# Two-dimensional Overflow Queueing Systems 

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# Two-dimensional Overflow Queueing Systems 

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## Introduction

> Well informed people know that it is impossible to transmit the human voice over wires.
> - Boston Newspaper, 1867

Following Leonard Kleinrock [37], queueing theory is the study of the phenomena of standing, waiting, and serving. As simple as this definition might be, as manifold are the applications and real world problems that can be described by queueing models. The queueing theory started at the early beginning of the 20 th century with the pioneer work of Frederik Ferdinand Wilhelm Johannsen [33], Tore Olaus Engset [21,22] and Agner Krarup Erlang [23] as the study of telephone networks with limited capacity. Numerous refinements and new applications of queueing theory have arisen since then. Nowadays, the rich and fertile theory is applied to the analysis of communication networks and computer systems for internet and data traffic or bandwidth management, to health care systems, traffic control, insurance mathematics, machine plants and almost every area of everyday life in which "standing, waiting, and serving" takes precedence. Although the work of Johannsen, Engset and Erlang might seem old-fashioned from todays point of view - and impossible if one believes a Boston newspaper from 1867 -, their results are 100 years after the publication of Erlang's first paper still perfectly applicable and applied in modern teletraffic engineering (see Stordahl 61] and ITU [32]).

Whenever a queueing system offers only a finite number of service and/or waiting positions, some service demands might be declined and some not. From an economic point of view, it might be useful to know the system's long-run behavior in order to balance for example the cost of lost demands and the benefits from service. Quantities like the overall blocking or overflow probability, the average departure rate from the waiting room and the servers and the average occupation proportion of the waiting and service positions are amongst others of special interest. A system designer for example has to
know these characteristics in order to control and optimize a queueing system and to reach a certain cost or utilization level. However, these characteristics can only be calculated for a limited class of queueing systems and the more involved the system dynamics get, the more involved the analysis of the long run behavior usually becomes.

In this thesis, we present two fairly general classes of so called overflow queueing networks. These networks consist of two queues, where the capacity of the first queue is always finite. Customers arriving at the first queue have an overflow capability from the first to the second queue if the first queue operates at a certain fixed capacity, i.e., under certain conditions, demands arriving at the first queue are allowed to join the second queue. In every model, the dynamic of the first queue is or is at least similar to the famous Erlang and Engset loss systems. The overflow stream will additionally be weighted with a parameter $p \in[0,1]$. The parameter $p$ can be used as a control parameter or to model the customers' impatience.

The first chapter gives a brief overview of the general stochastic structure underlying these networks. In principle, each of the queues is fed by a Markovian Poisson process and the service times are exponential.

In the second chapter, we consider a generalization of a queueing model presented by Perel and Yechiali [56] and additionally append an overflow capability. In this generalized two-queue network, the arrival and service rates for the finite first queue are state-dependent and the customers in the first queue act as servers for the second queue, i.e., the service rate in the second queue depends on the state of the first queue. This is for example the case for file sharing or torrent systems, where customers receive data from customers that are already in service. The state-dependent rates in the first queue cover many prominent queueing systems. The first queue is considered to have finite capacity, the second queue has an infinite capacity. We further consider two variants of this model by allowing customers to jockey from the second queue to the first queue. In queueing theory, jockeying is called the possibility for waiting customers to move from one queue to another queue. These jockeying customers can then act as servers for the second queue. We reduce the number of unknown steady-state probabilities of this system in a considerable amount by a generating functions approach due to Avi-Itzhak and Mitrani [6]. Some steady-state quantities of interest are also derived.

In the third chapter, we cover a variety of different models of two-queue networks in which we equip each of the two queues with a finite number of servers and waiting positions. We consider different routines for the handling of arriving, blocked, overflowing and jockeying customers. These models can
be used to analyze for example call centers, telecommunication systems or traffic flows. The main difference between the queueing models presented in the second and third chapter is the finiteness of the capacity of the second queue. Moreover, the customers in the first queue no longer serve the customers in the second queue. The finiteness of the state space of these queueing systems gives - in contrast to the models presented in the second chapter - rise to additional boundary conditions. These boundary conditions and the special structure of the steady-state equations make it impossible to carry out the approach from the second chapter. Nevertheless, the number of steady-state equations that describe the system's behavior can be reduced substantially by exploiting a separation method due to Morrison [46, 48]. By using this separation technique, we will reduce the problem of solving the steady-state equations to the problem of solving a substantially smaller number of homogeneous linear equations in two sets of unknowns. With this approach, explicit formulas depending on these unknowns can be given for various steady-state quantities in an elegant form. The basic technique is to partition the state space into certain regions and boundaries and to separate the stationary probabilities within these regions. In every model, the separation leads to a set of eigenvalue problems for the separation constants. The eigenvalues are given by the roots of polynomial equations and are the pairwise distinct eigenvalues of real tridiagonal symmetric matrices as well. They possess an interlacing property, called the "Sturm sequence property", which reduces the computational complexity considerably. The desired probabilities are expressed as sums of eigenfunctions in terms of the eigenvalues. The number of eigenfunctions and therefore the number of coefficients to be determined in these representations is in general substantially smaller than the number of stationary probabilities. The coefficients are determined by the normalization condition and a set of linear equations that stems from the boundary conditions. The desired probabilities and steadystate quantities can be numerically determined once the coefficients and eigenvalues are numerically calculated. Some of the results were published in Sendfeld [59].

We give a detailed model and literature review and name main applications of our models at the beginning of every chapter. Due to the natural occurrence of overflow queueing problems, the related literature is vast, see for example Disney and König [17] for a broad overview. Two chapters devoted to the theory of queueing networks with restricted accessibility and overflow are found in Syski [63]. In Kosten [38], some aspects of networks with restricted accessibility are considered. Additionally, Koury et al. 39]
and Krieger et al. [40] give reviews of iterative numerical methods for overflow queueing models. A brief discussion of numerical methods for some two-queue overflow systems and further references are given in Ching and Ng [10].

## Introduction (german)

> Wohlunterrichete Menschen wissen, dass es unmöglich ist, die menschliche Stimme mit Kabeln zu übertragen. Bostoner Tageszeitung, 1867

Nach Leonard Kleinrock [37] ist die Warteschlangentheorie das Studium der Phänomene des Stehens, Wartens und Bedienens. So einfach diese Definition auch sein mag, so mannigfaltig sind die Anwendungen und realen Probleme, die als Warteschlangenmodelle beschrieben werden können. Die Grundsteine der Warteschlangentheorie wurden Anfang des frühen 20. Jahrhunderts gelegt. Die Pionierarbeiten von Frederik Ferdinand Wilhelm Johannsen [33], Tore Olaus Engset [21,22] und Agner Krarup Erlang [23] befassten sich mit dem Studium der Telefonnetzwerke mit begrenzten Kapazitäten. Seither entstehen unzählige Verallgemeinerungen und neue Anwendungen der Warteschlangentheorie. Noch heute zählen die Kommunika-tions- und Netzwerktechnik bei der Analyse von Telefon-, Internet- und Datenverkehr zu den wichtigsten Anwendungsbereichen der Warteschlangentheorie. Auch im Bandbreitenmanagement, im Gesundheitssystem, der Verkehrsregulierung, in der Versicherungsmathematik und bei der Analyse von Kundenströmen - also in fast jedem Bereich des alltäglichen Lebens, in dem „Stehen, Warten und Bedienen" eine Rolle spielen - ist die Warteschlangentheorie unerlässlich. Obwohl die Arbeiten von Johannsen, Engset und Erlang aus heutiger Sicht altmodisch erscheinen mögen - sogar nutzlos, sofern man einer Bostoner Zeitung von 1867 glaubt -, so sind deren Resultate auch 100 Jahre nach der Veröffentlichung von Erlangs erster Arbeit anwendbar und werden nach wie vor in der modernen Telekommunikation angewendet (siehe Stordahl [61 und ITU [32).

Verfügt ein Warteschlangensystem nur über eine begrenzte Kapazität an Service- oder Warteplätzen, so kann dies dazu führen, dass nur ein Teil der Serviceanfragen erfüllt wird. Aus ökonomischer Sicht ist es dann sinnvoll, das Langzeitverhalten des Systems zu studieren, um zum Beispiel die Kosten
abgelehnter Anfragen und die Einnahmen angenommener Anfragen auszubalancieren. In dieser Hinsicht sind unter anderem die erwartete Blockierwahrscheinlichkeit, die durchschnittlichen Abgangsraten von Kunden aus den Warteräumen oder Servern und die durchschnittliche Anzahl besetzter Warteplätze und Server von besonderem Interesse. Sind diese Größen unter gegebenen Voraussetzungen bekannt, so können Sie genutzt werden, um ein neues System optimal zu planen oder ein bestehendes zu optimieren. Hierbei können Zielvorgaben, wie die Sollauslastung oder die Einhaltung einer Kostenobergrenze, durch entsprechende Wahl der beeinflussbaren Systemparameter erfüllt werden. Die zur optimalen Kontrolle eines solchen Systems notwendigen Größen können jedoch im Allgemeinen nur für eine begrenzte Klasse von Warteschlangensystemen effizient berechnet werden. Je komplizierter die Abhängikeiten und Kundenströme in einem System sind, desto aufwändiger - wenn nicht unmöglich - ist dessen Analyse.

In dieser Arbeit präsentieren wir zwei allgemeine Klassen sogenannter Warteschlangennetzwerke mit Overflow. Die Netzwerke in diesen Klassen bestehen jeweils aus zwei Warteschlangen. Die Kapazität der ersten Warteschlange ist stets endlich. In den betrachteten Modellen entspricht die Dynamik der ersten Warteschlange den bekannten Erlang- oder EngsetVerlustsystemen oder ist diesen ähnlich. Wie in der Warteschlangentheorie üblich, werden wir im Folgenden den Begriff Kunde synonym für Anfrage und den Begriff Server für die Bedieneinheit verwenden. Ist die erste Warteschlange bis zu einer festgelegten Kapazitätsauslastung belegt, werden ankommenden Kunden bestimmte Wechselmöglichkeiten (Overflow) zur zweiten Warteschlange eingeräumt. Ein Kundenwechsel findet in den betrachteten Netzwerken zusätzlich mit einer Wahrscheinlichkeit $p \in[0,1]$ tatsäschlich statt, das heißt, der Wechselstrom wird mit dem Parameter $p$ gewichtet. Der Parameter $p$ kann als Steuer- oder Kontrollparameter verwendet werden oder den Grad der Ungeduld der wechselnden Kunden beschreiben.

Das erste Kapitel gibt einen kurzen Überblick über die allgemeine stochastische Struktur, die den betrachteten Netzwerken zu Grunde liegt. Prinzipiell verfügt jede der Warteschlangen über einen markovschen PoissonAnkunftsprozess und exponentialverteilte Servicezeiten.

Im zweiten Kapitel betrachten wir eine starke Verallgemeinerung eines Warteschlangenmodells von Perel und Yechiali [56], indem wir unter anderem Overflow zulassen. Bei dieser Verallgemeinerung eines Netzwerks aus zwei Warteschlangen sind die Ankunfts- und Serviceraten in der ersten Warteschlange variabel und abhängig vom Zustand dieser Warteschlange. Ferner werden die Kunden der zweiten Warteschlange von den Kunden der ersten
bedient, so dass die Servicerate der zweiten Warteschlange ebenfalls vom Zustand der ersten abhängt. Dies ist zum Beispiel der Fall bei Filesharing- oder Torrent-Systemen, in denen Kunden Daten von anderen Kunden herunterladen, die bereits Service erhalten. Die Variabilität der Ankunfts- und Serviceraten in der ersten Warteschlange ermöglicht die Analyse vieler prominenter Warteschlangensysteme. Die erste Warteschlange besitzt stets endliche Kapazität, die zweite hingegen unendliche Kapazität. Wir betrachten ferner zwei Varianten dieses Modells, in denen wir Kunden, die sich im Warteraum der zweiten Warteschlange befinden, die Möglichkeit geben, zur ersten Warteschlange zu wechseln. Diese Kunden wiederum können dann in der ersten Warteschlange als Server für die zweite Warteschlange fungieren. Wir reduzieren die Anzahl der unbekannten stationären Wahrscheinlichkeiten drastisch, indem wir die Struktur der erzeugenden Funktionen dieser Wahrscheinlichkeiten mit Hilfe eines Ansatzes von Avi-Itzhak und Mitrani [6] analysieren. Ferner leiten wir Formeln für einige der wichtigsten stationären Größen her.

Im dritten Kapitel untersuchen wir zahlreiche unterschiedliche Warteschlangennetzwerke aus zwei Warteschlangen, in denen jede der beiden Warteschlangen mit einem Warteraum mit endlicher Kapazität und einer begrenzten Anzahl an Servern ausstattet ist. Wir betrachten unterschiedliche Routinen für die Behandlung ankommender, blockierter und wechselnder Kunden. Diese Modelle können verwendet werden, um zum Beispiel Call Center, Telekommunikationssysteme oder Verkehrsflüsse zu analysieren. Der Hauptunterschied zwischen den Modellen im zweiten und dritten Kapitel ist daher die nun endliche Kapazität der zweiten Warteschlange. Ferner werden die Kunden in der zweiten Warteschlange nicht mehr von den Kunden in der ersten Warteschlange bedient. Im Gegensatz zu den Modellen aus dem zweiten Kapitel gibt die endliche Kapazität der zweiten Warteschlange hier Anlass zu zusätzlichen Randbedingungen. Diese Randbedingungen und die spezielle Struktur der Gleichgewichtsgleichungen machen es unmöglich, diese Systeme mit den Methoden aus dem zweiten Kapitel zu analysieren. Jedoch reduzieren wir die Anzahl der unbekannten stationären Wahrscheinlichkeiten wiederum drastisch, indem wir eine Separations-Technik von Morrison [46, 48] verwenden und verallgemeinern. Mit Hilfe dieser Technik reduzieren wir das Problem der Lösung der Gleichgewichtsgleichung auf das Problem der Lösung einer erheblich kleineren Anzahl an homogenen Gleichungen und eines Eigenwertproblems. Mit dieser Methode können explizite Formeln für die verschiedensten stationären Größen in Abhängigkeit von diesen Eigenwerten in eleganter Form angeben werden. Die grundlegende

Technik basiert auf einer Partition des Zustandsraums in Mengen bestimmter innerer Punkte und Randpunkte. In den inneren Bereichen werden die stationären Wahrscheinlichkeiten dann in eine Summen-Produkt-Form zerlegt. Diese Zerlegung führt zu Eigenwertproblemen für tridiagonale symmetrische Matrizen, zu deren Lösung die stationären Gleichungen in den Randpunkten verwendet werden. Die Eigenwerte werden dabei als Nullstellen polynomialer Gleichungen bestimmt und sind paarweise verschieden und reell. Sie besitzen die sogenannte Sturm-Folgen- oder Verzahnungseigenschaft, die den numerischen Aufwand bei deren Berechnung erheblich reduziert. Die gesuchten Wahrscheinlichkeiten werden als gewichtete Summen der Eigenfunktionen in Abhängigkeit der Eigenwerte dargestellt. Die Koeffizienten in dieser Darstellung werden mit Hilfe der Normierungsbedingung und der linearen Randbedingungen bestimmt. Die gesuchten Wahrscheinlichkeiten können mit numerischen Verfahren berechnet werden, sobald die Koeffizienten und Eigenwerte berechnet worden sind. Einige der Resultate wurden in Sendfeld [59] veröffentlicht.

Wir geben zu Beginn jedes Kapitels eine detaillierte Literaturübersicht und Anwendungsbeispiele für die präsentierten Modelle. Aufgrund des natürlichen Auftretens von Warteschlangennetzwerken mit Wechselmöglichkeiten ist die verwandte Literatur reichhaltig. Eine breite Übersicht ist zum Beispiel in Disney und König [17] gegeben. Syski widmet in [63] zwei Kapitel der Analyse von Warteschlangennetzwerken mit begrenzter Kapazität und Wechselmöglichkeiten. Weitere Aspekte von Netzwerken mit begrenzter Kapazität werden in Kosten [38] betrachtet. Koury et al. [39] und Krieger et al. [40] geben Überblicke über iterative numerische Methoden für Warteschlangen mit Wechselmöglichkeiten. Eine kurze Diskussion numerischer Methoden für Zwei-Server-Warteschlangen mit Wechselmöglichkeiten und weitere Referenzen finden sich in Ching und Ng [10].

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## Chapter 1

## Preliminaries

This chapter serves as a brief introduction to the theory of Markov jump processes in continuous time in order to classify the stochastic models presented in this thesis. The well-known classical results on general Markov chains in continuous time in this chapter can be found in various textbooks and are taken from Alsmeyer [3], Asmussen [5] and Bremaud [7], whereas the results on quasi birth and death chains are due to Latouche and Ramaswami [41] and Neuts [53].

### 1.1 Markov chains in continuous time

In this section, we recapitulate the definition and the basic structure of a continuous time Markov process or Markov jump process on a countable state space. Let $\mathfrak{S}$ be a countable nonempty set, called the state space, and let $X=\left(X_{t}\right)_{t \in[0, \infty)}$ be a stochastic process on the probability space $(\Omega, \mathfrak{A}, P)$ with values in $\mathfrak{S}$. Let $\mathscr{F}=\left(\mathscr{F}_{t}\right)_{t \in[0, \infty)}$ be the canonical filtration of $X$, that is, $\mathscr{F}_{t}$ is the $\sigma$-algebra generated by all $X_{s}$ for $s \leq t$. The process $X$ is called Markov process if the Markov property holds:

$$
P\left(X_{t} \in A \mid \mathscr{F}_{s}\right)=P\left(X_{t} \in A \mid X_{s}\right) \quad P \text {-a.s. }
$$

for all $s, t \in[0, \infty)$ with $s<t$ and every $A \subset \mathfrak{S}$. The process is called homogeneous if the transition kernels $\mathbb{P}_{s, t}\left(X_{s}, A\right)=P\left(X_{t} \in A \mid X_{s}\right)$ can be chosen so as to depend on $s$ and $t$ only through the difference $t-s$, i.e., if

$$
P\left(X_{t} \in A \mid X_{s}=x\right)=\mathbb{P}_{0, t-s}(x, A) \quad P^{X_{s}} \text {-a.s. }
$$

This will be assumed in the following. We will write $\mathbb{P}_{0, t}=\mathbb{P}_{t}$, where $\mathbb{P}_{0}(x, \cdot)$ is the Dirac measure $\delta_{x}$ in $x$, and call $\mathbb{P}_{t}$ the $t$-step transition kernel. The family $\left(\mathbb{P}_{t}\right)_{t \in[0, \infty)}$ of the $t$-step transition kernels satisfies the Kolmogorov-Chapman equations:

$$
\begin{equation*}
\mathbb{P}_{s+t}=\mathbb{P}_{s} \mathbb{P}_{t} \tag{1.1.1}
\end{equation*}
$$

or equivalently

$$
\mathbb{P}_{s+t}(x, A)=\int_{\mathfrak{S}} \mathbb{P}_{t}(y, A) \mathbb{P}_{s}(x, d y)
$$

for all $s, t \in[0, \infty)$ and every $A \subset \mathfrak{S}$.
In our setting with denumerable state space $\mathfrak{S}$, we may identify the transition kernel $\mathbb{P}_{t}$ with the matrix $\boldsymbol{P}(t)=\left(p_{i, j}(t)\right)_{i, j \in \mathfrak{S}}$ and interpret the Kolmogorov-Chapman equations in the sense of matrix multiplication. We have
(i) $\boldsymbol{P}(t)$ is a stochastic matrix,
(ii) $\boldsymbol{P}(0)$ is the identity matrix and
(iii) $\boldsymbol{P}(t+s)=\boldsymbol{P}(t) \boldsymbol{P}(s)$ for all $s, t \in[0, \infty)$.

Moreover, we suppose that $\lim _{t \rightarrow 0} \boldsymbol{P}(t)=\boldsymbol{P}(0)$. With this assumption we can show that $p_{i, j}(t)$ is continuously differentiable for $t>0$ and differentiable from the right at 0 , i.e., the limit

$$
q_{i, j}=\lim _{t \rightarrow \infty} \frac{p_{i, j}(t)-p_{i, j}(0)}{t}
$$

exists and is finite for $i \neq j$ but maybe infinite for $i=j$. We also have $q_{i, j} \geq 0$ for all $i \neq j$. The matrix $Q=\left(q_{i, j}\right)_{i, j \in \mathfrak{S}}$ is called the infinitesimal generator, rate matrix or $Q$-matrix of the process $X . Q$ is called conservative if

$$
\sum_{j \neq i} q_{i, j}=-q_{i, i}<\infty
$$

or equivalently, in matrix form, if $Q \mathbf{1}=0$ holds, where $\mathbf{1}$ is the vector with all entries equal to 1 . It is useful to let $q_{i}=-q_{i i}$ for $i \in \mathfrak{S}$. We assume in the following that $Q$ is conservative.

The process $X$ has a fundamental jump structure. Let $S_{0}=0<S_{1}<$ $S_{2}<\ldots$ be the times of successive jumps of $X$, let $T_{n}=S_{n+1}-S_{n}$ for $n \geq 0$ be the associated sojourn times and let the sequence of states visited be given by $Y_{n}=X_{S_{n}}$ for $n \geq 0$. There are two phenomena which desire
further attention: absorption and explosion. The process may be absorbed in the sense that there is a last finite $S_{n}$. In this case we may set $T_{k}=\infty$ and $Y_{k}=X_{S_{n}}$ for $k \geq n$. The case of an explosion of the process, i.e., an accumulation of infinitely many jumps in finite time is more involved. The following condition for a conservative $Q$-matrix prevents the process $X$ from being explosive.

Proposition 1.1.1. Let $Q$ be conservative. Then $X$ is nonexplosive if the condition $\sup _{i \in \mathfrak{S}} q_{i}<\infty$ holds. That is for example the case if $\mathfrak{S}$ is finite.

The basic structure of $X$ up to the time of explosion is very simple (see Theorem II 1.2 in (5):

Proposition 1.1.2. The joint distribution of the sequences $\left(Y_{n}\right)_{n \geq 0}$ and $\left(T_{n}\right)_{n \geq 0}$ before explosion is given by:
(i) The sequence $\left(Y_{n}\right)_{n \geq 0}$ of states visited is a Markov chain.
(ii) There exist $\lambda(i) \geq 0$ such that $T_{0}, T_{1}, \ldots$ are independent and $T_{k}$ is exponentially distributed with parameter $\lambda\left(Y_{k}\right)$ for $k \geq 0$ given $\left(Y_{n}\right)_{n \geq 0}$.

The transition matrix $\hat{P}=\left(\hat{p}_{i, j}\right)_{i, j \in \mathfrak{S}}$ of the embedded Markov chain $\left(Y_{n}\right)_{n \geq 0}$ is linked to the generator $Q$ by

$$
\hat{p}_{i, i}=0 \quad \text { and } \quad \hat{p}_{i, j}=\frac{q_{i, j}}{q_{i}},
$$

if $0<q_{i}<\infty$ and $\hat{p}_{i, j}=\delta_{i j}$ for all $j \in \mathfrak{S}$ if $q_{i}=0$, where $\delta_{i j}$ is the Kronecker function, i.e., $\delta_{i j}=1$ for $i=j$ and 0 otherwise. Thus, absorption is excluded if $0<q_{i}<\infty$ for all $i \in \mathfrak{S}$.

This basic structure suggests that one can construct a Markov jump process in continuous time by specifying a conservative $Q$-matrix: Start the chain at an arbitrary state and let it reside in state $i_{0}$ for an exponential holding time with parameter $\lambda(i)=\sum_{j \neq i_{0}} q_{i_{0}, j}=q_{i_{0}}$ and jump to the state $i_{1}$ with probability $\hat{p}_{i_{0}, i_{1}}=q_{i_{0}, i_{1}} / q_{i_{0}}$. Indeed, under the condition that the resulting process is nonexplosive, one gets a continuous time Markov chain with infinitesimal generator $Q$. The assumption that the process in nonexplosive can of course be avoided. Moreover, the resulting Markov process is called the minimal construction and is, in addition to the conditions given in Proposition 1.1.1, nonexplosive if the embedded Markov chain is recurrent.

The concepts of irreducibility, recurrence and transience can be defined via the embedded Markov chain. In this manner, a Markov jump process has the respective property if and only if the embedded Markov chain does. An
ergodic process is an irreducible recurrent process with stationary measure having finite mass. We have the following stability theorem (see Theorem II 4.3 in (5)

Theorem 1.1.3. An irreducible nonexplosive Markov jump process is ergodic if and only if one can find a probability vector $p$ with $p Q=0$. In that case $p$ is the stationary distribution, i.e., $p \boldsymbol{P}(t)=\boldsymbol{P}(t)$ for every $t \in[0, \infty)$.

All stochastic processes considered in this thesis arise from a minimal construction based on a conservative rate matrix $Q$. The resulting processes are nonexplosive since one of the properties from Proposition 1.1.1 holds in every case. In the second chapter, we consider queueing systems with infinite state space. Thus, we have to guarantee the ergodicity of the stochastic processes. The main ergodicity criterion that serves this purpose is stated in the next section. Once ergodicity is established, we can focus on finding the unique normalized solution of the equation $p Q=0$. In the third chapter, we consider finite queueing systems. Ergodicity is in this case given by the irreducibility of $Q$ and the finiteness of the state space. The main objective is again to find the unique normalized solution of the equation $p Q=0$.

### 1.2 Quasi birth and death processes

We consider a two-dimensional Markov jump process $L=\left(L_{1, t}, L_{2, t}\right)_{t \in[0, \infty)}$ in continuous time on the state space $\mathfrak{S}=\{0, \ldots, N\} \times \mathbb{N}_{0}$. The first component $\left(L_{1, t}\right)_{t \in[0, \infty)}$ is called the phase or the phase process. The second component $\left.\left(L_{2, t}\right)_{t \in[0, \infty}\right)$ is called the level or the level process. Let

$$
l(m)=\{0, \ldots, N\} \times\{m\}
$$

be the collection of states in level $m$ for $m \geq 0$. We will also call $l(m)$ the level $m$

Definition 1.2.1. A Markov chain on $\mathfrak{S}$ is called a quasi birth and death process ( $Q B P$ ) if transitions are restricted to one-step transitions from states in one level to states in the same level or to states in the neighboring levels.

By this definition, the transitions of $L$ are restricted to one-step transitions from states in one level to states in the same level or to states in the neighboring levels, i.e., transitions from the states in $l(m)$ to the states in $l\left(m^{\prime}\right)$ can occur if and only if $m^{\prime}=m-1, m$, or $m+1$. Furthermore, we assume that the transitions are level-independent, i.e., the transition rate
from states in $l(m)$ to states in $l\left(m^{\prime}\right)$ depends on $m$ and $m^{\prime}$ only through the difference $m-m^{\prime}$. The QBP is called homogeneous in this case.

The states can be ordered lexicographically with respect to the first component in the order

$$
(0,0),(1,0,), \ldots,(N, 0),(0,1),(1,1), \ldots,(N, 1), \ldots
$$

In this order, the transition rate matrix or infinitesimal generator $Q$ has a block-tridiagonal form of the following type:

$$
Q=\left(\begin{array}{ccccc}
B & A_{0} & 0 & 0 & \ldots \\
A_{2} & A_{1} & A_{0} & 0 & \ldots \\
0 & A_{2} & A_{1} & A_{0} & \ldots \\
0 & 0 & A_{2} & A_{1} & \ldots \\
\vdots & \vdots & \vdots & \vdots & \ddots
\end{array}\right)
$$

where $B, A_{0}, A_{1}$ and $A_{2}$ are quadratic matrices. The matrices $A_{0}$ and $A_{2}$ are nonnegative; the matrices $B$ and $A_{1}$ have nonnegative off-diagonal entries and negative entries on their diagonals. The sum of the elements in each row of $Q$ equals 0 . The matrix $B$ contains the transition rates from states in level $l(0)$ to states the level $l(0)$. The matrices $A_{0}, A_{1}$ and $A_{2}$ are generated by the rates from states in level $l(m)$ to states in level $l(m+1)$, $l(m)$ and $l(m-1)$, respectively, for $m \geq 0$.

Quasi birth and death processes belong to the class of stochastic processes that can be analyzed by the matrix-analytic method. The origins of these technique go back to Marcel Neuts (see [51], 52] and [53]), whose research originated the matrix-geometric distribution and phase-type processes. The term "geometric" stems from the generalization of the structure of the stationary measure of the one-dimensional birth and death process on the nonnegative integers to the two-dimensional case. The next theorem displays this geometric structure of the stationary probability measure of a positive-recurrent QBP. The theorem and its proof can be found in Latouche and Ramaswami [41] (see Theorem 6.4.1).

Theorem 1.2.2. Assume that the continuous time quasi birth and death process is positive-recurrent. Let the stationary distribution $p$ of the process be partitioned by levels into subvectors $p_{m}, m \geq 0$. Then, the stationary probability distribution is such that

$$
p_{m}=p_{0} R^{m} \quad \text { for } m \geq 0
$$

where the matrix $R$ records the rate of sojourn in the states of level $l(m+1)$ per unit of the local time of $l(m)$. Furthermore, we have that $R=A_{0} S$, where the matrix $S$ records the expected sojourn time in the states of $l(m)$, starting from $l(m)$, before the first visit to $l(m-1)$.

The following stability theorem taken from [41] (see Theorem 7.2.4) gives a necessary and sufficient condition for stability and the existence of a unique stationary probability measure. The proof will be omitted and can be found in 41]; an earlier reference is [53] (see Theorem 3.1.1).

Theorem 1.2.3. Consider an irreducible, continuous time QBP with a finite number of phases and assume that the matrix $A=A_{0}+A_{1}+A_{2}$ is irreducible. The process is positive recurrent if and only if

$$
\begin{equation*}
\pi A_{0} \mathbf{1}<\pi A_{2} \mathbf{1} \tag{1.2.1}
\end{equation*}
$$

where $\mathbf{1}=(1, \ldots, 1)^{\top} \in \mathbb{R}^{N+1}$ and $\pi=\left(\pi_{0}, \ldots, \pi_{N}\right)$ is the existing and unique solution of the equations $\pi A=0$ and $\pi \mathbf{1}=1$. The process is recurrent if $\pi A_{0} \mathbf{1}=\pi A_{2} \mathbf{1}$ and transient if $\pi A_{0} \mathbf{1}>\pi A_{2} \mathbf{1}$.

The stability condition $\pi A_{0} \mathbf{1}<\pi A_{2} \mathbf{1}$ states that the phase-averaged rate for level transitions from $l(m)$ to $l(m+1)$ must be smaller than the phase-averaged rate for level transitions from $l(m)$ to $l(m-1)$ for every $m \geq 1$.

## Chapter 2

## Overflow to an infinite queue and customers as servers

### 2.1 Model overview

In this chapter, we consider a fairly general open queueing network consisting of two queues having an overflow capability from the first to the second queue. The arrival and service rates are state-dependent and the customers in the first queue act as servers for the second queue, i.e., the service rate in the second queue depends on the state of the first queue.

Consider two queues $Q_{1}$ and $Q_{2}$. Let the total number of customers in queue $Q_{1}$ be given by the state of a finite birth and death chain in continuous time with state space $\{0, \ldots, N\}$, reflecting barriers and state-dependent birth and death rates. Let the birth rate be $\lambda_{1, n}>0$ if the chain is in state $n=0, \ldots, N-1$ and the death rate be $\mu_{1, n}>0$ if the chain is in state $n=1, \ldots, N$. The state of the birth and death chain corresponds to the total number of customers in the queue $Q_{1}$, i.e., the number of customers in the waiting room plus the number of customers in service. If this number is $n$, then the arrival rate of the queue is $\lambda_{1, n}>0, n=0, \ldots, N-1$, and service rate is $\mu_{1, n}>0, n=1, \ldots, N$. Let $\lambda_{1, N}>0$ be the potential birth rate in state $N$, i.e., the potential arrival rate if $Q_{1}$ is in state $N$. Let $\mu_{1,0}=0$ and let $L_{1}$ be the (stationary) total number of customers in $Q_{1}$, i.e., the number of occupied waiting positions plus the number of customers being served (under stationary conditions). If $Q_{1}$ is fully occupied, i.e., if $L_{1}=N$, then the arrival stream of $Q_{1}$ is weighted with $p \in[0,1]$ and directed to the second queue. This procedure is called overflow. Furthermore, the fraction $1-p$ of arriving customers is lost in this case. The parameter $p$ can be
used as a control parameter or to model the customers' impatience. Assume that the second queue has no other external arrivals. In other words, the arrival rate at $Q_{2}$ is $p \lambda_{1, N}$ in the case $L_{1}=N$ and 0 otherwise. $Q_{2}$ has one server with exponentially distributed service times and a waiting room with infinite capacity. The service rate is variable and depends on the number of customers $L_{1}$ in $Q_{1}$ and is $L_{1} \mu_{2}$, where $\mu_{2}>0$. The customers in each queue are served in their order of arrival. This basic model is discussed in the next section.

We further consider two variants of the basic model by allowing customers to jockey from the second queue to the first queue. In queueing theory, jockeying is called the possibility for waiting customers to move from one queue to another queue. In the first variant, the first customer or more generally the first $k$ customers, $1 \leq k \leq N-1$, from $Q_{2}$, if one is present, are forced to move to $Q_{1}$ as soon as $Q_{1}$ empties. These customers can then act as servers for the second queue. We call this model the model with limited jockeying because the number of jockeying customers limited to a fixed value smaller than the capacity of the first queue. Therefore, some customers might have to stay in $Q_{2}$ although $Q_{1}$ has not reached its capacity bound. In the second variant, as soon as $Q_{1}$ empties, it is filled with the customers of $Q_{2}$ until it reaches its capacity bound or $Q_{2}$ empties. This jockeying procedure is called unlimited jockeying. The transferral of customers does not result in service interruptions, because the service rate in $Q_{2}$ is 0 at the time points of customer transferrals. See Figure 2.1 for a schematic overview of the three models.

As indicated in Figure 2.1 it is possible to let $Q_{2}$ have a Poisson arrival stream being independent of $Q_{1}$ with intensity $\lambda_{2}>0$, so that the arrival rate to $Q_{2}$ is $\lambda_{2}$ for $L_{1}=0, \ldots, N-1$ and $p \lambda_{1}+\lambda_{2}$ for $L_{1}=N$. In order to simplify the presentation of the results, we omit the derivations in this case and discuss the solution in Section 2.2.5.

Our basic model covers for example the case of $Q_{1}$ being an $M_{(n)} / M_{(n)} / 1 / N-1$-queue with one server and $N-1$ waiting positions or any $M_{(n)} / M_{(n)} / K / N-K$ queueing system with $K$ servers and $N-K$ waiting positions, where $N \geq K>0$. The index ( $n$ ) indicates state-dependent rates. Following the notation of van Doorn [19], the queueing systems with one server and $N-1$ waiting positions in $Q_{1}$ for example might be labeled with $\left(M_{(n)} / M_{(n)} / 1 / N-1\right)_{p-\text { overflow }} / M_{(n)} / 1$.


Figure 2.1: Model overview: Customers acting as servers, state-dependent service and arrival rates.

## Solution approach

We are interested in the two-dimensional server and waiting room demand process of this model, embedded at the time instants of arrivals to $Q_{1}$ and $Q_{2}$ and departures from $Q_{1}$ and $Q_{2}$. This process is a Markov chain with state space $\mathfrak{S}=\{(n, m) \mid n=0, \ldots, N, m \geq 0\}$, where the first and second component of $(n, m) \in \mathfrak{S}$ correspond to the number of occupied servers and/or waiting positions in $Q_{1}$ and $Q_{2}$, respectively.

We solve the system of steady-state equations that describe the systems' dynamics by exploiting the probability generating functions of the number of customers in $Q_{2}$, see for example Avi-Itzhak and Mitrani [6] and Perel and Yechiali [56]. The solution of these equations is given in terms of only $N$ unknowns $p_{1,0}, \ldots, p_{N, 0}$, where $p_{n, m}$ is the steady-state probability of having $n$ customers in $Q_{1}$ and $m$ customers in $Q_{2}$. These unknowns can then be determined by $N$ linear equations in the unknowns given later. The model presented in [56] is an important special case of our birth and death queueing system for the case $p=0, \lambda_{1, n}=\lambda_{1}>0, \mu_{1, n+1}=\mu_{1}>0, n=0, \ldots, N-1$ and $\lambda_{2}>0$.

## Applications

Our models can be used to analyze for example file sharing or torrent systems. In these systems, data files are shared, exchanged and spread over
the internet or a private intranet by letting downloading customers act as servers for further customers. Due to cost and bandwidth limitations it is a common case to have a limited number of primary download channels (i.e., the servers in $Q_{1}$ ), a limited number of primary queueing positions for these channels (i.e., the waiting positions in $Q_{1}$ ) and an unlimited number of secondary download channels (i.e., the positions in $Q_{2}$ ). A server or host provides a data file via the servers in the primary channels. This file is typically divided into several data portions which are downloaded by the customers in service in the primary channels. When the primary download and queueing positions are occupied, an incoming customer can decide whether to join one of the secondary download channels in $Q_{2}$ or not - the probability of joining may equal $p$ for all potential secondary customers. Additionally, the parameter $p$ can be regarded as a control parameter. The arrival stream of the secondary channels may consist of the overflow from the primary channels and/or an independent arrival stream. The secondary download channels are served by the customers that are present in the primary channels, i.e., the secondary customers can download the requested file or the data portions of this file from the customers in the primary channels. These torrent systems make hosting of a file with a potentially unlimited number of downloaders affordable because the costs of the secondary channels can be assigned to the primary and secondary customers. The download speed in the primary and secondary download channels can increase with the number of customers who are receiving service (in the primary channel). This is due to the fact that when customers download the same file at the same time, they can upload portions of the file to each other. In addition, the data portions of the file can be downloaded in an arbitrary order. Thus, a secondary customer can finish the service earlier than a primary customer who arrived earlier and who participated in serving this secondary customer. The download speed in the primary channel may also decrease with the number of customers because of a limitation of the bandwidth.

Further applications of our model are to the SETI@home and the GIMPS project. The SETI@home project (Search for ExtraTerrestrial Intelligence), initiated by the Space Sciences Laboratory of the University of California, Berkeley, is described in Perel and Yechiali [56]. This project is searching for extraterrestrial intelligence by radio telescopes. The GIMPS project (Great Internet Mersenne Prime Search) launched by the Mersenne Research Incorporation is a prominent and popular project searching for Mersenne prime numbers. In both projects, huge amounts of data have to be processed. This is done by institutions or private persons who install special computer
programs on their computers. Whenever a computer equipped with this software is idle, these programs are activated and data can be processed.

Having these examples in mind, it is very natural to assume variable service and arrival rates due to bandwidth or CPU limitations. The models presented in this chapter achieve the purpose of having variable arrival and service rates.

## Further related literature

The models presented in this chapter can be represented as quasi birth and death processes (QBP) presented in Section 1.2, see for example Latouche and Ramaswami [41. Consequently, their analysis can be carried out using a matrix-geometric approach, see Neuts [53]. Furthermore, all models can be formulated as Markov-modulated queues. For related literature on both the matrix-geometric approach and Markov-modulated queues see for example Asmussen [4]. For the latter see Mahabhashyam et al. [43], Muscariello et al. 50 and Takine [64] and the references therein.

Overflow queueing models are widespread in literature. Van Doorn 19 and Parthasarathy and Sudhesh [55] study the interoverflow time distribution of a finite birth and death queue model as presented for $Q_{1}$. Koury et al. [39] and Krieger et al. [40] give reviews of iterative numerical methods for overflow queueing models. A brief discussion of numerical methods for some two-queue overflow systems and further references are given in Ching and Ng [10]. While most of these formulations are of primary interest when the focus is on numerical results, the method used in the following gives deep insight into the structure of the stability conditions and solutions. Nevertheless, we will use the stability theorem for QBP presented in Section 1.2 to derive necessary and sufficient conditions for stability in every model. Related overflow models are studied in Chapter 3 of this thesis, in van Doorn [19] and Guérin, Lien [27] and the referenced literature therein using a variety of different techniques. Further related literature is mentioned in Chapter 3, Section 3.1 of this thesis.

### 2.2 Customers as servers: Basic model

### 2.2.1 Model description and steady-state equations

Let the number of customers in $Q_{1}$ be given by the state of a finite birth and death chain with state space $\{0, \ldots, N\}$, birth rate $\lambda_{1, n}>0$ if the chain
is in state $n=0, \ldots, N-1$ and death rate $\mu_{1, n}>0$ if the chain is in state $n=1, \ldots, N$. Let $\lambda_{1, N}>0$ be the potential birth rate if $Q_{1}$ is in state $N$. Let $\mu_{1,0}=0$ and let $L_{1}$ be the (stationary) total number of customers in $Q_{1}$. Let $Q_{2}$ have one server with exponentially distributed service times and a waiting room with infinite capacity. The service rate in $Q_{2}$ is $L_{1} \mu_{2}$ if $L_{1}$ customers are present in $Q_{1}$, where $\mu_{2}>0 . Q_{2}$ is fed by a portion of the rejected customers of $Q_{1}$. If queue one is fully occupied, i.e., if $L_{1}=N$, then the arrival stream of $Q_{1}$ is weighted with $p \in(0,1]$ and directed to $Q_{2}$ while the fraction $1-p$ of arriving customers is lost in this case. The customers in each queue are served in their order of arrival. The underlying Markov process is irreducible since there exists a path with positive probability from the state $(0,0)$ to every other state. Furthermore, the Markov process in non-explosive by Proposition 1.1.1. The transition rate diagram is given in Figure 2.2, where $\lambda_{2}=0$ in this section.


Figure 2.2: Basic model: Transition rate diagram.

The first queue is a finite one-dimensional birth and death chain and the distribution of $L_{1}$ is therefore well known, see for example or Cohen [12].

Setting $\rho_{n}=\lambda_{1, n} / \mu_{1, n+1}, n=0, \ldots, N-1$, we get

$$
\begin{equation*}
P\left(L_{1}=n\right)=\frac{\prod_{i=0}^{n-1} \rho_{i}}{1+\sum_{j=1}^{N} \prod_{i=0}^{j-1} \rho_{i}} \text { and } E L_{1}=\frac{\sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}}{1+\sum_{n=1}^{N} \prod_{i=0}^{n-1} \rho_{i}} \tag{2.2.1}
\end{equation*}
$$

where the empty product is set to 1 . For example, with constant service rates $\lambda_{1, n}=\lambda_{1}$, arrival rates $\mu_{1, n}=\mu_{1}$ and $\lambda_{1} \neq \mu_{1}$ we arrive at the classical $M / M / 1 / N-1$ queue:

$$
P\left(L_{1}=n\right)=\frac{1-\frac{\lambda_{1}}{\mu_{1}}}{1-\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N+1}}\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{n} \text { and } E L_{1}=\frac{\lambda_{1}}{\mu_{1}-\lambda_{1}}-\frac{(N+1) \lambda_{1}^{N+1}}{\mu_{1}^{N+1}-\lambda_{1}^{N+1}} .
$$

In the case $\lambda_{1}=\mu_{1}$, we have $P\left(L_{1}=n\right)=1 /(N+1)$ and $E L_{1}=N / 2$. For service rates $\lambda_{1, n}=\lambda_{1}$ and arrival rates $\mu_{1, n}=n \mu_{1}$, the first queue is an Erlang loss system, i.e., the $M / M / N / 0$ queue:

$$
P\left(L_{1}=n\right)=\frac{\frac{\rho^{n}}{n!}}{1+\sum_{n=1}^{N} \frac{\rho^{n}}{n!}} \text { and } E L_{1}=\frac{\sum_{n=1}^{N} \frac{\rho^{n}}{(n-1)!}}{1+\sum_{n=1}^{N} \frac{\rho^{n}}{n!}} \text {, }
$$

where $\rho=\lambda / \mu$ (see Erlang [24]).
It is shown in van Doorn [19] that the overflow process from the first queue is a renewal process of hyperexponential type. An expression for the Laplace transform of the interoverflow time, i.e., the time between successive moments of overflow, can be given. The intensity $\Lambda$ of the interoverflow process is

$$
\begin{equation*}
\Lambda=\frac{p \lambda_{1, N} \prod_{n=0}^{N-1} \rho_{n}}{1+\sum_{n=1}^{N} \prod_{i=0}^{n-1} \rho_{i}} \tag{2.2.2}
\end{equation*}
$$

Parthasarathy and Sudhesh [55] express the interoverflow time distribution as a power series expansion and as a hyperexponential distribution in closed form. Additionally, a closed-form expression in terms of the system parameters for the $r$-th moment of the overflow process is given.

Let $L_{1}$ and $L_{2}$ be the (stationary) total number of customers in $Q_{1}$ and
$Q_{2}$, respectively, and

$$
p_{n, m}=P\left(L_{1}=n, L_{2}=m\right)
$$

for $n=0, \ldots, N$ and $m \geq 0$. The unknown quantities $P\left(L_{2}=m\right), m \geq 0$, $E L_{2}$ and $\operatorname{Cov}\left(L_{1}, L_{2}\right)$ are of special interest and will be derived in the next section.

The bivariate server and waiting room demand distribution $\left(p_{n, m}\right)_{n=0, \ldots, N, m \geq 1}$ in the case $\lambda_{2}=0$ is the unique nonnegative and normalized solution (i.e., $\sum_{n=0}^{N} \sum_{m \geq 0} p_{n, m}=1$ ) of the following steady-state equations:

$$
\begin{align*}
& \left(\lambda_{1, n}\left(1-\delta_{n N}\right)+p \lambda_{1, N} \delta_{n N}+\left(1-\delta_{n 0}\right) \mu_{1, n}+\left(1-\delta_{m 0}\right) n \mu_{2}\right) p_{n, m} \\
& \quad=\left(1-\delta_{n 0}\right) \lambda_{1, n-1} p_{n-1, m}+\left(1-\delta_{n N}\right) \mu_{1, n+1} p_{n+1, m} \\
& \quad \quad+\left(1-\delta_{m 0}\right) \delta_{n N} p \lambda_{1, N} p_{n, m-1}+\left(1-\delta_{n 0}\left(1-\delta_{m 0}\right)\right) n \mu_{2} p_{n, m+1} \tag{2.2.3}
\end{align*}
$$

for $n=0, \ldots, N$ and $m \geq 0$, where $\delta_{i j}$ is the Kronecker function, i.e., $\delta_{i j}=1$ for $i=j$ and 0 otherwise.

### 2.2.2 Necessary and sufficient stability condition

In this section, we derive the necessary and sufficient condition for the existence of a normalized solution of the steady-state equations (2.2.3). First, we give an intuitive argument for the stability condition and the possibility of reducing the infinite number of unknowns in these equations to only $N$ unknowns, namely $p_{1,0}, \ldots, p_{N, 0}$.

The equations (2.2.3) for $m=0$ and $n=0, \ldots, N$, i.e.,

$$
\begin{align*}
& \lambda_{1,0} p_{0,0}= \mu_{1,1} p_{1,0},  \tag{2.2.4}\\
&\left(\lambda_{1, n}+\mu_{1, n}\right) p_{n, 0}=\lambda_{1, n-1} p_{n-1,0}+\mu_{1, n+1} p_{n+1,0}+n \mu_{2} p_{n, 1}, \\
& n=1, \ldots, N-1,  \tag{2.2.5}\\
&\left(p \lambda_{1, N}+\mu_{1, N}\right) p_{N, 0}= \lambda_{1, N-1} p_{N-1,0}+N \mu_{2} p_{N, 1}, \tag{2.2.6}
\end{align*}
$$

give after summation over $n=0, \ldots, N$

$$
\begin{equation*}
p \lambda_{1, N} p_{N, 0}=\mu_{2} \sum_{n=1}^{N} n p_{n, 1} . \tag{2.2.7}
\end{equation*}
$$

The equations (2.2.3) for $m \geq 1$ and $n=0, \ldots, N$, i.e.,

$$
\begin{align*}
& \lambda_{1,0} p_{0, m}=\mu_{1,1} p_{1, m},  \tag{2.2.8}\\
&\left(\lambda_{1, n}+\mu_{1, n}+n \mu_{2}\right) p_{n, m}=\lambda_{1, n-1} p_{n-1, m}+\mu_{1, n+1} p_{n+1, m}+n \mu_{2} p_{n, m+1}, \\
& n=1, \ldots, N-1,  \tag{2.2.9}\\
&\left(p \lambda_{1, N}+\mu_{1, N}+N \mu_{2}\right) p_{N, m}=\lambda_{1, N-1} p_{N-1, m}+p \lambda_{1, N} p_{N, m-1}+N \mu_{2} p_{N, m+1}, \tag{2.2.10}
\end{align*}
$$

yield by summing over $n=0, \ldots, N$

$$
\begin{equation*}
p \lambda_{1, N} p_{N, m}+\mu_{2} \sum_{n=1}^{N} n p_{n, m}=p \lambda_{1, N} p_{N, m-1}+\mu_{2} \sum_{n=1}^{N} n p_{n, m+1} . \tag{2.2.11}
\end{equation*}
$$

From (2.2.7) and (2.2.11) we obtain

$$
\begin{equation*}
p \lambda_{1, N} p_{N, m}=\mu_{2} \sum_{n=1}^{N} n p_{n, m+1} \tag{2.2.12}
\end{equation*}
$$

for all $m \geq 0$ by induction. Summation over $m \geq 0$ gives

$$
\begin{equation*}
p \lambda_{1, N} \sum_{m \geq 0} p_{N, m}=\mu_{2}\left(\sum_{n=1}^{N} n \sum_{m \geq 0} p_{n, m}-\sum_{n=1}^{N} n p_{n, 0}\right) . \tag{2.2.13}
\end{equation*}
$$

The equations (2.2.12) and (2.2.13) can (under stationary conditions) be written as

$$
p \lambda_{1, N} p_{N, m}=\mu_{2} P\left(L_{2}=m+1\right) E\left(L_{1} \mid L_{2}=m+1\right)
$$

and

$$
\begin{equation*}
p \lambda_{1, N} P\left(L_{1}=N\right)=\mu_{2}\left(E L_{1}-P\left(L_{2}=0\right) E\left(L_{1} \mid L_{2}=0\right)\right) \tag{2.2.14}
\end{equation*}
$$

where $P\left(L_{2}=m\right) E\left(L_{1} \mid L_{2}=m\right)=\sum_{n=1}^{N} n p_{n, m}$ for $m \geq 0$. The only term in (2.2.14) involving unknowns is the term

$$
P\left(L_{2}=0\right) E\left(L_{1} \mid L_{2}=0\right)=\sum_{n=1}^{N} n p_{n, 0} .
$$

This suggests that the set of unknown probabilities in the system of steadystate equations (2.2.3) can be reduced to $p_{1,0}, \ldots, p_{N, 0}$. From equation (2.2.14) one might also predict from the following intuitive argument that
the system is stable if

$$
\begin{equation*}
p \lambda_{1, N}<\mu_{2} E L_{1} \tag{2.2.15}
\end{equation*}
$$

holds. Near saturation we have $P\left(L_{1}=N\right) \sim 1, P\left(L_{2}=0\right) \sim 0$ and $E\left(L_{1} \mid L_{2}=0\right)$ is bounded since the system is stable. Hence by (2.2.14)

$$
\frac{p \lambda_{1, N}}{\mu_{2} E L_{1}} \sim 1-\frac{P\left(L_{2}=0\right) E\left(L_{1} \mid L_{2}=0\right)}{\mu_{2} E L_{1} P\left(L_{1}=N\right)}<1
$$

Furthermore, the arrival and service rates in $Q_{1}$ have no influence on the stability of $Q_{1}$, since it is a loss system, but on the stability of $Q_{2}$. The arrival rate in $Q_{2}$ is $p \lambda_{1, N}$ while the average service rate is $\mu_{2} E L_{1}$. The fraction of both should be smaller than 1 since $Q_{2}$ behaves like an $M / M / 1$ queue in this case. Once again we arrive at (2.2.15). The stability condition should in general be weaker because the second queue has a non-zero arrival rate if and only if the first queue is fully occupied. By regarding the joint queue length process of $Q_{1}$ and $Q_{2}$ as a quasi birth and death process and using the stability Theorem 1.2 .3 of Section 1.2 , we will show that these presumptions are correct and that the necessary and sufficient stability condition is indeed the one given in the next proposition, i.e., $p \lambda_{1, N} P\left(L_{1}=N\right)<\mu_{2} E L_{1}$.

Proposition 2.2.1. The system (2.2.3) has a unique nonnegative and normalized solution if and only if $p \lambda_{1, N} P\left(L_{1}=N\right)<\mu_{2} E L_{1}$ or equivalently

$$
\begin{equation*}
\frac{p \lambda_{1, N}}{\mu_{2}}<\frac{\sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}}{\prod_{n=0}^{N-1} \rho_{n}} \tag{2.2.16}
\end{equation*}
$$

holds, where $\rho_{n}=\lambda_{1, n} / \mu_{1, n+1}$ for $n=0, \ldots, N-1$.

Proof. By regarding $Q_{1}$ as the phase and $Q_{2}$ as the level, the joint queue length process of $Q_{1}$ and $Q_{2}$ is a quasi birth and death process. The stability condition $(2.2 .16)$ is then derived from Theorem 1.2 .3 in the following way. We assume a more general model for the moment by endowing $Q_{2}$ with an external arrival stream and service rates controlled by the state of $Q_{1}$. In this generalization of our basic model, the exponential arrival rate to $Q_{2}$ is $\lambda_{2, n}$ and the exponential service rate is $\mu_{2, n}$ given $L_{1}=n, n=0, \ldots, N$. The rates for arrivals and service in $Q_{1}$ remain unchanged. In order to ensure irreducibility, we assume that there exists an $i \in\{0, \ldots, N\}$ with $\lambda_{2, i}>0$ and a $j \in\{0, \ldots, N-1\}$ with $\mu_{2, j}>0$. With this setting, the matrices $A_{0}$,
$A_{1}$ and $A_{2}$ are given by

$$
A_{0}=\operatorname{diag}\left(\lambda_{2,0}, \ldots, \lambda_{2, N}\right), \quad A_{2}=\operatorname{diag}\left(\mu_{2,0}, \ldots, \mu_{2, N}\right)
$$

and $A_{1}=A-A_{0}-A_{2}$, where $A$ is the rate matrix of the phase process governing $Q_{1}$, i.e., the standard birth and death process on $\{0, \ldots, N\}$ with reflecting barriers and birth rates $\lambda_{1, n}$ in the states $n=0, \ldots, N-1$ and death rates $\mu_{1, n}$ in the states $n=1, \ldots, N$. The vector $\pi$ is the stationary probability measure of the phase process in $Q_{1}$ and therefore $\pi_{n}=P\left(L_{1}=\right.$ $n$ ), $n=0, \ldots, N$, is given by (2.2.1). The stability condition (1.2.1) is equivalent to

$$
\begin{equation*}
\sum_{n=0}^{N}\left(\mu_{2, n}-\lambda_{2, n}\right) \pi_{n}<0 . \tag{2.2.17}
\end{equation*}
$$

Setting $\lambda_{2, n}=0$ for $n=0, \ldots, N-1, \lambda_{2, N}=p \lambda_{1, N}$ and $\mu_{2, n}=n \mu_{2}>0$ we get stability if and only if (2.2.16) holds.

We continue this section with a discussion of the stability condition. The stability condition can be written as

$$
\begin{equation*}
p \lambda_{1, N} P\left(L_{1}=N\right)<\mu_{2} E L_{1} \tag{2.2.18}
\end{equation*}
$$

as one might expect, since (2.2.1) states

$$
\sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}=\frac{E L_{1}}{P\left(L_{1}=0\right)} \quad \text { and } \quad \prod_{n=0}^{N-1} \rho_{n}=\frac{P\left(L_{1}=N\right)}{P\left(L_{1}=0\right)} .
$$

The stability condition (2.2.18) can in this form be well interpreted: The system is stable if and only if the expected service rate $\mu_{2} E L_{1}$ in $Q_{2}$ suffices to handle the average arrival or overflow rate $p \lambda_{1, N} P\left(L_{1}=N\right)$ from $Q_{1}$ to $Q_{2}$.

Remark 2.2.2. The stability condition can be formulated independently of the arrival rate $\lambda_{1,0}$ in $Q_{1}$. This is explained by the fact that the arrival rate in $Q_{1}$ is $\lambda_{1,0}$ only in the case that no customers are present in the first queue. Therefore, the arrival rate $\lambda_{1,0}$ influences the length of the idle periods of the system (i.e., the periods in which no customers are served), but not the busy periods. The idle periods of this queueing system can be classified by the number of customers in $Q_{2}$ which are not served due to the lack of customers in $Q_{1}$.

When the first queue is an $M / M / 1 / N-1$ queue with one server, $N-1$
waiting positions and constant arrival and service rates, the stability condition simplifies:

Remark 2.2.3. In the case of constant service and arrival rates $\lambda_{1, n}=\lambda_{1}$ for $n=0, \ldots, N$ and $\mu_{1, n}=\mu_{1}$ for $n=1, \ldots, N$, the first queue is an $M / M / 1 / N-1$ queue and we get the stability condition

$$
\begin{equation*}
p \lambda_{1}<\mu_{2} \rho^{-N} \sum_{n=1}^{N} n \rho^{n} \tag{2.2.19}
\end{equation*}
$$

where $\rho=\lambda_{1} / \mu_{1}$ is the traffic intensity of $Q_{1}$. Observe that

$$
E L_{1}=\frac{\sum_{n=1}^{N} n \rho^{n}}{\left(1+\sum_{n=1}^{N} \rho^{n}\right)}<\rho^{-N} \sum_{n=1}^{N} n \rho^{n}
$$

holds in this case.

It follows from the condition (2.2.18) and (2.2.1) that a more restrictive and sufficient but not necessary condition for stability is

$$
p \lambda_{1, N}<\mu_{2} E L_{1}
$$

Observe that the state-dependent rates are involved in $E L_{1}$. (2.2.18) yields that the condition

$$
\Lambda<\mu_{2}
$$

where $\Lambda$ is given by (2.2.2), is necessary but not sufficient. Both results are not surprising: On the one hand, the arrival rate in $Q_{2}$ is only positive if $L_{1}=N$. Therefore $p \lambda_{1, N}$ is in general greater than or equal to the actual mean service rate. On the other hand, the service rate in $Q_{2}$ is in general greater than or equal to $\mu_{2}$ given that customers arrive to $Q_{2}$, i.e., given $L_{1}=N$. Observe that the condition (2.2.18) complies with $\Lambda<\mu_{2} E L_{2}$.

### 2.2.3 Generating functions and steady-state distribution

In this section, we exploit the steady-state equations (2.2.3) and derive recurrence equations for the probability generating functions

$$
G_{n}(z)=\sum_{m=0}^{\infty} p_{n, m} z^{m}, \quad|z| \leq 1
$$

These relations are then used to reduce the set of unknown steady-state probabilities to the unknowns $p_{1,0}, \ldots, p_{N, 0}$ and to derive equivalent formulations for the stability condition (2.2.16). The stability condition will be related to the existence of a certain number of zeros of a function originating from the recurrence equations for the probability generating functions. This relation will give a deeper insight into the existence and uniqueness of a solution of the steady-state equations (2.2.3) and into the nature of the unknowns $p_{1,0}, \ldots, p_{N, 0}$.

We can assume for the moment that the system is stable and that (2.2.16) holds. In this case, the generating functions are well defined for $|z| \leq 1$. Exploiting the recurrence relations and using an inductive argument and Cramer's rule we derive $N$ independent equations for the $N$ unknowns and the stability condition.

By multiplying the equations (2.2.4), (2.2.5), (2.2.6), (2.2.8), (2.2.9) and (2.2.10) by $z^{m}$ and summing over $m \geq 0$ we get, after simplifying,

$$
\begin{align*}
\lambda_{1,0} G_{0}(z)= & \mu_{1,1} G_{1}(z),  \tag{2.2.20}\\
\left(\left(\lambda_{1, n}+\mu_{1, n}\right) z-n \mu_{2}(1-z)\right) G_{n}(z)= & \lambda_{1, n-1} z G_{n-1}(z)+\mu_{1, n+1} z G_{n+1}(z) \\
& -n \mu_{2}(1-z) p_{n, 0} \tag{2.2.21}
\end{align*}
$$

for $n=1, \ldots, N-1$ and

$$
\begin{align*}
& \left(\mu_{1, N} z+\left(p \lambda_{1, N} z-N \mu_{2}\right)(1-z)\right) G_{N}(z) \\
& \quad=\lambda_{1, N-1} z G_{N-1}(z)-N \mu_{2}(1-z) p_{N, 0} . \tag{2.2.22}
\end{align*}
$$

In order to write these equations in matrix form, we define the vectors

$$
\begin{gathered}
G(z)=\left(G_{0}(z), \ldots, G_{N}(z)\right)^{\top}, \\
p=\left(0, p_{1,0}, 2 p_{2,0}, \ldots, N p_{N, 0}\right)^{\top}
\end{gathered}
$$

and the matrix $A(z) \in \operatorname{Mat}(N+1, N+1, \mathbb{R})$ by

$$
A(z)=\left(\begin{array}{cccccc}
\alpha_{0}(z) & -\mu_{1,1} & 0 & \ldots & \ldots & 0  \tag{2.2.23}\\
-\lambda_{1,0} z & \alpha_{1}(z) & -\mu_{1,2} z & \ddots & & \vdots \\
0 & \ddots & \alpha_{2}(z) & \ddots & \ddots & \vdots \\
\vdots & \ddots & \ddots & \ddots & \ddots & 0 \\
\vdots & & \ddots & \ddots & \ddots & -\mu_{1, N} z \\
0 & \ldots & \ldots & 0 & -\lambda_{1, N-1} z & \alpha_{N}(z)
\end{array}\right)
$$

where

$$
\begin{align*}
\alpha_{0}(z) & =\lambda_{1,0}  \tag{2.2.24}\\
\alpha_{n}(z) & =\left(\lambda_{1, n}+\mu_{1, n}\right) z-n \mu_{2}(1-z) \text { for } n=1, \ldots, N-1 \text { and }  \tag{2.2.25}\\
\alpha_{N}(z) & =\mu_{1, N} z+\left(p \lambda_{1, N} z-N \mu_{2}\right)(1-z) \tag{2.2.26}
\end{align*}
$$

In this notation, the equations (2.2.20), (2.2.21) and (2.2.22) are equivalent to

$$
\begin{equation*}
A(z) G(z)=-\mu_{2}(1-z) p \tag{2.2.27}
\end{equation*}
$$

Let $A_{n}(z)$ be the matrix obtained from $A(z)$ by replacing the $(n+1)$-th column by the vector $-\mu_{2}(1-z) p$ for $n=0, \ldots, N$. By Cramer's rule we may write

$$
\begin{equation*}
\operatorname{det}(A(z)) G_{n}(z)=\operatorname{det}\left(A_{n}(z)\right) \tag{2.2.28}
\end{equation*}
$$

for every $n=0, \ldots, N$ and all values $z$ such that $A(z)$ is invertible. Since the functions $A, A_{n}$ and $G_{n}$ are continuous and bounded with at most finitely many zeros in the interval $[0,1]$, equation (2.2.28) must hold for all $z \in$ $[0,1]$ and every $n=0, \ldots, N$. Hence, the generating functions $G_{0}, \ldots, G_{N}$ are uniquely determined by the equations (2.2.28) and $p_{1,0}, \ldots, p_{N, 0}$, since these are the only unknowns occurring in these equations (see also (2.2.20)(2.2.22)).
$\operatorname{det}(A(z))$ is a polynomial in $z$ of degree $N+1$. We will show in the following that $\operatorname{det}(A(z))$ has $N-1$ zeros in the open interval $(0,1)$ and one zero at $z=1$. Additionally, we will show that $\operatorname{det}(A(z))$ has another zero in the open interval $(1, \infty)$ if and only if the stability condition (2.2.16) holds. The $N-1$ zeros of $\operatorname{det}(A(z))$ in $(0,1)$ will provide us with $N-$ 1 linear homogeneous equations in the unknowns $p_{1,0}, \ldots, p_{N, 0}$. Another linear equation yielding a system of $N$ equations is (2.2.14). We are able to formulate this equation in terms of only $p_{1,0}, \ldots, p_{N, 0}$ and the system parameters. The result is stated in Proposition 2.2.8.

Let $q_{n}(z)$ be the $n$-th minor of $A(z)$ for $n=1, \ldots, N+1$ (the $n$-th minor of a matrix is the determinant of the $n$-th square sub-matrix). We have

$$
q_{1}(z)=\alpha_{0}(z), q_{2}(z)=\operatorname{det}\left(\begin{array}{cc}
\alpha_{0}(z) & -\mu_{1,1}  \tag{2.2.29}\\
-\lambda_{1,0} z & \alpha_{1}(z)
\end{array}\right), \ldots, q_{N+1}(z)=\operatorname{det}(A(z))
$$

By Laplace expansion of the determinants we get

$$
\begin{align*}
& q_{1}(z)=\alpha_{0}(z) q_{0}(z), \\
& q_{2}(z)=\alpha_{1}(z) q_{1}(z)-\lambda_{1,0} \mu_{1,1} z q_{0}(z) \quad \text { and }  \tag{2.2.30}\\
& q_{n}(z)=\alpha_{n-1}(z) q_{n-1}(z)-\lambda_{1, n-2} \mu_{1, n-1} z^{2} q_{n-2}(z)
\end{align*}
$$

for $n=3, \ldots, N+1$, where $q_{0}(z)=1$.
In order to proceed with investigating the number and the location of the zeros of $\operatorname{det}(A)=q_{N+1}$, we have to study the system (2.2.30) in order to find the algebraic properties and to determine the shape of the functions $q_{1}, \ldots, q_{N+1}$. The relevant properties are stated in the next proposition.

Proposition 2.2.4. The function $q_{n}$ is a polynomial in $z$ of degree $n-1$ for $n=1, \ldots, N$ and of degree $N+1$ for $n=N+1$. The functions $q_{0}, \ldots, q_{N+1}$ and $\alpha_{0}, \ldots, \alpha_{N}$ have the following properties:
(i) $q_{n}$ and $q_{n+1}$ have no common root in $(0,1)$ for $n=0, \ldots, N$.
(ii) $\operatorname{sgn}\left(\alpha_{0}(0)\right)=1$ and $\operatorname{sgn}\left(\alpha_{n}(0)\right)=-1$ for $n=1, \ldots, N$.
(iii) $\operatorname{sgn}\left(q_{n}(0)\right)=(-1)^{n+1}$ for $n=1, \ldots, N+1$.
(iv) $q_{n}(1)=\prod_{i=0}^{n-1} \lambda_{1, i}$ for $n=0, \ldots, N$ and $q_{N+1}(1)=0$.
(v) For $n=1, \ldots, N$ the following implication holds: If $\tilde{z}>0$ with $q_{n}(\tilde{z})=$ 0 , then

$$
\operatorname{sgn}\left(q_{n-1}(\tilde{z}) q_{n+1}(\tilde{z})\right)=-1
$$

(vi) $q_{n}$ has $n-1$ pairwise distinct zeros in $(0,1)$ for $n=1, \ldots, N$.
(vii) $\lim _{z \rightarrow \infty} q_{n}(z)=\infty$ for $n=2, \ldots, N$ and $\lim _{z \rightarrow \infty} q_{N+1}(z)=-\infty$.

Proof. By the recursive definition (2.2.29), $q_{n}$ is a polynomial in $z$ of degree $n-1$ for $n=1, \ldots, N$ and of degree $N+1$ for $n=N+1$. We will prove the remaining properties.
(i) $q_{0}$ and $q_{1}$ have no zero since $\lambda_{1,0}>0$. Suppose $q_{n}(z)=q_{n+1}(z)=0$ for some $z>0$ and some $n=2, \ldots, N$. Then $q_{n-1}(z)=0$ follows from (2.2.30). We get $q_{1}(z)=0$ by induction which is a contradiction.
(ii) $\operatorname{sgn}\left(\alpha_{0}(0)\right)=\operatorname{sgn}\left(\lambda_{1,0}\right)=1$ since $\lambda_{1,0}>0, \operatorname{sgn}\left(\alpha_{n}(0)\right)=\operatorname{sgn}\left(-n \mu_{2}\right)=$ -1 for $n=1, \ldots, N$ since $\mu_{2}>0$.
(iii) By (iil) we have $\operatorname{sgn}\left(q_{1}(0)\right)=\operatorname{sgn}\left(\alpha_{0}(0)\right)=1$ and

$$
\operatorname{sgn}\left(q_{2}(0)\right)=\operatorname{sgn}\left(\alpha_{1}(0) q_{1}(0)\right)=-1
$$

By induction we get from (iii) that

$$
\operatorname{sgn}\left(q_{n}(0)\right)=\operatorname{sgn}\left(\alpha_{n-1}(0)\right) \operatorname{sgn}\left(q_{n-1}(1)\right)=(-1)^{n+1}
$$

for $n=0, \ldots, N$. The functions $\alpha_{n}$ and therefore the functions $q_{n}$ are continuous in 0 .
(iv) Clearly, $q_{0}(1)=1$ and $q_{1}(1)=\lambda_{1,0}$. For $n=2, \ldots, N$ we get from (2.2.30) by induction that

$$
\begin{aligned}
q_{n}(1) & =\alpha_{n-1}(1) q_{n-1}(1)-\lambda_{1, n-2} \mu_{1, n-1} q_{n-2}(1) \\
& =\left(\lambda_{1, n-1}+\mu_{1, n-1}\right) \prod_{i=0}^{n-2} \lambda_{1, i}-\lambda_{1, n-2} \mu_{1, n-1} \prod_{i=0}^{n-3} \lambda_{1, i}=\prod_{i=0}^{n-1} \lambda_{1, i}
\end{aligned}
$$

$q_{N+1}(1)=0$ follows from inserting $q_{N-1}(1)=\prod_{i=0}^{N-2} \lambda_{1, i}$ and $q_{N}(1)=$ $\prod_{i=0}^{N-1} \lambda_{1, i}$ into (2.2.30) and simplifying.
(v) Let $n=1, \ldots, N$ and $\tilde{z}>0$ with $q_{n}(\tilde{z})=0$. By (2.2.30) we get

$$
\begin{aligned}
& \operatorname{sgn}\left(q_{n+1}(\tilde{z}) q_{n-1}(\tilde{z})\right)= \operatorname{sgn}( \\
&\left(\alpha_{n}(\tilde{z}) q_{n}(\tilde{z})\right. \\
&\left.\left.-\lambda_{1, n-1} \mu_{1, n-1} \tilde{z}^{2} q_{n-1}(\tilde{z})\right) q_{n-1}(\tilde{z})\right) \\
&=\operatorname{sgn}\left(-\lambda_{1, n-2} \mu_{1, n-1} \tilde{z}^{2} q_{n-1}^{2}(\tilde{z})\right)=-1
\end{aligned}
$$

In the last line we used $\lambda_{1, n-2}, \mu_{1, n-1}>0$ and the fact that $q_{n}$ and $q_{n-1}$ have no common zero which was shown in (ii).
(vi) Once again we use an inductive argument to show this property. $q_{1}$ has no root in $(0,1) . q_{2}$ has exactly one root in $(0,1)$ since $\lambda_{1,1}, \mu_{2}>0$, namely $z=\frac{\mu_{2}}{\lambda_{1,1}+\mu_{2}}$. Let $n=3, \ldots, N$ and

$$
0<z_{1}<\ldots<z_{n-2}<1
$$

be the pairwise distinct zeros of $q_{n-1}$. We show that (iiil), (iv) and (葍) imply that $q_{n}$ has $n-1$ pairwise distinct zeros $z_{n, 1}, \ldots, z_{n, n-1}$ in $(0,1)$ and that these zeros satisfy the interlacing property

$$
\begin{equation*}
z_{0}=0<z_{n, 1}<z_{1}<z_{n, 2}<\ldots<z_{n-2}<z_{n, n-1}<1=z_{n-1} \tag{2.2.31}
\end{equation*}
$$

Let $k=1, \ldots, n-2$. We obtain from (2.2.30) and (v) inductively that

$$
\operatorname{sgn}\left(q_{n}\left(z_{k}\right)\right)=-\operatorname{sgn}\left(q_{n-2}\left(z_{k}\right)\right)=(-1)^{k+n+1}
$$

By (iii) and (iiv) this is also true for $z_{0}=0$ and $z_{n-1}=1$. Thus, $q_{n}$ changes its sign $n-1$ times in the interval $(0,1)$. Therefore, the $n-1$ zeros of $q_{n}$ are real, pairwise distinct and must be located in the interval $(0,1)$. Furthermore, $q_{n}$ has alternating signs on the sequence of zeros of $q_{n-1}$ since

$$
\operatorname{sgn}\left(q_{n}\left(z_{k}\right)\right)=(-1)^{k+n+1}
$$

for $k=1, \ldots, n-2$. Together with $\operatorname{sgn}\left(q_{n}(0)\right)=-\operatorname{sgn}\left(q_{n-1}(0)\right)$, which follows from (iiii), this finally yields (2.2.31).
(vii) Let $n=2, \ldots, N$. From (iv), (vii) and since $q_{n}$ is a polynomial in $z$ of degree $n-1$, we obtain immediately that $\lim _{z \rightarrow \infty} q_{n}(z)=\infty$. For $n=N+1$ we get $\lim _{z \rightarrow \infty} q_{N+1}(z)=-\infty$ from equation (2.2.30) since $\lim _{z \rightarrow \infty} \alpha_{N}(z)=-\infty, \lim _{z \rightarrow \infty} q_{N}(z)=\lim _{z \rightarrow \infty} q_{N-1}(z)=\infty$ and since the system parameters are positive.

The proof of Proposition 2.2.4 (vii) shows that the $n-2$ pairwise distinct zeros $z_{n-1,1}, \ldots, z_{n-1, n-2}$ of $q_{n-1}$ and the $n-1$ pairwise distinct zeros $z_{n, 1}, \ldots, z_{n, n-1}$ of $q_{n}$ satisfy the interlacing property

$$
\begin{equation*}
0<z_{n, 1}<z_{n-1,1}<z_{n, 2}<\ldots<z_{n-1, n-2}<z_{n, n-1}<1 \tag{2.2.32}
\end{equation*}
$$

for every $n=1, \ldots, N$. We can expect a similar behavior for the zeros of $q_{N}$ and $q_{N+1}$. In fact, it will turn out that an interlacing property holds for these zeros and that the sign of the slope of $q_{N+1}$ at the point $z=1$ plays an important role for the existence of roots of $q_{N+1}$ in combination with the stability condition (2.2.16). Note that $q_{N+1}$ has a zero at the point 1 according to Proposition [2.2.4 (iv). In this context, it seems natural to determine the function $h_{N+1}$ with $q_{N+1}(z)=(1-z) h_{N+1}(z)$ for all $z>0$.

Choose continuous functions $h_{1}, \ldots, h_{N+1}$ on $[0, \infty)$ such that

$$
\begin{align*}
q_{n}(z) & =z^{n-1} \prod_{i=0}^{n-1} \lambda_{1, i}+(1-z) h_{n}(z), n=1, \ldots, N,  \tag{2.2.33}\\
q_{N+1}(z) & =(1-z) h_{N+1}(z) .
\end{align*}
$$

A system of recursive equations for $h_{1}(z), \ldots, h_{N+1}(z)$ is given in the
next proposition.
Proposition 2.2.5. The functions $h_{1}, \ldots, h_{N+1}$ defined by (2.2.33) are given by $h_{1}(z)=0, h_{2}(z)=-\lambda_{1,0} \mu_{2}$,

$$
\begin{align*}
h_{n}(z)= & -(n-1) \mu_{2} z^{n-2} \prod_{i=0}^{n-2} \lambda_{1, i}+\alpha_{n-1}(z) h_{n-1}(z)  \tag{2.2.34}\\
& -\lambda_{1, n-2} \mu_{1, n-1} z^{2} h_{n-2}(z)
\end{align*}
$$

for $n=3, \ldots, N$ and

$$
\begin{align*}
h_{N+1}(z)= & \left(p \lambda_{1, N} z-N \mu_{2}\right) z^{N-1} \prod_{i=0}^{N-1} \lambda_{1, i}+\alpha_{N}(z) h_{N}(z)  \tag{2.2.35}\\
& -\lambda_{1, N-1} \mu_{1, N} z^{2} h_{N-1}(z)
\end{align*}
$$

for all $z \in[0, \infty)$.
Proof. Let $z>0$ in the following and let the functions $q_{n}, n=0, \ldots, N+1$, and $h_{n}, n=1, \ldots, N+1$, be defined by (2.2.30) and (2.2.33), respectively, i.e.,

$$
\begin{aligned}
& q_{0}(z)=1 \\
& q_{1}(z)=\alpha_{0}(z) q_{0}(z), \\
& q_{2}(z)=\alpha_{1}(z) q_{1}(z)-\lambda_{1,0} \mu_{1,1} z q_{0}(z) \quad \text { and } \\
& q_{n}(z)=\alpha_{n-1}(z) q_{n-1}(z)-\lambda_{1, n-2} \mu_{1, n-1} z^{2} q_{n-2}(z)
\end{aligned}
$$

for $n=3, \ldots, N+1$ and

$$
\begin{aligned}
q_{n}(z) & =z^{n-1} \prod_{i=0}^{n-1} \lambda_{1, i}+(1-z) h_{n}(z), n=1, \ldots, N, \\
q_{N+1}(z) & =(1-z) h_{N+1}(z)
\end{aligned}
$$

for all $z \in[0, \infty)$. We will prove (2.2.34) by induction over $n$. We obtain immediately from the definition of $q_{1}$ and $q_{2}$ that $h_{1}(z)=0$ and $h_{2}(z)=$ $-\lambda_{1,0} \mu_{2}$. Now let $n=3, \ldots, N$ and $h_{n-1}$ and $h_{n-2}$ be given by (2.2.34). By replacing $\alpha_{n-1}(z)$ given by (2.2.25) we get

$$
\begin{aligned}
q_{n}(z) & =\alpha_{n-1}(z) q_{n-1}(z)-\lambda_{1, n-2} \mu_{1, n-1} z^{2} q_{n-2}(z) \\
& =\left(\left(\lambda_{1, n-1}+\mu_{1, n-1}\right) z-(n-1) \mu_{2}(1-z)\right)\left(z^{n-2} \prod_{i=0}^{n-2} \lambda_{1, i}\right)
\end{aligned}
$$

$$
\begin{aligned}
& +\alpha_{n-1}(z)(1-z) h_{n-1}(z) \\
& -\lambda_{1, n-2} \mu_{1, n-1}\left(z^{n-1} \prod_{i=0}^{n-3} \lambda_{1, i}+(1-z) h_{n-3}(z)\right) \\
= & z^{n-1} \prod_{i=0}^{n-1} \lambda_{1, i}+(1-z)\left(\alpha_{n-1}(z) h_{n-1}(z)\right. \\
& \left.\quad-(n-1) \mu_{2} z^{n-2} \prod_{i=0}^{n-2}-\lambda_{1, n-2} \mu_{1, n-1} z^{2} h_{n-2}(z)\right) \\
= & z^{n-1} \prod_{i=0}^{n-1} \lambda_{1, i}+(1-z) h_{n}(z) .
\end{aligned}
$$

The identity for $h_{N+1}(z)$ can be proved by inserting $q_{N-1}(z)$ and $q_{N}(z)$ into

$$
q_{N+1}(z)=\alpha_{N}(z) q_{N}(z)-\lambda_{1, N-1} \mu_{1, N} z^{2} q_{N-1}(z)
$$

and simplifying.

From Proposition $[2.2 .4$ we obtain the main condition that relates the number of zeros of the function $\operatorname{det}(A(z))$ for $z \in(0,1)$ to $h_{N+1}(1)$. This condition will be shown to be the link between the number of $\operatorname{zeros}$ of $\operatorname{det}(A)$ and the stability condition (2.2.16).

Theorem 2.2.6. $\operatorname{det}(A(z))$ is a polynomial of degree $N+1$ and has $N-1$ distinct zeros in the interval $(0,1)$ and one zero at $z=1$. Additionally, $\operatorname{det}(A(z))$ has another zero in the interval $(1, \infty)$ if and only if $h_{N+1}(1)<0$, where $h_{N+1}$ is defined by $\operatorname{det}(A(z))=(1-z) h_{N+1}(z)$ for all $z>0$.

Proof. By (2.2.29) we have to prove that $q_{N+1}(z)=\operatorname{det}(A(z))$ has the claimed properties. Clearly, $q_{N+1}$ has degree $N+1$. Without loss of generality, let $N$ be odd in the following. The case of an even $N$ can be handled analogously. By Proposition 2.2.4 (vii) the $N-1$ zeros $z_{1}, \ldots, z_{N-1}$ of $q_{N}$ can be labeled such that

$$
0<z_{1}<\ldots<z_{N-1}<1
$$

It follows from (2.2.30) and Proposition [2.2.4 (v) by induction that

$$
\begin{equation*}
\operatorname{sgn}\left(q_{N+1}\left(z_{k}\right)\right)=-\operatorname{sgn}\left(q_{N-1}\left(z_{k}\right)\right)=(-1)^{k+N}=(-1)^{k+1} \tag{2.2.36}
\end{equation*}
$$

for $k=1, \ldots, N-1$ since $N$ is odd. Hence, by continuity and

$$
\operatorname{sgn}\left(q_{N+1}(0)\right)=-\operatorname{sgn}\left(q_{N}(0)\right)
$$

(see Proposition 2.2.4 (iiii)), there exist $N-1$ zeros $z_{N+1,1}, \ldots, z_{N+1, N-1}$ of $q_{N+1}$ in $(0,1)$ with the interlacing property

$$
0<z_{N+1,1}<z_{1}<z_{N+1,2}<\ldots<z_{N+1, N-1}<z_{N-1}<1
$$

regardless of the value of $h_{N+1}(1) \cdot q_{N+1}(1)=0$ was shown in Proposition 2.2.4 (iv). Now we show that $q_{N+1}$ has one zero in $(1, \infty)$ if and only if $h_{N+1}(1)<0$. By (2.2.36) and since $q_{N+1}$ has at most $N+1$ zeros, the $(N+1)$-th zero has to be located in the interval $\left(z_{N-1}, \infty\right)$. By Proposition 2.2.4 (vii) every other case would give a set of more than $N+1$ zeros which is impossible since $q_{N+1}$ has degree $N+1$. Let $q_{N+1}^{\prime}$ be the derivative of $q_{N+1}$. Since $q_{N+1}(1)=0$ by Proposition 2.2.4 (iv)) and

$$
\operatorname{sgn}\left(q_{N+1}\left(z_{N-1}\right)\right)=-\operatorname{sgn}\left(q_{N-1}\left(z_{N-1}\right)\right)=(-1)^{N}=-1
$$

we must have $q_{N+1}^{\prime}(1)>0$. Each of the cases $q_{N+1}^{\prime}(1)<0$ and $q_{N+1}^{\prime}(1)=0$ would give a set of $N+1$ pairwise distinct zeros for $q_{N+1}^{\prime}$ which is impossible since $q_{N+1}^{\prime}$ is a polynomial of degree $N$. By $h_{N+1}(1)=-q_{N+1}^{\prime}(1)$ we get that the $(N+1)$-th zero lies in $(1, \infty)$ if and only if $h_{N+1}(1)<0$. We derive at the same condition in the case of an even $N$.

By Theorem 2.2.6 we have to find $h_{N+1}(1)$. Evaluation of (2.2.34) and (2.2.35) at $z=1$ gives a recursive system for $h_{1}(1), \ldots, h_{N+1}(1)$. We get $h_{1}(1)=0, h_{2}(1)=-\lambda_{1,0} \mu_{2}$,

$$
\begin{align*}
h_{n}(1)= & -(n-1) \mu_{2} \prod_{i=0}^{n-2} \lambda_{1, i}+\left(\lambda_{1, n-1}+\mu_{1, n-1}\right) h_{n-1}(1)  \tag{2.2.37}\\
& -\lambda_{1, n-2} \mu_{1, n-1} h_{n-2}(1)
\end{align*}
$$

for $n=3, \ldots, N$ and

$$
\begin{align*}
h_{N+1}(1)= & \left(p \lambda_{1, N}-N \mu_{2}\right) \prod_{i=0}^{N-1} \lambda_{1, i}+\mu_{1, N} h_{N}(1)  \tag{2.2.38}\\
& -\lambda_{1, N-1} \mu_{1, N} h_{N-1}(1)
\end{align*}
$$

This system can be solved explicitly in terms of the system parameters. The solution is stated in the next proposition.

Proposition 2.2.7. The functions $h_{1}, \ldots, h_{N+1}$ defined by (2.2.33) satisfy

$$
\begin{equation*}
h_{n}(1)=-\mu_{2}\left(\prod_{i=0}^{n-1} \lambda_{1, i}\right)\left(\sum_{k=1}^{n-1} \frac{1}{\mu_{1, k+1}}\left(\prod_{i=0}^{k} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right)\left(\sum_{j=1}^{k} j \prod_{i=0}^{j-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right)\right) \tag{2.2.39}
\end{equation*}
$$

for $n=1, \ldots, N$ and

$$
\begin{equation*}
h_{N+1}(1)=p \prod_{i=0}^{N} \lambda_{1, i}-\mu_{2}\left(\prod_{i=0}^{N-1} \mu_{1, i+1}\right)\left(\sum_{n=1}^{N} n \prod_{i=0}^{n-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right), \tag{2.2.40}
\end{equation*}
$$

where the empty sum and the empty product are defined to be 0 and 1, respectively.

Proof. (2.2.40) follows from (2.2.39) by inserting $h_{N}(1)$ and $h_{N-1}(1)$ into (2.2.38) and collecting terms. We will show (2.2.39) by induction over $n$ using the recursion (2.2.37). Evaluating (2.2.39) for $n=1$ and $n=2$ we get $h_{1}(1)=0$ and $h_{2}(1)=-\mu_{2} \lambda_{1,0}$. Let $n=3, \ldots, N$. We set

$$
g_{n}=\sum_{k=1}^{n-1} \frac{1}{\mu_{1, k+1}}\left(\prod_{i=0}^{k} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right)\left(\sum_{j=1}^{k} j \prod_{i=0}^{j-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right)
$$

so that $h_{n}(1)=-\mu_{2} g_{n} \prod_{i=0}^{n-1} \lambda_{1, i}$. Let $h_{n-1}(1)$ and $h_{n-2}(1)$ be given by (2.2.39). Substituting these terms in (2.2.37) yields

$$
\begin{align*}
h_{n}(1)= & -(n-1) \mu_{2} \prod_{i=0}^{n-2} \lambda_{1, i}-\mu_{2}\left(\lambda_{1, n-1}+\mu_{1, n-1}\right) g_{n-1} \prod_{i=0}^{n-2} \lambda_{1, i} \\
& +\mu_{2} \lambda_{1, n-2} \mu_{1, n-1} g_{n-2} \prod_{i=0}^{n-3} \lambda_{1, i} \\
= & -\mu_{2}\left(g_{n-1}+(n-1) \frac{1}{\lambda_{1, n-1}}+\frac{\mu_{1, n-1}}{\lambda_{1, n-1}}\left(g_{n-1}-g_{n-2}\right)\right) \prod_{i=0}^{n-1} \lambda_{1, i} . \tag{2.2.41}
\end{align*}
$$

Due to the additive structure of $g_{n}$ we have

$$
\frac{\mu_{1, n-1}}{\lambda_{1, n-1}}\left(g_{n-1}-g_{n-2}\right)=\frac{1}{\lambda_{1, n-1}}\left(\prod_{i=0}^{n-2} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right)\left(\sum_{j=1}^{n-2} j \prod_{i=0}^{j-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right)
$$

$$
\begin{equation*}
=\frac{1}{\mu_{1, n}}\left(\prod_{i=0}^{n-1} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right)\left(\sum_{j=1}^{n-2} j \prod_{i=0}^{j-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right) \tag{2.2.42}
\end{equation*}
$$

Writing

$$
(n-1) \frac{1}{\lambda_{1, n-1}}=\frac{1}{\mu_{1, n}}\left(\prod_{i=0}^{n-1} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right)\left((n-1) \prod_{i=0}^{n-2} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right)
$$

the equation (2.2.42) gives

$$
\begin{aligned}
&(n-1) \frac{1}{\lambda_{1, n-1}}+\frac{\mu_{1, n-1}}{\lambda_{n-1}}\left(g_{n-1}-g_{n-2}\right) \\
&=\frac{1}{\mu_{1, n}}\left(\prod_{i=0}^{n-1} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right)\left(\sum_{j=1}^{n-1} j \prod_{i=0}^{j-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right)
\end{aligned}
$$

Substituting this in (2.2.41) and once again exploiting the additive structure of $g_{n}$ we arrive at (2.2.39).

We can derive $h_{N+1}(1)$ also in closed form in terms of the system parameters and $p_{1,0}, \ldots, p_{N, 0}$. First we determine $G_{n}(1)$ for $n=0, \ldots, N$ in terms of $G_{0}(1)$. Setting $z=1$ in (2.2.20)-(2.2.22) gives

$$
\begin{align*}
\lambda_{1,0} G_{0}(1) & =\mu_{1,1} G_{1}(1)  \tag{2.2.43}\\
\left(\lambda_{1, n}+\mu_{1, n}\right) G_{n}(1) & =\lambda_{1, n-1} G_{n-1}(1)+\mu_{1, n+1} G_{n+1}(1) \quad \text { and }  \tag{2.2.44}\\
\mu_{1, N} G_{N}(z) & =\lambda_{1, N-1} G_{N-1}(1) \tag{2.2.45}
\end{align*}
$$

where $n=1, \ldots, N-1$ in the second equation. The solution of these equations in terms of $G_{0}(1)$ is clearly

$$
\begin{equation*}
G_{n}(1)=G_{0}(1) \prod_{i=0}^{n-1} \rho_{i} \text { for } n=0, \ldots, N \tag{2.2.46}
\end{equation*}
$$

where we have written $\rho_{i}=\lambda_{1, i} / \mu_{1, i+1}$ for $i=0, \ldots, N-1$. By

$$
\sum_{n=0}^{N} \sum_{m \geq 0} p_{n, m}=\sum_{n=0}^{N} G_{n}(1)=1
$$

summation over $n=0, \ldots, N$ yields

$$
\begin{equation*}
G_{0}(1)=\left(\sum_{n=0}^{N} \prod_{i=0}^{n-1} \rho_{i}\right)^{-1} \tag{2.2.47}
\end{equation*}
$$

This can also be deduced directly from (2.2.1). By (2.2.14), (2.2.46) and since $P\left(L_{1}=N\right)=G_{N}(1), E L_{1}=\sum_{n=1}^{N} n G_{n}(1)$ and

$$
P\left(L_{2}=0\right) E\left(L_{1} \mid L_{2}=0\right)=\sum_{n=1}^{N} n p_{n, 0}
$$

hold, we get

$$
p \lambda_{1, N} G_{N}(1)=\mu_{2}\left(G_{0}(1) \sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}-\sum_{n=1}^{N} n p_{n, 0}\right) .
$$

By (2.2.46) the latter equation gives

$$
\begin{equation*}
\mu_{2} \sum_{n=1}^{N} n p_{n, 0}=G_{0}(1)\left(\mu_{2} \sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}-p \lambda_{1, N} \prod_{i=0}^{N-1} \rho_{i}\right) . \tag{2.2.48}
\end{equation*}
$$

After rearranging terms and using (2.2.47) we get

$$
\begin{equation*}
h_{N+1}(1)=-\mu_{2}\left(\prod_{n=0}^{N-1} \mu_{i+1}\right)\left(\sum_{n=0}^{N} \prod_{i=0}^{n-1} \rho_{i}\right) \sum_{n=1}^{N} n p_{n, 0} \tag{2.2.49}
\end{equation*}
$$

from (2.2.40). By setting the right side of equation (2.2.40) and equation (2.2.49) equal, we get again the equation (2.2.14) in terms of $\sum_{n=1}^{N} n p_{n, 0}$ and the system parameters. This equation can also be derived from (2.2.40) and (2.2.48). The following proposition states the result.

Proposition 2.2.8. Let a solution $p_{n, m}$ for $n=0, \ldots, N$ and $m \geq 0$ of the equations (2.2.3) be given. Then the condition $\sum_{n=1}^{N} \sum_{m \geq 0} p_{n, m}=1$ implies the equivalence of (2.2.14) and

$$
\begin{equation*}
\sum_{n=1}^{N} n p_{n, 0}=\frac{\mu_{2} \sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}-p \lambda_{1, N} \prod_{i=0}^{N-1} \rho_{i}}{\mu_{2} \sum_{n=0}^{N} \prod_{i=0}^{n-1} \rho_{i}}>0 \tag{2.2.50}
\end{equation*}
$$

Proposition 2.2.8 provides a useful numerical check for the stationary probabilities. Now we give equivalent formulations for the stability criterion
(2.2.16).

Proposition 2.2.9. Let $\rho_{n}=\lambda_{1, n} / \mu_{1, n+1}$ for $n=0, \ldots, N-1$. The following conditions are equivalent:
(i) The system of equations (2.2.3) has a unique nonnegative and normalized solution.
(ii) $h_{N+1}(1)<0$ holds, where $h_{N+1}(1)$ is given by (2.2.40).
(iii) $p \lambda_{1} P\left(L_{1}=N\right)<\mu_{2} E L_{1}$ holds, where $P\left(L_{1}=N\right)$ and $E L_{1}$ are given by (2.2.1).
(iv) The system of equations (2.2.3) has a solution with $\sum_{n=1}^{N} p_{n, 0}>0$.

Proof. The equivalence of (ii) and (iii) was shown in Proposition [2.2.1. (iii) is a reformulation of (iii). Also by Proposition [2.2.1, the system is irreducible and positive recurrent under condition (iiii). We obtain immediately that $p_{n, m}>0$ for $n=0, \ldots, N$ and $m \geq 0$, and hence (iv) follows.

We show now that (iv) implies (iii). Let (iv) hold. Note that it cannot be assumed that the Markov chain defined by the equations (2.2.3) is positive recurrent and hence the stationary measure (the solution of (2.2.3)) can have an infinite mass. Therefore, we cannot use the equations (2.2.13), (2.2.14) or (2.2.50) under condition (iv) in order to ensure (iiii). Let the functions $\alpha_{n}(z)$ and $G_{n}(z), n=0, \ldots, N$, be defined by (2.2.24)-(2.2.26) and (2.2.28), respectively, for $|z| \leq 1$. This ensures that (2.2.43)-(2.2.45) hold and that the functions $q_{n}(z)$ and $h_{n}(z), n=1, \ldots, N+1$, are well defined for $|z| \leq 1$. By (2.2.28), (2.2.29) and (2.2.33) we get

$$
\begin{align*}
\operatorname{det}\left(A_{0}(z)\right) & =\operatorname{det}(A(z)) G_{0}(z)=q_{N+1}(z) G_{0}(z)  \tag{2.2.51}\\
& =(1-z) h_{N+1}(1) G_{0}(z)
\end{align*}
$$

for all $|z| \leq 1$. By writing down a recursive formula for $\operatorname{det}\left(A_{0}(z)\right)$ by means of Cramer's rule and performing an inductive argument, it can be shown that

$$
\left.\frac{\operatorname{det}\left(A_{0}(z)\right)}{1-z}\right|_{z=1}=-\mu_{2}\left(\prod_{n=0}^{N-1} \mu_{i+1}\right) \sum_{n=1}^{N} n p_{n, 0} .
$$

The calculations are straightforward but tedious and hence omitted. Using the above and (2.2.51) yields

$$
\begin{equation*}
G_{0}(1) h_{N+1}(1)=-\mu_{2}\left(\prod_{n=0}^{N-1} \mu_{i+1}\right) \sum_{n=1}^{N} n p_{n, 0} . \tag{2.2.52}
\end{equation*}
$$

Together with $G_{0}(1)>0$ and $\sum_{n=1}^{N} p_{n, 0}>0$ this shows that $h_{N+1}(1)<0$ and therefore (iii) holds.

The following consequence of Proposition 2.2 .9 and the irreducibility of the system might be well known but is interesting to notice.

Corollary 2.2.10. The system of equations (2.2.3) has a unique normalized solution if and only if it is stable. In this case $p_{n, m}>0$ holds for all $n=$ $0, \ldots, N$ and all $m \geq 0$. The system is unstable if and only if $p_{n, m}=0$, $n=0, \ldots, N, m \geq 0$, is the only solution.

Setting $h_{N+1}(1)<0$ in (2.2.40) and simplifying gives the stability condition

$$
\begin{equation*}
p \lambda_{1, N} \prod_{n=0}^{N-1} \rho_{n}<\mu_{2} \sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i} \tag{2.2.53}
\end{equation*}
$$

The following statement links the number of zeros of $\operatorname{det}(A)$ to the stability condition and is an immediate consequence of Theorem 2.2.6 and Proposition 2.2.9.

Theorem 2.2.11. $\operatorname{det}(A(z))$ is a polynomial of degree $N+1$ and has $N-1$ zeros in the interval $(0,1)$ and one zero at $z=1$. Additionally, $\operatorname{det}(A(z))$ has another zero in the interval $(1, \infty)$ if and only if the system (2.2.3) is stable, i.e., if and only if

$$
\begin{equation*}
\frac{p \lambda_{1, N}}{\mu_{2}}<\frac{\sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}}{\prod_{n=0}^{N-1} \rho_{n}} \tag{2.2.54}
\end{equation*}
$$

holds, where $\rho_{n}=\lambda_{1, n} / \mu_{1, n+1}, n=0, \ldots, N-1$.
Remark 2.2.12. By the condition given in Theorem 2.2.11, when the system is stable, meaning (2.2.54) holds, we can use the $N-1$ zeros $z_{1}, \ldots, z_{N-1}$ of $\operatorname{det}(A(z))=q_{N+1}(z)$ in $(0,1)$ to find the $N$ unknown probabilities $p_{1,0}, \ldots, p_{N, 0}$. This can be done by inserting these values into (2.2.28) for $n=0$. The $N-1$ pairwise distinct zeros then deliver $N-1$ equations for these unknowns, namely

$$
\operatorname{det}\left(A_{0}\left(z_{1}\right)\right)=0, \ldots, \operatorname{det}\left(A_{0}\left(z_{N-1}\right)\right)=0
$$

One more equation relating the unknowns is (2.2.14) which is equivalent to (2.2.50). This provides us with $N$ (independent) equations in the $N$
unknowns $p_{1,0}, \ldots, p_{N, 0}$. Observe that $\operatorname{det}\left(A_{0}\left(z_{i}\right)\right)$ and $\operatorname{det}\left(A_{n}\left(z_{i}\right)\right)$ for $i=$ $0, \ldots, N$ differ from each other by (2.2.28) only by a multiplicative constant, i.e.,

$$
\operatorname{det}\left(A_{0}\left(z_{i}\right)\right)=\frac{G_{n}\left(z_{i}\right)}{G_{0}\left(z_{i}\right)} \operatorname{det}\left(A_{n}\left(z_{i}\right)\right)
$$

for $i=0, \ldots, N$

A crucial point is to show that these equations are indeed linearly independent. For $N=2,3$ it is possible to show this analytically. Numerical calculations support this conjecture for larger values of $N$. It is conjectured that these equations are indeed independent (see also Avi-Itzhak and Mitrani [6], Levy and Yechiali [42], Perel and Yechiali [56], Yechiali [66]).

### 2.2.4 Stationary quantities and numerical aspects

Once $p_{1,0}, \ldots, p_{N, 0}$ are determined, $G_{0}, \ldots, G_{N}$ are given and we can calculate $E L_{2}$, the mean number of customers in $Q_{2}$, by (2.2.14), i.e.,

$$
\begin{equation*}
E L_{2}=\sum_{n=0}^{N} G_{n}^{\prime}(1) \tag{2.2.55}
\end{equation*}
$$

Furthermore, $\operatorname{Cov}\left(L_{1}, L_{2}\right)$ is of special interest because of the dependence of $Q_{2}$ on $Q_{1}$. By summing (2.2.20)-(2.2.22) over $n=0, \ldots, N$ we get

$$
\begin{equation*}
p \lambda_{1, N} z G_{N}(1)-\mu_{2} \sum_{n=1}^{N} G_{n}(z)=-\mu_{2} \sum_{n=1}^{N} n p_{n, 0} \tag{2.2.56}
\end{equation*}
$$

Differentiating this equation with respect to $z$ at $z=1$ leads to

$$
p \lambda_{1, N}\left(G_{N}(1)+G_{N}^{\prime}(1)\right)=\mu_{2} \sum_{n=1}^{N} n G_{n}^{\prime}(1)
$$

and thus

$$
\mu_{2} E\left(L_{1} L_{2}\right)=p \lambda_{1, N}\left(G_{N}(1)+G_{N}^{\prime}(1)\right)
$$

since $\sum_{n=1}^{N} n G_{n}^{\prime}(1)=E\left(L_{1} L_{2}\right)$. The latter equation is equivalent to

$$
\begin{equation*}
\mu_{2} E\left(L_{1} L_{2}\right)=p \lambda_{1, N}\left(P\left(L_{1}=N\right)+E\left(L_{2} \mid L_{1}=N\right) P\left(L_{1}=N\right)\right) \tag{2.2.57}
\end{equation*}
$$

since $G_{N}(1)=P\left(L_{1}=N\right)$ and $G_{N}^{\prime}(1)=E\left(L_{2} \mid L_{1}=N\right) P\left(L_{1}=N\right)$. Observe that $P\left(L_{1}=N\right)$ is known.

Remark 2.2.13. Numerical analysis shows that in general the covariance of $L_{1}$ and $L_{2}$ changes signs when the system parameters are varied. The numerical analysis rather suggests that $\operatorname{Cov}\left(L_{1}, L_{2}\right)$ as a function of $p$ on the interval $[0,1]$ is convex or concave and shows that it can be monotone decreasing, increasing or both with one or two zeros (where one zero is at $p=0$ ). Furthermore, there exists a threshold parameter $p^{*}$ for the overflow weight $p$ that separates the case of positive and nonnegative correlation from each other as one might expect. See Figure 2.3 for the shape of the function $p \mapsto \operatorname{Cov}\left(L_{1}, L_{2}\right)$ for $N=2$ with $\lambda_{1,0}=\lambda_{1,1}=\lambda_{1}, \mu_{1,1}=\mu_{1,2}=\mu_{1}$ and various parameter selections. An interesting case is $N=1$, where we can show that $\operatorname{Cov}\left(L_{1}, L_{2}\right)=0$ for all choices of these system parameters. $Q_{1}$ and $Q_{2}$ are in general not independent for $N=1$. The analysis of this case is carried out in the next section.

Further numerical analysis suggests that $\operatorname{Cov}\left(L_{1}, L_{2}\right) \geq 0$ for every choice of the system parameters for which the stability condition (2.2.54) holds and for which $Q_{1}$ is stable, meaning $\prod_{n=0}^{N-1} \rho_{n}<1$ (see Figure 2.3 and Table (2.1). This can be explained intuitively. Due to the overflow mechanism, the queue length in $Q_{2}$ can on the one hand only increase when the queue length in $Q_{1}$ increases and $Q_{1}$ reaches its capacity maximum. On the other hand, the service rate in $Q_{2}$ increases and therefore the queue length in $Q_{2}$ decreases simultaneously. Suppose that these two effects would overall lead to a decreasing queue length in $Q_{2}$. In this case it would be possible to raise $p \lambda_{1, N}$ slightly without leaving the stable regime in $Q_{2}$ and without affecting the queue length in $Q_{1}$. This would lead to a rising queue length in $Q_{2}$ and to $\operatorname{Cov}\left(L_{1}, L_{2}\right)>0$, which would contradict the assumption. Consequently, $\operatorname{Cov}\left(L_{1}, L_{2}\right)$ must be nonnegative if the system and $Q_{1}$ are stable. Due to the mentioned "stability reserve", this result will also hold if $Q_{1}$ is slightly unstable. This intuitive argument can be reversed in the case of a highly unstable $Q_{1}$, i.e., $\prod_{n=0}^{N-1} \rho_{n} \gg 1$. In this case, the first queue is on average fully occupied and the probability of overflow is close to 1 . Therefore, the queue length in $Q_{2}$ has a tendency to shorten if the queue length in $Q_{1}$ rises because otherwise $Q_{2}$ would become unstable.

In the case that (2.2.54) and $\prod_{n=0}^{N-1} \rho_{n}<1$ hold, equation (2.2.57) provides an upper bound for $E L_{2}$ in terms of the unknowns $p_{1,0}, \ldots, p_{N, 0}$. (2.2.54) and $\prod_{n=0}^{N-1} \rho_{n}<1$ imply that $\operatorname{Cov}\left(L_{1}, L_{2}\right) \geq 0$ and therefore $E\left(L_{1} L_{2}\right) \geq E L_{1} E L_{2}$. Then (2.2.57) yields

$$
\begin{equation*}
E L_{2} \leq \frac{p \lambda_{1, N} P\left(L_{1}=N\right)}{\mu_{2} E L_{1}}\left(1+E\left(L_{2} \mid L_{1}=N\right)\right), \tag{2.2.58}
\end{equation*}
$$



Figure 2.3: $\quad p \mapsto \operatorname{Cov}\left(L_{1}, L_{2}\right)$ for $N=2, \lambda_{1,0}=\lambda_{1,1}=\lambda_{1}, \mu_{1,1}=$ $\mu_{1,2}=\mu_{1}$ and $\left(\lambda_{1}, \mu_{1}, \mu_{2}\right)=(2.5,4,0.5),(3.8,4,2.5),(5.3,4,2.5),(4,2,1.2)$, $(5.4,1,2.5),(5.6,4,2.2)$ (line by line from left to right, the gray shaded region is the stability region, i.e., all $p$ that satisfy (2.2.19) ).
where $P\left(L_{1}=N\right)$ and $E L_{1}$ are known and

$$
E\left(L_{2} \mid L_{1}=N\right) P\left(L_{1}=N\right)=G_{N}^{\prime}(1)
$$

has to be determined. The additional benefit of this bound compared to formula (2.2.55) is that there is only one unknown quantity, namely $G_{N}^{\prime}(1)$. In case of heavy traffic for the first queue, i.e., $\prod_{n=0}^{N-1} \rho_{n} \rightarrow 1$, the first queue is always occupied implying $P\left(L_{1}=N\right)=1$ and $E L_{2}=E\left(L_{2} \mid L_{1}=N\right)$. Then (2.2.58) gives the bound

$$
\begin{equation*}
E L_{2} \leq \frac{p \lambda_{1, N}}{N \mu_{2}-p \lambda_{1, N}}, \tag{2.2.59}
\end{equation*}
$$

| $\lambda_{1}$ | $\mu_{1}=0.1$ | $\mu_{1}=0.2$ | $\mu_{1}=0.3$ | $\mu_{1}=0.4$ | $\mu_{1}=0.5$ | $\mu_{1}=0.6$ | $\mu_{1}=0.7$ |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| 0.1 | $\mathbf{0 . 0 2}$ | 0.056 | 0.051 | 0.041 | 0.033 | 0.027 | 0.022 |
| 0.2 | -0.061 | $\mathbf{0 . 0 3 8}$ | 0.069 | 0.074 | 0.07 | 0.063 | 0.056 |
| 0.3 | -0.152 | -0.013 | $\mathbf{0 . 0 5 5}$ | 0.083 | 0.092 | 0.091 | 0.087 |
| 0.4 | -0.259 | -0.085 | 0.018 | $\mathbf{0 . 0 7 2}$ | 0.098 | 0.109 | 0.111 |
| 0.5 | -0.385 | -0.176 | -0.04 | 0.041 | $\mathbf{0 . 0 8 8}$ | 0.113 | 0.125 |
| 0.6 | -0.532 | -0.287 | -0.118 | -0.008 | 0.062 | $\mathbf{0 . 1 0 5}$ | 0.129 |
| 0.7 | -0.705 | -0.421 | -0.217 | -0.075 | 0.019 | 0.082 | $\mathbf{0 . 1 2 2}$ |
| 0.8 | -0.908 | -0.58 | -0.338 | -0.164 | -0.041 | 0.043 | 0.101 |
| 0.9 | -1.148 | -0.77 | -0.485 | -0.274 | -0.122 | -0.012 | 0.066 |
| 1 | -1.436 | -0.998 | -0.663 | -0.412 | -0.224 | -0.086 | 0.014 |
| 1.1 | -1.787 | -1.275 | -0.882 | -0.581 | -0.354 | -0.183 | -0.055 |
| 1.2 | -2.221 | -1.617 | -1.151 | -0.792 | -0.517 | -0.307 | -0.148 |
| 1.3 | -2.774 | -2.048 | -1.488 | -1.056 | -0.722 | -0.465 | -0.267 |
| 1.4 | -3.498 | -2.606 | -1.921 | -1.392 | -0.984 | -0.668 | -0.422 |
| 1.5 | -4.487 | -3.353 | -2.493 | -1.833 | -1.325 | -0.931 | -0.625 |
| 1.6 | -5.918 | -4.405 | -3.28 | -2.431 | -1.782 | -1.281 | -0.894 |
| 1.7 | -8.171 | -5.991 | -4.43 | -3.281 | -2.42 | -1.764 | -1.261 |
| 1.8 | -12.237 | -8.655 | -6.261 | -4.582 | -3.366 | -2.464 | -1.783 |
| 1.9 | -21.772 | -14.05 | -9.622 | -6.809 | -4.903 | -3.556 | -2.574 |
| 2 | -70.383 | -30.757 | -17.789 | -11.477 | -7.82 | -5.485 | -3.899 |
| 2.1 |  |  | -66.638 | -27.383 | -15.431 | -9.77 | -6.54 |
| 2.2 |  |  |  |  | -84.64 | -27.299 | -14.299 |
| 2.3 |  |  |  |  |  |  | -426.843 |
| 2.4 |  |  |  |  |  |  |  |

Table 2.1: $\operatorname{Cov}\left(L_{1}, L_{2}\right)$ for $N=2, \lambda_{1,0}=\lambda_{1,1}=\lambda_{1}, \mu_{1,1}=\mu_{1,2}=\mu_{1}$, $p=\mu_{2}=1, \lambda_{1}=0.1,0.2 \ldots, 2.4$ and $\mu_{1}=0.1,0.2 \ldots, 0.7$ (bold numbers mark the diagonal $\lambda_{1}=\mu_{1}$, the cells are left blank in the cases where the system is unstable).
which is tight since there are always $N$ customers present in $Q_{1}$ serving $Q_{2}$ with rate $N \mu_{2}$ and the arrival rate of $Q_{2}$ is $p \lambda_{1, N}$. In the case $N=1$, equality also holds in (2.2.59) for arbitrary choice of the system parameters. This is shown in Section 2.2.6.

### 2.2.5 Model extension: External arrivals to second queue

The model can be generalized by equipping $Q_{2}$ with an arrival stream having exponentially distributed interarrival times with intensity $\lambda_{2}>0$. The arrival stream is independent of $Q_{1}$ and the arrival stream of $Q_{2}$ (see Figure 2.2 for the corresponding transition rate diagram). In this case, the generating functions $(\overline{2.2 .20})-(\sqrt{2.2 .22})$ and the auxiliary $\alpha$-functions $(\overline{2.2 .24})-(2.2 .26)$ are given by

$$
\left(\lambda_{1,0}+\lambda_{2}(1-z)\right) G_{0}(z)=\mu_{1,1} G_{1}(z)
$$

$$
\begin{aligned}
\left(\left(\lambda_{1, n}+\mu_{1, n}\right) z+\left(\lambda_{2} z-n \mu_{2}\right)(1-z)\right) G_{n}(z) & =\lambda_{1, n-1} z G_{n-1}(z) \\
+\mu_{1, n+1} z G_{n+1}(z) & -n \mu_{2}(1-z) p_{n, 0} \\
\left(\mu_{1, N} z+\left(\left(p \lambda_{1, N}+\lambda_{2}\right) z-N \mu_{2}\right)(1-z)\right) G_{N}(z)= & \lambda_{1, N-1} z G_{N-1}(z) \\
& -N \mu_{2}(1-z) p_{N, 0}
\end{aligned}
$$

where the second equation holds for $n=1, \ldots, N-1$ and

$$
\begin{aligned}
\alpha_{0}(z) & =\lambda_{1,0}+\lambda_{2}(1-z) \\
\alpha_{n}(z) & =\left(\lambda_{1, n}+\mu_{1, n}\right) z+\left(\lambda_{2} z-n \mu_{2}\right)(1-z) \text { for } n=1, \ldots, N-1 \\
\alpha_{N}(z) & =\mu_{1, N} z+\left(\left(p \lambda_{1, N}+\lambda_{2}\right) z-N \mu_{2}\right)(1-z)
\end{aligned}
$$

A procedure similar to the one for the basic model can be carried out, showing that $\operatorname{det}(A(z))$ is a polynomial of degree $2 N+1$ with $N-1$ zeros in $(0,1)$, one zero at $z=1$ and all other zeros in $(1, \infty)$ if and only if $h_{N+1}(1)<0$, where
$h_{N+1}(1)=p \prod_{i=0}^{N} \lambda_{1, i}-\left(\prod_{i=0}^{N-1} \mu_{1, i+1}\right)\left(\mu_{2} \sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}-\lambda_{2}\left(1+\sum_{n=1}^{N} \prod_{i=0}^{n-1} \rho_{i}\right)\right)$
with $\rho_{i}=\lambda_{1, i} / \mu_{1, i+1}$ for $i=0, \ldots, N-1$. The stability condition $h_{N+1}(1)<$ 0 (see also (1.2.1)) is equivalent to

$$
\frac{p \lambda_{1, N} \prod_{n=0}^{N-1} \rho_{n}}{\mu_{2} \sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}}+\frac{\lambda_{2}}{\mu_{2} E L_{1}}<1
$$

or $p \lambda_{1, N} P\left(L_{1}=N\right)+\lambda_{2}<\mu_{2} E L_{1}$ where $P\left(L_{1}=N\right)$ and $E L_{1}$ are given by (2.2.1). Setting $p=0$, the stability condition becomes $\lambda_{2}<\mu_{2} E L_{1}$ which corresponds to the result in [56. By letting $p=0, \lambda_{1, n}=\lambda_{1}$ and $\mu_{1, n+1}=\mu_{2}$ for $n=0, \ldots, N-1$ in our model, both models coincide.

Remark 2.2.14. It is possible to further generalize the model by assuming that the service rate in $Q_{2}$ is $\mu_{2, n}\left(\right.$ instead of $\left.n \mu_{2}\right)$ when $Q_{1}$ is in state $n$ as done in Section 2.2.2 and suggested in Perel and Yechiali 56. It has to be assumed that $\mu_{2, n} \geq 0$ for all $n \geq 0$ and $\mu_{2, n}>0$ for at least one value $n=1, \ldots, N$. The stability condition in this case is given by (2.2.17). This will lead to a modified construction of the generating functions, the auxiliary functions and the stability condition in this sequel (we omit the details). In order to proceed and investigate the existence and number of zeros of $\operatorname{det}(A(z))$, it is necessary to specify $\mu_{2, n}$.

### 2.2.6 Closed-form solution in the case of capacity one

For the case $N=1$, an analytic solution is available. While the case $N=2$ is also analytically solvable, the solutions are lengthy and we focus on the case $N=1$. Let $\lambda_{0,0}, \lambda_{1,0}, \mu_{2}>0, \mu_{1,1}=\mu_{1}>0$ and $p \in[0,1]$. For simplicity, we choose $\lambda_{0,0}=\lambda_{1,0}=\lambda_{1}$; by letting $p \in[0, \infty)$ the adequate choice of the parameter $p$ can be used to treat the case $\lambda_{1,0} \neq \lambda_{1,1}$ because $p \in[0,1]$ is no necessary condition. The balance equations are

$$
\begin{align*}
\lambda_{1} p_{0, m} & =\mu_{1} p_{1, m} \text { for } n=0, m \geq 0,  \tag{2.2.60}\\
\left(p \lambda_{1}+\mu_{1}\right) p_{1,0} & =\lambda_{1} p_{0,0}+\mu_{2} p_{1,1} \text { for } n=1, m=0 \text { and } \\
\left(p \lambda_{1}+\mu_{1}+\mu_{2}\right) p_{1, m} & =\lambda_{1} p_{0, m}+\mu_{2} p_{1, m+1}+p \lambda_{1} p_{1} \text { for } n=1 \text { and } m \geq 1 .
\end{align*}
$$

The equation (2.2.14) yields $p \lambda_{1} P\left(L_{1}=1\right)=\mu_{2}\left(E L_{1}-p_{1,0}\right)$ and from

$$
P\left(L_{1}=1\right)=E L_{1}=\frac{\lambda_{1}}{\lambda_{1}+\mu_{1}}
$$

and (2.2.60) we get

$$
\begin{align*}
& p_{1,0}=\frac{\mu_{2}-p \lambda_{1}}{\mu_{2}} \cdot E L_{1}=\frac{\lambda_{1}}{\lambda_{1}+\mu_{1}} \cdot \frac{\mu_{2}-p \lambda_{1}}{\mu_{2}} \quad \text { and }  \tag{2.2.61}\\
& p_{0,0}=\frac{\mu_{1}}{\lambda_{1}} p_{1,0}=\frac{\mu_{1}}{\lambda_{1}+\mu_{1}} \cdot \frac{\mu_{2}-p \lambda_{1}}{\mu_{2}} \tag{2.2.62}
\end{align*}
$$

The generating functions $G_{0}$ and $G_{1}$ satisfy

$$
\begin{aligned}
\lambda_{1,0} G_{0}(z) & =\mu_{1} G_{1}(z) \quad \text { and } \\
\left(z \mu_{1}+\left(z p \lambda_{1}-\mu_{2}\right)(1-z)\right) G_{1}(z) & =z \lambda_{1} G_{0}(z)-\mu_{2} p_{1,0}(1-z)
\end{aligned}
$$

for $|z| \leq 1$. These equations give

$$
G_{1}(z)=\frac{z \lambda_{1} G_{0}(z)-\mu_{2} p_{1,0}(1-z)}{z \mu_{1}+\left(z p \lambda_{1}-\mu_{2}\right)(1-z)}=\frac{z \mu_{1} G_{1}(z)-\mu_{2} p_{1,0}(1-z)}{z \mu_{1}+\left(z p \lambda_{1}-\mu_{2}\right)(1-z)}
$$

which yields

$$
\begin{aligned}
& G_{1}(z)=\frac{\mu_{2} p_{1,0}(1-z)}{\left(\mu_{2}-z p \lambda_{1}\right)(1-z)}=\frac{\mu_{2}-p \lambda_{1}}{\mu_{2}-z p \lambda_{1}} \cdot \frac{\lambda_{1}}{\lambda_{1}+\mu_{1}} \quad \text { and } \\
& G_{0}(z)=\frac{\mu_{1}}{\lambda_{1}} G_{1}(z)=\frac{\mu_{2}-p \lambda_{1}}{\mu_{2}-z p \lambda_{1}} \cdot \frac{\mu_{1}}{\lambda_{1}+\mu_{1}} .
\end{aligned}
$$

Observe that the stability condition is $p \lambda_{1}<\mu_{2}$ by Proposition 2.2.1, where $\lambda_{1}=\lambda_{1,1}$ if the condition $\lambda_{1,0}=\lambda_{1,1}=\lambda_{1}$ is dropped. Then, $\mu_{2}-z p \lambda_{1} \neq 0$ for all $|z| \leq 1$. Differentiating and setting $z=1$ leads to

$$
\begin{aligned}
G_{0}^{\prime}(1) & =\frac{p \lambda_{1} \mu_{1}}{\left(\lambda_{1}+\mu_{1}\right)\left(\mu_{2}-p \lambda_{1}\right)} \quad \text { and } \\
G_{1}^{\prime}(1) & =\frac{p \lambda_{1}^{2}}{\left(\lambda_{1}+\mu_{1}\right)\left(\mu_{2}-p \lambda_{1}\right)} .
\end{aligned}
$$

Finally, we can calculate the expected queue length of the second queue:

$$
E L_{2}=G_{0}^{\prime}(1)+G_{1}^{\prime}(1)=\frac{p \lambda_{1}}{\mu_{2}-p \lambda_{1}}
$$

Since $E\left(L_{1} L_{2}\right)=G_{1}^{\prime}(1)$, we get

$$
E\left(L_{1} L_{2}\right)=\frac{\lambda_{1}}{\lambda_{1}+\mu_{1}} \cdot \frac{p \lambda_{1}}{\mu_{2}-p \lambda_{1}}=E L_{1} E L_{2}
$$

Surprisingly on first sight, we get $\operatorname{Cov}\left(L_{1}, L_{2}\right)=0$ although $L_{1}$ and $L_{2}$ are in general not independent. The service rate in $Q_{2}$ for example equals 0 for $L_{1}=0$ and $\mu_{2}$ for $L_{1}=1$. An explanation for this phenomenon is the following. Under stationary conditions, $Q_{1}$ is in state 1 for a fraction $\lambda_{1} /\left(\lambda_{1}+\mu_{1}\right)$ of the time. In this case, $Q_{2}$ is busy for a fraction $p \lambda_{1} / \mu_{2}$ of the time. Since $Q_{2}$ is stable and only fed with customers or served if $L_{1}=1$ and remains unchanged if $L_{1}=0$, the queue length has to stay essentially the same and is not influenced by the queue length of $Q_{1}$. As mentioned in Remark 2.2.13 $\operatorname{Cov}\left(L_{1}, L_{2}\right)=0$ does not hold in general for $N>1$ because in this case the service rate in $Q_{2}$ is positive for $L_{1} \geq 1$, whereas the arrival rate in $Q_{2}$ is positive only if $L_{1}=N$. The service rate in $Q_{2}$ increases linearly with the queue length in $Q_{1}$ and the arrival rate of $Q_{2}$ is zero except in the case $L_{1}=N$. Therefore, the queue length in $Q_{2}$ tends to shorten with increasing $L_{1}$ up to the point $L_{1}=N$ where customers arrive to $Q_{2}$. In this case, the queue length in $Q_{2}$ can increase or decrease or stay unchanged in equilibrium, depending on the system parameters and the combination of these factors. This leads to a generally almost unpredictable behavior of $\operatorname{Cov}\left(L_{1}, L_{2}\right)$ (see also Remark (2.2.13).

### 2.3 Customers as servers: Models with jockeying

In this section, we discuss two variants of the basic model from Section 2.2, We let the assumptions be as for the basic model from Section 2.2 with the
following exception. In the first variant, the first customer of $Q_{2}$, if one is present, is forced to move to $Q_{1}$ as soon as $Q_{1}$ empties. This jockeying customer can then act as a server for the second queue. The stochastic processes that describes the queue lengths in this model is a quasi birth and death process ( QBP ). This model can be further generalized by letting a fixed number $1 \leq k \leq N-1$ jockey from $Q_{2}$ to $Q_{1}$ if $Q_{1}$ becomes empty (see Remark (2.3.9). In this case, it cannot be represented as a QBP because the transitions from states in the level process, i.e., the number of customers in $Q_{2}$, are not restricted to transitions to the states in the two adjacent levels. In the second variant of the basic model, as soon as $Q_{1}$ empties, $Q_{1}$ is filled with the customers from $Q_{2}$ until it reaches its capacity bound or $Q_{2}$ empties. We called this jockeying procedure unlimited jockeying. The stochastic processes that describes the queue lengths in this model is again a QBP.

The approach for reducing the number of unknowns in the first model can be carried over to these model variations. In turn, the steady-state probabilities of the vector of the queue lengths are functions of $p_{1,0}, \ldots, p_{N, 0}$ and we can give a set of $N$ (independent) equations for these unknowns. Steady-state quantities of interest can be computed by means of the generating functions. For the model with unlimited jockeying a complete analytic solution is available, since in this case, $Q_{1}$ and $Q_{2}$ can be regarded as a special case of a single queue with infinite capacity and state-dependent service and arrival rates. The analysis of this model is carried out in Section 2.3.4.

### 2.3.1 Model description and steady-state equations

Let $Q_{1}$ be as in the basic model and let $N \geq 2$. We consider the first variant of the basic model. In this variant, the first customer of $Q_{2}$, if one is present, is forced to move to $Q_{1}$ as soon as $Q_{1}$ empties. The case $N=1$ is equivalent to the second model variant. Let $Q_{2}$ be fed by the $p$-weighted overflow stream from $Q_{1}$, where $p \in[0,1]$, and let the service rate in $Q_{2}$ be $n \mu_{2}$ if $L_{1}=n$ where $\mu_{2}>0$. Modify $Q_{1}$ in the following way: If $Q_{1}$ becomes empty, then the first customer of $Q_{2}$ is instantly transferred to $Q_{1}$. Let $p_{n, m}$ be the steady-state probability of having $n$ customers present in $Q_{1}$ and $m$ customers in $Q_{2}, n=0, \ldots, N, m \geq 0$. We must have $p_{0, m}=0$ for $m \geq 1$, because in this case, one customer from $Q_{2}$ will be transferred instantly to $Q_{1}$. Let $\delta_{i j}$ be the Kronecker function. The balance equations for this model
for $n=0, \ldots, N$ and $m=0$ are

$$
\begin{align*}
\lambda_{1,0} p_{0,0}= & \mu_{1,1} p_{1,0}  \tag{2.3.1}\\
\left(\lambda_{1, n}+\mu_{1, n}\right) p_{n, 0}= & \lambda_{1, n-1} p_{n-1,0}+\mu_{1, n+1} p_{n+1,0}+\left(n \mu_{2}+\delta_{1 n} \mu_{1,1}\right) p_{n, 1} \\
& \quad \text { for } n=1, \ldots, N-1 \text { and }  \tag{2.3.2}\\
\left(p \lambda_{1, N}+\mu_{1, N}\right) p_{N, 0}= & \lambda_{1, N-1} p_{N-1,0}+N \mu_{2} p_{N, 1} . \tag{2.3.3}
\end{align*}
$$

The balance equations for $n=0, \ldots, N$ and $m \geq 1$ are

$$
\begin{align*}
p_{0, m}= & 0  \tag{2.3.4}\\
\left(\lambda_{1, n}+\mu_{1, n}+n \mu_{2}\right) p_{n, m}= & \lambda_{1, n-1} p_{n-1, m}\left(1-\delta_{1, n}\right)+\mu_{1, n+1} p_{n+1, m} \\
& +\left(n \mu_{2}+\delta_{1 n} \mu_{1,1}\right) p_{n, m+1} \tag{2.3.5}
\end{align*}
$$

for $n=1, \ldots, N-1$ and

$$
\begin{align*}
\left(p \lambda_{1, N}+\mu_{1, N}+N \mu_{2}\right) p_{N, m}= & \lambda_{1, N-1} p_{N-1, m}+p \lambda_{1, N} p_{N, m-1} \\
& +N \mu_{2} p_{N, m+1} \tag{2.3.6}
\end{align*}
$$

The Markov chain defined by these balance equations is irreducible if the state space is restricted to $\{(0,0)\} \cup\{(n, m) \mid n=1, \ldots, N, m \geq 1\}$; the states $(0, m), m \geq 1$, will not be visited. The transition rate diagram for this model is depicted in Figure 2.4, where $\lambda_{2}=0$. Observe that the states $(0, m), m \geq 1$, will not be reached in this diagram if $\lambda_{2}=0$.

The main ideas for deriving the stability condition and reducing the number of unknowns can be carried over from the basic model without jockeying. One might predict that the possibility of jockeying influences the queue length of the first queue and therefore could inhibit the determination of a closed-form expression for the queue length probabilities in the first queue without using further unknown quantities. Nevertheless, this complication can be avoided. The approach is carried out in the next sections.

### 2.3.2 Necessary and sufficient stability condition

The balance equations for $m=0$ and $n=0, \ldots, N$ give

$$
\begin{equation*}
p \lambda_{1, N} p_{N, 0}=\mu_{2} \sum_{n=1}^{N} n p_{n, 1}+\mu_{1,1} p_{1,1} \tag{2.3.7}
\end{equation*}
$$



Figure 2.4: Limited jockeying: Transition rate diagram.
after summation over $n=0, \ldots, N$. The balance equations for $m \geq 1$ and $n=0, \ldots, N$ yield

$$
\begin{align*}
p \lambda_{1, N} p_{N, m}+\mu_{1,1} p_{1, m} & +\mu_{2} \sum_{n=1}^{N} n p_{n, m} \\
& =p \lambda_{1, N} p_{N, m-1}+\mu_{1,1} p_{1, m+1}+\mu_{2} \sum_{n=1}^{N} n p_{n, m+1} \tag{2.3.8}
\end{align*}
$$

by summing over $n=0, \ldots, N$. From (2.3.7) and (2.3.8) we obtain analogously to the derivations for the basic model that

$$
\begin{equation*}
p \lambda_{1, N} P\left(L_{1}=N\right)=\mu_{1,1} \sum_{m \geq 1} p_{1, m}+\mu_{2}\left(E L_{1}-\sum_{n=1}^{N} n p_{n, 0}\right) . \tag{2.3.9}
\end{equation*}
$$

All terms in the above equation (except the system parameters) are unknown for this model. Nevertheless, the set of unknown probabilities in the system of steady-state equations can be reduced to $p_{1,0}, \ldots, p_{N, 0}$. This is done in the next section.

From Theorem 1.2 .3 we get the necessary and sufficient stability condition.

Proposition 2.3.1. The system of equations (2.3.1) -(2.3.6) has a unique nonnegative and normalized solution if and only if

$$
\begin{equation*}
p \lambda_{1, N}<\frac{\mu_{2} \sum_{n=1}^{N} n \prod_{i=1}^{n-1} \rho_{i}+\mu_{1,1}}{\prod_{n=1}^{N-1} \rho_{n}} \tag{2.3.10}
\end{equation*}
$$

holds, where $\rho_{n}=\lambda_{1, n} / \mu_{1, n+1}$ for $n=0, \ldots, N-1$.

Proof. The proof is an almost verbatim repetition of the proof of Proposition 2.2.1. We regard $Q_{1}$ as the phase and $Q_{2}$ as the level of the underlying quasi birth and death process with phase and level given by $L_{1}$ and $L_{2}$, respectively. The stability condition (2.3.10) is then derived from Theorem 1.2 .3 in the following way. The exponential arrival rate in $Q_{2}$ is $p \lambda_{1, N}$ if $L_{1}=N$ and 0 if $L_{1}<N$. The exponential service rate in $Q_{2}$ is $n \mu_{2}$ given $L_{1}=n, n=0, \ldots, N$. The rates for arrivals in $Q_{1}$ are $\lambda_{1, n}$ given $L_{1}=n$ for $n=0$ and $n=2,3, \ldots, N-1$ and $\lambda_{1,1}+\mu_{1,1}$ for $L_{1}=1$. Observe that the arrival stream in $Q_{1}$ consists of the original external arrival stream and the jockeying customers from $Q_{2}$. The service rate in $Q_{1}$ is $\mu_{1, n}$ given $L_{1}=n$ for $n=0, \ldots, N$, where $\mu_{1,0}=0$. Observe that the states $(0, m)$ for $m \geq 1$ are not visited by the quasi birth and death process. Therefore, we can restrict the state space of the process to $\{(n, m) \mid n=1, \ldots, N, m \geq 1\}$ and consequently, the phase process has the state space $\{1, \ldots, N\}$. This ensures the irreducibility of the whole process. With this setting, the $(N \times N)$-matrices $A_{0}, A_{1}$ and $A_{2}$ from Theorem 1.2 .3 are given by

$$
A_{0}=\operatorname{diag}\left(0, \ldots, 0, p \lambda_{1, N}\right), \quad A_{2}=\operatorname{diag}\left(\mu_{2}, 2 \mu_{2}, \ldots, N \mu_{2}\right)
$$

and $A_{1}=A-A_{0}-A_{2}$, where $A$ is the rate matrix of the phase process governing $Q_{1}$, i.e., the standard birth and death process on $\{1, \ldots, N\}$ with birth rate $\lambda_{1,1}+\mu_{2}$ in state 1 , birth rates $\lambda_{1, n}$ in the states $n=2, \ldots, N-$ 1 and death rates $\mu_{1, n}$ in the states $n=2, \ldots, N$. The vector $\pi$ is the stationary probability measure of the phase process in $Q_{1}$, i.e., $\pi_{n}=P\left(L_{1}=\right.$ $n), n=1, \ldots, N$, is given by (2.2.1) with a suitable normalization factor. The stability condition (1.2.1) is then easily computed as (2.3.10).

One might predict from equation 2.3 .9 and

$$
P\left(L_{1}=1\right)=\sum_{m \geq 1} p_{1, m}=\sum_{m \geq 0} p_{1, m}-p_{1,0}
$$

that the system is stable if $p_{1,0}>0$ or $\sum_{n=1}^{N} p_{n, 0}>0$. In this case a stability condition could be

$$
\begin{equation*}
p \lambda_{1, N} P\left(L_{1}=N\right)<\mu_{1,1} P\left(L_{1}=1\right)+\mu_{2} E L_{1} \tag{2.3.11}
\end{equation*}
$$

In addition to this condition, the stability condition has an interesting interpretation in view of the results of Section 2.2 and the proof of Proposition 2.3.1. Let $L_{1}^{*}$ be the stationary queue length of the first queue in the basic model, i.e., the queue without arrivals from jockeying customers from $Q_{2}$ and distribution given by (2.2.1). By multiplying the inequality (2.3.10) with $\prod_{n=1}^{N-1} \rho_{n}$ we can rephrase the stability condition as

$$
\begin{equation*}
p \lambda_{1, N} P\left(L_{1}^{*}=N\right)<\mu_{1,1} P\left(L_{1}^{*}=1\right)+\mu_{2} E L_{1}^{*} \tag{2.3.12}
\end{equation*}
$$

where we have used $\sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}=E L_{1}^{*} / P\left(L_{1}^{*}=0\right), \prod_{n=0}^{N-1} \rho_{n}=P\left(L_{1}^{*}=\right.$ $N) / P\left(L_{1}^{*}=0\right)$ and $\rho_{0}=P\left(L_{1}^{*}=1\right) / P\left(L_{1}^{*}=0\right)$. This equation states that the average arrival rate in $Q_{2}$ should be smaller than the average departure rate in $Q_{2}$ in the case of no jockeying. The average departure rate is the sum of the average service rate delivered by the servers in $Q_{2}$ and the average departure rate of jockeying customers. One might expect that the conditions (2.3.11) and (2.3.12) are equivalent. We will show in the next section that they differ only by a positive constant and are therefore indeed equivalent. As for the basic model, one can show that a unique normalized solution of the steady-state equations (2.3.1)-(2.3.6) exists if and only if an $n \in\{0, \ldots, N\}$ and an $m \geq 0$ exist with $p_{n, m}>0$ (see also Proposition 2.2 .9 and Corollary 2.2.10).

### 2.3.3 Generating functions and steady-state distribution

In this section, we derive the recurrence equations for the probability generating functions

$$
G_{n}(z)=\sum_{m=0}^{\infty} p_{n, m} z^{m}, \quad|z| \leq 1
$$

As in the basic model, these relations are used to reduce the set of unknown steady-state probabilities to the unknowns $p_{1,0}, \ldots, p_{N, 0}$. The stability condition will then be related to the existence of $N-1$ pairwise distinct zeros
of a function originating from the recurrence equations for the probability generating functions.

We can assume for the moment that the system is stable and that (2.3.10) holds. In this case, the generating functions are well defined on $|z| \leq 1$. By multiplying the equations (2.3.1)-(2.3.6) by $z^{m}$ and summing over $m \geq 0$ we get, after simplifying,

$$
\begin{align*}
\lambda_{1,0} G_{0}(z) & =\mu_{1,1} p_{1,0}=\lambda_{1,0} p_{0,0}  \tag{2.3.13}\\
\left(\lambda_{1,1} z-\left(\mu_{1,1}+\mu_{2}\right)(1-z)\right) G_{1}(z) & =\mu_{1,2} z G_{2}(z)  \tag{2.3.14}\\
& -\left(\mu_{1,1}+\mu_{2}\right)(1-z) p_{1,0} \\
\left(\left(\lambda_{1, n}+\mu_{1, n}\right) z-n \mu_{2}(1-z)\right) G_{n}(z)= & \lambda_{1, n-1} z G_{n-1}(z) \\
& +\mu_{1, n+1} z G_{n+1}(z)  \tag{2.3.15}\\
& -n \mu_{2}(1-z) p_{n, 0} \\
\left(\mu_{1, N} z+\left(p \lambda_{1, N} z-N \mu_{2}\right)(1-z)\right) G_{N}(z)= & \lambda_{1, N-1} z G_{N-1}(z) \\
& -N \mu_{2}(1-z) p_{N, 0} \tag{2.3.16}
\end{align*}
$$

where the third equation holds for $n=2, \ldots, N-1$ and vanishes in the case $N=2$. In order to write these equations in matrix from, we define the vectors

$$
\begin{gathered}
G(z)=\left(G_{1}(z), \ldots, G_{N}(z)\right)^{\top} \\
p=\left(\left(\mu_{1,1}+\mu_{2}\right) p_{1,0}, 2 \mu_{2} p_{2,0}, \ldots, N \mu_{2} p_{N, 0}\right)^{\top}
\end{gathered}
$$

and the matrix $A(z) \in \operatorname{Mat}(N, N, \mathbb{R})$ by

$$
A(z)=\left(\begin{array}{cccccc}
\alpha_{1}(z) & -\mu_{1,2} z & 0 & \cdots & \cdots & 0  \tag{2.3.17}\\
-\lambda_{1,0} z & \alpha_{2}(z) & -\mu_{1,3} z & \ddots & & \vdots \\
0 & \ddots & \alpha_{3}(z) & \ddots & \ddots & \vdots \\
\vdots & \ddots & \ddots & \ddots & \ddots & 0 \\
\vdots & & \ddots & \ddots & \ddots & -\mu_{1, N} z \\
0 & \ldots & \ldots & 0 & -\lambda_{1, N-1} z & \alpha_{N}(z)
\end{array}\right)
$$

where

$$
\begin{align*}
\alpha_{1}(z) & =\lambda_{1,1} z-\left(\mu_{1,1}+\mu_{2}\right)(1-z)  \tag{2.3.18}\\
\alpha_{n}(z) & =\left(\lambda_{1, n}+\mu_{1, n}\right) z-n \mu_{2}(1-z) \text { for } n=2, \ldots, N-1 \text { and }  \tag{2.3.19}\\
\alpha_{N}(z) & =\mu_{1, N} z+\left(p \lambda_{1, N} z-N \mu_{2}\right)(1-z) \tag{2.3.20}
\end{align*}
$$

For simplicity let $\alpha_{0}(z)=0$. The dimension of the matrix is $N$ whereas it was $N+1$ in (2.2.23). We dropped the first line of the matrix in (2.2.23) because the generating function $G_{0}$ does only depend on $p_{1,0}$ which itself is determined by the functions $G_{1}, \ldots, G_{N}$.

The equations (2.3.14), (2.3.15) and (2.3.16) are equivalent to

$$
\begin{equation*}
A(z) G(z)=-(1-z) p \tag{2.3.21}
\end{equation*}
$$

By Cramer's rule we have

$$
\begin{equation*}
\operatorname{det}(A(z)) G_{n}(z)=\operatorname{det}\left(A_{n}(z)\right) \tag{2.3.22}
\end{equation*}
$$

for all values $z$ such that $A(z)$ is invertible, where $A_{n}(z)$ is the matrix obtained from $A(z)$ by replacing the $n$-th column with the vector $-(1-z) p$. The generating functions $G_{0}, \ldots, G_{N}$ are uniquely determined by the equations (2.3.22) and $p_{1,0}, \ldots, p_{N, 0}$, since these are the only unknowns occurring in these equations.
$\operatorname{det}(A(z))$ is a polynomial in $z$ of degree $N+1$. We will show in the following that $\operatorname{det}(A(z))$ has $N-1$ zeros in the open interval $(0,1)$ and one zero at $z=1$. Additionally, we will show that $\operatorname{det}(A(z))$ has another zero in the open interval $(1, \infty)$ if and only the stability condition (2.3.10) holds. The $N-1$ zeros of $\operatorname{det}(A(z))$ in $(0,1)$ will provide us with $N-1$ linear homogeneous equations in the unknowns $p_{1,0}, \ldots, p_{N, 0}$. Another linear equation yielding a system of $N$ linear equations can be derived from (2.3.9) and (2.3.41) in a similar manner as for the basic model from Section 2.2.

Observe that (2.3.9) can be written as

$$
\begin{equation*}
p \lambda_{1, N} G_{N}(1)=\mu_{1,1} G_{1}(1)+\mu_{2} \sum_{n=1}^{N} n G_{n}(1)-\mu_{1,1} p_{1,0}-\mu_{2} \sum_{n=1}^{N} n p_{n, 0} \tag{2.3.23}
\end{equation*}
$$

and does not involve $G_{0}(1)$. The normalization condition which is

$$
\begin{equation*}
\sum_{n=0}^{N} \sum_{m \geq 0} p_{n, m}=\sum_{n=0}^{N} G_{n}(1)=1 \tag{2.3.24}
\end{equation*}
$$

in terms of the generating functions is equivalent to

$$
\begin{equation*}
G_{1}(1)=\frac{\lambda_{1,0}-\mu_{1,1} p_{1,0}}{\lambda_{1,0} \sum_{n=1}^{N} \prod_{i=1}^{n-1} \rho_{i}} \tag{2.3.25}
\end{equation*}
$$

where $\mu_{1,1} p_{1,0}=\lambda_{1,0} p_{0,0}=\lambda_{1,0} G_{0}(1)$ and where we have written $\rho_{i}=$ $\lambda_{1, i} / \mu_{1, i+1}$ for $i=0, \ldots, N-1$. This can be shown by solving the equations (2.3.13)- (2.3.16) for $z=1$ in terms of $G_{0}(1)$ and $G_{1}(1)$ and using the normalization condition. These equations are

$$
\begin{align*}
\lambda_{1,0} G_{0}(1) & =\mu_{1,1} p_{1,0}  \tag{2.3.26}\\
\lambda_{1,1} G_{1}(1) & =\mu_{1,2} G_{2}(1)  \tag{2.3.27}\\
\left(\lambda_{1, n}+\mu_{1, n}\right) G_{n}(1) & =\lambda_{1, n-1} G_{n-1}(1)+\mu_{1, n+1} G_{n+1}(1) \quad \text { and }  \tag{2.3.28}\\
\mu_{1, N} G_{N}(z) & =\lambda_{1, N-1} G_{N-1}(1) \tag{2.3.29}
\end{align*}
$$

where $n=2, \ldots, N-1$ in the third equation. The solution of the equations (2.3.27)- (2.3.29) in terms of $G_{1}(1)$ is given by

$$
\begin{equation*}
G_{n}(1)=G_{1}(1) \prod_{i=1}^{n-1} \rho_{i} \text { for } n=1, \ldots, N . \tag{2.3.30}
\end{equation*}
$$

By the normalization condition (2.3.24), summation of (2.3.26) and (2.3.30) over $n=1, \ldots, N$ yields (2.3.25). Substituting (2.3.25) and (2.3.30) in (2.3.23) and collecting terms finally yields the additional equation

$$
\begin{equation*}
\frac{\mu_{1,1} p_{1,0}+\mu_{2} \sum_{n=1}^{N} n p_{n, 0}}{\lambda_{1,0}-\mu_{1,1} p_{1,0}}=\frac{\mu_{1,1}+\mu_{2} \sum_{n=1}^{N} n \prod_{i=1}^{n-1} \rho_{i}-p \lambda_{1, N} \prod_{i=1}^{N-1} \rho_{i}}{\lambda_{1,0} \sum_{n=1}^{N} \prod_{i=1}^{n-1} \rho_{i}} \tag{2.3.31}
\end{equation*}
$$

that relates the unknowns $p_{1,0}, \ldots, p_{N, 0}$ to the system parameters.
Let $G_{n}^{*}(1)$ be the generating function of the phase in the basic model at $z=1$, i.e., let $G_{0}^{*}(1)$ be given by (2.2.47) and let $G_{n}^{*}(1)$ be given by (2.2.46) for $n=1, \ldots, N$. Equation (2.3.30) shows that $G_{n}(1)$ and $G_{n}^{*}(1)$ differ only by a constant independent of $n$, namely

$$
\begin{equation*}
G_{n}(1)=G_{n}^{*}(1)\left(\frac{G_{1}(1)}{G_{0}^{*}(1) \rho_{0}}\right) \tag{2.3.32}
\end{equation*}
$$

for $n=1, \ldots, N$. We have shown the following result.
Remark 2.3.2. The equations (2.3.11) and (2.3.12) are equivalent given $p_{1,0}>0$.

The equation (2.3.32) has another consequence that yields a lower bound for $E L_{1}$. A real-valued random variable $X$ is called stochastically larger than the real-valued random variable $Y$, if $P(X>t) \geq P(Y>t)$ for all $t \in \mathbb{R}$.

Proposition 2.3.3. Under stationary conditions, i.e., if (2.3.10) is fulfilled, the queue length $L_{1}$ of the first queue with jockeying is stochastically larger than the queue length $L_{1}^{*}$ of the first queue without jockeying. Thus, a lower bound for $E L_{1}$ is given by

$$
E L_{1} \geq E L_{1}^{*}=\frac{\sum_{n=1}^{N} n \prod_{i=0}^{n-1} \rho_{i}}{1+\sum_{n=1}^{N} \prod_{i=0}^{n-1} \rho_{i}}
$$

We will not give a formal proof but point out the idea. Under stationary conditions, the equation (2.3.32) implies that $G_{n}(1) \geq G_{n}^{*}(1)$ for all $n=$ $1, \ldots, N$ if this is the case for at least one $n \in\{1, \ldots, N\}$ or if $G_{0}(1) \leq G_{0}^{*}(1)$ which implies the first condition. Therefore,

$$
E L_{1}=\sum_{n=1}^{N} n G_{n}(1) \geq \sum_{n=1}^{N} n G_{n}^{*}(1)=E L_{1}^{*}
$$

holds.
Now we turn our attention to the derivation of the auxiliary functions $q_{1}, \ldots, q_{N}$ and their properties. This leads to the stability theorem concerning the zeros of $\operatorname{det}(A)$. For $n=1, \ldots, N+1$ we let

$$
q_{1}(z)=\alpha_{1}(z), q_{2}(z)=\operatorname{det}\left(\begin{array}{cc}
\alpha_{1}(z) & -\mu_{1,2} z  \tag{2.3.33}\\
-\lambda_{1,1} z & \alpha_{2}(z)
\end{array}\right), \ldots, q_{N}(z)=\operatorname{det}(A(z)) .
$$

By Laplace expansion we get

$$
\begin{align*}
& q_{1}(z)=\alpha_{1}(z) q_{0}(z) \quad \text { and } \\
& q_{n}(z)=\alpha_{n}(z) q_{n-1}(z)-\lambda_{1, n-1} \mu_{1, n} z^{2} q_{n-2}(z) \tag{2.3.34}
\end{align*}
$$

for $n=2, \ldots, N$, where $q_{0}(z)=1$.
The crucial properties of the functions $q_{1}, \ldots, q_{N}$ and $\alpha_{1}, \ldots, \alpha_{N}$ are derived in the same manner as in Proposition 2.2.4,

Proposition 2.3.4. The function $q_{n}$ is a polynomial in $z$ of degree $n$ for $n=1, \ldots, N-1$ and of degree $N+1$ for $n=N$. The functions $q_{n}$ and $\alpha_{n}$, $n=1, \ldots, N$, satisfy the following properties:
(i) $q_{n}$ and $q_{n+1}$ have no common root in $(0,1)$ for $n=0, \ldots, N-1$.
(ii) $\operatorname{sgn}\left(\alpha_{0}(0)\right)=1$ and $\operatorname{sgn}\left(\alpha_{n}(0)\right)=-1$ for $n=1, \ldots, N$.
(iii) $\operatorname{sgn}\left(q_{n}(0)\right)=(-1)^{n+1}$ for $n=1, \ldots, N$.
(iv) $q_{n}(1)=\prod_{i=0}^{n-1} \lambda_{1, i}$ for $n=0, \ldots, N-1$ and $q_{N}(1)=0$.
(v) For $n=1, \ldots, N-1$ the following implication holds: If $\tilde{z}>0$ with $q_{n}(\tilde{z})=0$, then

$$
\operatorname{sgn}\left(q_{n-1}(\tilde{z}) q_{n+1}(\tilde{z})\right)=-1
$$

(vi) $q_{n}$ has $n-1$ pairwise distinct zeros in $(0,1)$ for $n=1, \ldots, N-1$.
(vii) $\lim _{z \rightarrow \infty} q_{n}(z)=\infty$ for $n=2, \ldots, N-1$ and $\lim _{z \rightarrow \infty} q_{N}(z)=-\infty$.

As for the basic model, we can determine functions $h_{1}, \ldots, h_{N}$ on $[0, \infty)$ satisfying

$$
\begin{align*}
& q_{n}(z)=z^{n} \prod_{i=1}^{n} \lambda_{1, i}+(1-z) h_{n}(z) \text { for } n=1, \ldots, N-1 \text { and }  \tag{2.3.35}\\
& q_{N}(z)=(1-z) h_{N}(z)
\end{align*}
$$

inductively from (2.3.34). These functions are given by

$$
\begin{align*}
& h_{1}(z)=-\left(\mu_{1,1}+\mu_{2}\right)  \tag{2.3.36}\\
& h_{2}(z)=\alpha_{2}(z) h_{n}(z)(1-z)  \tag{2.3.37}\\
& h_{n}(z)=-n \mu_{2} z^{n-1} \prod_{i=1}^{n-1} \lambda_{1, i}+\alpha_{n}(z) h_{n-1}(z)-\lambda_{1, n-1} \mu_{1, n} z^{2} h_{n-2}(z) \tag{2.3.38}
\end{align*}
$$

for $n=3, \ldots, N-1$ and

$$
\begin{array}{r}
h_{N}(z)=\left(p \lambda_{1, N} z-N \mu_{2}\right) z^{N-1} \prod_{i=0}^{N-1} \lambda_{1, i}+\alpha_{N}(z) h_{N-1}(z)  \tag{2.3.39}\\
-\lambda_{1, N-1} \mu_{1, N} z^{2} h_{N-2}(z)
\end{array}
$$

for all $z \in[0, \infty)$. See Proposition 2.2.5 in the previous section for a similar derivation for the basic model. From the previous proposition we can derive the stability condition in terms of $h_{N}(1)$ in the same manner as in Theorem 2.2.6.

Theorem 2.3.5. $\operatorname{det}(A(z))$ is a polynomial of degree $N+1$ and has $N-1$ distinct zeros in the interval $(0,1)$ and one zero at $z=1$. Additionally, $\operatorname{det}(A(z))$ has another zero in the interval $(1, \infty)$ if and only if $h_{N}(1)<0$, where $h_{N}$ is defined by $\operatorname{det}(A(z))=(1-z) h_{N}(z)$ for all $z>0$.

It remains to determine $h_{N}(1)$. Evaluating (2.3.36)-(2.3.38) and (2.3.39) at $z=1$ gives a recursive system of equations for $h_{1}(1), \ldots, h_{N}(1)$. We omit the details and state the solution:

$$
\begin{align*}
h_{n}(1)= & -\mu_{2}\left(\prod_{i=1}^{n} \lambda_{1, i}\right)\left(\sum_{k=1}^{n} \frac{1}{\mu_{1, k+1}}\left(\prod_{i=1}^{k} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right)\left(\sum_{j=1}^{k} j \prod_{i=1}^{j-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right)\right) \\
& -\mu_{1,1}\left(\prod_{i=1}^{n} \lambda_{1, i}\right)\left(\sum_{k=1}^{n} \frac{1}{\mu_{1, k+1}} \prod_{i=1}^{k} \frac{\mu_{1, i+1}}{\lambda_{1, i}}\right) \tag{2.3.40}
\end{align*}
$$

for $n=1, \ldots, N-1$, where the empty sum and empty product are defined to be 0 and 1, respectively. Inserting $h_{N-1}(1)$ and $h_{N-2}(1)$ from (2.3.40) into (2.3.39) and simplifying finally leads to

$$
\begin{equation*}
h_{N}(1)=p \prod_{i=1}^{N} \lambda_{1, i}-\mu_{2}\left(\prod_{i=1}^{N-1} \mu_{1, i+1}\right)\left(\sum_{n=1}^{N} n \prod_{i=1}^{n-1} \frac{\lambda_{1, i}}{\mu_{1, i+1}}\right)-\prod_{i=0}^{N-1} \mu_{1, i+1} \tag{2.3.41}
\end{equation*}
$$

Setting $h_{N}(1)<0$ and simplifying gives the stability condition from Proposition 2.3.1. The next theorem combines Proposition 2.3.1, Remark 2.3.2 and Theorem 2.3.5 and relates the number of zeros of $\operatorname{det}(A(z))$ to this stability condition.

Theorem 2.3.6. $\operatorname{det}(A(z))$ is a polynomial of degree $N+1$ and has $N-1$ distinct zeros in the interval $(0,1)$ and one zero at $z=1$. Additionally, $\operatorname{det}(A(z))$ has another zero in the interval $(1, \infty)$ if and only if the system of equations $(2.3 .1)-(\sqrt{2.3 .6})$ has a unique nonnegative and normalized solution, i.e., if and only if $p \lambda_{1, N} P\left(L_{1}=N\right)<\mu_{1,1} P\left(L_{1}=1\right)+\mu_{2} E L_{2}$ or equivalently

$$
\begin{equation*}
p \lambda_{1, N}<\frac{\mu_{2} \sum_{n=1}^{N} n \prod_{i=1}^{n-1} \rho_{i}+\mu_{1,1}}{\prod_{n=1}^{N-1} \rho_{n}} \tag{2.3.42}
\end{equation*}
$$

holds, where $\rho_{n}=\lambda_{1, n} / \mu_{1, n+1}, n=0, \ldots, N-1$.
Remark 2.3.7. The stability condition is independent of the arrival rate $\lambda_{1,0}$ in $Q_{1}$. This is on the one hand due to the fact that the stability is derived from the generating functions $G_{1}, \ldots, G_{N}$ which are independent of $\lambda_{1,0}$. On the other hand, the arrival rate in $Q_{1}$ is $\lambda_{1,0}$ only in the case that no customers are present in both queues. Therefore, the arrival rate $\lambda_{1,0}$ influences the length of the idle period of the system, but not the busy
period. Due to the jockeying discipline, $Q_{1}$ is in state 0 if and only if $Q_{2}$ is empty.

Remark 2.3.8. By the condition given in Theorem 2.2.11, when the system is stable, meaning $(2.2 .54)$ holds, we can use the $N-1$ zeros of $\operatorname{det}(A(z))=$ $q_{N}(z)$ in $(0,1)$ to find the $N$ unknown probabilities $p_{1,0}, \ldots, p_{N, 0}$. The $N-1$ pairwise distinct zeros deliver $N-1$ equations for these unknowns. One more equation relating the unknowns is (2.3.31). This provides us with $N$ (independent) equations in the $N$ unknowns $p_{1,0}, \ldots, p_{N, 0}$. For further comments on the procedure see Remark 2.2.12.

Once $p_{1,0}, p_{2,0}, \ldots, p_{N, 0}$ are determined, $G_{0}, \ldots, G_{N}$ are given and we can calculate $E L_{1}, E L_{2}$ and $\operatorname{Cov}\left(L_{1}, L_{2}\right)$ :

$$
E L_{1}=\sum_{n=1}^{N} n G_{n}(1), E L_{2}=\sum_{n=0}^{N} G_{n}^{\prime}(1) \text { and } E\left(L_{1} L_{2}\right)=\sum_{n=1}^{N} n G_{n}^{\prime}(1)
$$

An equation similar to (2.2.57) for the basic model relating these terms is the equation

$$
\begin{aligned}
& \mu_{2} E\left(L_{1} L_{2}\right)+\mu_{1,1} E\left(L_{2} \mid L_{1}=1\right) P\left(L_{1}=1\right) \\
& =p \lambda_{1, N}\left(P\left(L_{1}=N\right)+E\left(L_{2} \mid L_{1}=N\right)\left(P\left(L_{1}=N\right)\right)\right)
\end{aligned}
$$

where $E\left(L_{1} L_{2}\right)=\sum_{n=1}^{N} n G_{n}^{\prime}(1), E\left(L_{2} \mid L_{1}=n\right) P\left(L_{1}=n\right)=G_{n}^{\prime}(1)$ for $n=1, \ldots, N$ and $P\left(L_{1}=N\right)=G_{N}(1)$. The equation is shown by summing the equations (2.3.13)-(2.3.16) and differentiating the result at $z=1$.
Remark 2.3.9. It is possible to further generalize the model by:

1. Equipping $Q_{2}$ with an exponential arrival stream with rate $\lambda_{2}>0$ (see Figure 2.4).
2. Assuming that the service rate in $Q_{2}$ is $\mu_{2, n}\left(\right.$ instead of $\left.n \mu_{2}\right)$ when $Q_{1}$ is in state $n$ (see also Remark 2.2.14).
3. Letting a fixed number $1 \leq k \leq N-1$ jockey to $Q_{1}$ if $Q_{1}$ becomes empty.

In all cases, this will lead to a modified construction of the generating functions, the auxiliary functions and the stability condition (we omit the details). In the second case, the stability condition involves additional terms depending on the system parameters (compare (2.2.40) with (2.3.41) and (2.2.54) with (2.3.42)). The steady-state equations simplify substantially for $k=N$. The solution in this case is given in the next section.

### 2.3.4 Closed-from solution in the case of unlimited jockeying

For this model a complete analytic solution is available. In order to simplify the resulting formulas we assume constant arrival and service rates in $Q_{1}$. Let $N>1$ and let $Q_{1}$ be an $M / M / 1 / N-1$ queue with arrival rate $\lambda_{1}>0$ and service rate $\mu_{1}>0$. Let $Q_{2}$ be fed by the $p$-weighted overflow stream from $Q_{1}, p \in[0,1]$, and let the service rate in $Q_{2}$ be $n \mu_{2}$ if $L_{1}=n$ where $\mu_{2}>0$. Modify $Q_{1}$ in the following way: If $Q_{1}$ becomes empty, then the first customers waiting in $Q_{2}$ are instantly transferred to $Q_{1}$ until $Q_{1}$ is fully occupied or $Q_{2}$ empties, whatever happens first. Let $p_{n, m}$ be the steadystate probability of having $n$ customers present in $Q_{1}$ and $m$ customers in $Q_{2}, n=0, \ldots, N, m \geq 0$. We must have $p_{n, m}=0$ for $m \geq 1$ and $n<N$, because in this case, some customers will be transferred instantly from $Q_{2}$ to $Q_{1}$. The balance equations for this model are

$$
\begin{align*}
\lambda_{1} p_{0,0} & =\mu_{1} p_{1,0}  \tag{2.3.43}\\
\left(\lambda_{1}+\mu_{1}\right) p_{n, 0} & =\lambda_{1} p_{n-1,0}+\mu_{1} p_{n+1,0}, n=1, \ldots, N-1,  \tag{2.3.44}\\
\left(p \lambda_{1}+\mu_{1}\right) p_{N, 0} & =\lambda_{1} p_{N-1,0}+\left(\mu_{1}+N \mu_{2}\right) p_{N, 1}, n=N, m=0 \tag{2.3.45}
\end{align*}
$$

$$
\begin{equation*}
\left(p \lambda_{1}+\mu_{1}+N \mu_{2}\right) p_{N, m}=p \lambda_{1} p_{N, m-1}+\left(\mu_{1}+N \mu_{2}\right) p_{N, m+1}, n=N, m \geq 1 \tag{2.3.46}
\end{equation*}
$$

$$
\begin{equation*}
p_{n, m}=0, m \geq 1, n=0, \ldots, N-1 \tag{2.3.47}
\end{equation*}
$$

Solving (2.3.43) and (2.3.44) yields $p_{n, 0}=\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{n} p_{0,0}$ for $n=0, \ldots, N$. Inserting

$$
p_{N-1,0}=\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N-1} p_{0,0} \quad \text { and } \quad p_{N, 0}=\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N} p_{0,0}
$$

into (2.3.45) gives

$$
p_{N, 1}=\frac{\left(p \lambda_{1}+\mu_{1}\right)\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N}-\lambda_{1}\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N-1}}{\mu_{1}+N \mu_{2}} p_{0,0}=\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N}\left(\frac{p \lambda_{1}}{\mu_{1}+N \mu_{2}}\right) p_{0,0}
$$

This and (2.3.46) lead to

$$
p_{N, m}=\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N}\left(\frac{p \lambda_{1}}{\mu_{1}+N \mu_{2}}\right)^{m} p_{0,0}
$$

for all $m \geq 0$. Thus all non-zero probabilities are expressed in terms of $p_{0,0}$ and we can determine $p_{0,0}$ from the normalization condition

$$
\sum_{n=0}^{N} \sum_{m=0}^{\infty} p_{n, m}=1
$$

Substituting the appropriate terms in the normalization condition and simplifying we get

$$
p_{0,0}=\left(\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N} \frac{\mu_{1}+N \mu_{2}}{\mu_{1}+N \mu_{2}-p \lambda_{1}}+\frac{\mu_{1}-\lambda_{1}\left(\frac{\lambda_{1}}{\mu_{1}}\right)^{N-1}}{\mu_{1}-\lambda_{1}}\right)^{-1}
$$

The necessary and sufficient stability condition is clearly $p \lambda_{1}<\mu_{1}+N \mu_{2}$ (see also Theorem 1.2.3).

It is obvious that we can model the two queues as an ordinary single $M_{(n)} / M_{(n)} / 1$ queue, where the subscript $(n)$ stands for state-dependent service and arrival rates. The state space of this queue is $\mathbb{N}_{0}$, where the states $0, \ldots, N$ correspond to the states $(0,0), \ldots,(N, 0)$ and the states $m \geq N+1$ correspond to the states $(N, m), m \geq 1$. The arrival rate is $\lambda_{1}$ in the states $0, \ldots, N-1$ and $p \lambda_{1}$ in the states $m \geq N$. The service rate is $\mu_{1}$ in the states $1, \ldots, N$ and $\mu_{1}+N \mu_{2}$ in the states $m \geq N+1$. By classical formulas for the steady-state distribution of infinite birth and death chains (see for example Cohen [12]) we get the same result.

## Chapter 3

## Overflow to a finite queue with waiting room

### 3.1 Model overview

As in the previous chapter, we consider an open queueing network consisting of two queues $Q_{1}$ and $Q_{2}$ with an overflow capability from $Q_{1}$ to $Q_{2}$. The main difference between the queueing models presented in the following and those presented in the first part of this thesis is the finiteness of the capacity of the second queue. Moreover, customers in the first queue no longer serve the customers in the second queue and the service and arrival rates are not variable. The finiteness gives rise to additional boundary conditions. These boundary conditions and the special structure of the steady-state equations make it impossible to carry out the approach from the second chapter. Nevertheless, the number of steady-state equations that describe the system's behavior can be reduced substantially. The steady-state probabilities and quantities of interest can be stated in an elegant way that reveals the underlying structure of the solutions. The separation approach used in this chapter is, due to the terms in the steady-state equations in the previous chapter that arise from the serving customers, not applicable to the models presented in the second chapter.

Now we give an overview over the models considered in this chapter and the methods used for deriving the quantities of interest. Each queue is equipped with a finite number of servers and a waiting room with finite capacity. We assume independent Poisson arrivals at each queue and exponential service times with server-dependent parameters. $Q_{i}, i=1,2$, possesses $n_{i} \geq 1$ servers and $q_{i} \geq 0$ waiting positions and is fed by a Poisson
arrival stream with intensity $\lambda_{i}>0$. In contrast to the previous chapter, we use lower case letters for the capacities of the queues in order to simplify the resulting formulas. The service requirement of a customer at $Q_{i}$ is exponentially distributed with mean $1 / \mu_{i}$, where $\mu_{i}>0$ for $i=1,2$. A server serves exactly one customer at a time, in case one is present. Within the queues, customers are served in their order of arrival.

Since there is a limited number of servers and the waiting rooms are finite in each queue, arriving customers may be blocked at one of the queues depending on whether all servers and/or waiting positions are occupied. Moreover, blocked customers may overflow to the other queue if its capacity is exhausted. Since blocked customers may be allowed to overflow to the second queue, it seems appropriate to also allow waiting $Q_{1}$-customers to jockey to the second queue if there is capacity available. We will consider a variety of different models by combining different blocking, overflow and jockeying rules. These rules are explained in the following.

## Blocking rules

We consider two blocking models for arriving $Q_{1}$-customers. In the first model, an arriving $Q_{1}$-customer is blocked if all $n_{1}$ servers are busy. In the second model, blocking takes place if all $n_{1}$ servers and all $q_{1}$ waiting positions are occupied.

A customer who arrives at $Q_{2}$ is served in $Q_{2}$ if less than $n_{2}$ servers are busy in $Q_{2}$ at the time of his arrival. The customer is queued in one of the waiting positions in $Q_{2}$ if all servers are busy and at least one waiting position is available in $Q_{2}$, otherwise the customer is blocked and cleared from the system.

## Overflow rules

Blocked customers from $Q_{1}$ are treated with respect to two different routines, which will be called overflow rules. In the first routine, a blocked customer is served by one of the servers in $Q_{2}$ if at least one is available. If no server is available in $Q_{2}$, then the blocked customer is rejected and queued at $Q_{1}$ if a waiting position is available. The customer leaves the system otherwise. In the second routine, the blocked stream from $Q_{1}$ is directed to the waiting room in $Q_{2}$. If no waiting position is available in $Q_{2}$, then a blocked customer is rejected at $Q_{2}$ and queued in $Q_{1}$ if a waiting position is available. The blocked $Q_{1}$-customers are lost if the waiting rooms are fully occupied in both $Q_{1}$ and $Q_{2}$. Note that the blocked traffic and the overflow traffic from $Q_{1}$
are in general not identical, because blocked customers are allowed to join $Q_{1}$ if they are rejected at $Q_{2}$.

The overflow stream is the fraction of those arriving $Q_{1}$-customers who are blocked at $Q_{1}$ and who can potentially join $Q_{2}$. This overflow stream will additionally be weighted with a parameter $p \in[0,1]$, i.e., an arriving customer who is blocked at $Q_{1}$ and overflows to $Q_{2}$ according to the blocking and overflow rules, joins $Q_{2}$ with probability $p$ and leaves the system with probability $1-p$. This overflow model with parameter $p$ will be called $p$ overflow model or weighted overflow model in the following. The model with $p=1$ will be called deterministic overflow model. Observe, that the queue sizes at $Q_{1}$ and $Q_{2}$ are in general not independent for every choice of the model configurations, not even for $p=0$.

## Jockeying rules

The possibility of waiting customers to move to another queue is called jockeying. Jockeying will be restricted to the customers of the first queue and therefore, $Q_{2}$-customers are not allowed to leave the second queue in order to receive service in the first queue. The treatment of waiting $Q_{1^{-}}$ customers can be classified according to three different cases. In the first case, a waiting $Q_{1}$-customer is placed in service in the second queue if at least one server and all waiting positions are available in $Q_{2}$. In the second case, waiting $Q_{1}$-customers must wait for a server in $Q_{1}$ to become free, i.e., jockeying is not allowed. In the third case, waiting $Q_{1}$ customers are placed in $Q_{2}$ if at least one waiting position or one server is available in $Q_{2}$.

## Basic model notation

In order to distinguish between the different models and rules for blocking, overflow and jockeying, we denote an overflow system with this characteristics by a string

$$
\alpha / \beta / \gamma
$$

similar to Kendell's notation. In our notation, $\alpha$ refers to the blocking rule, $\beta$ to the treatment of overflowing customers and $\gamma$ is the indicator for the jockeying rule. The blocking rule indicator $\alpha$ is either S or W , where S indicates that arriving $Q_{1}$-customers are blocked if all servers in $Q_{1}$ are busy and W indicates that the customers are blocked if all servers and all waiting positions in $Q_{1}$ are occupied. The second indicator $\beta$ is S or W , if the blocked customers overflow to the servers or the waiting rooms, respectively,
in $Q_{2}$. Finally, $\gamma$, the jockeying rule indicator, is S or W if the waiting $Q_{1^{-}}$ customers are allowed to jockey to the servers or to the waiting room in $Q_{2}$, respectively. $\gamma$ is N if no jockeying is allowed. The models $\mathrm{S} / \mathrm{S} / \gamma$ and $\mathrm{W} / \mathrm{S} / \gamma$ for $\gamma=\mathrm{S}, \mathrm{W}, \mathrm{N}$ are depicted in Figure 3.1. A schematic overview of the models $\mathrm{S} / \mathrm{W} / \gamma$ and $\mathrm{W} / \mathrm{W} / \gamma$ for $\gamma=\mathrm{S}, \mathrm{W}, \mathrm{N}$ is shown in Figure 3.2.


Figure 3.1: One-way $p$-overflow, $p \in[0,1]$ : Overflow to the servers in $Q_{2}$ (routine S)

## Solution approach

We are interested in the two-dimensional server and waiting room demand process of the one-way $p$-overflow model, embedded at the time instants of arrivals to $Q_{1}$ and $Q_{2}$ and departures from $Q_{1}$ and $Q_{2}$. This process is a Markov chain with state space $\mathfrak{S}=\left\{0, \ldots, k_{1}\right\} \times\left\{0, \ldots, k_{2}\right\}$, where $k_{i}=n_{i}+q_{i}$ for $i=1,2$ and where we have labeled the servers and waiting rooms in $Q_{i}$ with $1, \ldots, n_{i}$ and $n_{i}+1, \ldots, k_{i}$, respectively, for $i=1,2$.
Remark 3.1.1. For every choice of $\lambda_{1}, \lambda_{2}, \mu_{1}, \mu_{2}>0$ in each model described above, the state space is finite and the fact that there exists a path from any state to $(0,0)$ (and vice versa) having positive probability, yields irreducibility. Thus, the Markov chain is stationary with unique stationary distribution.

In some models it is possible to allow $q_{1}=\infty$. In these cases we give a necessary condition for stability. The stationary distribution is uniquely


Figure 3.2: One-way $p$-overflow, $p \in[0,1]$ : Overflow to the waiting room in $Q_{2}$ (routine W)
determined by the $\left(k_{1}+1\right)\left(k_{2}+1\right)$ balance equations and the normalization condition. To the best of our knowledge, no closed-form expression for the stationary distribution is known. By using a technique due to Morrison [46,48], we will reduce the problem of solving these $\left(k_{1}+1\right)\left(k_{2}+1\right)+1$ equations to the problem of solving a substantially smaller number of homogeneous linear equations in two sets of unknowns. With this approach, explicit formulas depending on these unknowns can be given for various steady-state quantities. Morrison considers the deterministic models $\mathrm{S} / \mathrm{S} / \mathrm{S}$, $\mathrm{S} / \mathrm{S} / \mathrm{N}, \mathrm{W} / \mathrm{S} / \mathrm{N}$ and $\mathrm{W} / \mathrm{W} / \mathrm{N}$ with $p=1$ and service rates $\mu_{1}=\mu_{2}$ and reduces the number of equations and unknowns by a separation method. The basic technique is to partition the state space into certain regions and boundaries and to separate the stationary probabilities within these regions. In every model, the separation leads to a set of eigenvalue problems for the separation constants. The eigenvalues are given by the roots of polynomial equations and are the pairwise distinct eigenvalues of real tridiagonal symmetric matrices as well. They possess an interlacing property, called the "Sturm sequence property", which reduces the computational complexity considerably. The desired probabilities are expressed as sums of eigenfunctions in terms of the eigenvalues. The number of eigenfunctions and therefore the number of coefficients to be determined in these representations is in general substantially smaller than the number of stationary probabilities.

The coefficients are determined by the normalization condition and a set of linear equations that stems from the boundary conditions. The desired probabilities and steady-state quantities can be numerically determined once the coefficients and eigenvalues are numerically calculated.

We will extend the results of Morrison in three directions. Firstly, we allow different service rates at the two queues, i.e., customers at $Q_{i}$ have service rate $\mu_{i}, i=1,2$, with arbitrary $\mu_{1}, \mu_{2}>0$. Secondly, the overflow process is weighted with a parameter $p \in[0,1]$. In one model it is also possible to weight the stream of jockeying customers with the same parameter $p$. Thirdly, we consider several variations of the blocking and overflow routines as described above. It should be mentioned that the resulting formulas do not scale in the system parameters in general, i.e., for example the case of arbitrary service rates cannot be reduced to the case $\mu_{1}=\mu_{2}=1$ by an adequate choice of the remaining parameters and the separation constants. Only the case $\mu_{1}=\mu_{2}$ can be reduced to $\mu_{1}=\mu_{2}=1$ by dividing the arrival rates and the separation constants by the service rate.

## Stationary quantities

Once the stationary probabilities are specified in terms of the eigenvalues, it is possible to derive formulas for various steady-state quantities depending on the unknowns, like the loss probabilities, i.e., the probability that an arriving $Q_{i}$-customer is lost, $i=1,2$, the overflow and jockeying probability, the probability that an arriving customer is queued upon arrival, the mean departure rate from the waiting rooms to the servers, the average number of customers waiting and in service, the average waiting times and many more. These formulas have the advantage that the desired stationary quantities can be calculated directly in terms of the unknowns without computing the steady-state probabilities. We will display this possibility for the two basic models described in Section 3.2 and Section 3.3.1. It is immediately seen from Figure 3.1 that $Q_{1}$ in isolation is an Erlang loss system in the models without waiting room in $Q_{1}$, i.e., $q_{1}=0$. In these cases, the overflow stream from $Q_{1}$ to $Q_{2}$ consists of the $p$-fraction of blocked customers at $Q_{1}$. Let $O_{12}$ be the expected (stationary) number of demands per unit time, which flow over from $Q_{1}$ to $Q_{2}$ (see Figure 3.1). With $\rho_{1}=\lambda_{1} / \mu_{1}$, we must have

$$
O_{12}=p \lambda_{1} \cdot \frac{\rho_{1}^{n_{1}}}{n_{1}!\sum_{i=0}^{n_{1}} \frac{\rho_{1}^{i}}{i!}} .
$$

This is of course the loss probability for the famous Erlang loss system with capacity $n_{1}$ (see Erlang [24] and (2.2.1)). In general, the loss probabilities for $Q_{1}$ and $Q_{2}$, the blocking probabilities for customers at $Q_{2}$ and therefore the expected number of demands per unit time that flow over from $Q_{1}$ to $Q_{2}$ are unknown.

It is known from Cohen [11], Hordijk and Ridder (30] and van Marion [44] that the loss probabilities are in general sensitive with respect to the servicetime distribution. In the literature, probabilistic and numerical bounds for the loss probabilities are known in some special cases. Van Dijk [15] derives upper and lower bounds for the loss probabilities by approximating the one-way overflow model by models with modified input and interconnection characteristics. Hordijk and Ridder [30, 31] and Ridder [57, 58] derive upper and lower bounds for weighted stationary probabilities by constructing approximating reversible Markov chains. These bounds are insensitive with respect to the service time distributions and are used to derive upper and lower bounds for the loss probabilities.

## Applications

The most apparent application of this type of overflow queueing models is of course to telecommunication systems. As an example for a feasible application, the models presented in this chapter might be used to model call centers. In a specific call center telecommunication system, incoming telephone calls are answered by a limited number of operators (i.e., the servers). An incoming call is put on hold if all operators are busy. The number of calls on hold are in most practical situations finite and the on hold positions correspond to the waiting rooms in our queueing systems. Consider two incoming lines for calls. The first and second line are the queues $Q_{1}$ and $Q_{2}$, respectively. It may be permitted that customers from the first line have a certain type of priority, i.e., they may be allowed to overflow to the second line under one of the overflow rules specified earlier. The different rules for blocking, overflow and jockeying can be used to represent a certain constellation given in practice. Furthermore, the parameter $p$ can be viewed as a control parameter or can be used to model the impatience rate of the overflowing calls. In another considerable application, one could assign costs or a premium to the parameter $p$ that has to be paid by the priority customers from the first queue in order to receive service at the second queue. An increase of $p$ results in a higher probability of acceptance at the second queue and a higher premium should be paid in order to receive a higher value for $p$. Thus, a customer or a system designer has to find a
cost optimal value for $p$ depending on his point of view.
Another application that can be slightly generalized to match our models is described in Altman et al. [2] and Hassin [28]. Consider two gas stations located at the same main road. Every gas station has a finite number of gas pumps and waiting positions. The two gas stations are located one after another, where the first and second station correspond to $Q_{1}$ and $Q_{2}$, respectively. Consider that some customers prefer the first and some customers prefer the second station. Additionally, if we consider one-way traffic, which is for example the case on an highway, then we get an arrival stream for each station that is independent from the other one. In the one-way setting, customers that find the first station occupied, can drive by in order to be served at the second station. Some of these overflowing customers might also reject service at the second station and drive by to another station they prefer. In this case, the overflow parameter $p$ reflects this behavior. It should be noted, that some of the blocking, overflow and jockeying rules are unnatural in this context. Nevertheless, due to the diversity of the models presented in the following, many applications for example to communication, computer or traffic control systems are conceivable.

## Further related literature

Some of the results developed in the following sections where published by the author in [59]. In the literature, numerical computations for some of the deterministic overflow models, i.e., with $p=1$, are available. Kaufman 35] (model W/W/S and W/W/N) uses block iterative techniques and successive overrelaxation techniques for the numerical derivation of the steady-state probabilities. Chan [8, 9] (model W/W/N and W/W/S) uses block iterative techniques and preconditioned conjugate gradient methods for the numerical computation of the steady-state probabilities. The deterministic model with arbitrary service rates is treated in [9] in the context of overflow queueing networks with an arbitrary but finite number of queues. For some models considered in Chan [8], Kaufman [35] and Morrison [46], i.e., S/S/S, S/S/N and $\mathrm{W} / \mathrm{W} / \mathrm{N}$, numerical results are presented in Kaufman et al. 34. The overflow model with deterministic overflow and without waiting rooms is known as the one-way overflow queueing model (van Dijk [15, 16], Doremalen [18], Hordijk and Ridder [30]). In telecommunication theory, it is a special type of an asymmetric grading (Kosten [36], van Marion 44] and Syski [63]).

Overflow queueing models are widespread in literature. We already mentioned Disney and König [17] for a broad overview and Koury et al. [39]
and Krieger et al. [40] for reviews of iterative numerical methods for overflow queueing models. A brief discussion of numerical methods for some two-queue overflow systems and further references are given in Ching and Ng [10]. Related overflow models are studied in van Doorn [19], El-Taha and Heath [20], Parthasarathy and Sudhesh [55] and the referenced literature therein using a variety of different techniques (for some comments on [19] and [55] see page [13). An overflow model with multiple primary queues, finite waiting rooms and a shared secondary overflow queue is numerically studied in Guérin and Lien [27] and related models are reviewed. Early references are given in van Marion [44, Morrison (46] and Morisson [47]. Further related literature is mentioned in Chapter 2. Section [2.1] of this thesis.

### 3.2 Overflow without waiting rooms

### 3.2.1 Steady-state equations and separation approach

In this section, we consider the basic $p$-overflow model without waiting rooms, that is, the capacity $q_{i}$ of the waiting room in $Q_{i}$ is 0 and the total capacity of $Q_{i}$ is $k_{i}=n_{i}$ for $i=1,2$. In this case, all model variations coincide. This basic model and the derivation of the steady-state probabilities will be the starting point for the derivations in the more sophisticated models. Their analysis is a generalization of this basic approach.

Let $Q_{i}, i=1,2$, be fed by a Poisson arrival stream with intensity $\lambda_{i}>0$ and let the service times at $Q_{i}$ be independently identically and exponentially distributed with mean $1 / \mu_{i}>0$. The arrival stream from $Q_{1}$ is blocked and directed to $Q_{2}$ if all servers in the first queue are busy. Overflowing $Q_{1}$ customers and arriving $Q_{2}$-customers are lost if all servers in $Q_{2}$ are busy. The bivariate server and waiting room demand distribution is the unique nonnegative and normalized solution of the following equilibrium equations:

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i n_{1}}\right)+p \lambda_{1} \delta_{i n_{1}}\left(1-\delta_{j n_{2}}\right)+\lambda_{2}\left(1-\delta_{j n_{2}}\right)+i \mu_{1}+j \mu_{2}\right) p_{i, j} \\
& \quad=\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+\left(1-\delta_{i n_{1}}\right)(i+1) \mu_{1} p_{i+1, j}  \tag{3.2.1}\\
& \quad+\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \delta_{i n_{1}}+\lambda_{2}\right) p_{i, j-1}+\left(1-\delta_{j n_{2}}\right)(j+1) \mu_{2} p_{i, j+1}
\end{align*}
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$, where $\delta_{i j}$ is the Kronecker symbol, i.e., $\delta_{i j}=1$ if $i=j$ and $\delta_{i j}=0$ otherwise. Equation (3.2.1) for $i \neq n_{1}$ characterizes the flow into and out of states which is not due to an overflow of customers. In contrast to that, equation (3.2.1) for $i=n_{1}$ characterizes the flow which is caused by the overflow mechanism.

For $i \neq n_{1}$ the equations (3.2.1) are given by

$$
\begin{align*}
\left(\lambda_{1}+\lambda_{2}\left(1-\delta_{j n_{2}}\right)+i \mu_{1}\right. & \left.+j \mu_{2}\right) p_{i, j}=\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+(i+1) \mu_{1} p_{i+1, j} \\
& +\left(1-\delta_{j 0}\right) \lambda_{2} p_{i, j-1}+\left(1-\delta_{j n_{2}}\right)(j+1) \mu_{2} p_{i, j+1} \tag{3.2.2}
\end{align*}
$$

for $i=0, \ldots, n_{1}-1$ and $j=0, \ldots, n_{2}$. From these equation one might predict that the queue lengths of $Q_{1}$ and $Q_{2}$ behave independent from each other given that $Q_{1}$ is not fully occupied. In fact, it turns out that the variables in (3.2.2) can be separated, in the sense that there are solutions of the form $p_{i, j}=\alpha_{i} \beta_{j}$ for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$ as we will show in the following.

Setting $p_{i, j}=\alpha_{i} \beta_{j}$ for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$ in (3.2.2) and evaluating the result in $j=0$ results in the equations

$$
\begin{equation*}
\left(\lambda_{1}+i \mu_{1}+c\right) \alpha_{i}=\lambda_{1}\left(1-\delta_{i 0}\right) \alpha_{i-1}+(i+1) \mu_{1} \alpha_{i+1} \tag{3.2.3}
\end{equation*}
$$

for $i=0, \ldots, n_{1}-1$ with the separation constant $c=\lambda_{2}-\mu_{2} \frac{\beta_{1}}{\beta_{0}}$. On the other hand, the evaluation of the result in $i=0$ leads to

$$
\begin{equation*}
\left(\lambda_{2}\left(1-\delta_{j n_{2}}\right)+j \mu_{2}-c\right) \beta_{j}=\lambda_{2}\left(1-\delta_{j 0}\right) \beta_{j-1}+\left(1-\delta_{j n_{2}}\right)(j+1) \mu_{2} \beta_{j+1} \tag{3.2.4}
\end{equation*}
$$

for $j=0, \ldots, n_{2}$ with the separation constant $c=-\lambda_{1}+\mu_{1} \frac{\alpha_{1}}{\alpha_{0}}$. Note that

$$
\begin{equation*}
c=\lambda_{2}-\mu_{2} \frac{\beta_{1}}{\beta_{0}}=-\lambda_{1}+\mu_{1} \frac{\alpha_{1}}{\alpha_{0}} \tag{3.2.5}
\end{equation*}
$$

which follows from (3.2.2) with $i=j=0$. By comparing (3.2.3) and (3.2.4) we get the following Lemma.

Lemma 3.2.1. If $\alpha_{i}=s_{i}\left(c, \lambda_{1}, \mu_{1}\right), i=0, \ldots, n_{1}$, is a solution of (3.2.3) for $i=0, \ldots, n_{1}-1$, then a solution of (3.2.4) for $j=0, \ldots, n_{2}-1$ is given by $\beta_{j}=s_{j}\left(-c, \lambda_{2}, \mu_{2}\right), j=0, \ldots, n_{2}$.

By this lemma it is only necessary to find a solution of the equations (3.2.3) for $\lambda_{1}, \mu_{1}>0$ and $c \in \mathbb{R}$. Their solution instantaneously yields a solution of the equations (3.2.4). In fact, the solution can be given in closed form in terms of the system parameters as a polynomial in the separation variable $c$. We will prove this fact and state the solution and some useful properties of these polynomials in the next section.

### 3.2.2 Solution for the separation variables

We derive the solution and some useful properties of the solutions of the basic recurrence equation (3.2.3) in this section. Let $c \in \mathbb{R}$ and $\lambda, \mu>0$ be arbitrary and define $\alpha_{i}=s_{i}(c)=s_{i}(c, \lambda, \mu), i \geq 0$, recursively by $s_{0}(c)=1$ and (3.2.3), i.e.,

$$
\begin{equation*}
(\lambda+i \mu+c) s_{i}(c)=\lambda\left(1-\delta_{i 0}\right) s_{i-1}(c)+(i+1) \mu s_{i+1}(c), \quad i \geq 0 \tag{3.2.6}
\end{equation*}
$$

Multiplying this recurrence relation by $z^{i}, z \in \mathbb{R}$, and summing over $i \geq 0$ yields the differential equation

$$
\begin{equation*}
\mu(1-z) f_{c}^{\prime}(z)=(\lambda(1-z)+c) f_{c}(z) \tag{3.2.7}
\end{equation*}
$$

where $f_{c}$ is the generating function of the sequence $\left(s_{i}(c, \lambda, \mu)\right)_{i \geq 0}$, that is,

$$
\begin{equation*}
f_{c}(z)=\sum_{i=0}^{\infty} s_{i}(c, \lambda, \mu) z^{i} \tag{3.2.8}
\end{equation*}
$$

for $z \in \mathbb{R}$ with $|z|<1$. The condition $s_{0}(c, \lambda, \mu)=1$ yields the boundary condition $f_{c}(0)=1$ for the differential equation (3.2.7). The solution of (3.2.7) with $f_{c}(0)=1$ is given by

$$
\begin{equation*}
f_{c}(z)=e^{\frac{\lambda}{\mu} z}(1-z)^{-\frac{c}{\mu}} . \tag{3.2.9}
\end{equation*}
$$

Using the power series expansions

$$
\begin{gathered}
e^{a z}=\sum_{n=0}^{\infty} \frac{a^{n}}{n!} z^{n} \quad \text { and } \\
(1-z)^{b}=\sum_{n=0}^{\infty}\binom{b}{n}(-1)^{n} z^{n},
\end{gathered}
$$

where $\binom{b}{n}=\frac{b(b-1) \cdot \ldots \cdot(b-n+1)}{n!}$ for $b \in \mathbb{R}$ is the generalized binomial coefficient, we get a closed-form expression for the coefficients $s_{i}(c, \lambda, \mu), i \geq 0$, by the identity theorem for power series:

Proposition 3.2.2. Let $c \in \mathbb{R}, \lambda>0$ and $\mu>0$ be arbitrary, then the solution of (3.2.6) with $s_{0}(c, \lambda, \mu)=1$ is given by

$$
\begin{equation*}
s_{i}(c, \lambda, \mu)=\frac{1}{i!}\left(\frac{\lambda}{\mu}\right)^{i} \sum_{k=0}^{i}\binom{i}{k} \lambda^{-k} \prod_{j=0}^{k-1}(j \mu+c) \tag{3.2.10}
\end{equation*}
$$

$$
\begin{equation*}
=\left(\frac{\lambda}{\mu}\right)^{i} \sum_{k=0}^{i} \frac{\lambda^{-k}}{(i-k)!k!}\left(\frac{c}{\mu}\right)_{k} \tag{3.2.11}
\end{equation*}
$$

for $i \geq 1$, where $(\alpha)_{0}=1$ and $(\alpha)_{k}=\prod_{j=0}^{k-1}(j+\alpha), k \geq 1$, is the Pochhammer symbol.

Note that we have assumed that the sequence $\left(s_{i}(c, \lambda, \mu)\right)_{i \geq 0}$ is indeed absolutely summable. This is immediately clear if we define $f_{c}$ for $|z|<1$ by (3.2.9) and develop $f_{c}$ into the power series (3.2.8). The functions $s_{i}(c, \lambda, \mu)$, $i \geq 0$, are known as the Poisson-Charlier polynomials in a more general setting (see Abramowitz and Stegun [1]) and are connected to a family of special functions, the confluent hypergeometric functions of the second kind. For $a, b, z \in \mathbb{C}$ with $b \neq 0,-1,-2, \ldots$ and $|\arg (z)|<\pi$ the confluent hypergeometric function of the second kind $U(a, b, z)$ is a solution of the Kummer differential equation

$$
z y^{\prime \prime}+(b-z) y^{\prime}-a y=0
$$

with boundary conditions $y(a, b, 0)=1$ and $\left.\frac{\partial}{\partial z} y(a, b, z)\right|_{z=0}=a / b$. For $\operatorname{Re}(a)>0$ and $\operatorname{Re}(z)>0$, it has the integral representation

$$
U(a, b, z)=\frac{1}{\Gamma(a)} \int_{0}^{\infty} e^{-z t} t^{a-1}(1+t)^{b-a-1} d t
$$

(see Abramowitz and Stegun [1], Chapter 13 and Srivastava and Kashyap [60], Chapter II.3). The following proposition shows the connection between the functions $s_{i}(c, \lambda, \mu), i \geq 0$, the confluent hypergeometric function and the signless Stirling numbers of the first kind.

Proposition 3.2.3. Let $c \in \mathbb{R}, \lambda>0$ and $\mu>0$ be arbitrary and $s_{i}(c, \lambda, \mu)$, $i \geq 0$, be defined by (3.2.10).
a) For every $c \notin\left\{\mu(n+1-i) \mid n \in \mathbb{N}_{0}\right\}$ it holds

$$
s_{i}(c, \lambda, \mu)=\frac{1}{i!} U\left(-i, 1-i-\frac{c}{\mu}, \frac{\lambda}{\mu}\right)
$$

b) For $0 \leq n \leq k$ let $\sigma(k, n)$ equal the number of permutations of $k$ elements which contain exactly $n$ permutation cycles. Then

$$
\begin{equation*}
s_{i}(c, \lambda, \mu)=\frac{1}{i!} \sum_{n=0}^{i}\left(\frac{c}{\mu}\right)^{n} \sum_{k=n}^{i}\binom{i}{k}\left(\frac{\lambda}{\mu}\right)^{i-k} \sigma(k, n) \tag{3.2.12}
\end{equation*}
$$

holds. The numbers $\sigma(k, n), 0 \leq n \leq k$, are called the signless Stirling number of the first kind.

Proof. For (a) see Abramowitz and Stegun [1], Chapter 13. To show (b) we use (3.2.10) and $\prod_{j=0}^{k-1}(j-b)=\sum_{n=0}^{k} \sigma(k, n) b^{n}, b \in \mathbb{R}$, (see Abramowitz and Stegun [1], Chapter 24.1.3)

The next Proposition states some properties of the functions $s_{i}(c, \lambda, \mu)$ for $i \geq 0$. These properties will be useful in the sequel.

Proposition 3.2.4. Let $\lambda, \mu>0$ and $c \in \mathbb{R}$. The solution $s_{i}(c, \lambda, \mu), i \geq 0$, of the equations (3.2.6) satisfies the recurrence relations

$$
\begin{gather*}
\mu(i+1) s_{i+1}(c)=\lambda s_{i}(c)+c s_{i}(c+\mu) \quad \text { and }  \tag{3.2.13}\\
s_{i}(c)=s_{i}(c+\mu)-s_{i-1}(c+\mu)\left(1-\delta_{i 0}\right) \tag{3.2.14}
\end{gather*}
$$

for $i \geq 0$. Moreover, for every $n \geq 0$, the equations

$$
\begin{gather*}
\sum_{i=0}^{n} s_{i}(c)=s_{n}(c+\mu),  \tag{3.2.15}\\
\sum_{i=0}^{n}(n-i) s_{i}(c)=\left(1-\delta_{n 0}\right) s_{n-1}(c+2 \mu) \quad \text { and }  \tag{3.2.16}\\
\sum_{i=1}^{n} i s_{i}(c)=n s_{n}(c+\mu)-\left(1-\delta_{n 0}\right) s_{n-1}(c+2 \mu) \tag{3.2.17}
\end{gather*}
$$

hold.

Proof. Rewriting the differential equation (3.2.7) using (3.2.9) yields

$$
f_{c}^{\prime}(z)=\frac{\lambda}{\mu} f_{c}(z)+\frac{c}{\mu} f_{c+\mu}(z) .
$$

The power series expansions of $f_{c}$ and $f_{c+\mu}$ imply

$$
\sum_{i=1}^{\infty} i s_{i}(c) z^{i-1}=\frac{\lambda}{\mu} \sum_{i=0}^{\infty} s_{i}(c) z^{i}+\frac{c}{\mu} \sum_{i=0}^{\infty} s_{i}(c+\mu) z^{i}
$$

for all $|z|<1$. Therefore, by multiplying this equation with $\mu$ and using the identity theorem for power series, we conclude that

$$
\mu(i+1) s_{i+1}(c)=\lambda s_{i}(c)+c s_{i}(c+\mu), \quad i \geq 0,
$$

i.e., that (3.2.13) holds. In the same manner, (3.2.14) follows from

$$
f_{c}(z)=(1-z) f_{c+\mu}(z)
$$

Summing (3.2.14) over $i=0, \ldots, n$ gives immediately (3.2.15). (3.2.16) follows easily from (3.2.13) by summing over $i=0, \ldots, n$ and using (3.2.15). Finally, (3.2.17) follows from (3.2.15) together with (3.2.16).

### 3.2.3 Boundary condition and separation constants

We have shown in the previous section in Lemma 3.2.1 and Proposition 3.2.2 that a solution of (3.2.3) for $i=0, \ldots, n_{1}$ and (3.2.4) for $j=0, \ldots, n_{2}-1$ with $\alpha_{0}=\beta_{0}=1$ is given by $\alpha_{i}=s_{i}\left(c, \lambda_{1}, \mu_{1}\right)$ for $i=0, \ldots, n_{1}$ and $\beta_{j}=s_{j}\left(-c, \lambda_{2}, \mu_{2}\right)$ for $j=0, \ldots, n_{2}$. Consequently, a natural candidate for the solution of the steady-state equations (3.2.1) is

$$
p_{i, j}=\alpha_{i} \beta_{j}=s_{i}\left(c, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c, \lambda_{2}, \mu_{2}\right)
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$.
It still remains to determine the separation constant $c$, which can in view of (3.2.5) be reformulated as

$$
c=\lambda_{2}-\mu_{2} s_{1}\left(-c, \lambda_{2}, \mu_{2}\right)=-\lambda_{1}+\mu_{1} s_{1}\left(c, \lambda_{1}, \mu_{1}\right)
$$

This equation is a tautology as one could expect and therefore gives no additional information about $c$. In the previous sections, all steady-state equations despite the equation for $j=n_{2}$ have been utilized to find $\alpha_{i}$ and $\beta_{j}$. By taking the boundary equation for $j=n_{2}$ into account, we will derive the crucial condition in the following lemma that will be used to determine the separation constant $c$ - or more precisely as we will see - the feasible separation constants $c_{1}, \ldots, c_{n_{2}}$.

Lemma 3.2.5. Let $\beta_{j}=s_{j}\left(-c, \lambda_{2}, \mu_{2}\right)$ for $j \geq 0$ and $\lambda_{2}, \mu_{2}>0$ and $c \in \mathbb{R}$ be given by (3.2.10). Then the equation (3.2.4) for $j=n_{2}$ is equivalent to

$$
\begin{equation*}
c s_{n_{2}}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)=0 \tag{3.2.18}
\end{equation*}
$$

Proof. Setting $\beta_{j}=s_{j}\left(-c, \lambda_{2}, \mu_{2}\right)$ for $j=n_{2}-1$ and $j=n_{2}$ in (3.2.4), $\lambda=\lambda_{2}, \mu=\mu_{2}$ and substituting $c$ with $-c$ immediately yields

$$
\begin{equation*}
\left(n_{2} \mu_{2}-c\right) s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right)=\lambda_{2} s_{n_{2}-1}\left(-c, \lambda_{2}, \mu_{2}\right) \tag{3.2.19}
\end{equation*}
$$

Additionally, setting $i=n_{2}$ in (3.2.13) gives

$$
n_{2} \mu_{2} s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right)=\lambda_{2} s_{n_{2}-1}\left(-c, \lambda_{2}, \mu_{2}\right)-c s_{n_{2}-1}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)
$$

Substituting $n_{2} \mu_{2} s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right)$ from the latter equation on the right side of equation (3.2.19) yields

$$
c\left(s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right)+s_{n_{2}-1}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)\right)=0
$$

Together with (3.2.14) this finally shows (3.2.18). The argument can simply be reversed to show the other direction of the equivalence.

Equation (3.2.18) provides a condition from which the feasible candidates for the separation constant $c$ can be derived. In fact, it is immediately seen from Proposition 3.2.2 that $s_{n_{2}}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)$ is a polynomial of degree $n_{2}$ in $c$ for every $n_{2} \geq 1$. Furthermore, it follows from Proposition 3.2.2 that $s_{n_{2}}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)$ has $n_{2}$ negative and pairwise distinct roots $c_{1}, \ldots, c_{n_{2}}$ for every $n_{2} \geq 1$, as we will show in the following. For the moment we will suppress the indexes of the system parameters $\lambda_{i}, \mu_{i}$ and $n_{i}$ for $i=1,2$ and substitute $\mu_{2}-c$ with $c$.

As before let $s_{i}(c, \lambda, \mu)$ for $i \geq 0$ be given by $s_{0}(c, \lambda, \mu)=1$ and the recurrence relations (3.2.6). Now we investigate the connection between the zeros of $s_{i}(c, \lambda, \mu)$ and those of $s_{i+1}(c, \lambda, \mu)$ as functions of $c$ for $i \geq 1$. We will show that these two sets of zeros satisfy an interlacing property similar to the one observed in the previous chapter in (2.2.32) and the proof of Theorem 1.2.3. Although the two sets of zeros satisfy a similar interlacing property, the proof of this property has to be carried out in a different way. This can be done by using the so called Sturm sequence property or roots separation theorem. For the sake of completeness, we will state the theorem with a short proof (see Theorem 8.4.1 in Golub and Loan [26] and Chapter $5, \S 37$ in Wilkinson (65]).

Theorem 3.2.6 (Sturm sequence property). For every $n \geq 1$ let

$$
A_{n}=\left(a_{i, j}\right)_{i, j=1, \ldots, n}
$$

be an $n \times n$ symmetric tridiagonal matrix with real entries. Further suppose that $A_{n}$ is unreduced, i.e., $A_{n}$ has nonnegative elements on the secondary diagonals. Then $A_{n}$ has $n$ pairwise distinct real eigenvalues. Let the eigenvalues

$$
\lambda_{1}\left(A_{n}\right), \ldots, \lambda_{n}\left(A_{n}\right)
$$

of $A_{n}$ be ordered in increasing order, then the eigenvalues of $A_{n}$ strictly separate the eigenvalues of $A_{n+1}$ :
$\lambda_{1}\left(A_{n+1}\right)<\lambda_{1}\left(A_{n}\right)<\lambda_{2}\left(A_{n+1}\right)<\ldots<\lambda_{n}\left(A_{n+1}\right)<\lambda_{n}\left(A_{n}\right)<\lambda_{n+1}\left(A_{n+1}\right)$.

This relation is called the strict interlacing property.

Proof. By Chapter 2, $\S 47$ in Wilkinson [65], the eigenvalues of $A_{n}$ separate the eigenvalues of $A_{n+1}$ at least in the weak sense, meaning
$\lambda_{1}\left(A_{n+1}\right) \leq \lambda_{1}\left(A_{n}\right) \leq \lambda_{2}\left(A_{n+1}\right) \leq \ldots \leq \lambda_{n}\left(A_{n+1}\right) \leq \lambda_{n}\left(A_{n}\right) \leq \lambda_{n+1}\left(A_{n+1}\right)$.
Let $\chi_{n}$ be the characteristic polynomial of $A_{n}$, i.e.,

$$
\chi_{n}(t)=\operatorname{det}\left(t E_{n}-A_{n}\right)
$$

for $t \in \mathbb{R}$ and set $b_{n}=a_{n, n+1}=a_{n+1, n}$ for $n \geq 1$. By Laplace expansion of $\chi_{n}$ we get the recursive equations

$$
\begin{equation*}
\chi_{n}(t)=\left(t-a_{n, n}\right) \chi_{n-1}(t)-b_{n-1}^{2} \chi_{n-2}(t) \tag{3.2.22}
\end{equation*}
$$

for $n \geq 2$, where we have set $\chi_{0}(t)=1$. Suppose that $\chi_{n}\left(t_{0}\right)=\chi_{n-1}\left(t_{0}\right)=0$ for some $t_{0} \in \mathbb{R}$ and $n \geq 2$. Then it follows from (3.2.21) and the assumption that $A_{k}$ is unreduced, i.e., $b_{k} \neq 0$ for all $k \geq 1$, by induction that

$$
\chi_{0}\left(t_{0}\right)=\chi_{1}\left(t_{0}\right)=\ldots=\chi_{n}\left(t_{0}\right)=0
$$

This is a contradiction to $\chi_{0}\left(t_{0}\right)=1$ and thus we must have strict inequalities in (3.2.21).

By the roots separation theorem, the goal is to describe the recurrence relations (3.2.6) by a symmetric tridiagonal matrix. Therefore, we start with symmetrizing these relations.

Lemma 3.2.7. Set

$$
\begin{equation*}
u_{i}(c)=u_{i}(c, \lambda, \mu)=\left(\frac{i!\mu^{i+1}}{\lambda^{i}}\right)^{\frac{1}{2}} s_{i}(c, \lambda, \mu) \tag{3.2.23}
\end{equation*}
$$

for $i \geq 0$. Then $s_{0}(c, \lambda, \mu)=1$ and the equations (3.2.6) are equivalent to
$u_{0}(c)=\mu^{\frac{1}{2}}$ and the symmetrized equations

$$
\begin{equation*}
\left(1-\delta_{i 0}\right) i^{\frac{1}{2}} u_{i-1}(c)-\left(i+\frac{\lambda+c}{\mu}\right)\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} u_{i}(c)+(i+1)^{\frac{1}{2}} u_{i+1}(c)=0 \tag{3.2.24}
\end{equation*}
$$

for $i \geq 0$.

In order to write the symmetrized recurrence relations (3.2.24) from the previous Lemma in matrix form, we introduce the matrices

$$
M_{n}=M_{n}(\lambda, \mu)=\left(m_{i, j}\right)_{i, j=1, \ldots, n}
$$

for $n \geq 1$ defined by

$$
\begin{array}{rlr}
m_{i, i} & =-(i-1)\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} & \text { for } i=1, \ldots, n \\
m_{i, i+1}=m_{i+1, i} & =i^{\frac{1}{2}} & \text { for } i=1, \ldots, n-1 \text { and }  \tag{3.2.25}\\
m_{i, j} & =0 & \text { for }|i-j|>1
\end{array}
$$

$M_{n}$ is given in matrix form by

$$
\left(\begin{array}{cccccc}
0 & 1 & 0 & \cdots & \cdots & 0 \\
1 & -\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} & \sqrt{2} & \ddots & & \vdots \\
0 & \sqrt{2} & -2\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} & \ddots & \ddots & \vdots \\
\vdots & \ddots & \ddots & \ddots & \ddots & 0 \\
\vdots & & \ddots & \ddots & -(n-2)\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} & \sqrt{n-1} \\
0 & \cdots & \cdots & 0 & \sqrt{n-1} & -(n-1)\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}}
\end{array}\right)
$$

for $n \geq 1$. $M_{n}$ is a real symmetric matrix. Thus $M_{n}$ is diagonalizable and possesses only real eigenvalues. Furthermore, $M_{n}$ is tridiagonal, that is $m_{i, j}=0$ for $|i-j|>1$, and has nonzero elements on the secondary diagonals, i.e., $m_{i, i+1}, m_{i+1, i} \neq 0$ for $i=1, \ldots, n-1$. Then it follows from Theorem 3.2.6 that $M_{n}$ has $n$ pairwise distinct eigenvalues.

Let $u(c)^{\top}=\left(u_{0}(c), \ldots, u_{n-1}(c)\right)^{\top} \in \mathbb{R}^{n}$ be given by (3.2.23) and consider the following system of linear equations:

$$
\begin{equation*}
\left(M_{n}-\frac{\lambda+c}{\mu}\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} E_{n}\right) u(c)=0 \tag{3.2.26}
\end{equation*}
$$

where $E_{n}$ is the identity matrix in $\operatorname{Mat}(n, n, \mathbb{R})$. From (3.2.24) and (3.2.26) we immediately obtain the following result.

Lemma 3.2.8. Let $u(c)^{\top}=\left(u_{0}(c), \ldots, u_{n-1}(c)\right)^{\top} \in \mathbb{R}^{n}$ and $u_{n}(c)=0$. Then the following statements are equivalent:
(i) $u(c)$ is a nontrivial solution of (3.2.26).
(ii) $u_{0}(c), \ldots, u_{n}(c)$ is a solution of (3.2.24) for $i=0, \ldots, n-1$ with $u_{0} \neq$ 0 .

From the roots separation theorem and Lemma 3.2.8 we get the desired statement about the zeros of $s_{n}(c, \lambda, \mu)$ :

Theorem 3.2.9. For $n \geq 1, c \in \mathbb{R}, \lambda>0$ and $\mu>0$ let $s_{n}(c, \lambda, \mu)$ be defined by (3.2.10). Then the equation

$$
s_{n}(c, \lambda, \mu)=0
$$

has $n$ negative and distinct solutions $c_{1}, \ldots, c_{n}$ in the variable $c$ given by

$$
\begin{equation*}
c_{i}=(\lambda \mu)^{\frac{1}{2}} \varepsilon_{i}-\lambda \tag{3.2.27}
\end{equation*}
$$

for $i=1, \ldots, n$, where $\varepsilon_{1}, \ldots, \varepsilon_{n}$ are the pairwise distinct eigenvalues of the matrix $M_{n}(\lambda, \mu)$.

Proof. Let $n \geq 1$. The matrix $M_{n}$ satisfies the assumptions of the roots reparation theorem and therefore has $n$ real distinct eigenvalues $\varepsilon_{1}, \ldots, \varepsilon_{n}$. Thus, the equation

$$
\left(M_{n}-\frac{\lambda+c}{\mu}\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} E_{n}\right) u=0
$$

with $c \in \mathbb{R}$ and $u^{\top}=\left(u_{0}, \ldots, u_{n-1}\right)^{\top} \in \mathbb{R}^{n}$ has $n$ real and linearly independent solutions. The solutions are of the form $\left(c_{j}, u^{j}\right)$ for $j=1, \ldots, n$, where

$$
c_{j}=(\lambda \mu)^{\frac{1}{2}} \varepsilon_{j}-\lambda
$$

and $u^{j}=\left(u_{0}^{j}, \ldots, u_{n-1}^{j}\right)$ is an eigenvector of $M_{n}$ to the eigenvalue $\varepsilon_{j}$ for $j=1, \ldots, n$. We define $u_{n}^{j}=0$. By Remark 3.2 .8 (i) $u_{0}^{j}$ is nonzero for all $j$. Thus, we can normalize the eigenvector $u^{j}$ such that $u_{0}^{j}=\mu^{\frac{1}{2}}$. We further obtain from Lemma 3.2.8 that $c_{j}, u_{0}^{j}, \ldots, u_{n}^{k}$ is a solution of (3.2.24) for all $j=1, \ldots, n$. If we define $s_{i}\left(c_{j}, \lambda, \mu\right)$ by (3.2.23) with $u_{0}, \ldots, u_{n}$ replaced
by $u_{0}^{j}, \ldots, u_{n}^{j}$, then $s_{0}\left(c_{j}, \lambda, \mu\right)=1$ and the equations (3.2.6) are fulfilled. Then Proposition 3.2.2 yields that $s_{i}\left(c_{j}, \lambda, \mu\right)$ is of the form (3.2.10) for $i=1, \ldots, n$. In particular $s_{n}\left(c_{j}, \lambda, \mu\right)$ is of the form (3.2.10), a polynomial of degree $n$ and $u_{n}^{j}=0$ gives $s_{n}\left(c_{j}, \lambda, \mu\right)=0$ for $j=1, \ldots, n$. It follows from (3.2.10) that $s_{n}(c, \lambda, \mu)>0$ for $c \geq 0$ so that the numbers $c_{1}, \ldots, c_{n}$ must be negative.

Using the Sturm sequence property one can show another result that reduces the computational requirements for the derivation of the eigenvalues of a symmetric tridiagonal matrix with nonzero elements on the secondary diagonals substantially: Consider the conditions of Theorem 3.2.6 and let $s(t)$ be the number of sign changes of the sequence $\chi_{0}(t), \ldots, \chi_{n}(t)$, where $\chi_{k}(t)$ is defined by (3.2.22) with the additional convention that the sign of $\chi_{k}(t)$ is $-\operatorname{sgn}\left(\chi_{k-1}(t)\right)$ if $\chi_{k}(t)=0$. Then $s(t)$ equals the number of eigenvalues of $A_{n}$ that are less than $t$. Based on this observation, fast and accurate algorithms - like bisection, multisection, polysection or Godunovinverse iteration - for the numerical computation of the eigenvalues have been developed. A huge literature is available for real symmetric tridiagonal eigenvalue problems (see Cullum [14], Chapter 3, Section 3.5, Golub and Loan [26], §8.4, Matsekh [45] and Swarztrauber [62]).

We summarize the results for the number and location of the separation constants described by equation (3.2.18) in the following conclusion of Theorem 3.2.9.

Corollary 3.2.10. For $n_{2} \geq 0, \lambda_{2}>0$ and $\mu_{2}>0$ let $s_{n_{2}}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)$ be defined by (3.2.10). Then the equation

$$
c s_{n_{2}}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)=0
$$

has $n_{2}+1$ and distinct solutions $c_{0}, \ldots, c_{n_{2}}$ in the variable c given by $c_{0}=0$ and

$$
\begin{equation*}
c_{i}=\lambda_{2}+\mu_{2}-\left(\lambda_{2} \mu_{2}\right)^{\frac{1}{2}} \varepsilon_{i} \tag{3.2.28}
\end{equation*}
$$

for $i=1, \ldots, n$, where $\varepsilon_{1}, \ldots, \varepsilon_{n_{2}}$ are the pairwise distinct eigenvalues of the matrix $M_{n}\left(\lambda_{2}, \mu_{2}\right)$ defined by (3.2.25). Moreover, $c_{1}, \ldots, c_{n_{2}}$ are positive.

### 3.2.4 Steady-state probabilities

We have shown in Corollary 3.2.10 that $s_{n_{2}}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)$ is a polynomial in $c$ of degree $n_{2}$ with $n_{2}$ positive and distinct zeros which can be found by
solving the eigenvalue problem (3.2.26), i.e.,

$$
\left(M_{n}\left(\lambda_{2}, \mu_{2}\right)-\frac{\lambda_{2}+c}{\mu_{2}}\left(\frac{\lambda_{2}}{\mu_{2}}\right)^{-\frac{1}{2}} E_{n}\right) u=0
$$

By denoting the zeros by $c_{0}, \ldots, c_{n_{2}}$ as in Corollary 3.2.10, we have

$$
\begin{equation*}
c_{m} s_{n_{2}}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right)=0 \tag{3.2.29}
\end{equation*}
$$

for $m=0, \ldots, n_{2}$. For every $m=0, \ldots, n_{2}$, we get a solution of the steadystate equations (3.2.1) by setting

$$
\begin{equation*}
p_{i, j}=s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \tag{3.2.30}
\end{equation*}
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$. Therefore, we can represent the existing and unique stationary probabilities $p_{i, j}$ as linear combinations of the solutions (3.2.30) of the separation approach in the form

$$
\begin{equation*}
p_{i, j}=\sum_{m=0}^{n_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \tag{3.2.31}
\end{equation*}
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$. The constants $a_{0}, \ldots, a_{n_{2}}$, have to be chosen such that the boundary conditions in (3.2.1) corresponding to $i=n_{1}$ and the normalization condition are satisfied.

The boundary conditions for $i=n_{1}, j=0, \ldots, n_{2}-1$ are

$$
\begin{align*}
& \left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+j \mu_{2}\right) p_{n_{1}, j} \\
& \quad=\lambda_{1} p_{n_{1}-1, j}+\left(1-\delta_{j 0}\right)\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, j-1}+(j+1) \mu_{2} p_{n_{1}, j+1} \tag{3.2.32}
\end{align*}
$$

The boundary condition for $i=n_{1}$ and $j=n_{2}$ is

$$
\begin{equation*}
\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, n_{2}}=\lambda_{1} p_{n_{1}-1, n_{2}}+\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, n_{2}-1} \tag{3.2.33}
\end{equation*}
$$

Inserting (3.2.31) into (3.2.32) and (3.2.33) and simplifying with the help of the recurrence relations (3.2.13) and (3.2.14) leads to

$$
\begin{equation*}
\sum_{m=0}^{n_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{j}\left(-c_{m}\right)+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{j}\left(-c_{m}-\mu_{2}\right)\right)=0 \tag{3.2.34}
\end{equation*}
$$

for $i=n_{1}, j=0, \ldots, n_{2}-1$ and

$$
\begin{equation*}
\sum_{m=0}^{n_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}-1}\left(-c_{m}+\mu_{2}\right)+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}-1}\left(-c_{m}\right)\right)=0 \tag{3.2.35}
\end{equation*}
$$

with the abbreviated notation $s_{n_{1}}(\cdot)=s_{n_{1}}\left(\cdot, \lambda_{1}, \mu_{1}\right)$ and $s_{j}(\cdot)=s_{j}\left(\cdot, \lambda_{2}, \mu_{2}\right)$ for $j=0, \ldots, n_{2}$. Together with the normalization condition there are $n_{2}+2$ linear equations for the unknowns $a_{0}, \ldots, a_{n_{2}}$. Summing (3.2.34) over $j=$ $0, \ldots, n_{2}-1$ and using (3.2.15) yields the redundancy of (3.2.35). Furthermore, inserting (3.2.31) into the normalization condition $\sum_{i=0}^{n_{1}} \sum_{j=0}^{n_{2}} p_{i, j}=1$ and using (3.2.29) and (3.2.15) leads to an explicit expression for $a_{0}$ :

$$
\begin{equation*}
a_{0}=\left(s_{n_{1}}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right)\right)^{-1} . \tag{3.2.36}
\end{equation*}
$$

Remark 3.2.11. It is seen from (3.2.31) and (3.2.34) that the probabilities $p_{i, j}$ depend on the parameter $p \in[0,1]$ only through the coefficients $a_{0}, \ldots, a_{n_{2}}$. This is due to the fact that the stationary probabilities have been calculated from the nonoverflow balance equations, i.e., (3.2.1) for $i \neq n_{1}$, which are independent of $p$. Whereas the equations for the coefficients $a_{0}, \ldots, a_{n_{2}}$ arise from the overflow balance equations, i.e., (3.2.1) for $i=n_{1}$, which depend on $p$.

We summarize the results in the following theorem.
Theorem 3.2.12. The unique nonnegative and normalized solution of the steady-state equations (3.2.1) is given by

$$
\begin{equation*}
p_{i, j}=\sum_{m=0}^{n_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \tag{3.2.37}
\end{equation*}
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$ if the coefficients $a_{0}, \ldots, a_{n_{2}}$ are determined by (3.2.34) and (3.2.36), where $c_{0}=0$ and $c_{1}, \ldots, c_{n_{2}}$ are the by Corollary 3.2.10 positive and pairwise distinct solutions of the equation $s_{n_{2}}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right)=0, m=1, \ldots, n_{2}$.

The problem of determining the $\left(n_{1}+1\right)\left(n_{2}+1\right)$ unknowns $p_{i, j}, i=$ $0, \ldots, n_{1}, j=0, \ldots, n_{2}$, has now been reduced to the problem of determining $n_{2}$ eigenvalues from (3.2.29) and $n_{2}+1$ unknowns from the homogeneous linear equations (3.2.34) and (3.2.36).

The separation approach is displayed in Figure 3.3. The single line characterizes the boundary conditions, the unfilled endpoint characterizes the redundant boundary condition at $(i, j)=\left(n_{1}, n_{2}\right)$.


Figure 3.3: Separation scheme: Overflow without waiting rooms

Remark 3.2.13. By setting $n_{2}=0$, we can include the case that the second queue has no servers, i.e., the Erlang loss system. In this case, we obtain from (3.2.36) and (3.2.10) the standard formula for the stationary probabilities of the number of busy servers in the Erlang loss system with $n_{1}$ servers and traffic intensity $\rho=\lambda_{1} / \mu_{1}$ :

$$
P\left(L_{1}=i\right)=\frac{\rho^{i} / i!}{1+\rho+\ldots+\rho^{n_{1}} / n_{1}!}, \quad i=0, \ldots, n_{1}
$$

Remark 3.2.14. For $n_{2}>0$ the steady-state probability that $i$ customers are served at $Q_{1}$ is $\sum_{j=0}^{n_{2}} p_{i, j}, i=0, \ldots, n_{1}$. Using (3.2.29), (3.2.36), (3.2.10) and (3.2.15) we obtain again the standard Erlang loss formula

$$
P\left(L_{1}=i\right)=\sum_{j=0}^{n_{2}} p_{i, j}=\frac{\rho^{i} / i!}{1+\rho+\ldots+\rho^{n_{1}} / n_{1}!}, \quad i=0, \ldots, n_{1}
$$

### 3.2.5 Stationary quantities and numerical results

At the end of this section, we show how the results from the separation approach can be used to derive steady-state quantities of interest. The main characteristics of a queueing network with loss and overflow are, amongst others, the average blocking and loss probabilities for arriving customers, the overflow probability and the average number of customers in service. These quantities can be given in closed form in terms of the system parameters and the unknowns described in Theorem 3.2.12,

Let $B_{i}, i=1,2$, be the probability that an arriving $Q_{i}$-customer is blocked at $Q_{i}$ and let $O_{12}$ be the expected stationary number of customers
per unit time, which flow over from $Q_{1}$ to $Q_{1}$, then

$$
\begin{aligned}
B_{1} & =\sum_{j=0}^{n_{2}} p_{n_{1}, j}=\sum_{j=0}^{n_{2}} \sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \\
B_{2} & =\sum_{i=0}^{n_{1}} p_{i, n_{2}}=\sum_{i=0}^{n_{1}} \sum_{m=0}^{n_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \\
O_{12} & =\sum_{j=0}^{n_{2}-1} p_{n_{1}, j}=\sum_{j=0}^{n_{2}-1} \sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right)
\end{aligned}
$$

By (3.2.15) we get

$$
\begin{aligned}
B_{1} & =\sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right) \\
B_{2} & =\sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(\mu_{1}+c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \\
O_{12} & =\sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}-1}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right)
\end{aligned}
$$

The loss probability $P_{\text {Loss, } 1}$ for arriving $Q_{1}$-customers is then given by

$$
P_{\mathrm{Loss}, 1}=p_{n_{1}, n_{2}}+(1-p) O_{12}=B_{1}-p O_{12}
$$

the loss probability $P_{\text {Loss }, 2}$ for arriving $Q_{2}$-customers is equal to $B_{2}$.
$E L_{i}$ is the expected average queue length of $Q_{i}$ for $i=1,2$. A closed-form expression for $E L_{1}$ can be deduced directly from Remark 3.2.14. Moreover, $E L_{2}$ is given by

$$
E L_{2}=\sum_{j=1}^{n_{2}} j \sum_{i=0}^{n_{1}} p_{i, j}=\sum_{m=0}^{n_{2}} a_{m} \sum_{i=0}^{n_{1}} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) \sum_{j=1}^{n_{2}} j s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right)
$$

In order to derive $E L_{2}$, we can use (3.2.15) and (3.2.17) and get

$$
\begin{aligned}
E L_{2}= & \sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(\mu_{1}-c_{m}, \lambda_{1}, \mu_{1}\right)\left(n_{2} s_{n_{2}}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right)\right. \\
& \left.-s_{n_{2}-1}\left(2 \mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right)\right) \\
= & n_{2}-\sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(\mu_{1}-c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}-1}\left(2 \mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right)
\end{aligned}
$$

where we have used Corollary 3.2.10 and equation (3.2.36) for the last equality. However, by Little's law (see for example Asmussen [5] Theorem 4.1), the expected number of customers in the second queue is equal to the expected arrival rate to that queue multiplied with the average time spent in the queue. Therefore, we have

$$
\begin{equation*}
E L_{2}=\frac{p \lambda_{1} O_{12}+\lambda_{2}\left(1-B_{2}\right)}{\mu_{2}} . \tag{3.2.38}
\end{equation*}
$$

Note that the numerical efforts for deriving one of the quantities $B_{1}, B_{2}$, $O_{12}, E L_{1}$ and $E L_{2}$ is the same as deriving one of the stationary probabilities (3.2.37). Furthermore, these quantities can be calculated directly without computing the stationary probabilities.

These results can for example be used to minimize the total average costs of lost customers subject to the overflow parameter $p$ : The expected number of lost $Q_{1}$ - and $Q_{2}$-customers is given by $\lambda_{1} P_{\text {Loss, }, 1}=\lambda_{1}\left(B_{1}-p O_{12}\right)$ and $P_{\text {Loss }, 1}=\lambda_{2} B_{2}$, respectively. The total average costs of lost customers $C(p)$ is then given by

$$
C(p)=C_{1} \lambda_{1}\left(B_{1}-p O_{12}\right)+C_{2} \lambda_{2} B_{2},
$$

where $C_{i} \in \mathbb{R}, i=1,2$, are cost parameters. It is evident that an increase of $p$ results on the one hand in a decrease of the number of lost $Q_{1}$-customers because more of these customers will join $Q_{2}$. On the other hand, these overflowing $Q_{1}$-customers increase the occupation rate of $Q_{2}$ and thus, more $Q_{2}$-customers will be lost. In this view, the total costs $C(p)$ of lost customers should be a convex function of $p$ and a cost-minimal overflow parameter $p^{*}$ should exist. A numerical example is displayed in Figure 3.4. The cost minimizing values $p^{*}$ in this example are $p^{*}=0.9887,0.4838,0.1899$ and 0 (line by line from left to right) with associated minimal costs of $C\left(p^{*}\right)=$ $1.3714,1.4288,1.4775$ and 1.52. Thus, by inspecting the maxima of the cost functions, it is seen that the optimal choice of the overflow parameter $p$ can reduce the costs in a considerable amount.

The mean queue length $E L_{2}$ of the second queue can also be determined numerically from (3.2.38). Some numerical examples are given in Tables 3.1]3.6. For $p=0$, both queues are independent and $E L_{2}$ is given by the Erlang loss formula:

$$
E L_{2}=\frac{\sum_{i=1}^{n_{1}} \frac{\rho_{2}^{i}}{(i-1)!}}{1+\sum_{i=1}^{n_{1}} \frac{\rho_{2}^{i}}{i!}},
$$



Figure 3.4: $p \mapsto C(p)$ for $\left(n_{1}, n_{2}, \lambda_{1}, \mu_{1}, \lambda_{2}, \mu_{2}, C_{1}\right)=(2,2,3,3,1,1,1.2)$ and $C_{2}=3.4,3.6,3.8$ and 4 (line by line from left to right)
where $\rho_{2}=\lambda_{2} / \mu_{2}$. The first row of each of the Tables 3.1+3.5 reflects this observation. By inspecting the numerical results displayed in Tables 3.13.6, one finds that $E L_{2}$ is increasing in $\lambda_{2}, n_{2}$ and $p$ and decreasing in $\mu_{2}$. Furthermore, $E L_{2}$ is increasing for $p>0$ in $\lambda_{1}$ and decreasing for $p>0$ in $n_{1}$ and $\mu_{1}$. This is on the one hand explained by the fact that letting all other values fixed, the mean queue length must increase if the capacity $n_{2}$ or the offered workload $\lambda_{2}$ increase and decreases if the service rate $\mu_{2}$ is raised. Besides the usual influence of the arrival and service rates and the capacity on the queue length, on the other hand, an overflow is more likely if $\lambda_{1}$ or $p$ increase or $n_{1}$ decreases. Thus, $E L_{2}$ increases in these cases.

| $p$ | $n_{2}=2$ | $n_{2}=4$ | $n_{2}=6$ | $n_{2}=8$ | $n_{2}=10$ | $n_{2}=12$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 1.71004 | 3.27076 | 4.59956 | 5.60233 | 6.22761 | 6.52748 |
| 0.2 | 1.73063 | 3.33039 | 4.72756 | 5.83362 | 6.58125 | 6.98453 |
| 0.4 | 1.74819 | 3.38066 | 4.83522 | 6.03102 | 6.89379 | 7.40765 |
| 0.6 | 1.76333 | 3.42349 | 4.92648 | 6.19970 | 7.16787 | 7.79408 |
| 0.8 | 1.77652 | 3.46035 | 5.00446 | 6.34428 | 7.40718 | 8.14308 |
| 1 | 1.78810 | 3.49235 | 5.07162 | 6.46874 | 7.61572 | 8.45559 |

Table 3.1: Table of $E L_{2}$ for $n_{1}=2$ and $\left(\lambda_{1}, \mu_{1}, \lambda_{2}, \mu_{2}\right)=(1,0.1,2,0.3)$

| $p$ | $n_{2}=2$ | $n_{2}=4$ | $n_{2}=6$ | $n_{2}=8$ | $n_{2}=10$ | $n_{2}=12$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 1.71004 | 3.27076 | 4.59956 | 5.60233 | 6.22761 | 6.52748 |
| 0.2 | 1.72637 | 3.31814 | 4.70143 | 5.78646 | 6.50874 | 6.88981 |
| 0.4 | 1.74043 | 3.35861 | 4.78854 | 5.94651 | 6.76133 | 7.22921 |
| 0.6 | 1.75265 | 3.39350 | 4.86349 | 6.08565 | 6.98673 | 7.54366 |
| 0.8 | 1.76337 | 3.42382 | 4.92838 | 6.20682 | 7.18698 | 7.83228 |
| 1 | 1.77285 | 3.45038 | 4.98490 | 6.31267 | 7.36445 | 8.09517 |

Table 3.2: Table of $E L_{2}$ for $n_{1}=4$ and $\left(\lambda_{1}, \mu_{1}, \lambda_{2}, \mu_{2}\right)=(1,0.1,2,0.3)$

| $p$ | $n_{2}=2$ | $n_{2}=4$ | $n_{2}=6$ | $n_{2}=8$ | $n_{2}=10$ | $n_{2}=12$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 1.71004 | 3.27076 | 4.59956 | 5.60233 | 6.22761 | 6.52748 |
| 0.2 | 1.72233 | 3.30649 | 4.67649 | 5.74141 | 6.43967 | 6.80011 |
| 0.4 | 1.73301 | 3.33738 | 4.74327 | 5.86430 | 6.63303 | 7.05812 |
| 0.6 | 1.74236 | 3.36429 | 4.80151 | 5.97283 | 6.80826 | 7.30014 |
| 0.8 | 1.75063 | 3.38790 | 4.85256 | 6.06875 | 6.96642 | 7.52542 |
| 1 | 1.75798 | 3.40876 | 4.89753 | 6.15369 | 7.10878 | 7.73374 |

Table 3.3: Table of $E L_{2}$ for $n_{1}=6$ and $\left(\lambda_{1}, \mu_{1}, \lambda_{2}, \mu_{2}\right)=(1,0.1,2,0.3)$

| $p$ | $n_{2}=2$ | $n_{2}=4$ | $n_{2}=6$ | $n_{2}=8$ | $n_{2}=10$ | $n_{2}=12$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 1.71004 | 3.27076 | 4.59956 | 5.60233 | 6.22761 | 6.52748 |
| 0.2 | 1.71866 | 3.29586 | 4.65365 | 5.70014 | 6.37656 | 6.71854 |
| 0.4 | 1.72621 | 3.31779 | 4.70124 | 5.78782 | 6.51411 | 6.90094 |
| 0.6 | 1.73287 | 3.33708 | 4.74327 | 5.86635 | 6.64047 | 7.07384 |
| 0.8 | 1.73879 | 3.35416 | 4.78052 | 5.93670 | 6.75611 | 7.23671 |
| 1 | 1.74409 | 3.36936 | 4.81367 | 5.99978 | 6.86166 | 7.38927 |

Table 3.4: Table of $E L_{2}$ for $n_{1}=8$ and $\left(\lambda_{1}, \mu_{1}, \lambda_{2}, \mu_{2}\right)=(1,0.1,2,0.3)$

| $p$ | $n_{2}=2$ | $n_{2}=4$ | $n_{2}=6$ | $n_{2}=8$ | $n_{2}=10$ | $n_{2}=12$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| 0 | 1.71004 | 3.27076 | 4.59956 | 5.60233 | 6.22761 | 6.52748 |
| 0.2 | 1.71553 | 3.28676 | 4.63408 | 5.66475 | 6.32256 | 6.64904 |
| 0.4 | 1.72037 | 3.30088 | 4.66481 | 5.72141 | 6.41119 | 6.76592 |
| 0.6 | 1.72467 | 3.31342 | 4.69226 | 5.77279 | 6.49357 | 6.87769 |
| 0.8 | 1.72852 | 3.32460 | 4.71683 | 5.81937 | 6.56988 | 6.98402 |
| 1 | 1.73198 | 3.33463 | 4.73891 | 5.86162 | 6.64037 | 7.08470 |

Table 3.5: Table of $E L_{2}$ for $n_{1}=10$ and $\left(\lambda_{1}, \mu_{1}, \lambda_{2}, \mu_{2}\right)=(1,0.1,2,0.3)$

| $\lambda_{2}$ | $\lambda_{1}=0.5$ | $\lambda_{1}=1$ | $\lambda_{1}=1.5$ | $\lambda_{1}=2$ | $\lambda_{1}=2.5$ | $\lambda_{1}=3$ |
| :--- | :---: | :---: | :---: | :---: | :---: | :---: |
| 0.5 | 4.94770 | 5.74973 | 6.94810 | 7.83830 | 8.40927 | 8.77512 |
| 1 | 7.86791 | 8.13140 | 8.51837 | 8.82503 | 9.04536 | 9.20428 |
| 1.5 | 8.84975 | 8.94551 | 9.09722 | 9.23094 | 9.33734 | 9.42111 |
| 2 | 9.24300 | 9.28876 | 9.36491 | 9.43657 | 9.49736 | 9.54799 |
| 2.5 | 9.44250 | 9.46869 | 9.51365 | 9.55769 | 9.59659 | 9.63022 |
| 3 | 9.56073 | 9.57756 | 9.60704 | 9.63670 | 9.66362 | 9.68750 |

Table 3.6: Table of $E L_{2}$ for $n_{1}=n_{2}=10$ and $\left(\mu_{1}, \mu_{2}, p\right)=(0.1,0.1,0.5)$

### 3.3 Overflow with waiting rooms: From servers to servers

### 3.3.1 Jockeying to servers

### 3.3.1.1 Steady-state equations and separation approach

Now we consider the $p$-overflow model $\mathrm{S} / \mathrm{S} / \mathrm{S}$ with waiting rooms. The waiting room at $Q_{i}$ has capacity $q_{i} \geq 0$ for $i=1,2$. In this model, an arriving $Q_{1}$-customer is blocked and directed to $Q_{2}$ if all $n_{1}$ servers are busy in $Q_{1}$. Blocked customers are served by one of the servers in $Q_{2}$ if at least one is available, are queued in $Q_{1}$ if all servers in $Q_{2}$ are busy and a waiting position is available in $Q_{1}$ and are lost otherwise. The jockeying discipline ascertains that waiting $Q_{1}$-customers will be served at $Q_{1}$ if a $Q_{1}$-server becomes available or at $Q_{2}$ if a $Q_{2}$-server becomes available, all $Q_{1}$-servers are busy and no customers are waiting at $Q_{2}$, whatever happens first. With $k_{i}=q_{i}+n_{i}$ for $i=1,2$, the balance equations are

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)\left(1-\chi_{i-n_{1}}\left(1-\chi_{j-n_{2}}\right)\right)+p \lambda_{1} \chi_{i-n_{1}}\left(1-\chi_{j-n_{2}}\right)\right) \\
& \left.\quad+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
& =\left(1-\chi_{i-n_{1}-1} \chi_{n_{2}-j-1}\right)  \tag{3.3.1}\\
& \quad \times\left(\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1}\right) \\
& \quad+\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \delta_{i n_{1}} \chi_{n_{2}-j}+\lambda_{2}\left(1-\chi_{i-n_{1}-1} \chi_{n_{2}-j}\right)\right) p_{i, j-1} \\
& \quad+\left(1-\delta_{i k_{1}}\right)\left(\left(1-\chi_{i-n_{1}} \chi_{n_{2}-1-j}\right)\left((i+1) \wedge n_{1}\right) \mu_{1}+n_{2} \mu_{2} \delta_{j n_{2}} \chi_{i-n_{1}}\right) p_{i+1, j}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$, where $i \wedge j=\min \{i, j\}$ and $\chi_{i-n}=$ $\mathbf{1}_{[n, \infty)}(i)$ for $i, j, n \in \mathbb{N}$.

For $i=n_{1}+1, \ldots, k_{1}$ and $j=0, \ldots, n_{2}-1$ these equations imply $p_{i, j}=0$,
since it is impossible that customers are waiting at $Q_{1}$ while there is at least one server available at $Q_{2}$.

It will turn out that a similar but more complex separation approach than the one for the model without waiting rooms can be carried out for this model. This separation approach is depicted in Figure 3.5. The single lines correspond to the resulting boundary conditions and the unfilled circle highlights a redundant boundary condition at $(i, j)=\left(n_{1}, n_{2}\right)$.


Figure 3.5: Separation scheme: Model S/S/S

For $i=0, \ldots, n_{1}-1$ and $j=0, \ldots, k_{2}$ the balance equations take the form

$$
\begin{align*}
& \left(\lambda_{1}+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+i \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j}=\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j} \\
& \quad+(i+1) \mu_{1} p_{i+1, j}+\left(1-\delta_{j 0}\right) \lambda_{2} p_{i, j-1}+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1} \tag{3.3.2}
\end{align*}
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, k_{2}$. By the results of Section 3.2, there are solutions of the form $p_{i, j}=\alpha_{i} \beta_{j}$ in the region $i=0, \ldots, n_{1}$ and $j=0, \ldots, k_{2}$, where $\alpha_{i}$ satisfies (3.2.3) for $i=0, \ldots, n_{1}$ and is given by $\alpha_{i}=s_{i}\left(c, \lambda_{1}, \mu_{1}\right)$ for $i=0, \ldots, n_{1}$. For $j=0, \ldots, n_{2}, \beta_{j}$ is given by or at least proportional to $\beta_{j}=s_{j}\left(-c, \lambda_{2}, \mu_{2}\right) . \quad c$ is the separation constant for this region and will be determined in the next section. Moreover, for $i=0, \ldots, n_{1}$ and $j=n_{2}, \ldots, k_{2}$, the approach $p_{i, j}=\alpha_{i} \beta_{j}$ gives $\alpha_{i}=s_{i}\left(c, \lambda_{1}, \mu_{1}\right)$ and $\beta_{j}$ satisfies

$$
\begin{equation*}
\left(\lambda_{2}\left(1-\delta_{j k_{2}}\right)+n_{2} \mu_{2}-c\right) \beta_{j}=\lambda_{2} \beta_{j-1}+n_{2} \mu_{2}\left(1-\delta_{j k_{2}}\right) \beta_{j+1} \tag{3.3.3}
\end{equation*}
$$

for $j=n_{2}, \ldots, k_{2}$. We assume $q_{1} \geq 1$ for the moment. In the region $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}+1, \ldots, k_{2}$, the steady-state probabilities $p_{i, j}$ can be separated into the form $p_{i, j}=\gamma_{i} \delta_{j}$. This leads to the equations

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)+n_{1} \mu_{1}+d\right) \gamma_{i}=\lambda_{1} \gamma_{i-1}+\left(1-\delta_{i k_{1}}\right) n_{1} \mu_{1} \gamma_{i+1} \text { and }  \tag{3.3.4}\\
& \left(\lambda_{2}\left(1-\delta_{j k_{2}}\right)+n_{2} \mu_{2}-d\right) \delta_{j}=\lambda_{2} \delta_{j-1}+\left(1-\delta_{j k_{2}}\right) n_{2} \mu_{2} \delta_{j+1}
\end{align*}
$$

for $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}+1, \ldots, k_{2}$ with the separation constant

$$
d=n_{2} \mu_{2}-\lambda_{2} \frac{\delta_{k_{2}-1}}{\delta_{k_{2}}}=-n_{1} \mu_{1}+\lambda_{1} \frac{\gamma_{k_{1}-1}}{\gamma_{k_{1}}} .
$$

We will derive the solution of (3.3.3) and (3.3.4) and determine the separation constants in the next sections.

### 3.3.1.2 Solution of the separation approach

We have already shown that there are solutions of the form $p_{i, j}=\alpha_{i} \beta_{j}$ for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$. $\alpha_{i}$ equals $s_{i}\left(c, \lambda_{1}, \mu_{1}\right)$ for $i=0, \ldots, n_{1}$ and $\beta_{j}$ is proportional to $s_{i}\left(-c, \lambda_{2}, \mu_{2}\right)$ for $j=0, \ldots, n_{2} . c$ is a separation constant. Moreover, $\beta_{j}$ has to satisfy (3.3.3) for $j=n_{2}, \ldots, k_{2}$. In the region $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}+1, \ldots, k_{2}$, the separation approach $p_{i, j}=\gamma_{i} \delta_{j}$ led to the equations (3.3.4). We now show that the solutions of equations (3.3.3) and (3.3.4) can be expressed in terms of the Chebyshev polynomials of the second kind. First, we give a formal definition of these polynomials (see for example Abramowitz and Stegun [1]).

Definition 3.3.1. The Chebyshev polynomials of the second kind are defined by the recurrence relation

$$
\begin{equation*}
2 x U_{l}(x)=U_{l+1}(x)+U_{l-1}(x), \quad l \geq 0 \tag{3.3.5}
\end{equation*}
$$

with the boundary conditions $U_{-1} \equiv 0$ and $U_{0} \equiv 1$.
It is well known (see Abramowitz and Stegun (1) that

$$
\begin{equation*}
U_{l}\left(\frac{1}{2}\left(z+\frac{1}{z}\right)\right)=\sum_{k=0}^{l} z^{2 k-l} \tag{3.3.6}
\end{equation*}
$$

for $z \neq 0$. The following proposition states the solution of (3.3.3).
Proposition 3.3.2. A solution of (3.3.3) is given by $\beta_{j}=\phi_{j}(c)$ for $j=$
$n_{2}-1, \ldots, k_{2}$, where the function $\phi_{i}(c)$ is defined by

$$
\begin{equation*}
\phi_{j}(c)=\Psi_{k_{2}-j}(c)-\Psi_{k_{2}-j-1}(c), \quad j=n_{2}-1, \ldots, k_{2} \tag{3.3.7}
\end{equation*}
$$

with

$$
\begin{equation*}
\Psi_{l}(c)=\Psi_{l}\left(c, \lambda_{2}, \mu_{2}\right)=\left(\frac{n_{2} \mu_{2}}{\lambda_{2}}\right)^{\frac{l}{2}} U_{l}\left(\frac{\lambda_{2}+n_{2} \mu_{2}-c}{2 \sqrt{\lambda_{2} n_{2} \mu_{2}}}\right) \tag{3.3.8}
\end{equation*}
$$

for $l=-1, \ldots, q_{2}$. Moreover, $\Psi_{-1}(c)=0, \Psi_{0}(c)=1$ and

$$
\begin{equation*}
\left(\lambda_{2}+n_{2} \mu_{2}-c\right) \Psi_{l}(c)=\lambda_{2} \Psi_{l+1}(c)+n_{2} \mu_{2} \Psi_{l-1}(c) \tag{3.3.9}
\end{equation*}
$$

holds for every $l \geq 0$

Proof. $\Psi_{-1}(c)=0, \Psi_{0}(c)=1$ and (3.3.9) follow directly from (3.3.5). With $\phi_{j}(c)$ defined by (3.3.7) for $j=n_{2}-1, \ldots, k_{2}$, we obtain (3.3.3), i.e.,

$$
\left(\lambda_{2}\left(1-\delta_{j k_{2}}\right)+n_{2} \mu_{2}-c\right) \phi_{j}(c)=\lambda_{2} \phi_{j-1}(c)+n_{2} \mu_{2}\left(1-\delta_{j k_{2}}\right) \phi_{j+1}(c)
$$

for $j=n_{2}, \ldots, k_{2}$ from (3.3.9).

It is immediately seen by comparing the equations (3.3.3) and (3.3.4) that a solution of (3.3.3) yields a solution of (3.3.4). For convenience, it is reasonable to set $\beta_{j}=\phi_{j}(c)$ for $j=n_{2}, \ldots, k_{2}$. The solution of the equations (3.3.4) can be deduced directly from Proposition 3.3.2,

Corollary 3.3.3. A solution of (3.3.4) is given by $\gamma_{i}=\theta_{i}(d)$ for $i=$ $n_{1}, \ldots, k_{1}$ and $\delta_{j}=\phi_{j}(d)$ for $j=n_{2}, \ldots, k_{2}$ and every $d \in \mathbb{R}$, where the function $\phi_{j}(d)$ is defined by (3.3.7) and $\theta_{i}(d)$ is similarly to $\phi_{j}(c)$ defined by

$$
\begin{equation*}
\theta_{i}(d)=\Omega_{k_{1}-i}(d)-\Omega_{k_{1}-i-1}(d), \quad i=0, \ldots, k_{1} \tag{3.3.10}
\end{equation*}
$$

with

$$
\begin{equation*}
\Omega_{l}(d)=\Omega_{l}\left(d, \lambda_{1}, \mu_{1}\right)=\left(\frac{n_{1} \mu_{1}}{\lambda_{1}}\right)^{\frac{l}{2}} U_{l}\left(\frac{\lambda_{1}+n_{1} \mu_{1}+d}{2 \sqrt{\lambda_{1} n_{1} \mu_{1}}}\right) \tag{3.3.11}
\end{equation*}
$$

for $l=-1, \ldots, q_{1}$.

The solution of the steady-state equations can be deduced from Proposition 3.3 .2 and Corollary 3.3 .2 by taking the specific values of the solutions of the separation approach at the borders of the boundary region into ac-
count. This is done in the next section. We conclude this section with a note concerning $\phi_{k_{2}-1}(c)$ and a helpful relation between $\theta_{n_{1}}$ and $\theta_{n_{1}+1}$.

Lemma 3.3.4. (i) The function $\phi_{k_{2}-1}$ defined by (3.3.7) satisfies

$$
\begin{equation*}
\phi_{k_{2}-1}(c)=\frac{n_{2} \mu_{2}-c}{\lambda_{2}} \tag{3.3.12}
\end{equation*}
$$

for every $c \in \mathbb{R}$.
(ii) The functions $\theta_{n_{1}}$ and $\theta_{n_{1}+1}$ defined by (3.3.10) satisfy

$$
\begin{equation*}
\lambda_{1} \theta_{n_{1}}(d)-n_{1} \mu_{1} \theta_{n_{1}+1}(d)=d \Omega_{q_{1}-1}(d) \tag{3.3.13}
\end{equation*}
$$

for every $d \in \mathbb{R}$.

Proof. $U_{1}(x)=2 x$ follows from (3.3.5) and that shows

$$
\begin{aligned}
\phi_{k_{2}-1}(c) & =\Psi_{1}(c)-\Psi_{0}(c)=\left(\frac{n_{2} \mu_{2}}{\lambda_{2}}\right)^{\frac{1}{2}} U_{1}\left(\frac{\lambda_{2}+n_{2} \mu_{2}-c}{2 \sqrt{\lambda_{2} n_{2} \mu_{2}}}\right)-1 \\
& =\frac{n_{2} \mu_{2}-c}{\lambda_{2}}
\end{aligned}
$$

Now we show (iii). It follows analogously to the proof of Proposition (3.3.2) that $\theta_{l}(d), l \geq 0$, defined by (3.3.10) and (3.3.11) satisfies

$$
\begin{equation*}
\left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)+n_{1} \mu_{1}+d\right) \theta_{i}(d)=\lambda_{1} \theta_{i-1}(d)+n_{1} \mu_{1}\left(1-\delta_{i k_{1}}\right) \theta_{i+1}(d) \tag{3.3.14}
\end{equation*}
$$

for $i=n_{1}, \ldots, k_{1}$. Then, (3.3.13) follows from (3.3.11) and (3.3.14).

### 3.3.1.3 Boundary condition and separation constants

By Proposition 3.3.2, $\beta_{j}$ must by proportional to $s_{j}\left(-c, \lambda_{2}, \mu_{2}\right)$ for $j=$ $0, \ldots, n_{2}$ and to $\phi_{j}(c)$ for $j=n_{2}-1, \ldots, k_{2}$ so that we can choose

$$
\beta_{j}= \begin{cases}s_{j}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}(c), & j=0, \ldots, n_{2} .  \tag{3.3.15}\\ s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{j}(c), & j=n_{2}-1, \ldots, k_{2} .\end{cases}
$$

The separation constant $c$ has to be chosen such that the definitions of $\beta_{n_{2}-1}$ and $\beta_{n_{2}}$ match, i.e.,

$$
\begin{equation*}
s_{n_{2}-1}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}(c)=s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}-1}(c) \tag{3.3.16}
\end{equation*}
$$

has to be fulfilled. The following lemma gives an equivalent formulation of this condition.

Lemma 3.3.5. Let $s_{j}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right)$ for $j=n_{2}-1, n_{2}$ be given by (3.2.11) and let $\Psi_{j}(c)$ for $j=q_{2}-1, q_{2}$ be given by (3.3.8). Then the equation (3.3.16) is equivalent to

$$
\begin{equation*}
c\left[s_{n_{2}}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}}(c)-s_{n_{2}-1}\left(\mu_{2}-c, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}-1}(c)\right]=0 \tag{3.3.17}
\end{equation*}
$$

Proof. We write $s_{n}(c)$ for $s_{n}\left(c, \lambda_{2}, \mu_{2}\right)$ in the following. From the definition (3.3.7) of $\phi_{n_{2}-1}(c)$ and (3.3.9) we get

$$
\begin{aligned}
\phi_{n_{2}-1}(c) & =\Psi_{q_{2}+1}(c)-\Psi_{q_{2}}(c) \\
& =\frac{\lambda_{2}+n_{2} \mu_{2}-c}{\lambda_{2}} \Psi_{q_{2}}(c)-\frac{n_{2} \mu_{2}}{\lambda_{2}} \Psi_{q_{2}-1}(c)-\Psi_{q_{2}}(c) \\
& =\frac{n_{2} \mu_{2}-c}{\lambda_{2}} \Psi_{q_{2}}(c)-\frac{n_{2} \mu_{2}}{\lambda_{2}} \Psi_{q_{2}-1}(c) .
\end{aligned}
$$

Inserting this and $\phi_{n_{2}}(c)=\Psi_{q_{2}}(c)-\Psi_{q_{2}-1}(c)$ in (3.3.16) gives after simplifying

$$
\begin{align*}
& \Psi_{q_{2}}(c)\left(\lambda_{2} s_{n_{2}-1}(-c)-\left(n_{2} \mu_{2}-c\right) s_{n_{2}}(-c)\right) \\
& \quad=\Psi_{q_{2}-1}(c)\left(\lambda_{2} s_{n_{2}-1}(-c)-n_{2} \mu_{2} s_{n_{2}}(-c)\right) \tag{3.3.18}
\end{align*}
$$

With (3.2.13) and (3.2.14) we get expressions for the terms in the brackets in the equation above, namely

$$
\begin{aligned}
\lambda_{2} s_{n_{2}-1}(-c)-\left(n_{2} \mu_{2}-c\right) s_{n_{2}}(-c) & =c s_{n_{2}}\left(\mu_{2}-c\right) \quad \text { and } \\
\lambda_{2} s_{n_{2}-1}(-c)-n_{2} \mu_{2} s_{n_{2}}(-c) & =c s_{n_{2}-1}\left(\mu_{2}-c\right)
\end{aligned}
$$

Hence, (3.3.18) is equivalent to

$$
c s_{n_{2}}\left(\mu_{2}-c\right) \Psi_{q_{2}}(c)=c s_{n_{2}-1}\left(\mu_{2}-c\right) \Psi_{q_{2}-1}(c)
$$

and therefore equivalent to (3.3.17), too.
The function in the square brackets in equation (3.3.17) is a polynomial of degree $k_{2}$ in $c$ with $k_{2}$ positive and distinct zeros. These zeros are given by an eigenvalue problem for a tridiagonal real symmetric matrix (see Theorem 3.3.6 in the following). In order to prove this fact, it is necessary to symmetrize the recurrence equations (3.2.6) that define $s_{i}\left(c, \mu_{2}, \lambda_{2}\right)$ in a suitable way. We did this in a similar way in Section 3.2.3 for the basic model without waiting rooms. The corresponding result was stated
in Lemma 3.2.7. We will start by symmetrizing the equations (3.2.6) and define the corresponding tridiagonal matrix.

We suppress the indexes of $\lambda_{2}, \mu_{2}, n_{2}$ and $q_{2}$ for the moment. Let $\lambda, \mu>0, c \in \mathbb{R}$ and $n \geq 1, q \geq 0$ and set

$$
v_{i}(c)= \begin{cases}u_{i}(\mu-c), & i=0, \ldots, n,  \tag{3.3.19}\\ v_{n+q-1}(c) U_{n+q-j-1}\left(\frac{\lambda+n \mu-c}{2 \sqrt{\lambda n \mu}}\right), & j=n-1, \ldots, n+q,\end{cases}
$$

where $u_{i}(c), i \geq 0$, is given by (3.2.23). It follows immediately from the definition of $u_{i}(c), i \geq 0$, that $v_{n+q-1}(c) \neq 0$ and furthermore $v_{n+q}(c)=0$ follows from $U_{-1} \equiv 0$. By considering the equations for $v_{n}(c)$ and $v_{n-1}(c)$ and the definition of $\Psi_{q}$ and $\Psi_{q-1}$ it is seen that the existence of $v_{0}(c), \ldots, v_{n+q}(c)$ fulfilling (3.3.19) is equivalent to

$$
s_{n}(\mu-c, \lambda, \mu) \Psi_{q}(c)-s_{n-1}(\mu-c, \lambda, \mu) \Psi_{q-1}(c)=0
$$

We get from (3.2.24) and (3.3.5) the symmetrized set of equations

$$
\begin{equation*}
\left(1-\delta_{i 0}\right) i^{\frac{1}{2}} v_{i-1}(c)-\left(i+\frac{\lambda-c}{\mu}\right)\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} v_{i}(c)+(i+1)^{\frac{1}{2}} v_{i+1}(c)=0 \tag{3.3.20}
\end{equation*}
$$

for $i=0, \ldots, n-1$ and

$$
\begin{equation*}
n^{\frac{1}{2}} v_{i-1}(c)-\left(n+\frac{\lambda-c}{\mu}\right)\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} v_{i}(c)+n^{\frac{1}{2}} v_{i+1}(c)=0 \tag{3.3.21}
\end{equation*}
$$

for $i=n, \ldots, n+q-1$. Define the matrix $R_{n, q}=R_{n, q}(\lambda, \mu)=\left(r_{i, j}\right)_{i, j=1, \ldots, n+q}$ for $\lambda, \mu>0$ by $r_{i, j}=m_{i, j}$ for $i, j=1, \ldots, n$ (see (3.2.25)) and

$$
\begin{array}{rlr}
r_{i, i} & =-(n-1)\left(\frac{\lambda}{\mu}\right)^{-\frac{1}{2}} & \text { for } i=n, \ldots, n+q, \\
r_{i, i+1}=r_{i+1, i} & =n^{\frac{1}{2}} & \text { for } i=n, \ldots, n+q-1 \text { and }  \tag{3.3.22}\\
r_{i, j} & =0 & \text { for }|i-j|>1 .
\end{array}
$$

$R_{n, q}$ is a real symmetric tridiagonal matrix and has therefore $n+q$ real and pairwise distinct eigenvalues. We get the following result concerning the solutions of (3.3.17) by generalizing Theorem 3.2.9, where the latter follows itself from Theorem 3.3.6 with $q=0$ and $c$ replaced by $\mu-c$.

Theorem 3.3.6. For $n \geq 1, q \geq 0, c \in \mathbb{R}, \lambda>0$ and $\mu>0$. Then the
equation

$$
s_{n}(\mu-c, \lambda, \mu) \Psi_{q}(c)-s_{n-1}(\mu-c, \lambda, \mu) \Psi_{q-1}(c)=0
$$

has $n+q$ positive and distinct solutions $c_{1}, \ldots, c_{n+q}$ in the variable $c$ given by

$$
\begin{equation*}
c_{i}=\lambda+\mu-(\lambda \mu)^{\frac{1}{2}} \eta_{i}, \quad i=1, \ldots, n+q \tag{3.3.23}
\end{equation*}
$$

where $\eta_{1}, \ldots, \eta_{n+q}$ are the pairwise distinct eigenvalues of the matrix $R_{n, q}(\lambda, \mu)$.

Proof. The first part of the proof is essentially the same as the proof of Theorem 3.2.9, Let

$$
p_{n, q}(c)=p_{n, q}(c, \lambda, \mu)=s_{n}(\mu-c, \lambda, \mu) \Psi_{q}(c)-s_{n-1}(\mu-c, \lambda, \mu) \Psi_{q-1}(c)
$$

It remains to show that the zeros of $p_{n, q}(c, \lambda, \mu)=0$ are positive for $q \geq 0$. We show $p_{n, q}(c, \lambda, \mu) \neq 0$ for $c \leq 0$. Let $c \leq 0$. For $q=0, \Psi_{-1}(c)=0$ and $\Psi_{0}(c)=1$ yield $p_{n, 0}(c)=s_{n}(\mu-c, \lambda, \mu)$. From (3.2.10) it follows $s_{n}(\mu-c, \lambda, \mu)>0$ and $s_{n}(-c, \lambda, \mu)>0$. The latter yields together with (3.2.14) that

$$
\begin{equation*}
\frac{s_{n}(\mu-c, \lambda, \mu)}{s_{n-1}(\mu-c, \lambda, \mu)}>1 \tag{3.3.24}
\end{equation*}
$$

holds. Now let $q \geq 1$. With

$$
\begin{equation*}
z_{0}=\frac{\lambda-c+n \mu}{2 \sqrt{\lambda n \mu}}+\left(\left(\frac{\lambda-c+n \mu}{2 \sqrt{\lambda n \mu}}\right)^{2}-1\right)^{\frac{1}{2}}>0 \tag{3.3.25}
\end{equation*}
$$

we get from (3.3.8) and (3.3.6) that

$$
\Psi_{l}(c)=\left(\frac{n \mu}{\lambda}\right)^{\frac{l}{2}} U_{l}\left(\frac{1}{2}\left(z_{0}+\frac{1}{z_{0}}\right)\right)=\left(\frac{n \mu}{\lambda}\right)^{\frac{l}{2}} \sum_{k=1}^{l} z_{0}^{2 k-l}>0
$$

for $l=q-1, q$. This yields

$$
\begin{align*}
\frac{\Psi_{q}(c)}{\Psi_{q-1}(c)} & =\left(\frac{n \mu}{\lambda}\right)^{\frac{1}{2}} \frac{U_{q}\left(\frac{1}{2}\left(z_{0}+\frac{1}{z_{0}}\right)\right)}{U_{q-1}\left(\frac{1}{2}\left(z_{0}+\frac{1}{z_{0}}\right)\right)}=z_{0}\left(\frac{n \mu}{\lambda}\right)^{\frac{1}{2}} \frac{\sum_{k=0}^{q} z_{0}^{2 k-q}}{\sum_{k=1}^{q} z_{0}^{2 k-q}} \\
& >z_{0}\left(\frac{n \mu}{\lambda}\right)^{\frac{1}{2}}>1 \tag{3.3.26}
\end{align*}
$$

by (3.3.25). Now (3.3.24) and (3.3.26) give $p_{n, q}(c, \lambda, \mu) \neq 0$ for $c \leq 0$ and
$n, q \geq 1$.

### 3.3.1.4 Steady-state probabilities

By Theorem 3.3.6, the separation equation (3.3.17) has $k_{2}+1$ distinct solutions $c_{0}=0, c_{1}, \ldots, c_{k_{2}}$, where $c_{1}, \ldots, c_{k_{2}}$ are positive. We can represent the stationary probabilities $p_{i, j}$ for $i=0, \ldots, n_{1}$ and $j=0, \ldots, k_{2}$ as linear combinations of all feasible solutions of the separation approach $p_{i, j}=\alpha_{i} \beta_{j}$ where $\alpha_{i}$ is given by $s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right)$ for $m=0, \ldots, k_{2}$ and $\beta_{j}$ is given by (3.3.15). We get

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} b_{m}^{\prime} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2} .  \tag{3.3.27}\\ \sum_{m=0}^{k_{2}} b_{m}^{\prime} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2}\end{cases}
$$

The constants $b_{m}^{\prime}, m=0, \ldots, k_{2}$, have to be chosen such that the boundary conditions in (3.3.1) corresponding to $i=n_{1}, j=0, \ldots, k_{2}$ and the normalization condition are satisfied.

For $i=n_{1}, \ldots, k_{1}$ and $j=n_{2}, \ldots, k_{2}$, we examined the approach $p_{i, j}=$ $\gamma_{i} \delta_{j}$ and showed in Corollary 3.3 .3 that $\gamma_{i}$ must be proportional to $\theta_{i}(d)$ defined by (3.3.10) for $i=n_{1}, \ldots, k_{1}$ and that $\delta_{j}$ must be proportional to $\phi_{j}(d)$ defined by (3.3.7) for $j=n_{2}, \ldots, k_{2}$. Furthermore, $p_{i, j}=\gamma_{i} \delta_{j}$ has to be chosen such that it matches (3.3.27) on the boundary $i=n_{1}, j=n_{2}, \ldots, k_{2}$. For this reason we must ensure $\theta_{n_{1}}(d)=0$. The equation $\theta_{n_{1}}(d)=0$ on its part has $q_{1}$ negative and distinct zeros as the following theorem shows. Once again, we drop the indexes of $\lambda_{1}, \mu_{1}, n_{1}$ and $q_{1}$ for the moment.

Theorem 3.3.7. For $n \geq 1, q \geq 0, d \in \mathbb{R}, \lambda>0$ and $\mu>0$ let $\theta_{n}(d, \lambda, \mu)$ be defined by (3.3.29). Then the equation

$$
\theta_{n}(d, \lambda, \mu)=0
$$

has $q$ negative and distinct solutions in the variable $d$ given by the eigenvalues of the matrix $T_{n, q}(\lambda, \mu)=\left(t_{i, j}\right)_{i, j=1, \ldots, q}$, where

$$
\begin{align*}
t_{i, i} & =-\left(\left(1-\delta_{i 1}\right) \lambda+n \mu\right) & \text { for } i=1, \ldots, q, \\
t_{i, i+1}=r_{i+1, i} & =\sqrt{\lambda n \mu} & \text { for } i=1, \ldots, q-1 \text { and }  \tag{3.3.28}\\
t_{i, j} & =0 & \text { for }|i-j|>1 .
\end{align*}
$$

Proof. For $\lambda, \mu>0$ and $n, q \geq 1$ recall the definitions (3.3.10) and (3.3.11):

$$
\begin{equation*}
\theta_{i}(d)=\theta_{i}(d, \lambda, \mu)=\Omega_{n+q-i}(d)-\Omega_{n+q-i-1}(d), \quad i=0, \ldots, n+q \tag{3.3.29}
\end{equation*}
$$

and

$$
\begin{equation*}
\Omega_{l}(d)=\Omega_{l}(d, \lambda, \mu)=\left(\frac{n \mu}{\lambda}\right)^{\frac{l}{2}} U\left(\frac{\lambda+n \mu+d}{2 \sqrt{\lambda n \mu}}\right) \tag{3.3.30}
\end{equation*}
$$

for $l=-1, \ldots, q$. Define

$$
\begin{equation*}
w_{i}(d)=w_{i}(d, \lambda, \mu)=U_{i}\left(\frac{\lambda+n \mu+d}{2 \sqrt{\lambda n \mu}}\right)-\left(\frac{\lambda}{n \mu}\right)^{\frac{1}{2}} U_{i-1}\left(\frac{\lambda+n \mu+d}{2 \sqrt{\lambda n \mu}}\right) \tag{3.3.31}
\end{equation*}
$$

for $i \geq 0$. Obviously, $\theta_{n}(d)=0$ is equivalent to $w_{q}(d)=0$ by (3.3.29). Additionally, (3.3.5) yields

$$
\left(1-\delta_{i 0}\right) \sqrt{\lambda n \mu} w_{i-1}(d)-\left(\left(1-\delta_{i 0}\right) \lambda+n \mu+d\right) w_{i}(d)+\sqrt{\lambda n \mu} w_{i+1}(d)
$$

for $i \geq 0 . T_{n, q}=T_{n, q}(\lambda, \mu)$ is defined such that the solutions of $\theta_{n}(d)=0$ are exactly the eigenvalues of $T_{n, q} . T_{n, q}$ itself is a real symmetric tridiagonal matrix of rank $q$ and has therefore $q$ real and pairwise distinct eigenvalues by Theorem 3.2.6. In remains to show that the eigenvalues of $T_{n, q}$ are negative. Using the equivalence of $\theta_{n}(d)=0$ and $w_{n}(d)=0$, (3.3.31) and (3.3.26) it is seen analogously to the proof of the positivity argument in Theorem 3.3.6 that $\theta_{n}(d) \neq 0$ for $d \geq 0$.

We denote the solutions of $\theta_{n_{1}}(d)=0$ by $d_{1}, \ldots, d_{q_{1}}$. By Theorem 3.3.7 these solutions are negative and distinct. The roots of (3.3.17) are nonnegative and distinct so that $c_{0}, \ldots, c_{k_{2}}, d_{1}, \ldots, d_{q_{1}}$ are distinct. Thus, the probabilities $p_{i, j}$ for $i=n_{1}, \ldots, k_{1}$ and $j=n_{2}, \ldots, k_{2}$ are given by

$$
\begin{align*}
p_{i, j}=\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m},\right. & \left.\lambda_{2}, \mu_{2}\right) \theta_{i}\left(c_{m}\right) \phi_{j}\left(c_{m}\right)  \tag{3.3.32}\\
& +\sum_{l=1}^{q_{1}} e_{l} \theta_{i}\left(d_{l}\right) \phi_{j}\left(d_{l}\right)
\end{align*}
$$

for arbitrary $q_{1} \geq 0$. In order to match (3.3.27) and (3.3.32) it is convenient to choose $b_{m}^{\prime}=b_{m} \theta_{n_{1}}\left(c_{m}\right), m=0, \ldots, k_{2}$, so that (3.3.27) can be restated
as
$p_{i, j}=\left\{\begin{array}{l}\sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right), j=0, \ldots, n_{2}, \\ \sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{j}\left(c_{m}\right), j=n_{2}, \ldots, k_{2},\end{array}\right.$
for $i=0, \ldots, n_{1}$.
The constants $b_{0}, \ldots, b_{k_{2}}$ and $e_{1}, \ldots, e_{q_{1}}$ in (3.3.32) and (3.3.33) have to be chosen such that the boundary conditions in (3.2.1) for 1) $i=n_{1}$ and $j=0, \ldots, k_{2}$ and 2) $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}$ and the normalization condition are satisfied. The boundary conditions are

$$
\begin{align*}
&\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+j \mu_{2}\right) p_{n_{1}, j}=\lambda_{1} p_{n_{1}-1, j} \\
&+\left(1-\delta_{j 0}\right)\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, j-1} \\
&+(j+1) \mu_{2} p_{n_{1}, j+1}, \\
& i=n_{1}, j=0, \ldots, n_{2}-1,(3.3 .34  \tag{3.3.34}\\
&\left(\lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, n_{2}}= \lambda_{1} p_{n_{1}-1, n_{2}}+\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, n_{2}-1} \\
&+n_{2} \mu_{2} p_{n_{1}, n_{2}+1} \\
&+\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}+1, n_{2}}, \\
& i=n_{1}, j=n_{2},  \tag{3.3.35}\\
&\left(\lambda_{1}+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, j}= \lambda_{1} p_{n_{1}-1, j} \\
&+\left(1-\delta_{j k_{2}}\right) n_{2} \mu_{2} p_{n_{1}, j+1} \\
&+\lambda_{2} p_{n_{1}, j-1}+n_{1} \mu_{1} p_{n_{1}+1, j}, \\
& i=n_{1}, j=n_{2}+1, \ldots, k_{2},  \tag{3.3.36}\\
& \\
& \\
&\left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)+\lambda_{2}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i, n_{2}}= \lambda_{1} p_{i-1, n_{2}}+n_{2} \mu_{2} p_{i, n_{2}+1}  \tag{3.3.37}\\
&+\left(1-\delta_{i k_{1}}\right)\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i+1, n_{2}}, \\
& i=n_{1}+1, \ldots, k_{1}, j=n_{2} .
\end{align*}
$$

The equations (3.3.34) are identical to (3.2.32) so that (3.3.33), (3.3.34) and
(3.2.34) yield

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} b_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{j}\left(-c_{m}\right)\right. \\
&  \tag{3.3.38}\\
& \left.\quad+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{j}\left(-c_{m}-\mu_{2}\right)\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right)=0
\end{align*}
$$

for $i=n_{1}$ and $j=0, \ldots, n_{2}-1$. By substituting (3.3.32) and (3.3.33) into (3.3.36) and (3.3.37) and using (3.3.13) and the recurrence relations (3.2.13) and (3.2.14) we get

$$
\begin{array}{r}
\sum_{m=0}^{k_{2}} b_{m} c_{m} s_{n_{2}}\left(-c_{m}\right)\left(s_{n_{1}}\left(c_{m}+\mu_{1}\right) \theta_{n_{1}}\left(c_{m}\right)+s_{n_{1}}\left(c_{m}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right) \phi_{j}\left(c_{m}\right) \\
=n_{1} \sum_{l=1}^{q_{1}} e_{l} \theta_{n_{1}+1}\left(d_{l}\right) \phi_{j}\left(d_{l}\right) \tag{3.3.39}
\end{array}
$$

for $i=n_{1}, j=n_{2}+1, \ldots, k_{2}$ and

$$
\begin{gather*}
\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right)\left(\lambda_{2} \theta_{i}\left(c_{m}\right) \phi_{n_{2}-1}\left(c_{m}\right)-n_{2}\left(1-\delta_{i k_{1}}\right) \theta_{i+1}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right)\right) \\
=\sum_{l=1}^{q_{1}} e_{l}\left(n_{2}\left(1-\delta_{i k_{1}}\right) \theta_{i+1}\left(d_{l}\right) \phi_{n_{2}}\left(d_{l}\right)-\lambda_{2} \theta_{i}\left(d_{l}\right) \phi_{n_{2}-1}\left(d_{l}\right)\right) \tag{3.3.40}
\end{gather*}
$$

for $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}$. By summing (33.3.38), (3.3.39) and (3.3.40) and once again using the recurrence relations (3.2.13) and (3.2.14) it is seen that equation (3.3.35) is redundant.

Together with the normalization condition, we have $k_{2}+q_{1}+1$ linear equations for the unknowns $b_{0}, \ldots, b_{k_{2}}$ and $e_{1}, \ldots, e_{q_{1}}$. Inserting (3.3.32) and (3.3.33) into the normalization condition $\sum_{i=0}^{k_{1}} \sum_{j=0}^{k_{2}} p_{i, j}=1$ and using (3.2.16) results in an equation for $b_{0}$ :

$$
\begin{align*}
& b_{0} s_{n_{1}}(\mu) \theta_{n_{1}}(0)\left(s_{n_{2}}\left(\mu_{2}\right) \Psi_{q_{2}}(0)-s_{n_{2}-1}\left(\mu_{2}\right) \Psi_{q_{2}-1}(0)\right) \\
& \quad+\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right) \Omega_{q_{1}-1}\left(c_{m}\right) \Psi_{q_{2}}\left(c_{m}\right)  \tag{3.3.41}\\
& \quad+\sum_{l=1}^{q_{1}} e_{l} \Omega_{q_{1}-1}\left(d_{l}\right) \Psi_{q_{2}}\left(d_{l}\right)=1 .
\end{align*}
$$

Remark 3.3.8. It is seen that the probabilities $p_{i, j}$ depend on the parameter $p \in[0,1]$ only through the coefficients $b_{0}, \ldots, b_{k_{2}}$ and $e_{1}, \ldots, e_{q_{1}}$. This is due
to the fact that the stationary probabilities have been calculated from the non-overflow balance equations, i.e., (3.3.1) for $i=n_{1}, j=0, \ldots, k_{2}$ and $i=n_{1}+1, \ldots, k_{1}, j=n_{2}$, which are independent of $p$, whereas the equations for the coefficients arise from the $p$-dependent balance equations in (3.3.1).

We summarize the result in the following theorem.
Theorem 3.3.9. The unique nonnegative and normalized solution of the steady-state equations (3.3.1) is given by
$p_{i, j}=\left\{\begin{array}{l}\sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right), j=0, \ldots, n_{2}, \\ \sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{j}\left(c_{m}\right), j=n_{2}, \ldots, k_{2},\end{array}\right.$
for $i=0, \ldots, n_{1}$ and
$p_{i, j}=\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{i}\left(c_{m}\right) \phi_{j}\left(c_{m}\right)+\sum_{l=1}^{q_{1}} e_{l} \theta_{i}\left(d_{l}\right) \phi_{j}\left(d_{l}\right)$
for $i=n_{1}, \ldots, k_{1}, j=n_{2}, \ldots, k_{2}$ and every $q_{1} \geq 0$ if the coefficients $b_{0}, \ldots, b_{k_{2}}$ and $e_{1}, \ldots, e_{q_{1}}$ are determined by (3.3.38)-(3.3.40) and (3.3.41). $c_{1}, \ldots, c_{k_{2}}$ are the by Theorem 3.3.6 positive and pairwise distinct solutions of (3.3.17), $c_{0}=0$ and $d_{1}, \ldots, d_{q_{1}}$ are the by Theorem 3.3.7 negative and pairwise distinct solutions of $\theta_{n_{1}}\left(d_{l}\right)=0, l=1, \ldots, q_{1}$.

The problem of determining the $\left(k_{1}+1\right)\left(k_{2}+1\right)$ unknowns $p_{i, j}, i=$ $0, \ldots, k_{1}, j=0, \ldots, k_{2}$, has now been reduced to the problem of determining $k_{2}+q_{1}$ eigenvalues from (3.3.17) and the equation $\theta_{n_{1}}(d)=0$ and $k_{2}+$ $q_{1}+1$ unknowns from the linear equations (3.3.38)-(3.3.40) and (3.3.41). Alternative separation approaches for the case $\mu_{1}=\mu_{2}$ and $p=1$ can be found in Morrison [47] and Morrison and Wright [49]. The approach in [47] is preferable, when $q_{2}$ is large compared to $q_{1}, n_{1}$ and $n_{2}$ or infinite. The approach in 49] is preferable, when $q_{1}$ is large compared to $q_{2}$ and $n_{2}$ or infinite. Both approaches can be generalized to the case of arbitrary service rates and weighted overflow traffic, too.

Remark 3.3.10. In the case $q_{2}=0$, (3.3.17) reduces to (3.2.18) because of $\Psi_{-1} \equiv 0$ and $\Psi_{0} \equiv 1$. If in addition $q_{1}=0$, then (3.3.27) and (3.3.32) reduce to (3.2.31). Moreover, equation (3.3.38) reduces to (3.2.34) because of $\theta_{n_{1}} \equiv \phi_{n_{2}} \equiv 1$, while equations (3.3.39) and (3.3.40) vanish. Finally, (3.3.41) reduces to (3.2.36) because of $\Psi_{0} \equiv \Omega_{0} \equiv 1$ and $\Psi_{-1} \equiv \Omega_{-1} \equiv 0$.

### 3.3.1.5 Stationary quantities

The results of the separation approach can be used to derive steady-state quantities of interest. The characteristics of this queueing network can be calculated in a similar manner as for the model without waiting rooms. Most of these characteristics can be derived conformably to the results in Morrison [46]; therefore, we omit the details. The average blocking probabilities $B_{1}$ and $B_{2}$ for arriving customers, the loss probabilities $P_{\text {Loss,1 }}$ and $P_{\text {Loss,2 }}$ for arriving customers and the overflow probability $O_{12}$ are given by

$$
\begin{array}{rlrl}
B_{1} & =\sum_{j=0}^{n_{2}} p_{n_{1}, j}, & B_{2}=\sum_{i=0}^{k_{1}} p_{i, k_{2}} \\
P_{\mathrm{loss}, 1} & =\sum_{j=n_{2}}^{k_{2}} p_{k_{1}, j}+(1-p) O_{12}, & P_{\mathrm{loss}, 2}=B_{2} \\
O_{12} & =\sum_{j=0}^{n_{2}-1} p_{n_{1}, j}
\end{array}
$$

Additionally, let $P_{\text {queue, } 1}$ be the probability that an arriving $Q_{i}$-customer is queued in the waiting room in $Q_{i}, i=1,2$, then

$$
P_{\text {queue }, 1}=\sum_{i=n_{1}}^{k_{1}-1} \sum_{j=0}^{k_{2}} p_{i, j} \quad \text { and } \quad P_{\text {queue }, 2}=\sum_{i=0}^{k_{1}-1} \sum_{j=n_{2}}^{k_{2}-1} p_{i, j} .
$$

The mean departure rates $R_{11}, R_{12}$ and $R_{22}$ from the waiting room in $Q_{1}$ to the servers in $Q_{1}$ and $Q_{2}$, respectively, and from the servers in $Q_{2}$ to $Q_{2}$ are given by

$$
\begin{array}{ll}
R_{11} & =n_{1} \mu_{1} \sum_{i=n_{1}+1}^{k_{1}} \sum_{j=n_{2}}^{k_{2}} p_{i, j}, \\
R_{22} & =n_{2} \mu_{2} \sum_{i=0}^{k_{1}} \sum_{j=n_{2}+1}^{k_{2}} p_{n_{1}, j} .
\end{array}
$$

The mean number $E L_{i}$ of customers in $Q_{i}, i=1,2$, is given by

$$
\begin{aligned}
& E L_{1}=\sum_{i=0}^{n_{1}} \sum_{j=0}^{k_{2}} i p_{i, j}+n_{1} \sum_{i=n_{1}+1}^{k_{1}} \sum_{j=n_{2}}^{k_{2}} p_{i, j}+\sum_{i=n_{1}+1}^{k_{1}} \sum_{j=n_{2}}^{k_{2}}\left(i-n_{1}\right) p_{i, j} \\
& E L_{2}=\sum_{i=0}^{n_{1}} \sum_{j=0}^{n_{2}-1} j p_{i, j}+n_{2} \sum_{i=0}^{k_{1}} \sum_{j=n_{2}}^{k_{2}} p_{i, j}+\sum_{i=0}^{k_{1}} \sum_{j=n_{2}+1}^{k_{2}}\left(j-n_{2}\right) p_{i, j} .
\end{aligned}
$$

By (3.2.15), (3.3.7) and (3.3.10) we get

$$
\begin{aligned}
& B_{1}= \sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right), \\
& B_{2}= \sum_{m=0}^{k_{2}} b_{m} s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right)\left(s_{n_{1}}\left(c_{m}+\mu_{1}, \lambda_{1}, \mu_{1}\right) \theta_{n_{1}}\left(c_{m}\right)\right. \\
&\left.+s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right)+\sum_{l=1}^{q_{1}} e_{l} \Omega_{q_{1}-1}\left(d_{l}\right), \\
&+\sum_{l=1}^{q_{1}} e_{l} \Psi_{q_{2}}\left(d_{l}\right)+(1-p) O_{12}, \\
& P_{\text {loss }, 1}=\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(c_{m}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}}\left(c_{m}\right) \\
& \quad+\sum_{l=1}^{O_{12}}=\sum_{m=0}^{k_{2}} b_{m} \Omega_{q_{1}-1}\left(d_{l}\right) \Psi_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}-1}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right), \\
& R_{11}= n_{1} \mu_{1}\left(\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \Omega_{q_{1}-1}\left(c_{m}\right) \Psi_{q_{2}}\left(c_{m}\right)\right. \\
& R_{12}= n_{2} \mu_{2}\left(\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \Omega_{q_{1}-1}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right)\right. \\
&\left.\quad+\sum_{l=1}^{q_{1}} e_{l} \Omega_{q_{1}-1}\left(d_{l}\right) \phi_{n_{2}}\left(d_{l}\right)\right) \text { and } \\
& R_{22}= n_{2} \mu_{2} \sum_{j=n_{2+1}}^{k_{2}}\left(\frac{n_{2} \mu_{2}}{\lambda_{2}}\right)^{k_{2}-j} P_{\text {Loss,2 }} .
\end{aligned}
$$

where we have used $\phi_{k_{2}} \equiv \theta_{k_{1}} \equiv 1$. For the last equality, we exploited the sum of (3.3.1) over $i=0, \ldots, n_{1}$. It is also possible to give an expression for $E L_{1}$ and $E L_{2}$. We omit the details and refer to [46].

### 3.3.1.6 Alternative approach without waiting room in second queue

In the special case that the second queue has no waiting room, i.e., in the case $q_{2}=0$, another approach is feasible to derive the stationary proba-
bilities as functions of only $2 n_{2}+1$ unknowns (see 46] for the case $p=1$ and $\mu_{1}=\mu_{2}$ ). The approach from the previous section gives the probabilities as functions of $2\left(n_{2}+q_{1}\right)+1$ unknowns in this case. The approach is sketched in Figure 3.6, Compared with Figure 3.5, it is seen that the balance equations on the boundary line at $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}$, are no longer required to determine the separation constants in the region above the boundary line. Actually, these equations lead to a recursive formula for the stationary probabilities on the boundary line. The redundant boundary condition corresponds to the state $\left(n_{1}, n_{2}\right)$, which is represented by an unfilled circle in Figure 3.6


Figure 3.6: Separation scheme: Model $\mathrm{S} / \mathrm{S} / \mathrm{S}$ with $q_{2}=0$

Let $q_{2}=0$ in the following. The balance equations are given by (3.3.1) with $k_{2}$ replaced by $n_{2}$, i.e.,

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)\left(1-\chi_{i-n_{1}}\left(1-\delta_{j n_{2}}\right)\right)+p \lambda_{1} \chi_{i-n_{1}}\left(1-\delta_{j n_{2}}\right)\right) \\
& \left.\quad+\lambda_{2}\left(1-\delta_{j n_{2}}\right)+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
& =\left(1-\chi_{i-n_{1}-1}\left(1-\delta_{j n_{2}}\right)\right)  \tag{3.3.42}\\
& \quad \times\left(\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+\left(1-\delta_{j n_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1}\right) \\
& \quad+\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \delta_{i n_{1}} \chi_{n_{2}-j}+\lambda_{2}\left(1-\chi_{i-n_{1}-1} \chi_{n_{2}-j}\right)\right) p_{i, j-1} \\
& \quad+\left(1-\delta_{i k_{1}}\right)\left(\left(1-\chi_{i-n_{1}} \chi_{n_{2}-1-j}\right)\left((i+1) \wedge n_{1}\right) \mu_{1}+n_{2} \mu_{2} \delta_{j n_{2}} \chi_{i-n_{1}}\right) p_{i+1, j}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, n_{2}$. Observe that $\delta_{j n_{2}}=\chi_{j-n_{2}}$ and $1-\delta_{j n_{2}}=$ $\chi_{n_{2}-j-1}$ and note that $p_{i, j}=0$ for $i>n_{1}$ and $j<n_{2}$ because it is impossible for customers to wait at $Q_{1}$ while there is at least one server available at $Q_{2}$.

For $i=0, \ldots, n_{1}-1$ and $j=0, \ldots, n_{2}$ the balance equations are identical
to (3.3.2) with $k_{2}$ replaced by $n_{2}$ so that we establish from (3.2.31) that

$$
\begin{equation*}
p_{i, j}=\sum_{m=0}^{n_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \tag{3.3.43}
\end{equation*}
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$, where $c_{0}, \ldots, c_{n_{2}}$ are given by (3.2.29). It remains to determine $p_{i, n_{2}}$ for $i=n_{1}, \ldots, k_{1}$. Let $q_{1} \geq 1$ in the following. The balance equations

$$
\begin{align*}
\left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i, n_{2}} & =\lambda_{1} p_{i-1, n_{2}} \\
& +\left(1-\delta_{i k_{1}}\right)\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i+1, n_{2}} \tag{3.3.44}
\end{align*}
$$

for $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}$ yield

$$
\lambda_{1} p_{i-1, n_{2}}=\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i, n_{2}}
$$

for $i=n_{1}+1, \ldots, k_{1}$. It follows

$$
\begin{equation*}
p_{i, n_{2}}=\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{i-n_{1}} p_{n_{1}, n_{2}} \tag{3.3.45}
\end{equation*}
$$

for $i=n_{1}, \ldots, k_{1}$, where $p_{n_{1}, n_{2}}$ is given by (3.3.43).
It remains to satisfy the boundary conditions at $i=n_{1}$ and $j=0, \ldots, n_{2}$ and the normalization condition. For $i=n_{1}$ and $j=0, \ldots, n_{2}-1$ these boundary conditions are given by (3.2.32). We deduce from (3.3.43), (3.2.31) and (3.2.34) that

$$
\begin{equation*}
\sum_{m=0}^{n_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{j}\left(-c_{m}\right)+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{j}\left(-c_{m}-\mu_{2}\right)\right)=0 \tag{3.3.46}
\end{equation*}
$$

must hold for $i=n_{1}$ and $j=0, \ldots, n_{2}-1$. The boundary condition

$$
\begin{align*}
& \left(\lambda_{1}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, n_{2}}=\lambda_{1} p_{n_{1}-1, n_{2}} \\
& \quad+\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}+1, n_{2}}+\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, n_{2}-1} \tag{3.3.47}
\end{align*}
$$

for $i=n_{1}$ and $j=n_{2}$ reduces with the help of (3.3.45) to

$$
\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, n_{2}}=\lambda_{1} p_{n_{1}-1, n_{2}}+\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, n_{2}-1}
$$

which was shown to be equivalent to (3.2.35). The latter equation was seen to be redundant. Substituting (3.3.43) and (3.3.45) into the normalization
condition yields after simplification

$$
\begin{align*}
& a_{0} s_{n_{1}}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \\
& \quad+\left(\sum_{m=0}^{n_{2}} a_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right)\right) \sum_{l=1}^{q_{1}}\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{l}=1 \tag{3.3.48}
\end{align*}
$$

In summary, we can record the following result.
Theorem 3.3.11. The unique nonnegative and normalized solution of the steady-state equations (3.3.42) is given by

$$
p_{i, j}=\sum_{m=0}^{n_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right)
$$

for $i=0, \ldots, n_{1}$ and $j=0, \ldots, n_{2}$ and

$$
p_{i, n_{2}}=\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{i-n_{1}} p_{n_{1}, n_{2}}
$$

for $i=n_{1}, \ldots, k_{1}$ if the coefficients $a_{0}, \ldots, a_{n_{2}}$ are determined by (3.3.46) and (3.3.48), where $c_{0}=0$ and $c_{1}, \ldots, c_{n_{2}}$ are the by Corollary 3.2.10 positive and pairwise distinct solutions of the equation $s_{n_{2}}\left(\mu_{2}-c_{m}, \lambda_{2}, \mu_{2}\right)=0$, $m=1, \ldots, n_{2}$.

The number of unknowns has now been reduced from $\left(n_{1}+1\right)\left(n_{2}+1\right)+q_{1}$ to $2 n_{2}+1$.

Remark 3.3.12. For $q_{1}=0$ the results reduce to those from Section 3.2,
Remark 3.3.13. By (3.3.45) and (3.3.48) the results even hold in the case $q_{1}=\infty$ as long as the stability condition

$$
\begin{equation*}
\lambda_{1}<n_{1} \mu_{1}+n_{2} \mu_{2} \tag{3.3.49}
\end{equation*}
$$

is fulfilled.

The stability condition (3.3.49) states that the servers of $Q_{1}$ and $Q_{2}$ must together be able to handle the arrival stream of $Q_{1}$. Moreover, the stability condition is independent of $p$. This can be explained analytically by the fact that the stationary probabilities have been calculated from the $p$-independent balance equations, which led to the stationary probabilities (3.3.43) and (3.3.45). These probabilities and therefore equation (3.3.48)
are independent of $p$. Heuristically, this is due to the fact that on the one hand in the case $q_{1}=\infty$, the capacity of $Q_{2}$ is still limited by $n_{2}$ so that in the long-run average, the queue size of $Q_{2}$ is negligible and the arrival stream of $Q_{2}$ has no influence on the stability of the system. On the other hand, queued $Q_{1}$-customers overflow to the servers in $Q_{2}$ as soon as one is available. But once again in the long-run average, the queue size of $Q_{2}$ is negligible so that queued $Q_{1}$-customers must be handled by the servers of both $Q_{1}$ and $Q_{2}$ to guarantee stability. Furthermore, near saturation, there are always customers waiting in $Q_{1}$, who move to $Q_{2}$ if a server becomes available so that near saturation, the overflow stream is negligible, too.

### 3.3.2 No jockeying

We investigate the structure of the steady-state equations for the model $\mathrm{S} / \mathrm{S} / \mathrm{N}$ in this section. This model differs from the model S/S/S of Section 3.3.1 with jockeying only in the fact that waiting $Q_{1}$-customers are served exclusively by the $Q_{1}$-servers, i.e., no jockeying is allowed (see Figure 3.1). Recall that an arriving $Q_{1}$-customer is blocked and directed to $Q_{2}$ if all $n_{1}$ servers are busy in $Q_{1}$. Blocked customers are served by one of the servers in $Q_{2}$ if at least one is available, are queued in $Q_{1}$ if all servers in $Q_{2}$ are busy and a waiting position is available in $Q_{1}$ and are lost otherwise. We suppose that the number of waiting positions in $Q_{1}$ and $Q_{2}$ is positive, i.e. $q_{1} \geq 1$ and $q_{2} \geq 1$. This model is also treated in Morrison [46] for the case $p=1$ and $\mu_{1}=\mu_{2}$ and numerically in Kaufman et al. [34] for arbitrary service rates and $p=1$. In our case of arbitrary $\mu_{1}, \mu_{2}>0$ and $p$-weighted overflow traffic the balance equations are given by

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}} \chi_{j-n_{2}}\right)\left(1-\chi_{i-n_{1}}\left(1-\chi_{j-n_{2}}\right)\right)+p \lambda_{1} \chi_{i-n_{1}}\left(1-\chi_{j-n_{2}}\right)\right) \\
& \left.\quad+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
& =\left(1-\chi_{i-n_{1}-1} \chi_{n_{2}-j-1}\right)\left(\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}\right)  \tag{3.3.50}\\
& \quad+\left(1-\delta_{i k_{1}}\right)\left((i+1) \wedge n_{1}\right) \mu_{1} p_{i+1, j} \\
& \quad+\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \chi_{i-n_{1}} \chi_{n_{2}-j}+\lambda_{2}\right) p_{i, j-1}+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$. It is seen that the term $p \lambda_{1} \chi_{i-n_{1}}(1-$ $\chi_{j-n_{2}}$ ) on the left side of (3.3.50) and $p \lambda_{1} \chi_{i-n_{1}} \chi_{n_{2}-j}$ on the right side of (3.3.50) do not vanish in the regions $i=n_{1}, \ldots, k_{1}, j=0, \ldots, n_{2}-1$ and $i=n_{1}, \ldots, k_{1}, j=0, \ldots, n_{2}$, respectively. Consequently, in contrast to the previous models, the probabilities in these regions depend on the parameter $p$ not only through the coefficients determined by the boundary conditions (see
also (3.3.53) and (3.3.54)). The separation approach for model S/S/S (see Figure (3.5) can be extended for this model to the region $i=n_{1}+1, \ldots, k_{1}$, $j=0, \ldots, n_{2}-1$. The extended approach is illustrated in Figure 3.7 and leads to a reduction of the number of probabilities to be calculated in this region.


Figure 3.7: Separation scheme: Model S/S/N

Analogously to the model with jockeying of Section 3.3.1 the balance equations lead to
$p_{i, j}=\left\{\begin{array}{l}\sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right), j=0, \ldots, n_{2}, \\ \sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{j}\left(c_{m}\right), j=n_{2}, \ldots, k_{2},\end{array}\right.$
for $i=0, \ldots, n_{1}$ and
$p_{i, j}=\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{i}\left(c_{m}\right) \phi_{j}\left(c_{m}\right)+\sum_{l=1}^{q_{1}} e_{l} \theta_{i}\left(d_{l}\right) \phi_{j}\left(d_{l}\right)$
for $i=n_{1}, \ldots, k_{1}$ and $j=n_{2}, \ldots, k_{2}$. Again, $c_{0}, \ldots, c_{k_{2}}$ and $d_{1}, \ldots, d_{q_{1}}$ are the roots of (3.3.17) and $\theta_{n_{1}}\left(d_{l}\right)=0, l=1, \ldots, q_{1}$, respectively.

For $i=n_{1}+1, \ldots, k_{1}$ and $j=0, \ldots, n_{2}$ (3.3.50) yields

$$
\begin{equation*}
p_{i, j}=\sum_{r=i}^{k_{1}} p_{r, 0} \Pi_{r-i, j}, \tag{3.3.53}
\end{equation*}
$$

where $\Pi_{r, j}, r=0, \ldots, k_{1}, j=0, \ldots, n_{2}$, are the solutions of the $p$-dependent equations

$$
\begin{align*}
& \left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+j \mu_{2}\right) \Pi_{r, j} \\
& \quad=\left(p \lambda_{1}+\lambda_{2}\right)\left(1-\delta_{j 0}\right) \Pi_{r, j-1}+n_{1} \mu_{1}\left(1-\delta_{r 0}\right) \Pi_{r-1, j}+(j+1) \Pi_{r, j+1} \tag{3.3.54}
\end{align*}
$$

with boundary conditions $\Pi_{r 0}=\delta_{r 0}$ for $r=0, \ldots, k_{1}$, where $\delta_{r 0}$ is the Kronecker function. It is shown in Morrison [46] that further investigations of these quantities lead successively to formulas for $p_{n_{1}+1,0}, \ldots, p_{k_{1}, 0}$ depending on $b_{0}, \ldots, b_{k_{2}}$ and $e_{1}, \ldots, e_{q_{1}}$, while these coefficients can be calculated from the boundary conditions in (3.3.50). In particular $b_{0}, \ldots, b_{k_{2}}$, $e_{1}, \ldots, e_{q_{1}}$ and the probabilities $p_{n_{1}+1,0}, \ldots, p_{k_{1}, 0}$ are uniquely determined up to a multiplicative constant by the equations

$$
\begin{equation*}
\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right) \theta_{i}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right)+\sum_{l=1}^{q_{1}} e_{l} \theta_{i}\left(d_{l}\right) \phi_{n_{2}}\left(d_{l}\right)=\sum_{r=i}^{k_{1}} p_{r, 0} \Pi_{r-i, n_{2}} \tag{3.3.55}
\end{equation*}
$$

for $i=n_{1}+1, \ldots, k_{1}$,

$$
\begin{array}{r}
\sum_{m=0}^{k_{2}} b_{m}\left(c_{m} s_{n_{1}}\left(\mu_{1}+c_{m}\right) s_{j}\left(-c_{m}\right)+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{j}\left(-\mu_{2}-c_{m}\right)\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right) \\
=n_{1} \sum_{r=n_{1}+1}^{k_{1}} p_{r, 0} \Pi_{r-n_{1}-1, j} \tag{3.3.56}
\end{array}
$$

for $j=0, \ldots, n_{2}-1$,

$$
\begin{array}{r}
\sum_{m=0}^{k_{2}} b_{m} c_{m} s_{n_{2}}\left(-c_{m}\right)\left(s_{n_{1}}\left(\mu_{1}+c_{m}\right) \theta_{n_{1}}\left(c_{m}\right)+s_{n_{1}}\left(c_{m}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right) \phi_{j}\left(c_{m}\right) \\
=n_{1} \sum_{l=1}^{q_{1}} e_{l} \theta_{n_{1}+1}\left(d_{l}\right) \phi_{j}\left(d_{l}\right) \tag{3.3.57}
\end{array}
$$

for $j=n_{2}+1, \ldots, k_{2}$ and

$$
\begin{align*}
& \lambda_{2} \sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right) \theta_{i}\left(c_{m}\right) \phi_{n_{2}-1}\left(c_{m}\right) \\
& \quad+\lambda_{2} \sum_{l=1}^{q_{1}} e_{l} \theta_{i}\left(d_{l}\right) \phi_{n_{2}-1}\left(d_{l}\right)=\left(p \lambda_{1}+\lambda_{2}\right) \sum_{r=i}^{k_{1}} p_{r, 0} \Pi_{r-i, n_{2}-1} \tag{3.3.58}
\end{align*}
$$

for $i=n_{1}+1, \ldots, k_{1}$. It can be shown that the boundary equation for $i=n_{1}$ and $j=n_{2}$ is redundant. The normalization condition yields

$$
\begin{align*}
& b_{0} s_{n_{1}}(\mu) \theta_{n_{1}}(0)\left(s_{n_{2}}\left(\mu_{2}\right) \Psi_{q_{2}}(0)-s_{n_{2}-1}\left(\mu_{2}\right) \Psi_{q_{2}-1}(0)\right) \\
& \quad+\frac{\lambda_{1}}{n_{1}} \sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right)\left(\Omega_{q_{1}}\left(c_{m}\right)-1\right) \Psi_{q_{2}}\left(c_{m}\right)  \tag{3.3.59}\\
& \quad+\frac{\lambda_{1}}{n_{1}} \sum_{l=1}^{q_{1}} e_{l}\left(\Omega_{q_{1}}\left(d_{l}\right)-1\right) \Psi_{q_{2}}\left(d_{l}\right)=1
\end{align*}
$$

The result is outlined in the following theorem.

Theorem 3.3.14. The unique nonnegative and normalized solution of the steady-state equations (3.3.50) is given by
$p_{i, j}=\left\{\begin{array}{l}\sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{n_{2}}\left(c_{m}\right), j=0, \ldots, n_{2}, \\ \sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{n_{1}}\left(c_{m}\right) \phi_{j}\left(c_{m}\right), j=n_{2}, \ldots, k_{2},\end{array}\right.$
for $i=0, \ldots, n_{1}$,
$p_{i, j}=\sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \theta_{i}\left(c_{m}\right) \phi_{j}\left(c_{m}\right)+\sum_{l=1}^{q_{1}} e_{l} \theta_{i}\left(d_{l}\right) \phi_{j}\left(d_{l}\right)$
for $i=n_{1}, \ldots, k_{1}, j=n_{2}, \ldots, k_{2}$ and

$$
p_{i, j}=\sum_{r=i}^{k_{1}} p_{r, 0} \Pi_{r-i, j}
$$

for $i=n_{1}+1, \ldots, k_{1}$ and $j=0, \ldots, n_{2}$ if the coefficients $b_{0}, \ldots, b_{k_{2}}$ and $e_{1}, \ldots, e_{q_{1}}$ and the probabilities $p_{n_{1}+1,0}, \ldots, p_{k_{1}, 0}$ are determined by the $k_{2}+$ $2 q_{1}+1$ equations (3.3.55)-(3.3.58) and (3.3.59). $\Pi_{r, j}, r=0, \ldots, k_{1}, j=$ $0, \ldots, n_{2}$, are the solutions of (3.3.54). $c_{1}, \ldots, c_{k_{2}}$ are the by Theorem 3.3.6 positive and pairwise distinct solutions of (3.3.17), $c_{0}=0$ and $d_{1}, \ldots, d_{q_{1}}$ are
the by Theorem 3.3.7 negative and pairwise distinct solutions of $\theta_{n_{1}}\left(d_{l}\right)=0$, $l=1, \ldots, q_{1}$.

The problem of determining the $\left(k_{1}+1\right)\left(k_{2}+1\right)$ unknowns $p_{i, j}, i=$ $0, \ldots, k_{1}, j=0, \ldots, k_{2}$, has now been reduced to the problem of determining $k_{2}+q_{1}$ eigenvalues from (3.3.17) and the equation $\theta_{n_{1}}(d)=0$ and $k_{2}+2 q_{1}+1$ unknowns from the linear equations (3.3.55) to (3.3.59). An alternative separation approach for the case $\mu_{1}=\mu_{2}$ and $p=1$ that is preferable when $q_{2}$ is infinite or large compared to $q_{1}, n_{1}$ and $n_{2}$, can be found in Morrison [47]. This approach allows for a generalization to the case of arbitrary service rates and weighted overflow traffic.

### 3.3.3 Jockeying to the waiting room

Now we consider model $\mathrm{S} / \mathrm{S} / \mathrm{W}$. In this model, waiting customers from $Q_{1}$ move to $Q_{2}$ as soon as there is a position in the waiting room or a server available. An arriving $Q_{1}$-customer is blocked and directed to $Q_{2}$ if all $n_{1}$ servers are busy in $Q_{1}$. Blocked customers are served by one of the servers in $Q_{2}$ if at least one is available. They are queued in $Q_{1}$ if all servers in $Q_{2}$ are busy and a waiting position is available in $Q_{1}$ and are lost otherwise. Thus, $p_{i, j}=0$ for $i>n_{1}$ and $j=0, \ldots, k_{2}$, because it is impossible for customers to wait at $Q_{1}$ while there is at least one waiting position or server available at $Q_{2}$. The balance equations for this model are

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)\left(1-\chi_{i-n_{1}}\left(1-\chi_{j-n_{2}}\right)\right)+p \lambda_{1} \chi_{i-n_{1}}\left(1-\chi_{j-n_{2}}\right)\right) \\
& \left.\quad+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
& =\left(1-\chi_{i-n_{1}-1}\left(1-\delta_{j k_{2}}\right)\right) \\
& \quad \times\left(\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1}\right) \\
& \quad+\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \delta_{i n_{1}} \chi_{n_{2}-j}+\lambda_{1} \delta_{i n_{1}}\left(1-\chi_{n_{2}-j}\right)+\lambda_{2}\left(1-\chi_{i-n_{1}-1}\right)\right) p_{i, j-1} \\
& \quad+\left(1-\delta_{i k_{1}}\right)\left(\left(1-\chi_{i-n_{1}}\left(1-\delta_{j k_{2}}\right)\right)\left((i+1) \wedge n_{1}\right) \mu_{1}\right. \\
& \left.\quad+n_{2} \mu_{2} \delta_{j n_{2}} \chi_{i-n_{1}}\right) p_{i+1, j} \tag{3.3.60}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$. The balance equations are constructed so as to imply $p_{i, j}=0$ for $i>n_{1}$ and $j=0, \ldots, k_{2}$. The term $\lambda_{1} \delta_{i n_{1}}\left(1-\chi_{n_{2}-j}\right)$ in the second last line indicates that arriving $Q_{1}$-customers, who find all waiting positions in $Q_{1}$ unoccupied and all $n_{1}$ servers busy in $Q_{1}$, move instantaneously to $Q_{2}$ if there is a waiting position or server available. Hence, $\lambda_{1}$ is not weighted by $p$ since the customers are jockeying immediately to $Q_{2}$ after their arrival to $Q_{1}$ instead of overflowing. By definition, overflow
takes places in model-type $\mathrm{S} / \mathrm{S} / \gamma$ if and only if all $n_{1}$ servers in $Q_{1}$ are busy and there is at least one server available in $Q_{2}$.

Remark 3.3.15. It might be useful in applications to weight the traffic that is due to jockeying with $p$. This is done by replacing the terms in the first line of (3.3.60) by $\lambda_{1}\left(1-\delta_{i k_{1}}\right)\left(1-\chi_{i-n_{1}}\left(1-\delta_{j k_{2}}\right)\right)+p \lambda_{1} \chi_{i-n_{1}}\left(1-\delta_{j k_{2}}\right)$ and the term $\lambda_{1} \delta_{i n_{1}}\left(1-\chi_{n_{2}-j}\right)$ on the right side by $p \lambda_{1} \delta_{i n_{1}}\left(1-\chi_{n_{2}-j}\right)$. However, in this case, the model S/S/W is equivalent to model S/W/W from Section 3.4.2

Although the balance equations (3.3.60) are intricate at first glance, they are easily solved with an analog of the procedure of the alternative approach in Section 3.3.1.6. The separation approach for this model is shown in Figure 3.8


Figure 3.8: Separation scheme: Model S/S/W

For $i=0, \ldots, n_{1}-1$ and $j=0, \ldots, k_{2}$ the balance equations are identical to (3.3.2). It then follows from (3.3.27) that

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2},  \tag{3.3.61}\\ \sum_{m=0}^{k_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2},\end{cases}
$$

for $i=0, \ldots, n_{1}$ with $c_{0}, \ldots, c_{m}$ such that (3.3.17) is fulfilled. It remains to determine $p_{i, j}$ for $j=k_{2}$ and $i=n_{1}+1, \ldots, k_{1}$ because the other proba-
bilities vanish. The balance equations for $i=n_{1}+1, \ldots, k_{1}$ and $j=k_{2}$ are identical to (3.3.44) with $p \cdot{ }_{\cdot, n_{2}}$ replaced by $p_{\cdot, k_{2}}$. We obtain

$$
\begin{equation*}
p_{i, k_{2}}=\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{i-n_{1}} p_{n_{1}, k_{2}}, \quad i=n_{1}, \ldots, k_{1} \tag{3.3.62}
\end{equation*}
$$

as in (3.3.45), where $p_{n_{1}, k_{2}}$ is given by (3.3.61). As usual, the constants $a_{0}, \ldots, a_{k_{2}}$ can be determined from the normalization condition and the boundary conditions at $i=n_{1}$, i.e.,

$$
\begin{align*}
\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+j \mu_{2}\right) p_{n_{1}, j} & =\lambda_{1} p_{n_{1}-1, j}+\left(1-\delta_{j 0}\right)\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, j-1} \\
& +(j+1) \mu_{2} p_{n_{1}, j+1}, \quad j=0, \ldots, n_{2}-1,  \tag{3.3.63}\\
\left(\lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, n_{2}} & =\lambda_{1} p_{n_{1}-1, n_{2}}+n_{2} \mu_{2} p_{n_{1}, n_{2}+1} \\
& +\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, n_{2}-1}, \quad j=n_{2},  \tag{3.3.64}\\
\left(\lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, j} & =\lambda_{1} p_{n_{1}-1, j}+n_{2} \mu_{2} p_{n_{1}, j+1} \\
& +\left(\lambda_{1}+\lambda_{2}\right) p_{n_{1}, j-1}, \quad j=n_{2}+1, \ldots, k_{2}-1, \tag{3.3.65}
\end{align*}
$$

$$
\begin{align*}
\left(\lambda_{1}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, k_{2}} & =\lambda_{1} p_{n_{1}-1, n_{2}}+\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}+1, k_{2}} \\
& +\left(\lambda_{1}+\lambda_{2}\right) p_{n_{1}, k_{2}-1}, \quad j=k_{2} . \tag{3.3.66}
\end{align*}
$$

Inserting (3.3.61) and (3.3.62) into (3.3.63) to (3.3.66) and using (3.3.16) and the recurrence relations (3.2.13), (3.2.14) and additionally $\phi_{k_{2}} \equiv 1$ and (3.3.12) for the last equation gives the conditions for $a_{0}, \ldots, a_{k_{2}}$. These conditions are

$$
\begin{equation*}
\sum_{m=0}^{k_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{j}\left(-c_{m}\right)+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{j}\left(-c_{m}-\mu_{2}\right)\right) \phi_{n_{2}}\left(c_{m}\right)=0 \tag{3.3.67}
\end{equation*}
$$

for $j=0, \ldots, n_{2}-1$ which follow from (3.3.63) and

$$
\begin{align*}
\sum_{m=0}^{k_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}}\left(-c_{m}\right)-p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}-1}\left(-c_{m}\right)\right. &  \tag{3.3.68}\\
& \left.+\lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right)\right) \phi_{n_{2}}\left(c_{m}\right)=0
\end{align*}
$$

which follows from (3.3.64). The equations (3.3.65) yield

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}}\left(-c_{m}\right) \phi_{j}\left(c_{m}\right)\right.  \tag{3.3.69}\\
& \left.\quad+\lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right)\left(\phi_{j}\left(c_{m}\right)-\phi_{j-1}\left(c_{m}\right)\right)\right)=0
\end{align*}
$$

for $j=n_{2}+1, \ldots, k_{2}-1$ and finally (3.3.66) gives

$$
\begin{equation*}
\sum_{m=0}^{k_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}}\left(-c_{m}\right)-\lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right) \phi_{k_{2}-1}\left(c_{m}\right)\right)=0 \tag{3.3.70}
\end{equation*}
$$

Summing (3.3.67) over $j=0, \ldots, n_{2}-1$ and (3.3.69) over $j=n_{2}+1, \ldots, k_{2}-$ 1 using (3.3.7) and (3.2.15) gives

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}-1}\left(-c_{m}+\mu_{2}\right)\right. \\
&\left.+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}-1}\left(-c_{m}\right)\right) \phi_{n_{2}}\left(c_{m}\right)=0 \tag{3.3.71}
\end{align*}
$$

and

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} a_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}}\left(-c_{m}\right)\left(\Psi_{q_{2}-1}\left(c_{m}\right)-\Psi_{0}\left(c_{m}\right)\right)\right. \\
&\left.+\lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right)\left(\phi_{k_{2}-1}\left(c_{m}\right)-\phi_{n_{2}}\left(c_{m}\right)\right)\right)=0 \tag{3.3.72}
\end{align*}
$$

Finally, adding (3.3.68), (3.3.71) and (3.3.72) and simplifying by once again using (3.3.17) and the recurrence relations (3.2.13) and (3.2.14) results in (3.3.70). Consequently, condition (3.3.70) is redundant and $a_{0}, \ldots, a_{k_{2}}$ are determined by (3.3.67), (3.3.68), (3.3.69) and the normalization condition. The normalization condition can be expressed in terms of $a_{0}, \ldots, a_{k_{2}}$ and gives after substituting (3.3.61) and (3.3.62) and reduction with the help of (3.3.17) and (3.2.15)

$$
\begin{align*}
& a_{0} s_{n_{1}}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right)\left(s_{n_{2}}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}}(0)-s_{n_{2}-1}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}-1}(0)\right) \\
& \quad+\sum_{m=0}^{k_{2}} a_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \sum_{l=1}^{q_{1}}\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{l}=1 \tag{3.3.73}
\end{align*}
$$

The separation approach and the derivations of this section lead to the
following theorem that summarizes the results of this section.
Theorem 3.3.16. The unique nonnegative and normalized solution of the steady-state equations (3.3.60) is given by

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2}, \\ \sum_{m=0}^{k_{2}} a_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2},\end{cases}
$$

for $i=0, \ldots, n_{1}$ and

$$
p_{i, k_{2}}=\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{i-n_{1}} p_{n_{1}, k_{2}}
$$

for $i=n_{1}, \ldots, k_{1}$ if the coefficients $a_{0}, \ldots, a_{k_{2}}$ are determined by the $k_{2}+1$ equations (3.3.67), (3.3.68), (3.3.69) and (3.3.73). $c_{0}=0$ and $c_{1}, \ldots, c_{k_{2}}$ are the by Theorem 3.3.6 positive and pairwise distinct solutions of (3.3.17).

Remark 3.3.17. As in Section 3.3.1.6 the results remain valid even if $q_{1}=$ $\infty$ as long as the stability condition

$$
\lambda_{1}<n_{1} \mu_{1}+n_{2} \mu_{2}
$$

is fulfilled. The stability condition is again independent of $p$.
The problem of determining the $\left(n_{1}+1\right)\left(k_{2}+1\right)+q_{1}$ nonzero unknowns $p_{i, j}, i=0, \ldots, k_{1}, j=0, \ldots, k_{2}$, has now been reduced to the problem of determining $k_{2}+1$ eigenvalues and $k_{2}+1$ unknowns.

### 3.4 Overflow with waiting rooms: From servers to waiting room

### 3.4.1 Jockeying to servers and no jockeying

In this section, we consider the deterministic overflow models $\mathrm{S} / \mathrm{W} / \mathrm{S}$ and $\mathrm{S} / \mathrm{W} / \mathrm{N}$, i.e., we choose $p=1$, and show exemplarily for model $\mathrm{S} / \mathrm{W} / \mathrm{N}$ that the technique used throughout the previous sections does not succeed for this models. However, it is possible to treat the case S/W/W because in this case certain stationary probabilities vanish. This is done in the next section.

The number of servers and the waiting room capacity of $Q_{i}$ is $n_{i}$ and $q_{i} \geq 1$, respectively, for $i=1,2$. In these models with blocking rule S , an arriving $Q_{1}$-customer is blocked if all $n_{1}$ servers are busy. Blocked $Q_{1-}$ customers are treated with respect to overflow routine W . They are served by one of the servers in $Q_{2}$ if at least one is available and are queued in $Q_{2}$ if all $Q_{2}$-servers are busy and at least one waiting position is available. The blocked $Q_{1}$-customers who find $Q_{2}$ fully occupied are redirected to the waiting room in $Q_{1}$ and are lost if $Q_{1}$ is fully occupied. Hence, the overflow stream from $Q_{1}$ follows the same route through $Q_{2}$ as the $Q_{2}$-customers do. The jockeying discipline N ascertains that waiting $Q_{1}$-customers are served solely in $Q_{1}$. The balance equations for model $\mathrm{S} / \mathrm{W} / \mathrm{N}$ are given by

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}} \delta_{j k_{2}}\right)+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
& =\left(1-\delta_{i 0}\right)\left(1-\chi_{i-n_{1}-1} \delta_{j k_{2}}\right) \lambda_{1} p_{i-1, j}+\left(1-\delta_{i k_{1}}\right)\left((i+1) \wedge n_{1}\right) \mu_{1} p_{i+1, j} \\
& +\left(1-\delta_{j 0}\right)\left(\lambda_{1} \chi_{i-n_{1}}+\lambda_{2}\right) p_{i, j-1}+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1} \tag{3.4.1}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$. For $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}+$ $1, \ldots, k_{2}$ these equation reduce to

$$
\begin{align*}
&\left(\lambda_{1}\left(1-\delta_{i k_{1}} \delta_{j k_{2}}\right)+\right. \lambda_{2}(1- \\
&\left.\left.\delta_{j k_{2}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i, j} \\
&= \delta_{j k_{2}} \lambda_{1} p_{i-1, j}+\left(1-\delta_{i k_{1}}\right) n_{1} \mu_{1} p_{i+1, j}  \tag{3.4.2}\\
&+\left(\lambda_{1}+\lambda_{2}\right) p_{i, j-1}+\left(1-\delta_{j k_{2}}\right) n_{2} \mu_{2} p_{i, j+1} .
\end{align*}
$$

Using the separation idea in this region, i.e., trying again $p_{i, j}=\alpha_{i} \beta_{j}$, gives the equations

$$
\begin{equation*}
\left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}-\frac{\beta_{k_{2}-1}}{\beta_{k_{2}}}\left(\lambda_{1}+\lambda_{2}\right)\right) \alpha_{i}=\left(1-\delta_{i k_{1}}\right) n_{1} \mu_{1} \alpha_{i+1} \tag{3.4.3}
\end{equation*}
$$

for $i=n_{1}+1, \ldots, k_{1}$ and

$$
\begin{align*}
\left(\left(\lambda_{1}+\lambda_{2}\right)\left(1-\delta_{j k_{2}}\right)+n_{1} \mu_{1}\right. & \left.+n_{2} \mu_{2}-\left(1-\delta_{j k_{2}}\right) \frac{\alpha_{k_{1}-1}}{\alpha_{k_{1}}} \lambda_{1}\right) \beta_{j} \\
= & \left(\lambda_{1}+\lambda_{2}\right) \beta_{j-1}+\left(1-\delta_{j k_{2}}\right) n_{2} \mu_{2} \beta_{j+1} \tag{3.4.4}
\end{align*}
$$

for $j=n_{2}+1, \ldots, k_{2}$. It is seen that the separation approach leads to different separation constants in (3.4.3) and (3.4.4). Furthermore, a solution of (3.4.3) cannot be expressed in terms of a solution of (3.4.4). Instead of solving these equations simultaneously, equation (3.4.3) can be solved recursively, while (3.4.4) leads to another eigenproblem and consequently additional coefficients arise from the boundary conditions. Furthermore, it is
necessary to piece together the solutions at the boundaries of the separation regions. This results in additional conditions for the sets of eigenvalues. Moreover, a sequential approach as in Section 3.3.2 is not possible in this region and in the region $i=n_{1}+1, \ldots, k_{1}, j=0, \ldots, n_{2}$. Similar statements hold for the models $\mathrm{W} / \mathrm{W} / \mathrm{S}$ and $\mathrm{W} / \mathrm{W} / \mathrm{W}$ (see Section 3.6.1).

### 3.4.2 Jockeying to the waiting room

Now we consider the $p$-overflow model $\mathrm{S} / \mathrm{W} / \mathrm{W}$. The number of servers and the waiting room capacity of $Q_{i}$ is $n_{i}$ and $q_{i} \geq 1$, respectively, for $i=$ 1,2 . In this model with blocking rule S an arriving $Q_{1}$-customer is blocked and directed to $Q_{2}$ if all $n_{1}$ servers are busy in $Q_{1}$. Furthermore, blocked $Q_{1}$-customers are treated with respect to overflow routine W. Hence, the overflow stream from $Q_{1}$ follows the same route through the waiting room and servers in $Q_{2}$ as the $Q_{2}$-customers do. The jockeying discipline W ascertains that waiting $Q_{1}$-customers skip to $Q_{2}$ as soon as a waiting position or server is available in $Q_{2}$.

The balance equations are

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)\left(1-\chi_{i-n_{1}}\left(1-\delta_{j k_{2}}\right)\right)+p \lambda_{1} \chi_{i-n_{1}}\left(1-\delta_{j k_{2}}\right)\right. \\
& \left.\quad+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
& =\left(1-\chi_{i-n_{1}-1}\left(1-\delta_{j k_{2}}\right)\right) \\
& \quad \times\left(\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1}\right)  \tag{3.4.5}\\
& +\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \delta_{i n_{1}}+\lambda_{2}\left(1-\chi_{i-n_{1}-1}\right)\right) p_{i, j-1} \\
& +\left(1-\delta_{i k_{1}}\right)\left(\left(1-\chi_{i-n_{1}}\left(1-\delta_{j k_{2}}\right)\right)\left((i+1) \wedge n_{1}\right) \mu_{1}\right. \\
& \left.\quad+n_{2} \mu_{2} \delta_{j k_{2}} \chi_{i-n_{1}}\right) p_{i+1, j}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$. Once again, these equations are constructed such that $p_{i, j}=0$ for $i=n_{1}+1, \ldots, k_{1}$ and $j \neq k_{2}$ since it is impossible for customers to wait at $Q_{1}$ while there is at least one waiting position available in $Q_{2}$.

The separation approach for this model is identical to the one for model $\mathrm{S} / \mathrm{S} / \mathrm{W}$ depicted in Figure 3.8. For $i=0, \ldots, n_{1}-1$ and $j=0, \ldots, k_{2}$ the balance equations are identical to those in (3.3.2) so that $p_{i, j}$ for $i=0, \ldots, n_{1}$
and $j=0, \ldots, k_{2}$ can be chosen as in (3.3.27):

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2},  \tag{3.4.6}\\ \sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2},\end{cases}
$$

for $i=0, \ldots, n_{1}$, where $b_{0}, \ldots, b_{k_{2}}$ have to be determined from the boundary conditions in (3.4.5) and the normalization condition and $c_{0}, \ldots, c_{k_{2}}$ are the solutions of (3.3.17). The balance equations for $i=n_{1}+1, \ldots, k_{1}$ and $j=k_{2}$ are

$$
\begin{equation*}
\left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i, k_{2}}=\lambda_{1} p_{i-1, k_{2}}+\left(1-\delta_{i k_{1}}\right)\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i+1, k_{2}} \tag{3.4.7}
\end{equation*}
$$

These equations are equivalent to (3.3.44) with $p \cdot, n_{2}$ replaced by $p \cdot{ }_{\cdot, k_{2}}$. We get from (3.3.45) that

$$
\begin{equation*}
p_{i, k_{2}}=\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{i-n_{1}} p_{n_{1}, k_{2}}, \quad i=n_{1}, \ldots, k_{1} \tag{3.4.8}
\end{equation*}
$$

where $p_{n_{1}, k_{2}}$ is given by (3.4.6) (see also Section 3.3.3). Now we turn to the determination of the constants $b_{0}, \ldots, b_{k_{2}}$ with the help of the boundary conditions for $i=n_{1}$, i.e.,

$$
\begin{align*}
\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+j \mu_{2}\right) p_{n_{1}, j} & =\lambda_{1} p_{n_{1}-1, j}+\left(1-\delta_{j 0}\right)\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, j-1} \\
& +(j+1) \mu_{2} p_{n_{1}, j+1}, \quad j=0, \ldots, n_{2}-1,  \tag{3.4.9}\\
\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, n_{2}} & =\lambda_{1} p_{n_{1}-1, n_{2}}+n_{2} \mu_{2} p_{n_{1}, n_{2}+1} \\
& +\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, n_{2}-1}, \quad j=n_{2},  \tag{3.4.10}\\
\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, j} & =\lambda_{1} p_{n_{1}-1, j}+n_{2} \mu_{2} p_{n_{1}, j+1} \\
& +\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, j-1}, j=n_{2}+1, \ldots, k_{2}-1, \tag{3.4.11}
\end{align*}
$$

$$
\left(\lambda_{1}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}, k_{2}}=\lambda_{1} p_{n_{1}-1, n_{2}}+\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{n_{1}+1, k_{2}}
$$

$$
\begin{equation*}
+\left(p \lambda_{1}+\lambda_{2}\right) p_{n_{1}, k_{2}-1}, \quad j=k_{2} \tag{3.4.12}
\end{equation*}
$$

Using (3.4.6) and (3.4.8) and paralleling the simplifications that lead from (3.3.63) - (3.3.66) to (3.3.67) - (3.3.70) it is seen that (3.4.9) - (3.4.12) are
equivalent to the following set of equations:

$$
\begin{equation*}
\sum_{m=0}^{k_{2}} b_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{j}\left(-c_{m}\right)+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{j}\left(-c_{m}-\mu_{2}\right)\right) \phi_{n_{2}}\left(c_{m}\right)=0 \tag{3.4.13}
\end{equation*}
$$

for $j=0, \ldots, n_{2}-1$,

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} b_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}}\left(-c_{m}\right) \phi_{j}\left(c_{m}\right)\right.  \tag{3.4.14}\\
& \left.\quad+p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right)\left(\phi_{j}\left(c_{m}\right)-\phi_{j-1}\left(c_{m}\right)\right)\right)=0
\end{align*}
$$

for $j=n_{2}, \ldots, k_{2}-1$ and

$$
\begin{equation*}
\sum_{m=0}^{k_{2}} b_{m}\left(c_{m} s_{n_{1}}\left(c_{m}+\mu_{1}\right) s_{n_{2}}\left(-c_{m}\right)-p \lambda_{1} s_{n_{1}}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}\right) \phi_{k_{2}-1}\left(c_{m}\right)\right)=0 \tag{3.4.15}
\end{equation*}
$$

Note that both (3.4.10) and (3.4.11) reduce to the form of (3.4.14). Summing (3.4.13) over $j=0, \ldots, n_{2}-1$ and (3.4.14) over $j=n_{2}, \ldots, k_{2}-1$ and adding the resulting equations yields the redundancy of (3.4.15). Finally, the normalization condition, (3.4.6) and (3.4.8) lead to

$$
\begin{align*}
& b_{0} s_{n_{1}}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right)\left(s_{n_{2}}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}}(0)-s_{n_{2}-1}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}-1}(0)\right) \\
+ & \sum_{m=0}^{k_{2}} b_{m} s_{n_{1}}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \sum_{l=1}^{q_{1}}\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{l}=1 \tag{3.4.16}
\end{align*}
$$

as in model $\mathrm{S} / \mathrm{S} / \mathrm{W}$. The calculations of this section are summarized in the next theorem.

Theorem 3.4.1. The unique nonnegative and normalized solution of the steady-state equations (3.4.5) is given by

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2} \\ \sum_{m=0}^{k_{2}} b_{m} s_{i}\left(c_{m}, \lambda_{1}, \mu_{1}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2}\end{cases}
$$

for $i=0, \ldots, n_{1}$ and

$$
p_{i, k_{2}}=\left(\frac{\lambda_{1}}{n_{1} \mu_{1}+n_{2} \mu_{2}}\right)^{i-n_{1}} p_{n_{1}, k_{2}}
$$

for $i=n_{1}, \ldots, k_{1}$ if the coefficients $b_{0}, \ldots, b_{k_{2}}$ are determined by the $k_{2}+1$ equations (3.4.13), (3.4.14) and (3.4.16). $c_{0}=0$ and $c_{1}, \ldots, c_{k_{2}}$ are the by Theorem 3.3.6 positive and pairwise distinct solutions of (3.3.17).

The problem of determining the $\left(n_{1}+1\right)\left(k_{2}+1\right)+q_{1}$ nonzero unknowns has now been reduced to the problem of determining $k_{2}+1$ eigenvalues and $k_{2}+1$ unknowns.

Remark 3.4.2. The results remain valid even if $q_{1}=\infty$ as long as the stability condition $\lambda_{1}<n_{1} \mu_{1}+n_{2} \mu_{2}$ is fulfilled.

### 3.5 Overflow with waiting rooms: From waiting room to server

In the following two sections, we consider $p$-overflow models with blocking rule W and overflow routine S . In these models, arriving $Q_{1}$-customers are blocked if all $Q_{1}$-servers are busy and the $Q_{1}$-waiting room is fully occupied, i.e., no overflow is allowed as long as the waiting room in $Q_{1}$ is not fully occupied. The blocked customers overflow to $Q_{2}$ and are served by a server in $Q_{2}$ if at least one is idle. In these models the (unweighted) overflow stream is identical to the stream of blocked customers.

### 3.5.1 Jockeying to servers and to waiting room

At first, it should be mentioned that the configurations $\mathrm{W} / \mathrm{S} / \mathrm{S}$ and $\mathrm{W} / \mathrm{S} / \mathrm{W}$ make sense only if the capacity of the waiting room in $Q_{1}$ is set to zero. This can be explained exemplarily for model $\mathrm{W} / \mathrm{S} / \mathrm{S}$ by the following observations. On the one hand, in model $\mathrm{W} / \mathrm{S} / \mathrm{S}$ with $q_{1} \geq 1$, waiting $Q_{1}$-customers swap to $Q_{2}$ as soon as a $Q_{2}$-server becomes available so that $p_{i, j}=0$ for $i=n_{1}+1, \ldots, k_{1}$ and $j=0, \ldots, n_{2}-1$. On the other hand, arriving $Q_{1^{-}}$ customers overflow to $Q_{2}$ if all $Q_{1}$-servers are busy, the $Q_{1}$-waiting room is fully occupied and a server is available in $Q_{2}$, i.e., $i=k_{1}$ and $j=0, \ldots, n_{2}-1$, but in these cases $p_{i, j}=0$. Consequently, no overflow occurs in $\mathrm{W} / \mathrm{S} / \mathrm{S}$ for $q_{1} \geq 1$. The same observations can be made for jockeying discipline W , since it is a generalization of jockeying discipline S . Therefore, no overflow takes place in model $\mathrm{W} / \mathrm{S} / \mathrm{W}$ for the case $q_{1} \geq 1$. However, with $q_{1}=0$, model W/S/S and W/S/W correspond to model S/S/S and S/S/W, respectively, which were treated in Section 3.3.1 and Section 3.3.3,

### 3.5.2 No jockeying

We examine model $\mathrm{W} / \mathrm{S} / \mathrm{N}$ by generalizing the results in Morrison [48] to the case of $p$-overflow, $p \in[0,1]$, and arbitrary service rates $\mu_{1}, \mu_{2}>0$ (see also Kaufman et al. [34 for a numerical treatment for the case of arbitrary service rates and $p=1$ ). In this model, overflow is allowed if and only if $Q_{1}$ is fully occupied. The blocked customers overflow to $Q_{2}$ and are served by a server in $Q_{2}$ if at least one is idle. Waiting $Q_{1}$-customers are served exclusively at $Q_{1}$. The balance equations are given by

$$
\begin{align*}
\left(\lambda_{1}(1-\right. & \left.\delta_{i k_{1}}\right)\left(1-\delta_{i k_{1}}\left(1-\chi_{j-n_{2}}\right)\right)+p \lambda_{1} \delta_{i k_{1}}\left(1-\chi_{j-n_{2}}\right)+\lambda_{2}\left(1-\delta_{j k_{2}}\right) \\
& \left.+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
= & \lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+\left(1-\delta_{i k_{1}}\right)\left((i+1) \wedge n_{1}\right) \mu_{1} p_{i+1, j} \\
& +\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \delta_{i k_{1}} \chi_{n_{2}-j}+\lambda_{2}\right) p_{i, j-1} \\
& +\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1} \tag{3.5.1}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$. Figure 3.9 displays the separation approach used for this model.


Figure 3.9: Separation scheme: Model W/S/N

Separating the variables into $p_{i, j}=\alpha_{i} \beta_{j}$ for $i \neq k_{1}$ and $j=0, \ldots, k_{2}$ leads analogously to Section 3.3.1 to

$$
\beta_{j}= \begin{cases}s_{j}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}(c), & j=0, \ldots, n_{2},  \tag{3.5.2}\\ s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{j}(c), & j=n_{2}-1, \ldots, k_{2},\end{cases}
$$

where $c$ is a solution of $s_{n_{2}-1}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}(c)=s_{n_{2}}\left(-c, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}-1}(c)$ or equivalently (3.3.17). For $i=0, \ldots, n_{1}$ we get

$$
\begin{equation*}
\alpha_{i}=s_{i}\left(c, \lambda_{1}, \mu_{1}\right) \tag{3.5.3}
\end{equation*}
$$

For $i=n_{1}, \ldots, k_{1}$, the equations

$$
\begin{equation*}
\left(\lambda_{1}+n_{1} \mu_{1}+c\right) \alpha_{i}=\lambda_{1} \alpha_{i-1}+n_{1} \mu_{1} \alpha_{i+1} \tag{3.5.4}
\end{equation*}
$$

must be satisfied with the boundary conditions $\alpha_{n_{1}}=s_{n_{1}}\left(c, \lambda_{1}, \mu_{1}\right)$ and $\alpha_{n_{1}-1}=s_{n_{1}-1}\left(c, \lambda_{1}, \mu_{1}\right)$. It follows from (3.3.11) and (3.3.5) that the functions $\Omega_{i}(c), i \geq-1$, solve

$$
\left(\lambda_{1}+n_{1} \mu_{1}+c\right) \Omega_{i}=\lambda_{1} \Omega_{i+1}+n_{1} \mu_{1} \Omega_{i-1}
$$

for $i \geq 0$. Thus, a solution of (3.5.4) is given by

$$
\begin{equation*}
\alpha_{i}=\left(\frac{\lambda_{1}}{n_{1} \mu_{1}}\right)^{n_{1}-i}\left(s_{n_{1}}\left(c, \lambda_{1}, \mu_{1}\right) \Omega_{i-n_{1}}(c)-s_{n_{1}-1}\left(c, \lambda_{1}, \mu_{1}\right) \Omega_{i-n_{1}-1}(c)\right) . \tag{3.5.5}
\end{equation*}
$$

Plugging (3.5.3) and (3.5.4) together by an appropriate normalization yields the solutions for the first separation variable $\alpha_{i}$ :

$$
\alpha_{i}(c)= \begin{cases}\left(\frac{\lambda_{1}}{n_{1} \mu_{1}}\right)^{q_{1}} s_{i}\left(c, \lambda_{1}, \mu_{1}\right), & i=0, \ldots, n_{1}  \tag{3.5.6}\\ \left(\frac{\lambda_{1}}{n_{1} \mu_{1}}\right)^{k_{1}-i}\left(s_{n_{1}}\left(c, \lambda_{1}, \mu_{1}\right) \Omega_{i-n_{1}}(c)\right. & \\ \left.-s_{n_{1}-1}\left(c, \lambda_{1}, \mu_{1}\right) \Omega_{i-n_{1}-1}(c)\right), & i=n_{1}, \ldots, k_{1}\end{cases}
$$

Finally, together with (3.5.2), a solution of the balance equations (3.5.1) is given by

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2}  \tag{3.5.7}\\ \sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2}\end{cases}
$$

for $i=0, \ldots, k_{1}$. The boundary conditions at $i=k_{1}$ in (3.5.1) are

$$
\begin{aligned}
\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+j \mu_{2}\right) p_{k_{1}, j} & =\lambda_{1} p_{k_{1}-1, j} \\
& +\left(1-\delta_{j 0}\right)\left(p \lambda_{1}+\lambda_{2}\right) p_{k_{1}, j-1}
\end{aligned}
$$

$$
\begin{equation*}
+(j+1) \mu_{2} p_{k_{1}, j+1}, \quad j=0, \ldots, n_{2}-1 \tag{3.5.8}
\end{equation*}
$$

$$
\begin{align*}
\left(\lambda_{2}\left(1-\delta_{q_{2} 0}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{k_{1}, n_{2}} & =\lambda_{1} p_{k_{1}-1, n_{2}}+\left(1-\delta_{q_{2} 0}\right) n_{2} \mu_{2} p_{k_{1}, n_{2}+1} \\
& +\left(p \lambda_{1}+\lambda_{2}\right) p_{k_{1}, n_{2}-1}, \quad j=n_{2}, \quad(3.5 .9  \tag{3.5.9}\\
\left(\lambda_{2}\left(1-\delta_{j k_{2}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{k_{1}, j} & =\lambda_{1} p_{k_{1}-1, j}+\left(1-\delta_{j k_{1}}\right) n_{2} \mu_{2} p_{k_{1}, j+1} \\
& +\lambda_{2} p_{k_{1}, j-1}, \quad j=n_{2}+1, \ldots, k_{2} . \tag{3.5.10}
\end{align*}
$$

Substituting (3.5.7) into (3.5.8) and (3.5.10) gives

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} a_{m}\left[c_{m} s_{j}\left(-c_{m}\right)\left(s_{n_{1}}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}}\left(c_{m}\right)-s_{n_{1}-1}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right)\right. \\
+ & \left.p \lambda_{1} s_{j}\left(-c_{m}-\mu_{2}\right)\left(s_{n_{1}}\left(c_{m}\right) \Omega_{q_{1}}\left(c_{m}\right)-s_{n_{1}-1}\left(c_{m}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right)\right] \phi_{n_{2}}\left(c_{m}\right)=0 \tag{3.5.11}
\end{align*}
$$

for $j=0, \ldots, n_{2}-1$ and

$$
\begin{equation*}
\sum_{m=0}^{k_{2}} a_{m} c_{m} s_{n_{2}}\left(-c_{m}\right)\left(s_{n_{1}}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}}\left(c_{m}\right)-s_{n_{1}-1}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right) \phi_{j}\left(c_{m}\right) \tag{3.5.12}
\end{equation*}
$$

for $j=n_{2}+1, \ldots, k_{2}$. By summing (3.5.11) and (3.5.12) it is seen that (3.5.9) is redundant. The coefficients $a_{0}, \ldots, a_{k_{2}}$ are uniquely determined by (3.5.11), (3.5.12) and the normalization condition. This condition yields - after substituting (3.5.7) and simplifying - an explicit formula for $a_{0}$ in terms of the auxiliary functions. We arrive at the closed-form expression

$$
\begin{align*}
a_{0}= & \left(\left(s_{n_{1}}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right) \Omega_{q_{1}}(0)-s_{n_{1}-1}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right) \Omega_{q_{1}-1}(0)\right)\right. \\
& \left.\times\left(s_{n_{2}}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}}(0)-s_{n_{2}-1}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}-1}(0)\right)\right)^{-1} . \tag{3.5.13}
\end{align*}
$$

Theorem 3.5.1. The unique nonnegative and normalized solution of the steady-state equations (3.5.1) is given by

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2}, \\ \sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2},\end{cases}
$$

for $i=0, \ldots, k_{1}$ if the coefficients $a_{0}, \ldots, a_{k_{2}}$ are determined by (3.5.11),
(3.5.12) and (3.5.13). $c_{0}=0$ and $c_{1}, \ldots, c_{k_{2}}$ are the by Theorem 3.3.6 positive and pairwise distinct solutions of (3.3.17).

The number of unknowns has now been reduced from $\left(k_{1}+1\right)\left(k_{2}+1\right)$ to $2 k_{2}+1$.

### 3.6 Overflow with waiting rooms: From and to waiting room

### 3.6.1 Jockeying to servers and to waiting room

In this section, we consider the deterministic overflow models $\mathrm{W} / \mathrm{W} / \mathrm{S}$ and W/W/W, i.e., we choose $p=1$, and show exemplarily for model W/W/S that the technique used throughout the previous sections does not succeed for these models. Similar observations were made in Section 3.4.1 for models S/W/S and S/W/N. However, it is possible to treat the case W/W/N, because in this case certain stationary probabilities and inconsistencies at the boundaries of the balance equations vanish. This is done in the next section. Numerical methods can be found in Chan [9] and Kaufman [35] for model W/W/S and in Chan [8] and Kaufman [35] for model W/W/N

The number of servers and the waiting room capacity of $Q_{i}$ is $n_{i}$ and $q_{i} \geq 1$, respectively, for $i=1,2$. In these models with blocking rule W , an arriving $Q_{1}$-customer is blocked and directed to $Q_{2}$ if all $n_{1}$ servers are busy and all waiting positions in $Q_{1}$ are occupied. Blocked $Q_{1}$-customers are treated with respect to overflow routine W , i.e., they are served by one of the servers in $Q_{2}$ if at least one is available, are queued in $Q_{2}$ if all $Q_{2}{ }^{-}$ servers are busy and at least one waiting position is available and are lost otherwise. According to jockeying discipline S , waiting $Q_{1}$-customers are forced to move to $Q_{2}$ as soon as a server becomes available in $Q_{2}$. The balance equations for model W/W/S are given by

$$
\begin{align*}
& \left(\lambda_{1}\left(1-\delta_{i k_{1}} \delta_{j k_{2}}\right)+\lambda_{2}\left(1-\delta_{j k_{2}}\right)+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
& =\left(1-\chi_{i-n_{1}-1} \chi_{n_{2}-1-j}\right)\left(\lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}\right) \\
& \quad+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1}  \tag{3.6.1}\\
& \quad+\left(1-\delta_{j 0}\right)\left(\lambda_{1}\left(\delta_{i n_{1}} \chi_{n_{2}-j}+\delta_{i k_{1}} \chi_{j-n_{2}-1}\right)\right) \\
& \left.\quad+\lambda_{2}\left(1-\chi_{i-n_{1}-1} \chi_{n_{2}-j}\right)\right) p_{i, j-1} \\
& \quad+\left(1-\delta_{i k_{1}}\right)\left(\left(1-\chi_{i-n_{1}} \chi_{n_{2}-j-1}\right)\left((i+1) \wedge n_{1}\right) \mu_{1}\right. \\
& \left.\quad \quad+n_{2} \mu_{2} \delta_{j n_{2}} \chi_{i-n_{1}}\right) p_{i+1, j}
\end{align*}
$$

for $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$. Note that $p_{i, j}=0$ for $i=n_{1}+1, \ldots, k_{1}$ and $j=0, \ldots, n_{2}-1$. For $i=n_{1}+1, \ldots, k_{1}$ and $j=n_{2}+1, \ldots, k_{2}$ these equation reduce to

$$
\begin{align*}
\left(\lambda_{1}\left(1-\delta_{i k_{1}} \delta_{j k_{2}}\right)+\right. & \left.\lambda_{2}\left(1-\delta_{j k_{2}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{i, j} \\
= & \lambda_{1} p_{i-1, j}+\left(1-\delta_{j k_{2}}\right) n_{2} \mu_{2} p_{i, j+1} \\
& +\left(\lambda_{1} \delta_{i k_{1}}+\lambda_{2}\right) p_{i, j-1}+\left(1-\delta_{i k_{1}}\right) n_{1} \mu_{1} p_{i+1, j} \tag{3.6.2}
\end{align*}
$$

Separating the probabilities in this region into $p_{i, j}=\alpha_{i} \beta_{j}$ gives the equations

$$
\begin{align*}
\left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}-\frac{\beta_{k_{2}-1}}{\beta_{k_{2}}}\right. & \left.\left(\lambda_{1} \delta_{i k_{1}}+\lambda_{2}\right)\right) \alpha_{i} \\
= & \lambda_{1} \alpha_{i-1}+\left(1-\delta_{i k_{1}}\right) n_{1} \mu_{1} \alpha_{i+1} \tag{3.6.3}
\end{align*}
$$

for $i=n_{1}+1, \ldots, k_{1}$ and

$$
\begin{align*}
\left(\left(\lambda_{1}+\lambda_{2}\right)\left(1-\delta_{j k_{2}}\right)+n_{1} \mu_{1}\right. & \left.+n_{2} \mu_{2}-\frac{\alpha_{k_{1}-1}}{\alpha_{k_{1}}} \lambda_{1}\right) \beta_{j} \\
= & \left(\lambda_{1}+\lambda_{2}\right) \beta_{j-1}+\left(1-\delta_{j k_{2}}\right) n_{2} \mu_{2} \beta_{j+1} \tag{3.6.4}
\end{align*}
$$

for $j=n_{2}+1, \ldots, k_{2}$. As in Section 3.4.1 it is seen that the separation approach leads to different separation constants in (3.6.3) and (3.6.4) so that a solution of (3.6.3) cannot be expressed in terms of a solution of (3.6.4). Instead of solving these equations simultaneously, different eigenproblems have to be solved and consequently additional coefficients arise from the boundary conditions. Furthermore, it is necessary to match the solutions at the boundaries of the separation regions. This results in additional conditions for the sets of eigenvalues.

### 3.6.2 No jockeying

We apply a separation approach to determine the equations for the reduced system of balance equations for model $\mathrm{W} / \mathrm{W} / \mathrm{N}$ with $p$-overflow, $p \in[0,1]$, and arbitrary service rates $\mu_{1}, \mu_{2}>0$ in this section. In this model, overflow is allowed if and only if all servers and waiting positions in $Q_{1}$ are occupied. The blocked customers overflow to $Q_{2}$ and are queued in the waiting room or served by a server in $Q_{2}$ if at least one is available. Waiting $Q_{1}$-customers have to wait for service at $Q_{1}$. The deterministic model, i.e., the case $p=1$, with arbitrary service rates is also treated in Chan [8] and Kaufman [35]
with numerical methods. The balance equations for model $\mathrm{W} / \mathrm{W} / \mathrm{N}$ are

$$
\begin{align*}
\left(\lambda_{1}\left(1-\delta_{i k_{1}}\right)\right. & +p \lambda_{1} \delta_{i k_{1}}\left(1-\delta_{j k_{2}}\right)+\lambda_{2}\left(1-\delta_{j k_{2}}\right) \\
\quad & \left.+\left(i \wedge n_{1}\right) \mu_{1}+\left(j \wedge n_{2}\right) \mu_{2}\right) p_{i, j} \\
\quad= & \lambda_{1}\left(1-\delta_{i 0}\right) p_{i-1, j}+\left(1-\delta_{i k_{1}}\right)\left((i+1) \wedge n_{1}\right) \mu_{1} p_{i+1, j}  \tag{3.6.5}\\
\quad & +\left(1-\delta_{j 0}\right)\left(p \lambda_{1} \delta_{i k_{1}}+\lambda_{2}\right) p_{i, j-1}+\left(1-\delta_{j k_{2}}\right)\left((j+1) \wedge n_{2}\right) \mu_{2} p_{i, j+1}
\end{align*}
$$

or $i=0, \ldots, k_{1}$ and $j=0, \ldots, k_{2}$. The separation approach for model $\mathrm{W} / \mathrm{S} / \mathrm{N}$ can be used for this model. A schematic overview of this approach is given in Figure 3.9. For model W/W/N the redundant boundary equation is located at the state $\left(k_{1}, k_{2}\right)$.

Separating the variables into $p_{i, j}=\alpha_{i} \beta_{j}$ for $i \neq k_{1}$ and $j=0, \ldots, k_{2}$ leads analogously to the derivations in Section 3.5.2 to

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2}  \tag{3.6.6}\\ \sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2}\end{cases}
$$

for $i=0, \ldots, k_{1}$ with

$$
\alpha_{i}(c)= \begin{cases}\left(\frac{\lambda_{1}}{n_{1} \mu_{1}}\right)^{q_{1}} s_{i}\left(c, \lambda_{1}, \mu_{1}\right), & i=0, \ldots, n_{1},  \tag{3.6.7}\\ \left(\frac{\lambda_{1}}{n_{1} \mu_{1}}\right)^{k_{1}-i}\left(s_{n_{1}}\left(c, \lambda_{1}, \mu_{1}\right) \Omega_{i-n_{1}}(c)\right. & \\ & \left.-s_{n_{1}-1}\left(c, \lambda_{1}, \mu_{1}\right) \Omega_{i-n_{1}-1}(c)\right), \\ & i=n_{1}, \ldots, k_{1},\end{cases}
$$

for $c \in \mathbb{R}$. The boundary conditions at $i=k_{1}$ are

$$
\begin{align*}
\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+j \mu_{2}\right) p_{k_{1}, j}= & \lambda_{1} p_{k_{1}-1, j}+(j+1) \mu_{2} p_{k_{1}, j+1} \\
& +\left(1-\delta_{j 0}\right)\left(p \lambda_{1}+\lambda_{2}\right) p_{k_{1}, j-1}, \\
& j=0, \ldots, n_{2}-1,  \tag{3.6.8}\\
\left(\left(p \lambda_{1}+\lambda_{2}\right)\left(1-\delta_{q_{2} 0}\right)+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{k_{1}, n_{2}}= & \lambda_{1} p_{k_{1}-1, n_{2}} \\
& +\left(1-\delta_{q_{2} 0}\right) n_{2} \mu_{2} p_{k_{1}, n_{2}+1} \\
& +\left(p \lambda_{1}+\lambda_{2}\right) p_{k_{1}, n_{2}-1}, \\
& j=n_{2}, \tag{3.6.9}
\end{align*}
$$

$$
\begin{align*}
\left(p \lambda_{1}+\lambda_{2}+n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{k_{1}, j}= & \lambda_{1} p_{k_{1}-1, j}+n_{2} \mu_{2} p_{k_{1}, j+1} \\
& +\left(p \lambda_{1}+\lambda_{2}\right) p_{k_{1}, j-1},  \tag{3.6.10}\\
& j=n_{2}+1, \ldots, k_{2}-1, \\
\left(n_{1} \mu_{1}+n_{2} \mu_{2}\right) p_{k_{1}, k_{2}}= & \lambda_{1} p_{k_{1}-1, k_{2}} \\
& +\left(p \lambda_{1}+\lambda_{2}\right) p_{k_{1}, k_{2}-1}, \\
& j=k_{2} . \tag{3.6.11}
\end{align*}
$$

Substituting (3.6.6) into (3.6.8), (3.6.9) and (3.6.10) yield

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} a_{m}\left[c_{m} s_{j}\left(-c_{m}\right)\left(s_{n_{1}}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}}\left(c_{m}\right)-s_{n_{1}-1}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right)\right. \\
+ & \left.p \lambda_{1} s_{j}\left(-c_{m}-\mu_{2}\right)\left(s_{n_{1}}\left(c_{m}\right) \Omega_{q_{1}}\left(c_{m}\right)-s_{n_{1}-1}\left(c_{m}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right)\right] \phi_{n_{2}}\left(c_{m}\right)=0 \tag{3.6.12}
\end{align*}
$$

for $j=0, \ldots, n_{2}-1$ (see also (3.5.11)) and

$$
\begin{align*}
& \sum_{m=0}^{k_{2}} a_{m}\left[c _ { m } s _ { n _ { 2 } } ( - c _ { m } ) \left(s_{n_{1}}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}}\left(c_{m}\right)\right.\right. \\
& \left.\quad-s_{n_{1}-1}\left(c_{m}+\mu_{1}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right) \phi_{j}\left(c_{m}\right)+p \lambda_{1} s_{n_{2}}\left(-c_{m}\right)\left(s_{n_{1}}\left(c_{m}\right) \Omega_{q_{1}}\left(c_{m}\right)\right. \\
& \left.\left.\quad-s_{n_{1}-1}\left(c_{m}\right) \Omega_{q_{1}-1}\left(c_{m}\right)\right)\left(\phi_{j}\left(c_{m}\right)-\phi_{j-1}\left(c_{m}\right)\right)\right]=0 \tag{3.6.13}
\end{align*}
$$

for $j=n_{2}, \ldots, k_{2}-1$. By summing (3.6.12) and (3.6.13) it is seen that (3.6.11) is redundant. (3.6.12), (3.6.13) and the normalization condition determine the coefficients $a_{0}, \ldots, a_{k_{2}}$. Substituting (3.6.6) into the normalization condition and simplifying gives an explicit formula for $a_{0}$ in terms of the auxiliary functions. We arrive at the closed-form expression

$$
\begin{align*}
a_{0}=( & \left(s_{n_{1}}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right) \Omega_{q_{1}}(0)-s_{n_{1}-1}\left(\mu_{1}, \lambda_{1}, \mu_{1}\right) \Omega_{q_{1}-1}(0)\right) \\
& \left.\times\left(s_{n_{2}}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}}(0)-s_{n_{2}-1}\left(\mu_{2}, \lambda_{2}, \mu_{2}\right) \Psi_{q_{2}-1}(0)\right)\right)^{-1} \tag{3.6.14}
\end{align*}
$$

as in (3.5.13).
By the approach above, the number of unknowns reduces from ( $k_{1}+$ $1)\left(k_{2}+1\right)$ to $2 k_{2}+1$. The results in the case of blocking if all waiting positions are occupied, no jockeying and overflow to the waiting rooms of the second queue is summarized in the following theorem.

Theorem 3.6.1. The unique nonnegative and normalized solution of the
steady-state equations (3.6.5) is given by

$$
p_{i, j}= \begin{cases}\sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{j}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{n_{2}}\left(c_{m}\right), & j=0, \ldots, n_{2} \\ \sum_{m=0}^{k_{2}} a_{m} \alpha_{i}\left(c_{m}\right) s_{n_{2}}\left(-c_{m}, \lambda_{2}, \mu_{2}\right) \phi_{j}\left(c_{m}\right), & j=n_{2}, \ldots, k_{2}\end{cases}
$$

for $i=0, \ldots, k_{1}$ if the coefficients $a_{0}, \ldots, a_{k_{2}}$ are determined by (3.6.12), (3.6.13) and (3.6.14). $c_{0}=0$ and $c_{1}, \ldots, c_{k_{2}}$ are the by Theorem 3.3.6 positive and pairwise distinct solutions of (3.3.17).

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## Symbols

$A(z)$ matrix defined in (2.2.23) or (2.3.17)
$A_{0} \quad$ rate matrix for transitions from level $l(m)$ to level $l(m+1)$
$A_{1} \quad$ rate matrix for transitions from level $l(m)$ to level $l(m)$
$A_{2} \quad$ rate matrix for transitions from level $l(m)$ to level $l(m-1)$
$A_{n}(z)$ matrix obtained from $A(z)$ in Section 2.2 by replacing the $(n+1)$-th column by the vector $-\mu_{2}(1-z) p$ or in Section 2.3 by replacing the $n$-th column with the vector $-(1-z) p$
$B \quad$ rate matrix for transitions from level $l(0)$ to level $l(0)$
$B_{i} \quad$ probability that an arriving $Q_{i}$-customer is blocked at $Q_{i}, i=1,2$
$\chi_{i-n} \quad$ defined as 1 for $i \geq n$ and 0 otherwise
$\chi_{n}(t)=\operatorname{det}\left(t E_{n}-A_{n}\right)$, characteristic polynomial of $A_{n}$ in $t \in \mathbb{R}$
$\delta_{i j} \quad$ Kronecker function, defined by 1 for $i=j$ and 0 otherwise
diag diagonal matrix
$E_{n} \quad$ identity $(n \times n)$-matrix
$i \wedge j \quad$ minimum of $i$ and $j$
$k_{i} \quad=n_{i}+q_{i}$, total number of positions in $Q_{i}, i=1,2$
$L_{1} \quad$ phase of a quasi birth and death chain, (stationary) queue length of $Q_{1}$
$L_{2}$ level of a quasi birth and death chain, (stationary) queue length of $Q_{2}$
$\Lambda \quad$ intensity of the interoverflow process
$\lambda_{1} \quad$ arrival rate for $Q_{1}$
$\lambda_{1, n} \quad$ service rate for $Q_{1}$ if $Q_{1}$ is in state $n$
$\lambda_{2} \quad$ service rate for $Q_{2}$
$l(m)=\{0, \ldots, N\} \times\{m\}$, level of a quasi birth and death chain
$\operatorname{Mat}(n, n, C)(n \times n)$-matrix with entries in the set $C$
$M_{n}=M_{n}(\lambda, \mu)$, matrix defined in (3.2.25)
$\mu_{1} \quad$ service rate for $Q_{1}$
$\mu_{1, n} \quad$ service rate for $Q_{1}$ if $Q_{1}$ is in state $n$
$\mu_{2} \quad$ service rate for $Q_{2}$
$N \quad$ capacity of $Q_{1}$ in Chapter 2
$\mathbb{N}_{0} \quad$ set of the nonnegative integers
$n_{i} \quad$ number of servers in $Q_{i}, i=1,2$
$\mathrm{N} \quad$ in $\alpha / \beta / \mathrm{N}$ : indicator for "no jockeying", only in Chapter 3
$O_{12}$ expected (stationary) number of demands, which flow over from $Q_{1}$ to $Q_{2}$
$\Omega_{l} \quad$ defined in (3.3.11)
$\phi_{j} \quad$ defined in (3.3.7)
$\Pi_{r, j} \quad$ defined in (3.3.54)
$P_{\text {Loss }, i}$ probability that an arriving $Q_{i}$-customer is lost, $i=1,2$
$p_{n, m} \quad$ steady-state probability for $L_{1}=n$ and $L_{2}=m$
$P_{\text {queue }, i}$ probability that an arriving $Q_{i}$-customer is queued in the waiting room in $Q_{i}, i=1,2$
$\Psi_{l} \quad$ defined in (3.3.8)
$Q \quad=\left(q_{i, j}\right)_{i, j \in \mathfrak{S}}$, infinitesimal generator, rate matrix
$Q_{1} \quad$ first queue
$Q_{2} \quad$ second queue
$q_{i} \quad$ number of waiting positions in $Q_{i}, i=1,2$
$\mathbb{R} \quad$ set of the real numbers
$R_{i j} \quad$ mean departure rate from the waiting room in $Q_{i}$ to the servers in $Q_{j}$
$s_{i}(c, \lambda, \mu)=s_{i}(c)$, solution of (3.2.6) with $s_{0}(c, \lambda, \mu)=1$
$\mathrm{S} \quad$ in $\mathrm{S} / \beta / \gamma$ : indicator for "blocking if servers occupied", in $\alpha / \mathrm{S} / \gamma$ : indicator for "overflow to servers", in $\alpha / \beta / \mathrm{S}$ : indicator for "jockeying to servers"

S state space of a Markov chain
$\theta_{i} \quad$ defined in (3.3.10)
$u_{i}(c, \lambda, \mu)=u_{i}(c)$, defined in (3.2.23)
$U_{l} \quad l$-th Chebyshev polynomial of the second kind
$v_{i}(c)$ defined in (3.3.19)
$\mathrm{W} \quad$ in $\mathrm{W} / \beta / \gamma$ : indicator for "blocking if waiting room occupied", in $\alpha / \mathrm{W} / \gamma$ : indicator for "overflow to waiting room", in $\alpha / \beta / \mathrm{W}$ : indicator for "jockeying to waiting room"

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