}-T_{m}\left(u_{n}\right)\right)\right| d x d t+\int_{\Omega} U_{n}^{m}\left(x, u_{0 n}\right) d x
\end{align*}
$$

where $U_{n}^{m}(r)=\int_{0}^{u_{n}} \frac{\partial u_{n}}{\partial t} T_{1}\left(s-T_{m}(s)\right) d s$. In order to pass to the limit as $n$ tends to $+\infty$ in (36), we use $U_{n}^{m}\left(u_{n}(T)\right) \geq 0$, (12) and (21), we obtain that

$$
\begin{align*}
\lim _{n \longrightarrow \infty} & \int_{\left\{m \leq u_{n} \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} d x d t  \tag{37}\\
& \leq \int_{\left\{\left|u_{n}\right|>m\right\}}\left(|f|+C_{g}\right) d x d t+\int_{\left\{\left|u_{0}\right|>m\right\}}\left|u_{0}\right| d x .
\end{align*}
$$

Finally, by(12) and (37) we obtain (34).
Step 3: Almost everywhere convergence of the gradients. Fix $\ell>0$ and let $\phi(s)=s \exp \left(\delta s^{2}\right), \delta>0$. It is well known that when $\delta \geq\left(\frac{b(\ell)}{2 \alpha}\right)^{2}$ one has

$$
\begin{equation*}
\phi^{\prime}(s)-\frac{b(\ell)}{\alpha}|\phi(s)| \geq \frac{1}{2} \text { for all } s \in \mathbb{R} . \tag{38}
\end{equation*}
$$

Let $v_{j} \in \mathcal{D}(Q)$ be a sequence which converges to $u$ for the modular convergence in $W_{0}^{1, x} L_{\varphi}(Q)$ and let $\omega_{i} \in \mathcal{D}(Q)$ be a sequence which converges strongly to $u_{0}$ in $L^{2}(\Omega)$. Set $\omega_{i, j}^{\mu}=T_{\ell}\left(v_{j}\right)_{\mu}+\exp (-\mu t) T_{\ell}\left(w_{i}\right)$, where $T_{\ell}\left(v_{j}\right)_{\mu}$ is the mollification with respect to time of $T_{\ell}\left(v_{j}\right)$. Note that $\omega_{i, j}^{\mu}$ is a smooth function having the following properties:

$$
\begin{align*}
\frac{\partial}{\partial t}\left(\omega_{i, j}^{\mu}\right) & =\mu\left(T_{\ell}\left(v_{j}\right)-\omega_{i, j}^{\mu}\right), \omega_{i, j}^{\mu}(0)=T_{\ell}\left(\omega_{i}\right),\left|\omega_{i, j}^{\mu}\right| \leq \ell  \tag{39}\\
\omega_{i, j}^{\mu} & \rightarrow T_{\ell}(u)_{\mu}+\exp (-\mu t) T_{\ell}\left(w_{i}\right) \text { in } W_{0}^{1, x} L_{\varphi}(Q) \tag{40}
\end{align*}
$$

for the modular convergence as $j \rightarrow \infty$,

$$
\begin{equation*}
T_{\ell}(u)_{\mu}+\exp (-\mu t) T_{\ell}\left(w_{i}\right) \rightarrow T_{\ell}(u) \text { in } W_{0}^{1, x} L_{\varphi}(Q) \tag{41}
\end{equation*}
$$

for the modular convergence as $\mu \rightarrow \infty$. Let now the function $\rho_{m}$ on $\mathbb{R}$ with $m \geq \ell$ be defined by

$$
\rho_{m}(s)=\left\{\begin{align*}
1, & \text { if }|s| \leq m  \tag{42}\\
m+1-|s|, & \text { if } m \leq|s| \leq m+1 \\
0, & \text { if }|s| \geq m+1
\end{align*}\right.
$$

We set $\theta_{i, j}^{\mu, n}=T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}$. Using the admissible test function $Z_{i, j, n}^{\mu, m}=\phi\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}\left(u_{n}\right)$ as test function in $\left(\mathcal{P}_{n}\right)$ and since $g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) \phi\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}\left(u_{n}\right) \geq 0$ on $\left\{\left|u_{n}\right|>\ell\right\}$, we arrive at

$$
\begin{align*}
& \left\langle\frac{\partial u_{n}}{\partial t}, Z_{i, j, n}^{\mu, m}\right\rangle+\int_{Q} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot\left(\nabla T_{\ell}\left(u_{n}\right)-\nabla \omega_{i, j}^{\mu}\right) \phi^{\prime}\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}\left(u_{n}\right) d x d t \\
& +\int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} \phi\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}^{\prime}\left(u_{n}\right) d x d t \\
& +\int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} \Phi_{n}\left(u_{n}\right) \cdot \nabla u_{n} \phi\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}^{\prime}\left(u_{n}\right) d x d t  \tag{43}\\
& +\int_{Q} \Phi_{n}\left(u_{n}\right) \cdot\left(\nabla T_{\ell}\left(u_{n}\right)-\nabla \omega_{i, j}^{\mu}\right) \phi^{\prime}\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}\left(u_{n}\right) d x d t \\
& +\int_{\left\{\left|u_{n}\right| \leq \ell\right\}} g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) \phi\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}\left(u_{n}\right) d x d t \leq \int_{Q} f_{n} Z_{i, j, n}^{\mu, m} d x d t .
\end{align*}
$$

Denote by $\epsilon(n, j, \mu, i)$ any quantity such that $\lim _{i \rightarrow \infty} \lim _{\mu \rightarrow \infty} \lim _{j \rightarrow \infty} \lim _{n \rightarrow \infty} \epsilon(n, j, \mu, i)=0$.
The very definition of the sequence $\omega_{i, j}^{\mu}$ makes it possible to establish the following lemma.

Lemma 3.1 (cf.[2]) Let $Z_{i, j, n}^{\mu, m}=\phi\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}\left(u_{n}\right)$, we have for any $\ell \geq 0$

$$
\begin{equation*}
\left\langle\frac{\partial u_{n}}{\partial t}, Z_{i, j, n}^{\mu, m}\right\rangle \geq \epsilon(n, j, i) \tag{44}
\end{equation*}
$$

Concerning the right-hand of (43), by the almost everywhere convergence of $u_{n}$, we have $\phi\left(T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}\right) \rho_{m}\left(u_{n}\right) \rightharpoonup \phi\left(T_{\ell}(u)-\omega_{i, j}^{\mu}\right) \rho_{m}(u)$ weakly-* in $L^{\infty}(Q)$ as $n \rightarrow \infty$, and then

$$
\int_{Q} f_{n} \phi\left(T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}\right) \rho_{m}\left(u_{n}\right) d x d t \rightarrow \int_{Q} f \phi\left(T_{\ell}(u)-\omega_{i, j}^{\mu}\right) \rho_{m}(n) d x d t
$$

so that $\phi\left(T_{\ell}(u)-\omega_{i, j}^{\mu}\right) \rho_{m}(u) \rightharpoonup \phi\left(T_{\ell}(u)-T_{\ell}(u)_{\mu}-\exp (-\mu t) T_{\ell}\left(w_{i}\right)\right) \rho_{m}(u)$ weakly star in $L^{\infty}(Q)$ as $j \rightarrow \infty$, and finally,

$$
\phi\left(T_{\ell}(u)-T_{\ell}(u)_{\mu}-\exp (-\mu t) T_{\ell}\left(w_{i}\right)\right) \rho_{m}(u) \rightharpoonup 0 \text { weakly star as } \mu \rightarrow \infty .
$$

Then, we deduce that

$$
\begin{equation*}
\left\langle f_{n}, \phi\left(T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}\right) \rho_{m}\left(u_{n}\right)\right\rangle=\epsilon(n, j, \mu) . \tag{45}
\end{equation*}
$$

Similarly, Lebesgue's convergence theorem shows that

$$
\Phi_{n}\left(u_{n}\right) \rho_{m}\left(u_{n}\right) \rightarrow \Phi(u) \rho_{m}(u) \text { strongly in }\left(E_{\psi}(Q)^{N}\right) \text { as } n \rightarrow \infty
$$

and

$$
\Phi_{n}\left(u_{n}\right) \chi_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} \phi^{\prime}\left(T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}\right) \rightarrow \Phi(u) \chi_{\{m \leq u \leq m+1\}} \phi^{\prime}\left(T_{\ell}(u)-\omega_{i, j}^{\mu}\right)
$$

strongly in $\left(E_{\psi}(Q)^{N}\right)$. Then by virtue of $\nabla T_{\ell}\left(u_{n}\right) \rightharpoonup \nabla T_{\ell}(u)$ weakly star in $\left(L_{\varphi}(Q)^{N}\right)$, and $\nabla u_{n} \chi_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}}=\nabla T_{m+1}\left(u_{n}\right) \chi_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}}$ a. e. in $Q$, one has

$$
\begin{aligned}
& \int_{Q} \Phi_{n}\left(u_{n}\right) \cdot\left(\nabla T_{\ell}\left(u_{n}\right)-\nabla \omega_{i, j}^{\mu}\right) \phi^{\prime}\left(T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}\right) \rho_{m}\left(u_{n}\right) d x d t \\
& \quad \rightarrow \int_{Q} \Phi(u) \nabla\left(\nabla T_{\ell}(u)-\nabla \omega_{i, j}^{\mu}\right) \phi^{\prime}\left(T_{\ell}(u)-\omega_{i, j}^{\mu}\right) \rho_{m}(u) d x d t
\end{aligned}
$$

as $n \rightarrow \infty$, and

$$
\begin{aligned}
& \int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} \Phi_{n}\left(u_{n}\right) \phi\left(T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}\right) \nabla u_{n} \rho_{m}^{\prime}\left(u_{n}\right) d x d t \\
& \rightarrow \int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} \Phi(u) \phi\left(T_{\ell}\left(u_{n}\right)-\omega_{i, j}^{\mu}\right) \nabla u \rho_{m}^{\prime}(u) d x d t
\end{aligned}
$$

as $n \rightarrow+\infty$. Thus, by using the modular convergence of $\omega_{i, j}^{\mu}$ as $j \rightarrow+\infty$ and letting $\mu$ tend to infinity, we get

$$
\begin{equation*}
\int_{Q} \Phi_{n}\left(u_{n}\right) \cdot\left(\nabla T_{\ell}\left(u_{n}\right)-\nabla \omega_{i, j}^{\mu}\right) \phi^{\prime}\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}\left(u_{n}\right) d x d t=\epsilon(n, j, \mu) \tag{46}
\end{equation*}
$$

and

$$
\begin{equation*}
\int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} \Phi_{n}\left(u_{n}\right) \cdot \nabla u_{n} \phi\left(\theta_{i, j}^{\mu, n}\right) \rho_{m}^{\prime}\left(u_{n}\right) d x d t=\epsilon(n, j, \mu) \tag{47}
\end{equation*}
$$

Concerning the third term of the right-hand side of (43) we obtain that

$$
\begin{aligned}
& \left|\int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} \phi\left(\theta_{n, j}^{\mu, i}\right) \rho_{m}^{\prime}\left(u_{n}\right) d x d t\right| \\
& \leq \phi(2 k) \int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} d x d t
\end{aligned}
$$

Then by (34) we deduce that

$$
\begin{equation*}
\left|\int_{Q} a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} \phi\left(\theta_{n, j}^{\mu, i}\right) \rho_{m}^{\prime}\left(u_{n}\right) d x d t\right| \leq \epsilon(n, \mu, m) . \tag{48}
\end{equation*}
$$

Using the same technics as in the proof of Proposition 5.6 in [4], we obtain

$$
\begin{align*}
& \int_{Q}\left(a\left(x, t, T_{k}\left(u_{n}\right), \nabla T_{k}\left(u_{n}\right)\right)-a\left(x, t, T_{k}\left(u_{n}\right), \nabla T_{k}(u) \chi^{s}\right)\right)  \tag{49}\\
& \times\left(\nabla T_{k}\left(u_{n}\right)-\nabla T_{k}(u) \chi^{s}\right) d x d t \leq \varepsilon(n, j, \mu, i, s, m) .
\end{align*}
$$

To pass to the limit in (49) as $n j, m, s$ tend to infinity, we obtain

$$
\begin{align*}
& \lim _{s \rightarrow \infty} \lim _{n \rightarrow \infty} \int_{Q}\left(a\left(x, t, T_{k}\left(u_{n}\right), \nabla T_{k}\left(u_{n}\right)\right)-a\left(x, t, T_{k}\left(u_{n}\right), \nabla T_{k}(u) \chi^{s}\right)\right)  \tag{50}\\
& \times\left(\nabla T_{k}\left(u_{n}\right)-\nabla T_{k}(u) \chi^{s}\right) d x d t=0
\end{align*}
$$

And thus, as in the elliptic case (see [18]), there exists a subsequence also denoted by $u_{n}$ such that

$$
\begin{equation*}
\nabla u_{n} \rightarrow \nabla u \text { a.e. in } Q \tag{51}
\end{equation*}
$$

Then, for all $k>0$, one has

$$
\begin{array}{r}
a\left(x, t, T_{k}\left(u_{n}\right), \nabla T_{k}\left(u_{n}\right)\right) \rightharpoonup a\left(x, t, T_{k}(u), \nabla T_{k}(u)\right) \\
\quad \text { weakly star in }\left(L_{\psi}(Q)\right)^{N} \text { for } \sigma\left(\Pi L_{\psi}, \Pi E_{\varphi}\right) . \tag{52}
\end{array}
$$

Step 4: In this step we prove that u satisfies (15). According to (50), one can pass to the limit as $n$ tends to $+\infty$ for fixed $m \geq 0$ to obtain

$$
\begin{align*}
& \lim _{n \longrightarrow \infty} \int_{\left\{m \leq\left|u_{n}\right| \leq m+1\right\}} a\left(x, t, u_{n}, \nabla u_{n}\right) \nabla u_{n} d x d t \\
& =\int_{Q} a\left(x, t, T_{m+1}(u), \nabla T_{m+1}(u)\right) \nabla T_{m+1}(u) d x d t \\
& \quad-\int_{Q} a\left(x, t, T_{m}(u), \nabla T_{m}(u)\right) \nabla T_{m}(u) d x d t \\
& \quad=\int_{\{m \leq|u| \leq m+1\}} a(x, t, u, \nabla u) \cdot \nabla u d x d t \tag{53}
\end{align*}
$$

Taking the limit as $m \rightarrow+\infty$ in (53) and using the estimate (34) show that $u$ satisfies (15). Following the same technique as that used in [2], and by using (29), (50) and Vitali's theorem, we have

$$
\begin{equation*}
g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) \rightarrow g(x, t, u, \nabla u) \text { strongly in } L^{1}(Q) . \tag{54}
\end{equation*}
$$

Step 5: Passing to the limit. Let $S$ be a function in $W^{2, \infty}(\mathbb{R})$ such that $S^{\prime}$ has a compact support. Let $K$ be a positive real number such that $\operatorname{supp}\left(S^{\prime}\right) \subset[-K, K]$. Pointwise multiplication of the approximate equation $\left(\mathcal{P}_{n}\right)$ by $S^{\prime}\left(u_{n}\right)$ leads to

$$
\begin{align*}
\frac{\partial S\left(u_{n}\right)}{\partial t} & -\operatorname{div}\left(a\left(x, t, u_{n}, \nabla u_{n}\right) S^{\prime}\left(u_{n}\right)\right)+S^{\prime \prime}\left(u_{n}\right) a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} \\
& -\operatorname{div}\left(S^{\prime}\left(u_{n}\right) \Phi\left(u_{n}\right)\right)+S^{\prime \prime}\left(u_{n}\right) \Phi\left(u_{n}\right) \cdot \nabla u_{n} \\
& +g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) S^{\prime}\left(u_{n}\right) \\
& =f_{n} S^{\prime}\left(u_{n}\right) . \tag{55}
\end{align*}
$$

In what follows we pass to the limit as $n$ tends to $+\infty$ in each term of (55).

- Since $S$ is bounded and continuous, then the fact that $u_{n} \longrightarrow u$ a.e. in $Q$, implies that $S\left(u_{n}\right)$ converges to $S(u)$ a.e. in $Q$ and $L^{\infty}$ weakly-*. Consequently,

$$
\frac{\partial S\left(u_{n}\right)}{\partial t} \longrightarrow \frac{\partial S(u)}{\partial t} \quad \text { in } \mathcal{D}^{\prime}(Q) \text { as } n \text { tends to }+\infty
$$

- Since $\operatorname{supp}\left(S^{\prime}\right) \subset[-K, K]$, we have for $n \geq K$,

$$
a\left(x, t, u_{n}, \nabla u_{n}\right) S^{\prime}\left(u_{n}\right)=a\left(x, t, T_{K}\left(u_{n}\right), \nabla T_{K}\left(u_{n}\right)\right) S^{\prime}\left(u_{n}\right) \quad \text { a.e. in } Q .
$$

The pointwise convergence of $u_{n}$ to $u$ and (52) as $n$ tends to $\infty$ and the bounded character of $S^{\prime}$ permit us to conclude that
$a\left(x, t, T_{K}\left(u_{n}\right), \nabla T_{K}\left(u_{n}\right)\right) S^{\prime}\left(u_{n}\right) \longrightarrow a\left(x, t, T_{K}(u), \nabla T_{K}(u)\right) S^{\prime}(u)$ weakly star in $\left(L_{\psi}(Q)\right)^{N}$
as $n$ tends to infinity.

- Regarding the 'energy' term, we have for $n \geq K$

$$
S^{\prime \prime}\left(u_{n}\right) a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n}=S^{\prime \prime}\left(u_{n}\right) a\left(x, t, T_{K}\left(u_{n}\right), \nabla T_{K}\left(u_{n}\right)\right) \cdot \nabla T_{K}\left(u_{n}\right) \text { a.e. in } Q .
$$

The pointwise convergence of $S^{\prime}\left(u_{n}\right) \longrightarrow S^{\prime}(u)$ and (52) as $n$ tends to $+\infty$ and the bounded character of $S^{\prime \prime}$ permit us to conclude that
$S^{\prime \prime}\left(u_{n}\right) a\left(x, t, u_{n}, \nabla u_{n}\right) \cdot \nabla u_{n} \rightharpoonup S^{\prime \prime}(u) a\left(x, t, T_{K}(u), \nabla T_{K}(u)\right) \cdot \nabla T_{K}(u)$ weakly star in $L^{1}(Q)$.
Recall that $S^{\prime \prime}(u) a\left(x, t, T_{K}(u), \nabla T_{K}(u)\right) \cdot \nabla T_{K}(u)=S^{\prime \prime}(u) a(x, t, u, \nabla u) \cdot \nabla u \quad$ a.e. in $Q$.

- Since $\operatorname{supp}\left(S^{\prime}\right) \subset[-K, K]$, we have

$$
\begin{equation*}
S^{\prime}\left(u_{n}\right) \Phi_{n}\left(u_{n}\right)=S^{\prime}\left(u_{n}\right) \Phi_{n}\left(T_{K}\left(u_{n}\right)\right) \quad \text { a.e. in } Q . \tag{58}
\end{equation*}
$$

As a consequence of (11) and (29), it follows that

$$
\begin{equation*}
S^{\prime}\left(u_{n}\right) \Phi_{n}\left(u_{n}\right) \rightarrow S^{\prime}(u) \Phi\left(T_{K}(u)\right) \quad \text { a.e. in }\left(E_{\varphi}(Q)\right)^{N}, \tag{59}
\end{equation*}
$$

we have $\nabla S^{\prime \prime}\left(u_{n}\right)$ converges to $\nabla S^{\prime \prime}(u)$ weakly in $\left(L_{\varphi}(Q)\right)^{N}$ as $n$ tends to $+\infty$, while $\Phi_{n}\left(T_{K}\left(u_{n}\right)\right)$ is uniformly bounded with respect to $n$ and converges a. e. in $Q$ to $\Phi\left(T_{K}(u)\right)$ as $n$ tends to $+\infty$. Therefore

$$
\begin{equation*}
S^{\prime \prime}\left(u_{n}\right) \Phi_{n}\left(u_{n}\right) \nabla u_{n} \rightharpoonup S^{\prime \prime}(u) \Phi(u) \nabla u \quad \text { weakly in } L_{\varphi}(Q) \tag{60}
\end{equation*}
$$

- Since supp $S^{\prime} \subset[-K, K]$ and from (54), we have

$$
\begin{equation*}
S^{\prime}\left(u_{n}\right) g_{n}\left(x, t, u_{n}, \nabla u_{n}\right) \longrightarrow g(x, t, u, \nabla u) S^{\prime}(u) \quad \text { strongly in } L^{1}(Q) . \tag{61}
\end{equation*}
$$

- Due to $f_{n} \longrightarrow f$ in $L^{1}(Q)$ and the fact that $u_{n} \longrightarrow u$ a.e. in $Q$, we have

$$
\begin{equation*}
S^{\prime}\left(u_{n}\right) f_{n} \longrightarrow S^{\prime}(u) f \quad \text { strongly in } L^{1}(Q) \tag{62}
\end{equation*}
$$

As a consequence of the above convergence results, we are in a position to pass to the limit as $n$ tends to $+\infty$ in equation (55) and to conclude that

$$
\begin{align*}
\frac{\partial S(u)}{\partial t} & -\operatorname{div}\left(a(x, t, u, \nabla u) S^{\prime}(u)\right)+S^{\prime \prime}(u) a(x, t, u, \nabla u) \cdot \nabla u \\
& -\operatorname{div}\left(S^{\prime}(u) \Phi(u)\right)+S^{\prime \prime}(u) \Phi(u) \cdot \nabla u \\
& +g(x, t, u, \nabla u) S^{\prime}(u) \\
& =f S^{\prime}(u) \tag{63}
\end{align*}
$$

It remains to show that $S(u)$ satisfies the initial condition.
To this end, firstly note that, $S$ being bounded, $S\left(u_{n}\right)$ is bounded in $L^{\infty}(Q)$. Secondly, (55) and the above considerations on the behavior of the terms of this equation show that $\frac{\partial S\left(u_{n}\right)}{\partial t}$ is bounded in $L^{1}(Q)+V^{*}$. As a consequence, an Aubin's type lemma (see, e.g, [17]) implies that $S\left(u_{n}\right)$ lies in a compact set of $C^{0}\left([0, T], L^{1}(\Omega)\right)$. It follows that, on the one hand, $S\left(u_{n}\right)(t=0)=S\left(u_{0 n}\right)$ converges to $S(u)(t=0)$ strongly in $L^{1}(\Omega)$.

On the other hand, the smoothness of $S$ implies that

$$
S(u)(t=0)=S\left(u_{0}\right) \quad \text { in } \Omega .
$$

As a conclusion of step 1 to step 6 , the proof of Theorem 3.1 is complete.
Example 3.1 Let $\Omega$ be a bounded Lipschitz domain of $\mathbb{R}^{N}$ and $T>0$, we denote by $Q=\Omega \times[0, T]$, and let $\varphi$ and $\psi$ be two complementary Musielak functions. Moreover, we assume that $\varphi(x, t)$ decreases with respect to one of coordinates of $x$ (for example, $\varphi(x, t)=|t|^{p(x)} \log \left(1+t^{3}\right), p(x)=e^{\left(-x_{1}^{2}+x_{2}^{2}+\cdots+x_{N}^{2}\right)}$. We set

$$
\begin{gathered}
a(x, t, s, \zeta)=\left(3+\cos ^{2}(\varphi(x, s))\right) \psi_{x}^{-1}(\varphi(x,|\zeta|)) \frac{\zeta}{|\zeta|} \\
g(x, t, s, \zeta)=\frac{\varphi(x,|\zeta|)}{1+s^{2}}, \Phi(s)=\left(|s|^{r_{1}-1} s, \ldots,|s|^{r_{N}-1} s\right), \quad 1 \leq r_{1}, \ldots, R_{N}<\infty
\end{gathered}
$$

It is easy to show that $a(x, t, s, \zeta)$ is the Caratheodory function satisfying the growth condition (6), the coercivity (8) and the monotonicity condition, while the Caratheodory function $g(x, t, s, \zeta)$ satisfies the condition (9) and (10), Finally, the hypotheses of Theorem 3.1 are satisfied. Therefore, the following problem

$$
\left\{\begin{array}{l}
\lim _{m \xrightarrow[\longrightarrow]{ }} \int_{\{(x, t) \in Q ; m \leq|u(x, t)| \leq m+1\}} a(x, t, u, \nabla u) \cdot \nabla u \quad d x d t=0,  \tag{64}\\
\frac{\partial S(u)}{\partial t}-\operatorname{div}\left(a(x, t, u, \nabla u) S^{\prime}(u)\right)+S^{\prime \prime}(u) a(x, t, u, \nabla u) \cdot \nabla u \\
-\operatorname{div}\left(S^{\prime}(u) \Phi(u)\right)+S^{\prime \prime}(u) \Phi(u) \cdot \nabla u+g(x, t, u, \nabla u) S^{\prime}(u)=f S^{\prime}(u), \\
S(u)(t=0)=S\left(u_{0}\right) \text { in } \Omega, \\
\text { for every function } S \text { in } W^{2, \infty}(\mathbb{R}) \text { and such that } S^{\prime} \text { has a compact support in } \mathbb{R}
\end{array}\right.
$$

has at least one renormalised solution for any $f \in L^{1}(Q)$.

## References

[1] Aharouch, L., Bennouna, J. and Touzani, A. Existence of renormalized solutions of some elliptic problems in Orlicz spaces. Rev. Mat. Complut. 22 (1) (2009) 91-110.
[2] Ahmed Oubeid, M. L. , Benkirane, A. and Sidi El Vally, M. Strongly nonlinear parabolic problems in Musielak-Orlicz-Sobolev spaces. Bol. Soc. Paran. Mat. 33 (1) (2015) 191-223.
[3] Akdim, Y., Bennouna, J., Mekkour, M. and Redwane, H., Parabolic Equations with Measure Data and Three Unbounded Nonlinearities in Weighted Sobolev Spaces. Nonlinear Dynamics and Systems Theory 15 (2) (2015) 107-126.
[4] Azroul, E., Redwane, H. and Rhoudaf, M. Existence of a renormalized solution for a class of nonlinear parabolic equations in Orlicz spaces. Port. Math. 66 (1) (2009) 29-63.
[5] Bénilan, P., Boccardo, L., Gallouet, T., Gariepy, R., Pierre, M. and Vazquez, J.L. An $L^{1}$-theory of existence and uniqueness of solutions of nonlinear elliptic equations. Ann. Sc. Norm. Super. Pisa 22 (1995) 241-273.
[6] Benkirane, A. and Bennouna, J. Existence of solution for nonlinear elliptic degenrate equations. Nonlinear Analysis 54 (2003) 9-37.
[7] Benkirane, A., and Sidi El Vally, M. (Ould Mohamedhen Val). Some approximation properties in Musielak-Orlicz-Sobolev spaces. Thai. J. Math. 10 (2012) 371-381.
[8] Blanchard, D., and Murat, F. Renormalized solutions of nonlinear parabolic with $L^{1}$ data: existence and uniqueness. Proc. Roy. Soc. Edinburgh Sect, A 127 (1997) 1137-1152.
[9] Blanchard, D., Murat, F., and Redwane, H. Existence and uniqueness of renormalized solution for fairly general class of non linear parabolic problems. J. Differ. Equ. 177 (2001) 331-374.
[10] Boccardo, L. and Gallouet, T. On some nonlinear elliptic equations with right-hand side measures. Commun. Partial Differ. Equ. 17 (1992) 641-655.
[11] Boccardo, L., Giachetti, D., Diaz, J.I. and Murat, F. Existence and regularity of renormalized solutions for some elliptic problems involving derivation of nonlinear terms. J. Differ. Equ. 106 (1993) 215-237.
[12] DiPerna, R.J. and Lions, P.L. On the Cauchy problem for Boltzmann equations: global existence and weak stability. Ann. Math. 130 (1989) 321-366.
[13] Lions, P.L. Mathematical Topics in Fluid Mechanics, Incompressible Models. Oxford Univ. Press, 1996.
[14] Musielak, J. Modular spaces and Orlicz spaces. Lecture Notes in Math. (1983) 10-34.
[15] Porretta, A. Existence results for strongly nonlinear parabolic equations via strong convergence of truncations. Ann. Mat. Pura Appl. IV (177) (1999) 143-172.
[16] Serrin, J. Pathological solution of elliptic differential equations. Ann. Sc. Norm. Super. Pisa Cl. Sci. 18 (1964) 385-387.
[17] Simon, J. Compact sets in the space $L^{p}(0, T, B)$, Ann. Mat. Pura. Appl. 146 (1987) 65-96.
[18] Talha. A., Benkirane, A. Strongly nonlinear elliptic boundary value problems in MusielakOrlicz spaces, Monatsh. Math. 184 (2017) 1-32.
[19] Talha, A., Benkirane, A. and Elemine Vall, M.S.B. Entropy solutions for nonlinear parabolic inequalities involving measure data in Musielak-Orlicz-Sobolev spaces. Bol. Soc. Paran. Mat. 36 (2) (2018) 199-230.


[^0]:    * Corresponding author: mailto:calliera@dm.uba.ar

[^1]:    * Corresponding author: mailto:ezati@kiau.ac.ir

[^2]:    * Corresponding author: mailto:bhausaheb.desale@mathematics.mu.ac.in

[^3]:    * Corresponding author: mailto:eugene@fit.edu

[^4]:    * Corresponding author: mailto:naceur.benhadj@ept.rnu.tn

[^5]:    * Corresponding author: mailto:farehhannachi@yahoo.com

[^6]:    * Corresponding author: mailto:imron-its@matematika.its.ac.id

[^7]:    * Corresponding author: mailto:marwan04@just.edu.jo

[^8]:    * Corresponding author: mailto:talha.abdous@gmail.com

